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Table of Contents

Key Features and Dimensions of Climate Finance

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Abstract- International climate finance is the transfer of funds from the North to the South to help enable developing countries adapt to the unavoidable impacts of climate change (i.e. adaptation), reduce greenhouse gas emissions (i.e. mitigation) , and embark on clean energy development paths.

If we are to avoid the dangerous impact of climate change we must limit global mean temperature increase to 2^0C above pre industrial levels. This means stabilizing atmospheric GHG concentration below 450ppm carbon dioxide equivalent. Emission reductions required for a 450ppm pathway adapted from Mckinsey global GHG abatement cost curve. Failure to cut emissions on this kind of scale would result in serious risks of temperature increases of 3,4,5 deg. C and higher. Scientists tell us that to have a 50-50 chance of holding temperature below 2^0C global emissions would need to be below 35Gt CO_2e by 2030. The 2009 Copenhagen Accord pledged funds of $10 billion a year from 2010 to 2012, increasing to $100 billion per year by 2020 to combat climate change in developing countries through mitigation and adaptation. Even assuming ambitious GHG reductions by developed countries, large additional reductions in developing country emissions will be required in order to limit global warming to 2°C. This pathway requires global emissions to peak no later than 2015, and to fall 50% from 1990 levels by 2050, split so that developed nations shoulder the majority of the burden.

For mitigation in developing countries, approximately €55–80bn in financing from developed countries would likely be required annually in additional funds during the period 2010–2020 (an additional €10–20bn is required annually for adaptation). On the basis of the principle of compensation for incremental costs by developed countries, a total of €65 – 100 billion annually over the 2010 – 2020 period is needed to finance these reductions and meet developing countries' adaptation needs. . McKinsey estimates that in order to reach a desired 450 ppm pathway, €350 billion of incremental capital investment is needed between 2010 and 2020, and €595 billion between 2020 and 2030. Developing nations require €130 billion of capital investment between 2010 and 2020, and €280 billion between 2020 and 2030: China represents a large share of this (€60 billion or 44%).

A study by the World Bank estimates that $70-100 billion dollars per year will be needed for adaptation and indicates that costs may be even higher. For mitigation, estimates indicate that $139-175 billion/year will be needed in 2030.

Prominent climate funds include:[a] The World Bank's Climate Investment Funds (CIFs) [b] The GEF's Least Developed Country Fund (LDCF) and Special Climate Change Fund (SCCF) [c] Adaptation Fund (AF): [d] REDD+ funding from UN-REDD, implemented by FAO, UNEP, and UNDP.The Future sources of climate finance are [i] International Maritime Shipping, [ii] International Aviation, [iii] Direct Budget Contribution, [iv] Financial Transaction Tax (FTT): [v] Special Drawing Rights (SDR):

The Innovations of new finance are [i]A Trustee of 12 carbon funds and facilities, [ii]The green bonds, [iii]A Multi Cat program,[iv] Caribbean Catastrophe Risk Insurance Facility(CCRIF),[v] Certified Emission Reduction certificates for the UN Adaptation Fund, [vi]The International Finance Corporation.

Besides, the Durban decision on the Green Climate Fund , which was part of the Cancun Agreements ,stressed the need for an "early and adequate replenishment process", based on voluntary contributions.

The governance matters for institutionalization of Green Climate Fund under the accountability to its 'parent' body, the United Nations Framework Convention on Climate Change (UNFCCC) whose elements are,[i] ownership, [ii]participation, [iii]structure,[iv] mandate,[v]membership and representation.

The GCF's governance architecture provides to:
• identify clear responsibilities and lines of accountability within the Fund's structure;
• clarify the Fund's objectives and developing the standards and metrics to assess the progress
• develop environmental and social safeguard policies to ensure the Fund's activities
• identify, informing and empowering the constituencies of beneficiaries of the
• ensure the oversight system
Lastly, climate funds have continually neglected gender issues and failed to incorporate a gendered perspective into programmes and projects.

Index Terms- Climate finance, magnitude of finance, governance of finance, sources of finance

I. INTRODUCTION

What is climate finance

Climate finance is a key issue at the heart of the international climate negotiations. These include how much additional finance is needed for mitigation and adaptation, how will the money be raised to meet these needs, how will resources be transferred to diverse recipients in developing countries and how will progress and outcomes – reducing emissions of greenhouse gases (GHGs) and adapting to the impacts of climate change – be monitored? Climate finance can be a catalyst to leverage private and public resources, open new economic opportunities, promote technology deployment and transform development pathways.

To reduce worldwide greenhouse gas (GHG) emissions, increase GHG sequestration, and adapt to the harmful impacts of changing GHG levels, Climate finance is an effort to support

developing countries by providing funding from the developed world to assist these countries in mitigation and adaptation, and to embark on green development paths.

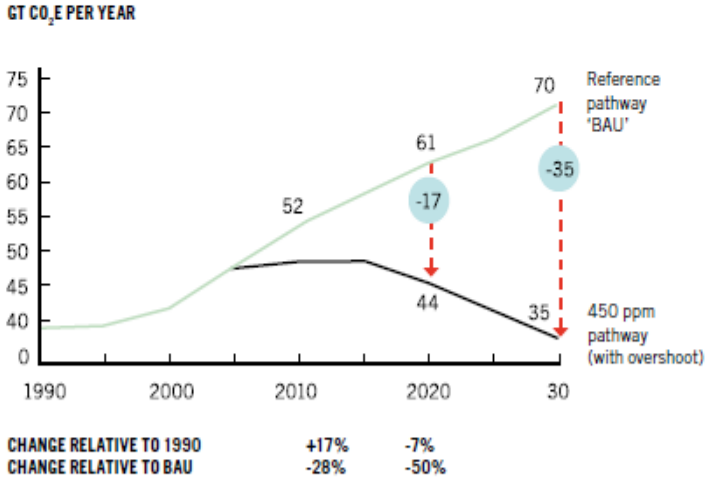
Climate finance, measured as annual commitments of funds or capital flows (disbursements of public funds or private investment) support climate action as well as cumulative amounts of investment or public fund capital over time. Measurements should be in a common currency and real year monetary units (e.g. real USD for the most recent year).Simply, Climate finance refers to funds that will be transferred to developing countries to cover their investments in mitigation and adaptation.

International climate finance is the transfer of funds from the North to the South to help enable developing countries adapt to the unavoidable impacts of climate change (i.e. adaptation), reduce greenhouse gas emissions (i.e. mitigation) , and embark on clean energy development paths.

Climate finance is further grounded in the concept of the North's historical responsibility for causing the climate crisis. The provision of climate finance can be seen as part of repayment of the climate debt that the North owes to the South. It is also a legal obligation under the United Nations Framework Convention on Climate Change (UNFCCC).

II. The need for Climate Finance

If we are to avoid the dangerous impact of climate change we must limit global mean temperature increase to 2^0C above pre industrial levels. This means stabilizing atmospheric GHG concentration below 450ppm carbon dioxide equivalent. To achieve this target we need to start now decarbonise the global economy. In absolute terms, this means a reduction in annual global emissions of 17 billion tones by 2030.Puting this into context, global GHG emission in 2005 were 49 billion tones CO_2e and they are projected to rise under business as usual scenarios to 61 billion tones in 2020 and 70 billion tones in 2030.To meet this targets we will need to raise finance at scale for climate change mitigation in developed and developing countries. In the figure, emission reductions required for a 450ppm pathway adapted from Mckinsey global GHG abatement cost curve which is shown below.(Fig.1).



Source: Oxford Institute for Energy Studies,GCP,2009

Fig.-1: GHG Abatement Cost Curve

The world must reduce the absolute level of global emissions of greenhouse gases (GHGs) by a factor of 2½ or more in 40 years and emissions per unit of output by a factor of 7 to 8 under reasonable growth assumptions; that is what is needed to give a 50-50 chance of holding to a 2 deg C increase in global temperatures relative to the 19th century. That will require substantial investment in both developed and developing countries and major technological advance. Failure to cut emissions on this kind of scale would result in serious risks of temperature increases of 3,4,5 deg C and higher. These temperatures would likely transform the relationship between humans and the planet: we have not seen 3 deg C for 3 million years and 5 deg C for 30 million years – homo sapiens has been around for approximately 200,000 years. The rich countries are not only wealthier and better equipped technologically than developing countries but they also emitted around 75% of cumulative global GHG emissions since the mid-19th century. The anthropogenic climate change which is occurring now and will occur over the next 20 years, which is largely the result of these past emissions of rich countries, will also require substantial investment in adaptation. For these reasons an equitable climate change agreement must involve substantial support by the rich countries for the mitigation and adaptation investment which is necessary in poorer countries. The conclusions of the UNFCCC COPs held in Durban in December 2011, and in Cancun in 2010, point in this direction: action on reducing emissions will need to be taken globally, but poorer countries need to be assured, through financial support, of equal access to sustainable development. The arithmetic is clear. The current total global emissions are nearing 50Gt CO_2 equivalent (CO_2e) p.a., with approximately 20Gt CO_2e p.a. in the rich world and the remaining 30 Gt CO_2e in the developing world. Taking into account the pledges in the Cancun agreement, by 2020 total emissions would be in the 48-52Gt CO_2e range, with rich countries 16-19Gt CO_2e, and developing countries at 32-33Gt CO_2e. If the rich countries accelerated their actions to reduce emissions they could potentially get down to 10Gt CO_2e by 2030. If poor countries managed to limit emission increases per capita to modest levels, they may hold their overall per annum emissions to approximately 40Gt CO_2e by 2030. This would mean a total global flow of emissions of approximately 50Gt

CO_2e p.a. by 2030. Scientists tell us that to have a 50-50 chance of holding temperature below 2^0C global emissions would need to be below 35Gt CO_2e by 2030.

In order to meet these targets there is need to raise new and additional finance at a global level. The 2009 Copenhagen Accord pledged funds of $10 billion a year from 2010 to 2012, increasing to $100 billion per year by 2020 to combat climate change in developing countries through mitigation and adaptation. Even assuming ambitious GHG reductions by developed countries, large additional reductions in developing country emissions will be required in order to limit global warming to 2°C. This pathway requires global emissions to peak no later than 2015, and to fall 50% from 1990 levels by 2050, split so that developed nations shoulder the majority of the burden.

In developing countries, some of these reductions have negative costs, such as energy efficiency in buildings, transport, and industry. Many areas have moderate positive costs (agriculture and forestry), and technology-intensive sectors (notably renewable energy) require significant funding.

III. Magnitude of Climate Finance Required

The magnitude of climate finance required for developing countries must be determined by climate protection objectives. There has been wide agreement that global mean temperatures should not rise more than about 2°C. To achieve this goal, global emissions will need to peak by 2020 and fall 50 percent from 1990 levels by 2050. It is increasingly believed that achieving these reductions will require developed countries to make reductions of 25–40 percent by 2020 and 80 percent by 2050. The magnitude of these emissions reductions and the costs of achieving them in turn determine the total level of financing required for mitigation actions. For mitigation in developing countries, approximately €55–80bn in financing from developed countries would likely be required annually in additional funds during the period 2010–2020 (an additional €10–20bn is required annually for adaptation). These amounts are in addition to what developing countries are expected to contribute and business-as-usual growth in both existing CDM offset credit markets and ODA programs. Even on optimistic assumptions about levels of private finance, mitigation funding from developed countries will have to include large public transfers as well as private transfers.

On the basis of the principle of compensation for incremental costs by developed countries, a total of €65 – 100 billion annually over the period of 2010 – 2020 is needed to finance these reductions and meet developing countries' adaptation needs. However, these cost figures do not capture the significant positive externalities throughout society from low-carbon investment such as increased employment, heightened energy security, improved agricultural productivity, and improved infrastructure.

The McKinsey Green House Gas Abatement Cost Curve (Fig.1) assesses the technical opportunities to abate CO_2 emissions that cost under €60/ton in the period to 2020. Abatement opportunities examined fall into three categories:[i] Energy efficiency (buildings, transport, industry), representing 5 Gt, [ii] Low-carbon energy supply, representing 4 Gt,[iii] Terrestrial carbon (forestry and agriculture), representing 10 Gt Investment in these sectors would start to turn these opportunities into real reductions. McKinsey estimates that in order to reach a desired 450 ppm pathway, €350 billion of incremental capital investment is needed between 2010 and 2020, and €595 billion between 2020 and 2030. Sector estimates are shown in Table 1.

Developed nations require €220 billion of capital investment per year between 2010 and 2020, and €315 billion between 2020 and 2030: this is mainly driven by replacement or upgrade of existing buildings (47% of the total capital need by 2020) and transportation stock (20% of the total capital need by 2020).

Developing nations require €130 billion of capital investment between 2010 and 2020, and €280 billion between 2020 and 2030: China represents a large share of this (€60 billion or 44%).

Table 1: Global climate finance requirement

	Global Investment Need		Developing Nation Investment Need	
Sector	2010-2020(€bn)	2020-2030(€bn)	2010-2020(€bn)	2020-2030(€bn)
Building(mainly energy efficiency)	125	155	25	45
Transportation(mainly energy efficiency)	70	215	25	100
Industry(mainly energy efficiency)	75	80	40	50
Power	65	125	30	70
Waste	10	10	5	5
Forestry and Agriculture(terrestrial carbon)	5	5	5	5

Source-Metz, 2009

A study by the World Bank estimates that $70-100 billion dollars per year will be needed for adaptation and indicates that costs may be even higher. For mitigation, estimates indicate that $139-175 billion/year will be needed in 2030. Developing countries have advocated for a goal as high as US $600 billion, or 1.5% of developed nations' GDP. In 2009, developed countries committed to provide Fast Start Finance (FSF) totaling US $30 billion between 2010 and 2012 and to mobilize US $100 billion/year by 2020. Given the context above, this commitment should be viewed as a starting point, not an end goal. However, an explicit plan for scaling up to that level of financing remains

elusive. Governments are struggling to effectively deliver their FSF commitments. Even more worrisome, little progress is being made to commit further public funding and to develop innovative financing methods that will reach the US $100 billion/year goal. But, the European Commission, African Group, World Bank and UNFCCC estimated the required climate finance which is not similar to others.

Table – 2: Recent Estimates of International Finance

Annual Funding Needs,2005(billions of dollars)			
Source	Year	Mitigation	Adaptation
EC,2009	2020	94	10-24
African group.2009	2020	200	>67
World Bank,2009	2030	139-175	20-100
UNFCCC,2008	2030	>65	28-59

Source-IIED,UK

It is to be noted that the developing countries of the Common Wealth Countries received the total climate fund as 400.00 million US dollar in which 106.8 million dollar for adaptation funds,264.8 million dollars for mitigation funds and 28.4 million dollar for REDD as on December 2010 which were far below the actual needs.(www.climatefundsupdate.org)

IV. Existing Sources of Climate Finance

The relationship between the Green Climate Fund and existing funds is unclear, though existing climate funds are meant to be harmonized and better coordinated after the operationalization of the GCF. A variety of financial sources already exist, though their scales are small compared with the financial flows anticipated under the GCF. Little collaboration and coordination has taken place between existing funds to date. The majority of these funds are managed by the World Bank, the UNFCCC Secretariat, or other UN agencies. The Global Environmental Facility (GEF), an independent financial organization, assists developing countries by providing grants in support of UN environmental agreements such as the UNFCCC. Since its founding in 1991, it has allocated a total of US $9.5 billion to environmental initiatives. Prominent climate funds include:

[a] **The World Bank's Climate Investment Funds (CIFs)**: Funds to help developing countries progress in low-emissions and climate-resilient development, on a scale of US $6.5 billion as pledged by participating countries in 2008. The funds were created with a sunset clause that will phase them out with the establishment of a UNFCCC financial mechanism.

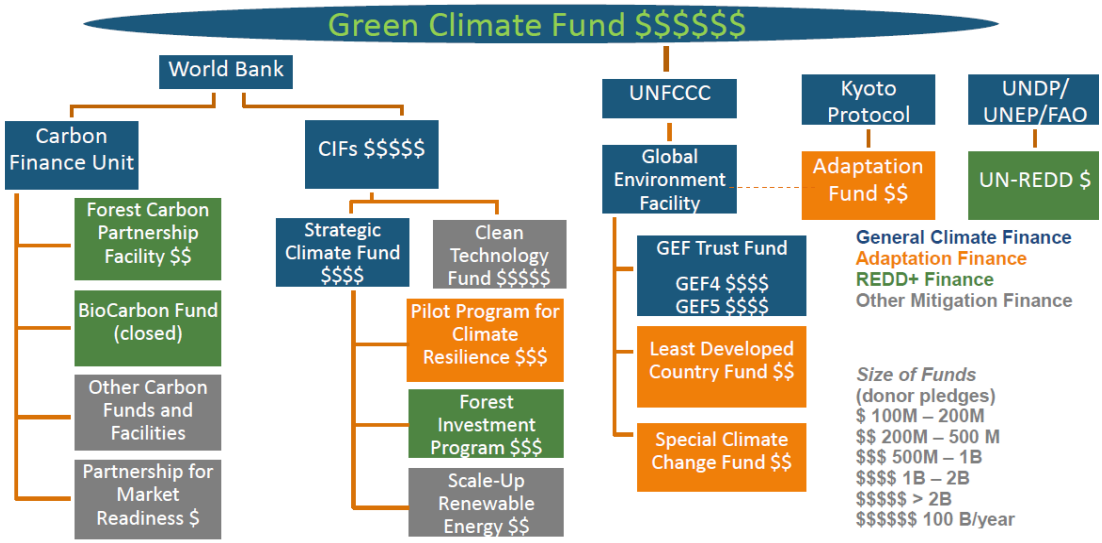
[b] **The GEF's Least Developed Country Fund (LDCF) and Special Climate Change Fund (SCCF)**: The LDC Fund promotes adaptation efforts for the least developed countries, specifically by financing mainly the design, but also the implementation, of National Adaptation Programs of Action (NAPAs), on a total scale of US $415 million. The SCCF provides support for long- and short-term adaptation efforts and technology transfer in all developing country parties to the UNFCCC on a total scale of $218 million. These funds will likely remain in place with the establishment of the Green Climate Fund.

[c] **Adaptation Fund (AF):** The Adaptation Fund was created within the Kyoto Protocol in order to provide funds for adaptation programs in developing countries. The Fund is financed with a small percentage of revenue from the trade of Certified Emission Reductions (CERs) allocated for the Clean Development Mechanism and other sources. Total value of assets stood at US $211 million as of January 31, 2011.

[d] **REDD+ funding from UN-REDD, implemented by FAO, UNEP, and UNDP**: Supports governments in designing and implementing REDD+ initiatives, on a scale of US $150.84 million. A chart mapping out the current climate finance sources is included in the **Chart 1**.

REDD+ funding from UN-REDD, implemented by FAO, UNEP, and UNDP

Chart 1: Existing Financial Sources



Source-*Rebecca* , 2011

V. FUTURE SOURCES OF CLIMATE FINANCE

[i] **International Maritime Shipping**: Regulating GHG emissions from the marine transportation sector could generate finance through levies, carbon offsetting, or the sale of allowances. This method could produce between US $2-19 billion per year.

[ii] **International Aviation**: A market-based mechanism, such as sectoral cap-and-trade, has been proposed to decrease aviation-related GHG emissions. It could also raise revenue; estimates project that such a mechanism could generate US $1-6 billion per year where the contributions of US was 41%,EU,22%,rest annex countries 14% and non-annex countries 23% respectively

[iii] **Direct Budget Contribution**: Developed nations pledge a specific contribution to climate finance from their national budget. Developing countries prefer this method, though it depends on developed nations' willingness to make and meet pledges, which can vary greatly depending on the current political and economic climate. Currently, the scale of this source is reflected in the scale of Fast Start Finance—US $30 billion over 3 years.

[iv] **Financial Transaction Tax (FTT):** This source would impose a global 0.005% tax on each trade of financial instruments such as currency and stocks. While implementing such a mechanism would require a device for global taxation and collection, estimates project that this could generate US $2-27 billion/year.

[v] **Special Drawing Rights (SDR):** At the December 2009 climate change conference in Copenhagen, philanthropist George Soros drew attention to another means to generate resources for climate change: the use of Special Drawing Rights (SDRs), which are "reserve assets" created by the International Monetary Fund (IMF). Soros suggested that an immediate infusion of SDRs could create a U.S. $100 billion "fast-start green fund" for climate finance that could be part of the answer to developing countries' adaptation and mitigation needs. At the World Economic Forum in January 2010, IMF Managing Director Dominique Strauss- Kahn echoed Soros's words, marking the first time the IMF has favorably acknowledged the possibility of using SDRs as a finance instrument. Strauss-Kahn was vague in his proposals and said the IMF will issue a paper sometime soon to elaborate. An immediate infusion of resources can happen if developed countries decide to convert their own idle SDRs into cash – up to U.S. $165 billion from the 2009 allocation. This funding can then be transferred to a United Nations Framework Convention on Climate Change (UNFCCC) fund or mechanism. Various options are described below for how the interest on those SDRs could be paid. However, while an immediate infusion of resources for climate adaptation and mitigation is needed, predictable and sustainable climate finance is essential. Therefore, developed country governments must also agree to support ongoing and regular allocations of SDRs for climate finance. For regular allocations of SDRs, both developed and developing countries could convert their SDR allocation into cash to be transferred to a UNFCCC fund. The fund would then make grants to developing countries for climate adaptation and mitigation, based on the rules established by its governing body. A key issue to be resolved relates to the interest charge which governments incur when they convert SDRs into hard currency. When SDRs are used for climate finance -- particularly for adaptation –developing country governments should not bear any of the costs involved, according to the "polluter pays" principle. Adaptation finance is a form of compensation for the measures developing countries are forced to take in the face of climate change they did not create. Therefore, all adaptation finance must be in the form of grants from developed countries.

There are various alternatives for covering the interest charge for SDRs, including:

1. Selling all or part of the IMF's enormous gold stocks (approximately 100 million ounces), which would likely cover several years of interest payments. (This option would require approval of the US Congress, among others.)

2. Requiring developed countries – those countries historically responsible for creating the climate crisis – to pay the interest charges. This could be seen as a contribution towards the climate debt that developed countries owe to countries in the global south.

3. Action by the IMF Board of Governors to "cancel" all SDRs which have been converted to hard currency for adaptation and mitigation purposes. Under this scenario, developing countries would not need to pay interest on or replenish their SDRs. This cancelation could be framed as a specific response to the global threat of climate change, and need not be seen as a precedent for the general use of SDRs.

The IMF Executive Board would eventually have to agree to use regular allocations of SDRs for climate finance and to transfer these resources to a UNFCCC fund. However, the first place of agreement on using SDRs for climate change should be through the UNFCCC, where all negotiations on climate change should take place. The G20 could help advance this process by supporting this non-traditional use of SDRs.

One risk is giving the IMF – an institution with an undemocratic governance structure and a history of attaching very harmful conditions to its loans – any role at all in climate finance. However, it is important to note that it is not the IMF, but the governing body of a UNFCCC fund, which controls the disbursement of SDRs. Furthermore, the IMF has no hand in how SDRs get used and the IMF cannot attach any conditions to the SDRs. The only role the IMF has is to actually create the SDRs. Another possible risk is that developed country governments could use SDRs as a way to get out of paying their full climate debt. This is because, depending on how repurchases and interest payments are handled, SDRs could generate significant sums of money without necessarily burdening the Treasuries of developed countries. In the face of this risk, it is incumbent on the climate justice community to insist that SDRs are one of the options to generate climate finance. However, in order to meet developing countries' staggering adaptation and mitigation needs, and to ensure that developed countries live up to their responsibility for providing public finance to developing

countries for climate change, a combination of innovative mechanisms to generate climate finance is needed.

The developed country governments could try to use SDRs as a way to get out of paying their full climate debt. Even with these potential risks, the many potential benefits of SDRs make them an option which must not be ignored. Given the current levels of attention from policymakers and the media to the potential for SDRs as a solution to climate finance, developing countries and the climate justice movement must deepen their understanding and debate on this complex topic. In the next couple of weeks the IMF will be issuing a paper on "out-of-the-box ideas" to generate revenue for climate change, including the use of SDRs. Developing countries and the climate justice movement must be ready to respond to the IMF's proposal and ideally come to agreement on whether and how to firmly place SDRs in the debate at the next round of UNFCCC negotiations. This brief aims to promote a deeper understanding and discussion on this complex topic.

Using SDRs for climate finance could be a way to help address the urgent need for the transfer of resources from north to south to address the climate crisis while having co-benefits for the global economy. SDRs should be thought of as part of the puzzle. In order to meet developing countries' adaptation and mitigation needs, a combination of innovative mechanisms is needed. Such mechanisms may include the use of current fossil-fuel subsidies for climate finance, creation of a financial transaction tax, and application of levies in the aviation and shipping industries.

VI. Innovations of new finance

[i]A Trustee of 12 carbon funds and facilities capitalized at $2.74 billion, which are directly helping developing countries finance climate action. Sixteen governments and 66 private companies from various sectors, have made financial contributions to these funds and facilities.

[ii]The green bonds introduced specifically to finance climate mitigation and adaptation work in developing countries. To date, over $2.3 billion in Green Bonds have been issued through 43 transactions in 16 currencies.

[iii]A Multi Cat program is a catastrophe bond issuance platform. It enables governments from developing countries to access affordable insurance coverage through the capital markets. Mexico used the platform to issue a $290 million series of notes in October 2009 to insure against earthquake and hurricane risks.

[iv]16 Caribbean countries establish the Caribbean Catastrophe Risk Insurance Facility (CCRIF). It offers insurance against major hurricanes and earthquakes. Similarly, we worked to develop a Central American Weather Risk Management Program in Honduras, Guatemala, and Nicaragua to help farmers hedge against weather risk.

[v]The Treasury monetizes Certified Emission Reduction certificates for the UN Adaptation Fund, with 9.5 million CERs sold, raising nearly $163 million to finance adaptation projects in developing countries (as of June 15, 2011)

[vi]The International Finance Corporation (our private sector arm) worked with Standard & Poors to develop the first Global Emerging Market Carbon Efficiency Index. Launched in December 2009, it gives carbon-efficient companies access to long-term investors.

VII. Options for improvement

One of the policy brief presents options for improving and expanding climate finance. These options include:

(1) Reforming the Clean Development Mechanism (CDM) offset market to leverage large-scale foreign direct investment in emission-reducing activities in developing countries, most importantly in technology transfer;

(2) Allocating emissions allowances in an international cap-and-trade scheme such that developing countries are (partly) compensated for their emission reductions;

(3) Establishing an international greenhouse gas charge or other mechanism in major developing countries that creates domestic streams of revenue;

(4) Reforming energy subsidies to free funds for government expenditure for climate mitigation and adaptation;

(5) Employing export credit agencies to leverage foreign direct investment in climate-related activities;

(6) Increasing bilateral and multilateral official development assistance for climate-related projects; and

(7) Providing large-scale financing for incremental costs contingent on implementation of emission reduction policies in developing countries.

VIII. The Green Climate Fund

The decision at the COP17 in Durban, South Africa, to adopt a governing instrument for the Green Climate Fund (GCF), as well as a transitional process for its full operationalisation by 2014, was a key component of the "Durban Package". Without agreement on the GCF, the "African COP" would have been considered a failure. The GCF was designed by a Transitional Committee (TC) in 2010 with the goal of becoming the main multilateral financing mechanism to support climate action in developing countries. Under a compromise decision reached at the last minute in Durban, the new GCF will be a legally independent institution with its own separate secretariat and the World Bank as its interim trustee, but functioning under the guidance of and accountable to the COP. Since the Durban Package did not reach any agreement on long-term sources of climate financing, the biggest challenge for the GCF will be to secure adequate and sustained funding. Otherwise, the GCF is in danger of becoming a beautifully articulated, but largely empty shell. Substantial, financial pledges made quickly by contributing countries would be necessary to show broad political support for the GCF and secure its viability; whether this actually happens in 2012 remains to be seen.

The GCF was conceived as the main multilateral financing mechanism to support climate action in developing countries at the Copenhagen COP in 2009, as parties pledged to mobilize $100 billion in long-term financing per year by 2020. The GCF is also supposed to channel "a significant share of new multilateral funding for adaptation," which is structurally underfunded in the current global climate finance architecture. The GCF will be an

operating entity of the financial mechanism of the Convention under Article 11, and will be "accountable to and function under the guidance of the COP". Although the Durban decision on the GCF stressed the need for an "early and adequate replenishment process", based on voluntary contributions, it does not address how to secure sufficient long term capitalization of the Fund. The Durban Package committed only to a very weak program of workshops until COP 18.

[i] The GCF Design Process and COP Decision: The decision to establish the GCF was part of the Cancun Agreements of COP 16 in Cancun, Mexico in 2010, which also set up a Transitional Committee (TC) – composed of 25 representatives from developing countries and 15 from developed countries – to work out design recommendations for the GCF and present them to the COP 17 in Durban. Over the course of a seven-month period, the TC held four official meetings and several information gathering workshops addressing comprehensive terms-of-reference for its work. The TC work was centered along four thematic work streams: on i) objectives and guiding principles of the GCF; ii) its governing arrangements; iii) operational modalities; and iv) monitoring and evaluation. Developing countries throughout the TC process had elaborated on their vision of a GCF with a capitalization of close to US$ 100 billion annually by 2020. By contrast, developed countries sketched the outlines of a GCF with reduced links to the COP and the UNFCCC principles and mandates. Limited public finance would primarily be used to catalyze and leverage private sector investment. Private money would form the core of the sum needed to help developing countries to make the transition to ambitious low-emission development pathways. GCF financing would be based on measurable and verifiable results, subject to implementing entities fulfilling high fiduciary standards. Ultimately, these opposing positions could not be reconciled in the TC.

[ii]The GCF Governing Instrument- Some of the key issues and key provisions of GCF board are listed below:[a] **Objectives and guiding principles:** To capture a high level of ambition for the GCF, as well as define its added-value to existing climate funds, the governing instrument refers to supporting a "paradigm shift towards low-emission and climate-resilient development pathways by providing support to developing countries to limit or reduce their greenhouse gas emissions and to adapt to the impacts of climate change." However, there is no explicit reference to equity considerations in conjunction with recognized Fund core objectives of efficiency and effectiveness. The objectives recognize the need for GCF actions to promote "environmental, social, economic and development co-benefits and taking a gender-sensitive approach" in all climate actions. With its reference to a gender-sensitive approach in all its funding, as well as calls to strive toward gender-balance in the GCF Board and Secretariat staff, the GCF is the first dedicated climate fund to include gender considerations from the outset.

[b]Governance Structure: While the governing instrument gives clear guidance for the representation of developing countries on the Board (including dedicated seats for the Least Developed Countries and Small Island Developing States), developed country representation on the GCF board may be decided on the basis of their financial contributions. The GCF will have a fully independent secretariat headed by an Executive Secretary to be determined based on merit in a transparent process. In the interim (until COP19), a temporary secretariat will be staffed by technical experts from the UNFCCC and the GEF, and the new Board in one of its first actions will have to decide on an Interim Secretariat head with expertise in the design and management of funds. [c] **GCF Relationship to the UNFCCC and the COP:** This was key point of contention during the TC process. Developing countries insisted on establishing the GCF with close ties to the UNFCCC, while some developed countries would have preferred to have no ties to the convention. The GCF is now designated as an operating entity of the UNFCCC's financial mechanism like the GEF – for some recipient countries this relationship is not close enough, and for other contributing countries, it is too close. This also explains why some disputed questions about the GCF-UNFCCC relationship were addressed in the cover note to the decision, for example that the COP will have to endorse the Board decision about the host country for the GCF. Probably the most important task for the new GCF Board in 2012 will be to define the arrangements between the COP and the GCF by COP18. Developing countries are likely to disagree on how close the oversight of the GCF by the COP should be in order to fulfill the mandate to be "accountable to and function under the guidance of the COP". Currently, the governing instrument dictates the Board will submit annual reports and take action responding to guidance it receives from the COP on its programs, policies and priorities. This is essentially the GEF-model. However, developing country board members may well aim for the GCF to go beyond the existing memorandum-of-understanding (MOU) between the UNFCCC and the GEF. [d] **Legal Personality:** Both the governing instrument and the cover note spell out that the GCF will have its own juridical personality and legal capacity, a key demand of developing countries wanting to avoid setting up the GCF under the auspices of an existing institution. Having legal personality will be an important attribute for the GCF should it wish to provide more than grant finance. Several countries, including South Korea and Germany have expressed interest, with the latter already hosting the UNFCCC Secretariat and having conferred judicial personality to the Kyoto Protocol Adaptation Fund. [e] **Operational Modalities:** The GCF will start out with only adaptation and mitigation funding windows, but the Board retains the flexibility to add other ones. Possible candidates would be a REDD+ window, one for technology transfer or a small-grants facility. Grants, concessional loans and other financial instruments, to be approved by the Board, will be used.

Why Governance Matters:

The GCF and its governance and institutional arrangements are crucial to its success: the architectural design must secure its credibility with potential donors, its legitimacy with recipient countries and their domestic social stakeholders, and its accountability to its 'parent' body, the United Nations Framework Convention on Climate Change (UNFCCC).

In these respects, climate finance governance essentially boils down to a discussion about power and authority: the diffusion of power amongst its various core stakeholder

communities – the donors who liquidate existing and new funds; the legal entities that govern the funds on behalf of the international community; the secretariats that execute mandates and policies; and the recipient countries and communities who benefit from the resources.

Accountable governance always matters. The core design elements of an accountable governance system are:

• *Ownership*: where does ultimate legal authority sit? Who are the 'owners' and how do they come to be 'owners'? Where is the balance of power?

• *Participation*: how, and to what extent, are other stakeholders provided with a vote and/or voice in the governance institutions?

• *Structures*: what are the mechanisms and bodies necessary to provide the initiative with the legitimate decision-making capabilities it needs?

• *Mandate*: what are the respective roles and responsibilities of the different bodies, and how do they relate to each other procedurally?

• *Membership and representation*: who serves on each body and how are they selected?

These elements must be coupled with a clearly identified purpose. In short, new and existing climate finance mechanisms:

• need to capture the attention of the private sector (diversity of sources);

• need to win the confidence of donor countries (to bring the capitalisation of the funds to scale);

• need to maintain the perception of equity and fairness of access (legitimacy); and

• need to deliver on the mitigation and adaptation targets set by the international community (effectiveness).

To accomplish this ambition, the GCF's governance and institutional arrangements will have to provide it with the following attributes:

• Global scope, credibility and expertise;

• Appropriate political balance and representation;

• Full transparency, with a state-of-the-art information disclosure regime;

• Strong fiduciary checks and balances;

• Nimbleness and efficiency (in administration and decision-making).

Thus, the primary institutional design and arrangements must all feed into these overall objectives. This incorporates those that concern the governing body (Board), the executive capacity (Secretariat) and accountability and administrative oversight (Trustee and Monitoring & Evaluation) bodies as well as those that encompass the relationship of these institutions to the 'parent' body (in the case of the GCF, the COP).

The remaining issues to be ironed out with respect to the GCF are core to governance architecture:

• identifying clear responsibilities and lines of accountability within the Fund's structure;

• clarifying the Fund's objectives and developing the standards and metrics to assess progress towards those objectives at the level of project, programme and portfolio;

• developing environmental and social safeguard policies designed to ensure the Fund's activities 'do no harm',

• identifying, informing and empowering the constituencies that are the intended beneficiaries of the Fund's activities so that they can participate in the process of independent oversight; and

• ensuring that the oversight system has the independence, expertise and capacities necessary to carry out its assigned functions.

IX. Concluding remarks

Notably, the 2011 emerged as yet another turbulent year for capital markets. Volatility increased for energy-related commodities, including carbon, with the onset of the Arab Spring, the shutdown of nuclear power stations in Japan and Germany in the wake of the Fukushima disaster, and the downgrade of the United States' AAA credit rating .Equally relevant was the crisis of confidence that ensued as the Greek debt crisis intensified, spurred by fears that it would spread to other European Union (EU) economies and lead to a double-dip recession. Foreign direct investment (FDI) is a key financing vector and can play an important role in support of the diffusion of low-carbon technologies. Until recently, however, the potentially important role of FDI has received little systematic attention in the climate change debate. In partnership with others, the OECD is working on how to define and measure green FDI, with a view to promoting a better understanding of the contribution FDI can make to the shift to a low-carbon, climate-resilient economy and the role policies may play in the greening of FDI.

Secondly, climate funds have continually neglected gender issues and failed to incorporate a gendered perspective into programmes and projects. The climate sector often presents women as passive victims of climate change, rather than effective agents of change, ignoring women's extensive knowledge and expertise with regard to climate change mitigation and adaptation strategies. If climate funds are to be used equitably and effectively to support the different needs of men, women, boys and girls, they must incorporate gender analysis throughout project design, implementation, monitoring and evaluation. Climate funds must also recognise that women are well positioned to be agents of change through mitigation and adaptation activities in their households, workplaces, communities and governments. Global efforts to address the challenges of climate change cannot afford to ignore them.

Thus, A comprehensive approach to gender mainstreaming is required. Women's and men's concerns and experiences should be integral to the design, implementation, monitoring and evaluation of policies and programmes in all political, economic and social spheres to ensure that inequality is not perpetuated. This means that the implications for women and men of any planned climate action, including legislation, policies or programmes, in all areas and at all levels, must be assessed. To reach those who need it most, the GCF must integrate gender considerations from top to bottom.

It should: [i]**Put gender balance at the heart of the governance structures of the fund**: The governance structures of the GCF should reflect principles of gender equity through the

ambition of equal gender representation in all decision-making bodies of the fund, from the board down, and all governance structures, including the board and secretariat, should include expertise in gender issues.

[ii]Specify gender equality as a guiding principle of the fund's work:

The full integration of gender considerations must be identified as a core objective of the fund, and gender-sensitive funding guidelines and criteria – both for allocation and evaluation, including the collection of sex-disaggregated data – should be developed for each of the thematic funding areas (for example, adaptation, mitigation and forestry).

[iii] Ensure gender equality and women's leadership in the central to the development and implementation of national strategies:

Gender-specific objectives and indicators should be core components of national climate change strategies, which should be developed on the basis of the full and meaningful participation of civil society, especially that of affected and marginalised communities, including women's organisations. Any national level co-ordinating entities should have the objective of equal gender representation. Where they exist, women's ministries and gender units within all ministries need to play a more central role in climate finance, and should establish climate change action as a core element of their mandate. A systematic capacity-building process, including the necessary funds, should be available to these departments and units, as well as to national women's organizations and gender experts.

ANNEXURE

Table 1: Estimating incremental mitigation costs

Sources of estimate ($bn per annum)	2010-2020	2020	2030	Assumptions
European Commission (2009)		118		In 2005 prices ($1.25 to €1 exchange rate, total net additional ("incremental") costs, assuming successful agreement -30% reduction for developed countries by 2020 compared to 1990, and NAMAs by developing countries.
McKinsey & Co (2009)	81-113		175	In 2005 prices, includes tech R&D
Pacific Northwest National Lab (2008)			139	Taken from World Bank World development Report (2010)
UNFCCC (2007)			92-97	In 2005 prices
Project Catalyst (2009)	69-100			In 2005 prices, 450 ppm stabilisation case
G77 + China (2009)		200-400		Estimate includes both Adaptation and Mitigation. Proposed at: 0.5% to 1% of GNP of Annex II Parties
African Group (2009)		200		0.5% of GDP of Annex II Parties

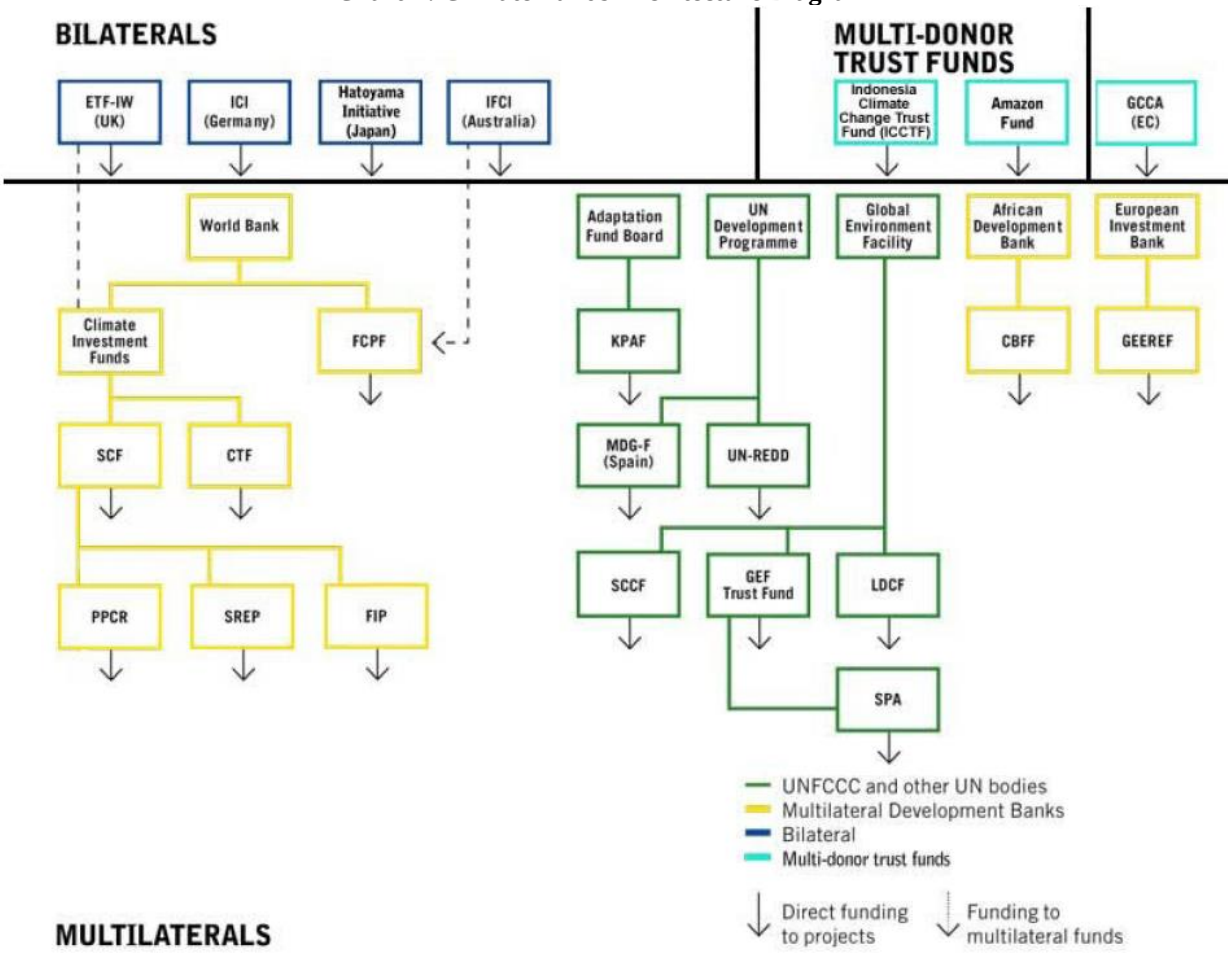
Source: Climate Funds update, available on: http://www.climatefundsupdate.org/resources/estimated-costs-climate-change

Table 2: Estimating incremental adaptation costs

Sources of estimate ($bn per annum)	2010-2015	2020	2030	Assumptions
European Commission (2009)		13-30		In 2005 prices ($1.25 to €1 exchange rate, total net additional ("incremental") costs, assuming successful agreement -30% reduction for developed countries by 2020 compared to 1990, and NAMAs by developing countries.
World Bank (2006)	9-41			450 ppm stabilisation case
Stern Review (2006)	4-37			450 ppm stabilisation case
UNDP HDR (2007)	83-105			450 ppm stabilisation case
UNFCCC (2007)			28-67	450 ppm stabilisation case
World Bank EACC (2010)			70-100	In 2005 prices, average annual costs between 2010-2050. Additional public sector costs, not costs incurred by private agents.
G77 + China (2009)		200-400		Estimate includes both Adaptation and Mitigation. Proposed at: 0.5% to 1% of GNP of Annex II Parties
African Group (2009)		>67		Estimate based on the programme for Adaptation Action under the AWG-LCA
IIED (2009)	No specific figures cited			Costs estimated to 2 or 3 times higher than UNFCCC figures

Source: Climate Funds update, available on http://www.climatefundsupdate.org/resources/estimated-costs-climate-change

Chart-1: Climate Funds Architecture Diagram



Source: Climate Funds Update, available on: http://www.climatefundsupdate.org/listing/architecture

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Examining the Relationship between Teachers' Beliefs and Perceptions of Motivation and the Students' Learning Achievements

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***Abstract*-** No doubt that motivation; on one hand, is personal and it is different from one student to another; it comes from within an individual, but it can be influenced by external factors. On the other hand, teachers as they are part and parcel of the learning process always have to bear in their minds that they can play a crucial role in positively motivating their students. Students of English sometimes reach a plateau of achievement where they feel frustrated and see no need for further improvements. Accordingly, the teachers need to do what can be done, at least by finding out what is important for their students as a means of encouraging them to step forward towards success in language learning. This research; therefore, attempts to explore how English teachers perceive motivation, to find out what factors that can increase or decrease the students' motivation according to the teachers' points of view, and find out if there is a significant relationship between motivation and the students' achievements. To find out these aims, a total number of 19 ELT teachers working in the college and Prep. School of the Near East University took part in this research by answering a questionnaire that consisted of three questions with 41 sub-questions. The questionnaire was distributed to teachers teaching in charge of English at the Near East College and Prep. School. All of the participants contributed to the research questions and none of them had any inquiry regarding the questionnaire questions. After obtaining questionnaires back within three days, the answers for each question were entered to SPSS program for analysis

Main results produced by this study are as follows:

According to the overall results, it could be concluded that motivation plays a fundamental role in language learning/teaching process. It was found that teachers had the same or similar points of view regarding the way they perceive motivation and the factors affecting students' motivation.

I. INTRODUCTION

The problem of student motivation is a universal problem for teachers. Some teachers find it more difficult to motivate some of their students. The researchers argue that being fully aware of the factors that can increase or decrease the students' motivation, will enable teachers to teach a target language more effectively.

Oxford (1994, p.12) considers motivation as one of the determining factors in success in developing a second language acquisition. Similarly, Gardner (1994, p. 361) claims that L2 motivation to be the "primary factor" in his socio-educational model. Susan and Larry (2008, p 426) considers motivation as a social-psychological factor frequently used to account for differential success in learning a second language

In a related context, the increasing spread of English in the world urges us to continuously look for effective instructional models. Motivation and its wonders are considered as one of these effective models.

Teachers' responses and interests in orientating English teaching are an essential step in the teaching process in order to meet the students' needs in language learning. Of the factors that influence student learning, motivation is surely one of the most potent.

Having noticed; on one hand, the tendency and interest of teachers to positively and effectively motivate their students by all possible means, the authors desire to explore this issue to identify the specific factors that raise and decrease motivation in language learning as well as to find out what techniques teachers use in order to increase their students' motivation. On the other hand, Cypriots increasingly learn English because of its importance as an international language, their continual negotiations with the other side of country, the increasing use of English in the daily life and as a means of coping with the international development. Meanwhile, little empirical evidence is available to understand motivational patterns in Cyprus. Therefore, motivation should be thoroughly investigated.

Teachers' responses and interests in orientating English teaching are an essential step in the teaching process in order to meet the students' needs in language learning. It is known that teachers can either positively or negatively affect the students' motivation.

The intent of this research is to emphasize the importance of motivation; and to explore the role of motivation in language learning, according to teachers. Bearing in mind this general aim, this research is set out to answer the following questions:

1.1 RESEARCH QUESTIONS:

2. How do English language teachers perceive motivation?
3. How should teachers act to explore for possible solutions to increase learner's motivation?
4. What are the factors that raise and decrease motivation in language learning?
5. Is there a significant relation between motivation and achievements of students in language learning studies?

II. LITERATURE REVIEW

2.1 Concept of motivation

Despite the importance of motivation that is emphasized by all the researchers and teachers since it is connected to the human nature, it is not easy to find a standardized definition for motivation. However, Pintrich & Schunk (2002, p.5) state that the word is of Latin root "movere" which means "to move". Such a clear meaning of the word root encourages various definitions of motivation in teaching/learning a language. Jeremy Harmer, in his book "the practice of English language teaching, (4th edition) defines motivation at its basic level as "some kind of internal drive which pushes someone to do things in order to achieve something".

Longman dictionary (third edition) defines motivation as the driving force in any situation that leads to action.

Moreover, motivation can be an internal state or condition that serves to activate or energize behavior giving it direction. It can also be a process that starts with a need that activates behavior which in turn moves someone towards achieving a goal. In an educational context, motivation is the combination of effort plus desire to achieve the goal of learning the target language. Motivation is the single most important factor in the practical realization of language learning. It also plays an important and crucial role in language learning. Furthermore motivation is one of the keys to successful language learning. Sucuoğlu (1997) states that motivation is an internal state or condition that activates behavior and gives it direction; a desire or want that energizes and directs goal-oriented behavior and an influence of needs and desires on the intensity and direction of behavior. In addition to these, motivation is typically defined as the forces that account for the arousal, selection, direction and continuation of behavior.

Furthermore, Pintrich & Schunk (2002, p.5) define motivation as " the process whereby goal-directed activity is instigated and sustained".

Grander (1994, p.10) defines motivation as the combination of effort plus desire to achieve the goal of learning the language. Thus, it can be concluded that a motivated individual shows "favorable attitudes toward learning the language" (Gradner,1994, p 361). Such definition indicates the essential influence of motivation on language learning.

Gamon, and Emeritus (2001) pointed out that motivation influences how and why people learn as well as how they perform. Dr. Ibrahim Elfiky (1999), in his book "ten keys to ultimate success, p. 17), states that motivation "the driving force of human behavior" is the first key that leads to success.

In a related context, Gardner (1985, p. 10) defines motivation as "the extent to which the individual works or strives to learn the language because of a desire to do so and satisfaction experienced in this activity". Gardner also perceives motivation as consisting of four elements: a goal, the desire to attain the goal, positive attitudes toward learning the language and effortful behavior to that effect. Moreover, Gardner (1994) states that a specific language motivation in the combination of effort plus desire to achieve the goal of learning language. Thus, motivated individual shows "favorable attitudes toward learning the target language". Similar to Gardner's approaches about motivation, Pintrich and Schunk (2002) refer to motivation as "the process whereby goal-directed activity is instigated and sustained".

In contrast, Oxford (1994, p. 14) thinks that "motivation reflects the power to attain the goal". Furthermore, she claims that "this power stems from the desire to attain the goal, positive attitudes toward learning the language and effortful behavior".

2.2 Types of motivation

Two types of motivation have to be distinguished; intrinsic "integrative" and extrinsic "instrumental" motivation.

Gardner (1985) and Gadner & Lambert (1972) distinguished two different types of motivation; integrative and instrumental. When a learner is interested to integrate himself/herself with people and their culture, this means that s/he is integratively motivated. But when a learner is interested in learning a language as a means of attaining a higher social or economic status, or getting a better career, this means s/he is instrumentally motivated. Ely (1986) argued that it is not always easy to distinguish between integrative and instrumental motivation claiming that it's possible to find other reasons. Sucuoglu (1997) proposed that some students; for example, study as a result of the pressure imposed on them by their parents, teachers, and so forth.

As quoted in Wei Ming's article (2007), according to Brown (2000, p. 162), instrumental motivation, refers to "acquiring language as a means for attaining instrumental goals", while integrative motivation, according to (Gardner et al., 1992, p. 198)"stems from a desire to understand the language and culture of another group for the purpose of interaction".

Similarly, (Dörnyei, 1994, p. 275) distinguished between the two kinds of motivation behaviors; extrinsically and intrinsically behaviors. First, extrinsically motivated behaviors are the ones that the individual performs to "receive some extrinsic award; for example, good grades or to avoid punishment; while intrinsically motivated behaviors are internal rewards; for example, the joy of doing a particular activity or satisfying one's curiosity" .

III. METHODOLOGY

3.1 Participants

There are 19 subjects in total, twelve female teachers and four male ones. All of the participants were randomly selected and asked to answer a questionnaire at their schools in the Near East University except three female teachers who answered the questionnaire by the internet. All questions were interested in motivation, in how teachers perceive motivation, in the factors that can affect the students' motivation and in what teachers can do to increase their students' motivation. Regarding the female teachers, their age varied from 23 to 50 years old. In regard to the male teachers, their age varied from 27 to 37 years old. Thirteen of the participants work in the secondary school of the NEU college, four of them work in the elementary school of the NEU and one of them work in both elementary and secondary school. While answering the questionnaire, all the participants, except three of them, were present and all contributed to and participated in answering the questionnaire.

3.2 Materials

A questionnaire consisting of three background questions and 41 main questions was administered to teachers. The questions (statements) were answered according to a Likert-scale, so that respondents pointed out their opinions depending on the scale. Two scales were used; one based on agreement level of respondents and the other one was a frequency scale.

3.3 procedures

First, the questionnaire designed, sent to a friend working as English teacher in the Near East University College and Prep. School and therefore, distributed to teachers teaching in charge of English at the Near East University College and Prep. School. After obtaining questionnaires back within three days via the email, the answers for each question were entered to SPSS program for analysis.

3.4 Data analysis

Quantitative data were collected through the questionnaire administered to teachers. The analysis process will consist of quantitative procedures in evaluating the data by using the SPSS program (15). The quantitative forms of data will consist of forms of data collected from the questionnaire and research questions answered and processed according to their relevance with each other, and in accordance with the research objectives and research questions. Both comparative and descriptive statistics were used to present the analyzed data. Mean, standard deviation and significance level of means were considered across the data analysis procedures.

Regarding age, in general all answers are homogenous and there is no significant difference between the mean values of age groups.

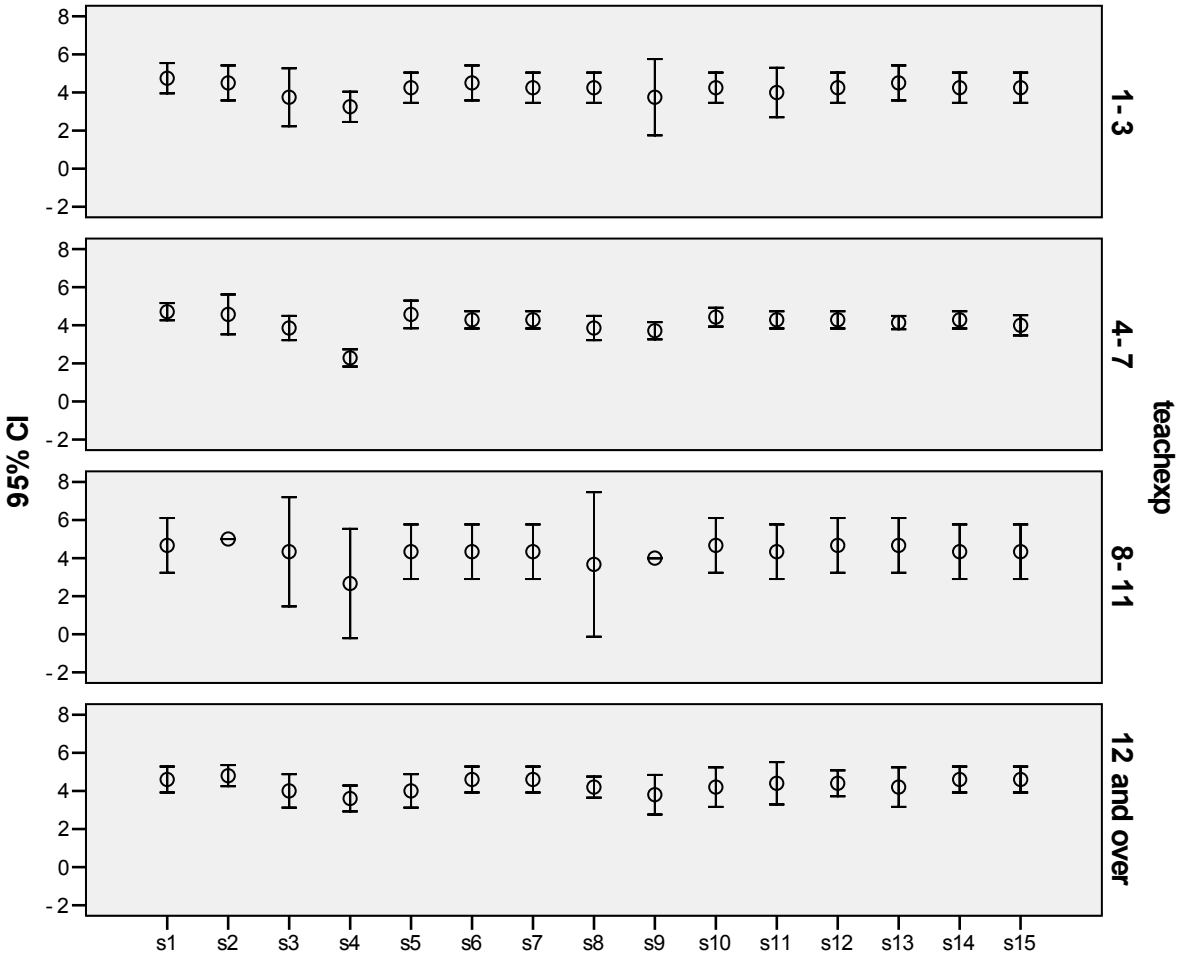
Teaching experience

Statement	Age Groups	N	Mean	Mean difference	Std. Deviation	Sig.
4	4-7 12-over	7 5	2.2857 3.6000	-1.31429*	.48795 .54772	.018

**P* value is significant at .050

In this study, statistical figures such as mean, mean difference, standard deviation are used to reveal the statistical values of responses.

ANOVA tests were carried out to find out whether there were any significant differences between the responses of respondents. The only significant difference was found out in the responses of respondents for Statement 4, between the teaching experience group 4-7 and 12-over ($p = .018$). This indicates that respondents who have a teaching experience between 4-7 years disagree that motivation is more important than intelligence, whereas the teachers having 12 years or more experience in teaching almost agree with this statement (3.6000).



The graph shows the responses of teachers to statements from 1 to 15. As is observable in the second row of Graph 1, the majority of the respondents agree about all the statements while those whose teaching experiences vary between (8-11 years) have differing opinions. For more illustration, it can be concluded from the table that groups' responses in (1-3), (4-7), and (12 and over) are consistent unlike group (8-11) whose responses show variations.

ANOVA test was processed to find out if there are any significant differences between the responses of respondents according to their age groups. However, the ANOVA test results revealed that there were no significant differences between the responses of teachers in terms of their age groups. In this case, it could be suggested that all teachers share the same or similar opinions about the issues mentioned in the statements.

Teaching experience

Statement	Teaching Experience	N	Mean	Mean difference	Std. Deviation	Sig.
4	4-7 12 and over	7 5	2.2857 3.6000	1.31429	.48795 .54772	.018

**P* value is significant at .050

According to the teaching experience, the only significant difference was found to be in statement4 between two groups. The group which has teaching experience between 4 and 7 years and the group which has teaching experience over 12 years. However, this difference does not affect the overall results as all other responses are in coherence. All responses more or less are very close to each other in terms of age and teaching experience, so that there is no statistically significant difference between the responses.

IV. Results and Discussion

In this section, research questions will be answered according to the findings obtained throughout the questionnaire. The findings, processed through SPSS 15, will help to answer the questions and carry out the discussion.
The first research question is as follows:

How do English teachers perceive motivation?

Table 1

Stat. No.	statement	N	1- Strongly disagree 2- Disagree 3- Neutral 4- Agree 5- Strongly agree	Frequency	Percentage	Mean	Std. D
1.	there is a significant relation between motivation and achievement of students in language learning	19	Agree Strongly agree Total	6 13 19	31.6 68.4 100.0	4.6842	.47757
2.	Motivation is important for success in language learning	19	Disagree Agree Strongly agree Total	1 3 15 19	5.3 15.8 78.9 100.0	4.6842	.74927
3.	People's past success in language learning is an important factor as it enhances their motivation for future studies in language learning	19	neutral Agree Strongly agree total	6 8 5 19	31.6 42.1 26.3 100.0	3.9474	.77986
4.	Motivation is more important that intelligence in language learning.	19	disagree Neutral Agree total	7 7 5 19	36.8 36.8 26.3 100.0	2.8947	.80930
5.	As a teacher, I think that motivation enhances language learning.	19	neutral Agree Strongly agree Total	2 9 8 19	10.5 47.4 42.1 100.0	43158	.67104
6.	Positive assessment motivates the students.	19	agree Strongly agree Total	11 8 19	57.9 42.1 100.0	4.4211	.50726
7.	Positive motivation makes language learning easier.	19	agree Strongly agree Total	12 7 19	63.2 36.8 100.0	4.3684	.49559
8.	Success of students always depends on their motivation and eagerness to learn.	19	disagree Neutral Agree Strongly agree Total	1 2 12 4 19	5.3 10.5 63.2 21.1 100.0	4.0000	.74536
9.	Group work can motivate students more in language learning.	19	disagree Neutral Agree	1 4 12	5.3 21.1 63.2	3.7895	.71328

			Strongly agree Total	2 19	10.5 100.0		
10.	Motivation plays an crucial and important role in language learning	19	neutral Agree Strongly agree Total	1 10 8 19	5.3 52.6 42.1 100.0	4.3684	.59726
11.	If students are not motivated, learning process becomes much less efficient.	19	neutral Agree Strongly agree Total	2 10 7 19	10.5 52.6 36.8 100.0	4.2632	.65338
12.	Motivation makes learning easier.	19	agree Strongly agree total	12 7 19	63.2 36.8 100.0	4.3684	.49559
13.	Motivation gives a sense of achievement.	19	neutral Agree Strongly agree total	1 11 7 19	5.3 57.9 36.8 100.0	4.3158	.58239
14.	Motivation increases learners' performance	19	Agree Strongly agree Total	12 7 19	63.2 36.8 100.0	4.3684	.49559
15.	Motivation makes learners happy.	19	neutral Agree Strongly agree Total	1 12 6 19	5.3 63.2 31.6 100.0	4.2632	.56195

Frequency test has been conducted in order to find out how teachers perceive motivation in language learning.

According to table 1, most of the teachers in statement1 share common definitions and ideas in the way they perceive motivation. 68.4% strongly agree that there is a significant relation between motivation and achievements of students in language learning and 31.6% agree and adopted the same idea. In addition, most of the teachers share the same or similar ideas when they have been told that motivation is important for success in language learning; 78.9% strongly agree and 15.8% agree. This finding supports the idea Gamon, and Emeritus (2001) discuss when they point out that motivation influences how and why people learn as well as how they perform. Regarding statement 3, the respondents' opinions varied about whether the people's past success in language learning is an important factor as it enhances their motivation for future studies in language learning. 31.6% prefer to be neutral and only 8% and 9% strongly agree and agree. However, none of them reject this idea. In statement 4 as shown in the table, 36.8% of the respondents disagree that motivation is more important than intelligence while other 36% are neutral. Since 5% of the respondents view the opposite and none of them strongly agree, it can be concluded that such an interesting finding shows that it is more probable that intelligence is more important that motivation. In statment5, 47.4% agree that motivation enhances language learning and 42.1% strongly agree about that. Such a percentage clearly shows that the majority of the respondents adopt the role of the motivation in enhancing the process of language learning, and evoking the learners' positive attitudes towards learning. Such a finding has been tackled before by Grander when he states that motivated individual shows "favorable attitudes toward learning the language" (Gradner, 1994, p 361). Therefore, it could be concluded that motivation has its essential influence on language learning. In addition, all the participants in statement 6 share the same or similar opinions. All of them believe that positive assessment motivates the students; 57.9% agree; 42.1% strongly agree in regard to the role positive assessment plays to motivate the students while none of the respondents disagree. Similarly, all the respondents, in statement 7, share the same point of view as well. All of them believe that positive motivation makes learning language easier. In statement 8, most of the respondents share the same idea; 63.2% and 21.1% of them agree and strongly agree that success of students always depends on their motivation and eagerness to learn. 10.5% of the respondents are neutral and only 5.3% disagree. Such a good response confirms what Grander (1994, p.10) tackles when he defines motivation as the combination of effort plus desire to achieve the goal of learning the language. As regards statement9, most of the respondents "63.2%" agree that work group motivates students more when learning a language. Only 5.3% of them disagree. In regard to statement 10, the majority of the respondents believe that motivation plays a crucial and important role in language learning where 52% of them agree and 42.1 strongly agree. Similar positive responses are found in statement 11 where 52% and 36.8% of the respondents agree and strongly agree that learning process becomes much less efficient if students are not motivated. It is worth notably here that none of the respondents denied this idea. Regarding statements 12,13,14, and 15, the majority of the respondents share the same or similar opinion. Most of them believe that motivation makes learning easier, gives a sense of achievement, increases the learners' performance and makes learners happy. Accordingly, such interesting findings have been tackled by Pintrich & Schunk (2002, p.5) when they define motivation as " the process whereby goal-directed activity is instigated and sustained". They have been also discussed in brief by Jeremy Harmer, in his book "the practice of English language teaching, p. 98, (4th edition) "some kind of internal drive which pushes someone to do things in order to achieve something".

The second question is as follows:

How should teachers act to explore for possible solutions to increase learner's motivation?

Table 2

Stat. NO.	statement	N	1.never 2.rarely 3.sometimes 4.often 5.always	Frequency	Percentage	Mean	Std. D
16.	Using visual materials	19	rarely Sometimes Often Always Total	3 10 4 2 19	15.8 52.6 21.1 10.5 100.0	3.2632	.87191
17.	Getting all the students involved in task learning	19	Sometimes Often Always Total	1 7 11 19	5.3 36.8 57.9 100.0	4.5263	.61178
18.	Choosing interesting topics	19	often always total	10 9 19	52.6 47.4 100.0	4.4737	.51299
19.	Using gestures to describe materials	19	rarely Sometimes Often Always Total	1 1 10 7 19	5.3 5.3 52.6 36.8 100.0	4.2105	.78733
20.	Using inductive ways of teaching such as indirect teaching of grammar	19	rarely Sometimes Often Always Total	3 7 7 2 19	15.8 36.8 36.8 10.5 100.0	3.4211	.90159
21.	Asking students to produce language on their own initiative	19	sometimes Often Always Total	12 6 1 19	63.2 31.6 5.3 100.0	3.4211	.60698
22.	Following a student-centered teaching/learning practice	19	rarely Sometimes Often Always Total	2 9 5 3 19	10.5 47.4 26.3 15.8 100.0	3.4737	.90483
23.	Following a teacher-centered practice	19	rarely Sometimes Often Always Total	1 8 8 2 19	5.3 42.1 42.1 10.5 100.0	3.5789	.76853
31.	I develop the students' cognitive or mental abilities	19	Sometimes Often Always Total	3 10 6 19	15.8 52.6 31.6 100.0	4.1579	.68825
32.	I raise the learners'	19	Rarely	1	5.3	3.8947	.80930

	general awareness about the different ways languages are learnt and the number of factors that can contribute to success	19	Sometimes Often Always total	4 10 4 19	21.1 52.6 21.1 100.0		
33.	I focus on motivational flow and not just the informational flow in my class	19	Rarely Sometimes Often Always total	1 8 4 6 19	5.3 42.1 21.1 31.6 100.0	3.7895	.97633
34.	I promote cooperation instead of competition	19	sometimes Often Always total	4 11 4 19	21.1 57.9 21.1 100.0	4.0000	.66667
35.	I monitor students' progress	19	sometimes Often Always total	4 4 11 19	21.1 21.1 57.9 100.0	4.3684	.83070
36.	I encourage learners to personalize the classroom	19	rarely Sometimes Often Always Total	2 5 6 6 19	10.5 26.3 31.6 31.6 100.0	3.8421	1.01451
37.	I indicate to my students that I believe in their efforts to learn and their capability to complete the tasks	19	sometimes Often Always Total	3 11 5 19	15.8 57.9 26.3 100.0	4.1053	.65784
38.	I try and promote interaction, cooperation and sharing of genuine personal information among learners.	19	rarely Sometimes Often Always total	1 4 7 7 19	5.3 21.1 36.8 36.8 100.0	4.0526	.91127
39.	I provide regular feedback about the areas which they should particularly concentrate on.	19	rarely Sometimes Often Always Total	1 5 6 7 19	5.3 26.3 31.6 36.6 36.8 100.0	4.0000	.94281
40.	I use an inductive way of teaching.	19	never Rarely Sometimes Often Always Total	1 2 5 9 2 19	5.3 10.5 26.3 47.4 10.5 100.0	3.4737	1.02026

In statement 16, according to this frequency test, 52.6% of the respondents claim that they sometimes use visual materials; 21.1 % of them often use visual materials; 10.5 always use them; while 15.8 rarely use the visual material. Therefore, it can be concluded that respondents share varying responses in regard to using visual materials as a means of increasing students' motivation.

According to the results revealed in table2, all the participants, in statement 17, share the same or similar ideas; they believe that getting all the students involved in task learning increases students' motivation. Similarly, in statement 18, similar positive responses have been found. All the participants share nearly the same idea that choosing interested topics increases the students' motivation; 52.6% and 47.4% often and always choose interesting topics when teaching. None of the participants disagree with this idea. As regards statement 19, 52.6% claim that they use gestures to describe the materials used in learning and 36.8% claim that they always do that while 5.3% claim that they rarely do that. In statement 20, respondents share varying point of views where 36.8% claim that they either sometimes or

often use inductive ways of teaching such as indirect grammar teaching while 15.8 % of them claim that they don't do that. In statement 21, respondents' response shows that they believe in asking students to produce language on their own initiative as a means of increasing their (students) motivation; 63.2% of them claim that they sometimes encourage their students to produce language on their own initiative while 31.6 claim that they often do that. It is worth notably according to the table that none of the respondents excludes this idea. Regarding statement 22 and 23, the respondents shared different opinions. As for statement 22, 47.4 of them claim that they follow a student- centered teaching/learning practice while 10.5 of them claim that they rarely follow that format "student-centered teaching". Similarly for statement 23, 42.1% of the respondents claim that they often and sometimes follow a teacher- centered teaching while 10.5% claim that they always follow this format "teacher-centered teaching". Although the percentage of the student-centered format percentage is not low, , it can be concluded that teacher-centered teaching is probably more used in the schools according to the findings in statement 22 and 23. That might be because none of the respondents work as university teachers; all of the teachers involved in this study work in the elementary, secondary, prep., and elementary schools of the Near East University. Furthermore, the majority of the participants in statement 31, adopt the idea that they develop the students' cognitive abilities; 31.6% in addition to 52.6% of them claim that they always and often develop their students' cognitive abilities. In statement 32, 52.6% of the participants claim that they often raise the learners' general awareness about the different ways languages are learned while a few of them (5.3%) claim that they rarely do that. One interesting finding is seen in statement 33 where 42.1% of the respondents claim that they sometimes focus on motivational flow and not just the informational flow in their classes. Regarding statement 34, 57.9% of the participants claim that they often promote cooperation rather than competition. In statement 35, the majority of the participants share nearly similar point of view where 57.9% of them claim that they monitor their students' progress, 21.1% of them claim that they often adopt this idea and other 21.1% of them claim that they sometimes do that. For sentence 36, most of the participants share similar points of view. According to the data analysis revealed in table3, 31.6 of the participants adopt the idea that they encourage their students to personalize the classroom; 31.6% of them claim that they always do that, other 31.6 of them claim that they sometimes do that, and 26.3% of them point out that they sometimes do that while a few participants (10.5%) claim that they rarely encourage their students to personalize the classroom. For statement 37, all participants share similar ideas; 57.9% of them often indicate to their students that they "teachers" believe in their students' efforts to learn the language and their capability to complete the learning tasks. In regard to statement 38, the participants shared similar opinions where 36.8% claim that they always promote interaction, cooperation and sharing of genuine personal information among their students. Similar responses are found in statement 39 where 36.8% of the participants claim that they provide feedback for their students about the areas that students should particularly focus on, 31.6% of the participants often do that, and 26% of them sometimes do that. For statement 40, the participant share varying opinions; 47.4% of the participants claim that they use inductive way of teaching; 26.3% claim that they sometimes do that; 10.5% of them claim that they always do that; other10.5% of them claim that they rarely do that; while 5.3% of them claim that they never do that.

The third research question is as follows:
What are the factors that raise and decrease motivation in language learning?

Frequency test has been conducted in order to find out what factors that raise and decrease motivation in language learning.

Table 3

Stat. No.	Statement	N	1.strongly disagree 2.disagree 3.agree 4.strongly agree	frequency	Percentage	Mean	Std. D
24.	Considering that psychology affects students' motivation	19	Sometimes Often Always Total	2 9 8 19	10.5 47.4 42.1 100.0	4.3158	.67104
25.	Considering the age as a factor in the students' motivation	19	rarely Sometimes Often Always Total	1 2 10 6 19	5.3 10.5 52.6 31.6 100.0	4.1053	.80930
26.	Considering that culture affects students' motivation	19	sometimes Often Always Total	6 10 3 19	31.6 52.6 15.8 100.0	3.8421	.68825

27.	Using authentic materials.	19	rarely Sometimes Often Always 44.00 Total	1 8 4 5 1 19	5.3 42.1 21.1 26.3 5.3 100.0	5.8421	9.28717
28.	Being friendly with the student	19	Sometimes Often Always total	2 8 9 19	10.5 42.1 47.4 100.0	4.3684	.68399
29.	Choosing interesting topics	19	Sometimes Often Always 44.00 total	4 7 7 1 19	21.1 36.8 36.8 5.3 100.0	6.2632	9.17025
30.	Considering students' psychological mood	19	sometimes Often Always Total	3 12 4 19	15.8 63.2 21.1 100.0	4.0526	.62126

According to the results revealed in Table 3, all the participants share the same or similar points of view in statement 24; 47.4% of the participants claimed that they often consider psychology as a factor influencing students' motivation in addition to 42.1% of them who stated that they always consider psychology as a factor influencing the students' motivation. It is worth notably hat none of the participants excluded or refused this factor. Analysis of data indicates that the majority of the respondents (52.6%) in addition to 31.6% are actually often or always considers age as a factor that affects students' motivation. In statement 27, the participants share varying opinions; most of them (42.1%) sometimes use authentic materials and 26.3% always adopt the same idea while a few other respondents (5.3%) rarely use the authentic materials in teaching. Moreover, it is noted in statement 28 that the majority of the participants believe in being friendly to the students; 47.4% of the participants claim that they are always friendly to their students and 42.1% claim that they are often friendly to the students. It is worth notably that none of the participants excludes the importance of this factor. Regarding statements 29 and 30, it is noted that the majority of the participants share the same or similar ideas about choosing interesting subjects and considering the students' psychological mood. 36.8% claim that they always choose interesting subjects and other 36.8% claim that they often choose interesting subjects. For statement 30, the majority of the participants (63.2%) claim that they often consider the students' psychological mood in addition to other (21.1%) of the participants who claim that they always consider that.

The fourth research question is as follows:
Is there a significant relation between motivation and achievements of students in language learning?

Frequency test has been conducted in order to find out if there is a significant relation between motivation and achievements of students in language learning.

Table 4

Stat. No.	Statement	N	1.never 2.rarely 3.sometimes 4.often 5.always	Frequency	Percentage	Mean	Std. D
41.	There is a significant relation between motivation and achievements of students in language learning.	19	sometimes Often Always total	4 4 11 19	21.1 21.1 57.9 100.0	4.3684	.83070

Stat. No.	statement	N	1.never 2.rarely 3.sometimes 4.often 5.always	Frequency	Percentage	Mean	Std. D
1.	There is a significant relation between motivation and achievements of students in language learning.	19	Agree Strongly agree Total	6 13 19	31.6 68.4 100.0	4.6842	.47757

As the data analysis revealed in table 4 show, 57.9% of the participants believe that there is always a significant relationship between motivation and achievements of students in language learning; 21.1% of them claim that often such relation is existed and other 21.1% of them claim that sometimes such a relation takes place.

Moreover, the data analysis of data revealed in table 1 indicates that most of the teachers in statement1 share common definitions and ideas in the way they perceive motivation. 68.4% strongly agree that there is a significant relation between motivation and achievements of students in language learning and 31.6% agree and adopted the same idea.

Since none of the participants disagree that the relation is existed, it can be concluded that there is a significant relationship between motivation and achievements of students in language learning.

V. Conclusion

Based on the findings section, most language teachers agree that motivation of the students is one of the most important factors influencing their success or failure in learning the language. Furthermore, they are aware of many useful steps towards increasing their students' motivation.

Learning and exploring is a continuous process where in every attempt some more issues relevant to research field can be discovered. Thus, further research is necessary and definitely beneficial to enrich the findings relevant to this area of education. However, so far carried out relevant research also reveals that motivation and success/achievement of students are strictly relevant to each other and usually, the higher the motivation the higher the achievement level of students.

As this study was carried out in a sufficient time, it could be claimed that all procedures and steps required for a successful research were carried out with nearly hundred per cent accuracy; however, all attention necessary for a successful study was paid to results and procedures to fulfill the purpose of the study.

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Digital Ordering System for Restaurant Using Android

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Abstract- Nowadays web services technology is widely used to integrate heterogeneous systems and develop new applications. Here an application of integration of hotel management systems by web services technology is presented. Digital Hotel Management integrates lots of systems of hotel industry such as Ordering System Kitchen Order Ticket (KOT), Billing System, Customer Relationship Management system (CRM) together. This integration solution can add or expand hotel software system in any size of hotel chains environment.

This system increases quality and speed of service. This system also increases attraction of place for large range of customers. Implementing this system gives a cost-efficient opportunity to give your customers a personalized service experience where they are in control choosing what they want, when they want it – from dining to ordering to payment and feedback.

We are implementing this system using android application for Tablet PC's. The front end will be developed using JAVA Android and the backend will work on MySQL database.

Index Terms-

DFD: Data Flow Diagram.
DOSRUA: Digital Ordering System for Restaurant Using
KOT: Kitchen Order Ticket
Android
UML: Unified Modeling Language.

I. INTRODUCTION

Restaurants are one of the favorite premises. With no regard to the actual reasons for visiting restaurants, customer will make orders and wait for the ordered meals. However, it is common if customers complain for not feeling satisfied about the services offered.

There are many reasons leading to the feeling of dissatisfaction including being entertained late in terms of order taking by the waiter and meals serving. The issue of being late entertained could be solved with help of the advancement in the technologies of communication. In accordance, this study initiates an integrated and networked system, with the focus is on its ability to solve the above described limitations in order taking.

This study names the system as Digital Ordering System for Restaurant Using Android (DOSRUA).In definition, DOSRUA is an integrated system, developed to assist restaurant management groups by enabling customers to immediately make orders on their own selves. This will minimize the number of minutes to wait for the meal serving.

Fig1.1

This project deals with Digital ordering system for restaurant. This topic includes scope of the project, project characteristics, Operating environments, Assumption and dependencies, design and implementation constraints. Scope of the project includes features that can be implemented. Design part includes the method and way of designing the product. It also explains certain constraints on designing and implementation.

II. RELATED WORK

The existing system is paper based. The traditional menu cards in the restaurants are paper based. Waiters use paper to write the order of customers. The records are stored on paper. As with anything paper based, it is so easy for things to get damaged by Coffee stains etc, or paper being lost due to fire or accidents or just generally lost. There is wastage of time, money, and paper. As traditional menu cards are paper based, any changes that need to be made in the menu card will lead to wastage. As it will require reprinting of all the menu cards. Also, for small changes it is not possible to print all the menu cards again and again. There is no power to dynamically make any changes in the menu card. To access a particular record from the stack of papers is not efficient. From the customer's point of view, this system is time consuming. As, one has to wait until the waiter comes to take the order, one has to call waiter number of times till he notices it, there can be misinterpretation while the waiter is writing your order on paper, and it might be possible that you are served with a wrong dish. There has been improvements in the management

of restaurants. Each waiter is assigned a group of tables, after taking orders for a table the waiters enter the orders (a list of dishes and drinks ordered by the diner or group of diners) into the system at the PC. The waiter usually knows of any dishes that are unavailable before taking an order. The system must confirm the availability of dishes. Should an item not be available the system must allow the waiter to change or even delete a customer's order. Dishes to be prepared are sent to the kitchen, drinks orders to the bar. Starters and main course orders are usually taken together. Drinks and desert orders may be taken separately. Kitchen staff sees the dish orders on their screen, prepare them in an appropriate sequence and confirm preparation to the system when complete, similarly with the bar. When a waiter sees the completion indications on his terminal he collects the items and takes them to the table. The waiter can also check on the status of dish and drink orders. At the end of the meal the waiter will have the system print a bill, and he will enter the details of payment for it. The management can give discounts. The system keeps track of the numbers of customers served by each waiter and the amount of money taken by each waiter. The management can view these statistics [2]. The next advancement was "QORDER": the portable ordering system for Android devices. Here the waiter no longer approaches the table with his notepad, but rather with the QOrder hand held device. He enters order information on the touch screen and then sends it to the kitchen in real time for processing. Simultaneously, your POS system receives the sales information for later billing. QOrder utilizes WIFI to easily reach to your most remote corner spot in your establishment. Once the guests wish to leave, the waiter prints the receipt out on his belt printer and processes payment with the handheld unit much like he would on the POS system [3].

But there are still many areas which are not closely looked at. Like, making dynamic changes in the menu card, to get rid away from the heap of paper based records, to assure the customer that he'll be served with what he has ordered, to get the customer feedback on record.

Some of the existing system's are mentioned below:

- PixelPoint

PAR PixelPoint Company uses this software for managing the restaurant. The system consists of the company's software and hardware. This network system is compatible to TCP/IP, enabling information sending through both wireless and conventional networks [4].

- LRS Restaurant Server Pager Starter Kit

This system improves the food-ordering service quality in restaurants and reduces the waiting time of clients. The on-site paging system is used at UHF frequency or the frequency range of 467 MHz for sending the order data [5].

- Billpro Pocket® and Billpro POS for Restaurant

This system receives a client's order and makes a list by means of the designed client's template in the kitchen. The food ordering device is portable. The waiter takes the client's order and sends it to the client's template in the cook room[6].

- Implementation of Network-based Smart Order

System The Smart Order System in Restaurants (SOSIR) has been modified to take order from the client's table through RS-232 signal, which is sent to the cashier counter. The cashier counter system is connected to a database. When the clients' orders are sent the cashier counter system will screen and prioritize the orders before sending the information to the kitchen for the chef to cook [1].

Table 2.1: Comparison of Existing Systems.

	Pixel Point	LRS	Smart Order	DOSRUA
Wireless Network	YES	YES	NO	YES
Touch Screen	YES	NO	NO	YES
Digital Menu	NO	NO	NO	YES
Status of Order	NO	NO	NO	YES
Group Order	NO	NO	YES	YES

III. PROJECT SCOPE

In current formal dining environments, some form of physical static menu is utilized to convey the available food and beverage choices to customers. Said menus are generally paper based and hence impose restrictions on the textual real estate available and the ability a restaurateur has to update them. This document specifies the requirements for a restaurant paper menu and ordering replacement strategy to alleviate the problems associated with the current archaic method. Three related concepts are encompassed by the general scope of the Restaurant Menu and Ordering System. The first pertains to the replacement of paper-based menus using an electronic format, the second relates to a complementary electronic strategy for the front of house handling of a customer's order and the third surrounds the process of transferring said electronic orders to the kitchen for preparation. It should be noted that while the suggested strategy incorporates the use of various hardware components, the primary focus of the presented SRS relates to the constituent software elements.

The following are the features which can be a part of the proposed system:

Ordering, Waiting, Billing, Table Reservation, Home Delivery, KOT, Advertisement.

3.1 User Classes and Characteristics

The end-users of the DOSRUA fall into three primary categories, unskilled, partly skilled and highly skilled. Unskilled user: The users of the tablets at the table are walk-in customers and should therefore be assumed to have no relevant prior skills

or education other than basic abilities to operate an automated system; no more complex than a mobile phone. Partly skilled user: The users of the tablets and displays are managers and chefs respectively and they should be able to use the system and further be able to train others with minimal training themselves. They must be able to explain all elements of the user interfaces except the server. Supervisors also fall into the same category, though they will have to learn other sections of the system (refunds etc); these should not be of notably greater complexity than the standard functions. This class of user would be expected to have a high-school certificate education or equivalent. Highly skilled user: The initial installation and configuration of hardware and the constituent proposed system components (especially the server) is guaranteed to require someone with notable computer experience, including extensive experience with network and operating systems to complete it. The software should not be needlessly complex, but it is still expected not to be entirely 'plug and play'. This class of user is expected to have a graduate certificate or equivalent, as well as extensive computer experience.

3.1.1 Operating Environment

Android Operating system is an open source operating system. There are thousands and thousands of developers are there at sites trying to make android a better a operating system. There are so many eyeballs looking over the code every day. So the loopholes are quickly patched and fixed. Therefore android is secured. It always encourages your creativity. Unlike the iphone OS, Android user interface has been constantly refining and over the years. With Android 4.0 Google has made the user interface much more polished and modern. Apple charges people who want to develop applications for the App store $100/year, while Google only charges Android developers $25. So android prevails.

3.1.2 Design and Implementation Constraint

The proposed system should be written in an object oriented language with strong GUI links and a simple, accessible network API. Front end can be designed by using Rapid Application Development Tool (Indigo Eclipse). The system must provide a capacity for parallel operation and system design should not introduce scalability issues with regard to the number of tablets or displays connected at any one time. The end system should also allow for seamless recovery, without data loss, from individual device failure. It is worth noting that this system is likely to conform to what is available. With that in mind, the most adaptable and portable technologies should be used for the implementation. The system has criticality in so far as it is a live system. If the system is down, then customers must not notice, or notice that the system recovers quickly (seconds). The system must be reliable enough to run crash and glitch free more or less indefinitely, or facilitate error recovery strong enough such that glitches are never revealed to its end-users.

3.1.3 Assumptions and Dependencies

The implication is that the target hardware will provide a capacity for standalone program/application deployment and not require customized embedded firmware to be written. It is further assumed that tablet PCs of sufficient processing capability and battery life will be utilized. The SRS assumes that none of the constituent system components will be implemented as embedded applications. The surface computers employed by the system should facilitate being utilized/left on for extended periods (sufficient for daily use) and that they are programmable in the same fashion asx86 architecture computers. Finally, it is further assumed that the deployment environment is capable of supporting an IEEE 802.11 wireless network for system communication. The maximum distance of transmission is within 50-100 meters, about the range of Wi-Fi.

3.2 System Features

Tablet on table:-

- There will be a tablet on each table.
- This will allow the customers to browse the food items for the time they wish.
- This will allow the customers to browse the food items the way the customer wish.

Customer feedback:-

- Customer can enter the feedback about the service and the food served.
- This helps the Restaurant owner to analyse the service and make necessary changes if needed.
- This also helps the Customer's to decide a particular food item with a positive feedback.

Searching Item:-

- Customer can search a particular food item according to name, price, category etc.
- This saves a lot of time of customer to order an item.

Offers for Customer:-

- The Restaurant owner can post various offers on tablet.
- This will help the customer as well as the restaurant owners.

Attractive Presentation:-

- The Menu is organized in an attractive way.
- There are images of every food item which will make the view of customers more clear about how the food will look like after delivery.
- here is an attractive use of Various themes and colour schemes.

Sorting an Item:-

- The food items will be sorted according to price, season and user ratings.
- This helps the customer to find or select a food item which has a good rating and which is liked by a many customers.
- This also helps the Restaurant owner to make changes in a particular food item if it has low ratings which improves the quality of food.

Time to Serve:-

- The menu includes the approximate time to be served of a particular food item.
- This will help the customer to select the food item accordingly.

Modifiable Menu:-

- The menu can be modified by the Kitchen manager.

- The items which are not available in a particular time period are not displayed on the menu card.

3.3 External interfaces Requirements

3.3.1 User Interfaces

User Tablets:

This type of the tablets is especially for the use of normal users coming in the restaurant. These tablets will consist of the whole menu of the restaurant. They will be enabled with the Wi-Fi connectivity. The items in the menu are non editable for these types of the tablets. So, the user can not interfere in the menu and make changes in it. The tablets should be able to display all the items of the menu with sufficient visibility. Customer from any layer of the society should be able to handle and operate all the functions easily.

Manager Tablets:

These tablets are especially for the use of the restaurant manager. The manager should be able to control the function of whole restaurant from a single tablet. He can access any tablet and should be able to make changes to the menu. Like he can change price of particular item or he can disable particular item which is not available at that particular time.

Display at Kitchen:

These are present at the kitchen near chef so that he should be able to see what a particular has ordered. All the ordered items are displayed on the screen giving the table number below. They should be sufficiently large to be seen by chef at a reasonable distance. Chef should be able to denote a particular item that is ready.

3.3.2 Hardware Interfaces
There are three external hardware devices used by the proposed system, each related to a user interface. These devices are the wireless tablets and the displays. All the devices must be physically robust and immune to liquid damage and stains. The devices(with the possible exception of displays) must also have good industrial design aesthetics, as they are to be used in place of normal restaurant tables and notepads and will be in direct contact with customers. The devices behave as 'terminals' in the sense that they never have a full system image, do not store data and are not used for the core logic of the system. However, they should be fully capable tablets that can use textual data from the server along with local UI/interpretation code to display UI elements and take input. All order and transaction records should be stored on the server, not these tablets. The performance of dumb terminals over an area the size of a restaurant is likely to be unacceptable. In all the cases, the hardware device takes information from the proposed system and processes the information to display. It also provides user input information to the proposed system.

3.3.3 Software Interfaces
We will require interface with a JSP/Servlet that stores the information necessary for our system to operate. The JSP/Servlet must be able to provide, on request and with low latency, data concerning the restaurant's menu, employees (and their passwords) and available dietary requirements. Additionally, it should take and archive data provided to it .This data will include records of all orders and transactions (system states and state changes) executed. JSP/Servlet must store all data such that it can be used for accounting, as well as accountability.

3.3.4 Communication Interfaces
The DOSRUA will interact with a WiFi to maintain communication with all its devices.

3.4 Non-Functional Requirements:
This subsection presents the identified non-functional requirements for the subject of proposed system. The subcategories of non-functional requirements given are performance safety, security and software quality attributes.

Table 3.1 Non-Functional Requirements

Description
A manger password used for tablet login must have a bit strength of at least 64bits.
A manager password used for tablet login must be changed every three months.
A manager shall only be able to log into one tablet at any given instance of time.
The display shall not require an user to log-in.

3.4.1 Performance
Performance requirements define acceptable response times for system functionality.
• The load time for user interface screens shall take no longer than two seconds.
• The log in information shall be verified within five seconds.
• Queries shall return results within five seconds.

3.4.2 Safety
Table presents the identified non-functional safety requirements that directly relate to the entire subject proposed system.

3.4.3 Security
Table 3.4.3 presents the identified non-functional security requirements that directly relate to the entire subject proposed system.

Table 3.2 Security

The system shall log every state and state change of every user tablet and display to provision recovery from system failure. .
The system shall be capable of restoring itself to its previous

state in the event of failure (e.g. a system crash or power loss)
The system shall be able to display a menu at all timesto facilitate Manual order taking should the need arise.
The system shall utilise periodic 30-second keep alive messages between tablets and the server to monitor tablet operational status.
The system shall flag tablets that fail to send timely keep alive messages as non-operational and disassociate the assigned waiter from the tablet

Software Quality Attributes:

1) Coding in Android is very beneficial from the developer's point of view. There is availability of large number of documentation. Also, it could be easily run on tablet as well as any Android phone.

2) Using Android is very flexible as the developed product could be deployed on tablets as well as android driven mobile phones which as available in abundance.

3) Though maintenance is required, it is negligible.

4) The devices on which Android run are highly portable.

3.5 Other Requirements

3.5.1 DB requirements:-

The database required for this system is SQLite database for storing details on the tablet itself. It also needs a database on the server which is handled by JSP and SQL.

So what basically is SQLite?

SQLite is a relational database management system contained in a small (~350 KiB) C programming library. In contrast to other database management systems, SQLite is not a separate process that is accessed from the client application, but an integral part of it. SQLite is ACID-compliant and implements most of the SQL standard, using a dynamically and weakly typed SQL syntax that does not guarantee the domain integrity.

SQLite is a popular choice as embedded database for local/client storage in application software such as web browsers. It is arguably the most widely deployed database engine, as it is used today by several widespread browsers, operating systems, and embedded systems, among others. OS like Android, Web browsers like Mozilla etc. SQLite has many bindings to programming languages.

Features Of SQLite

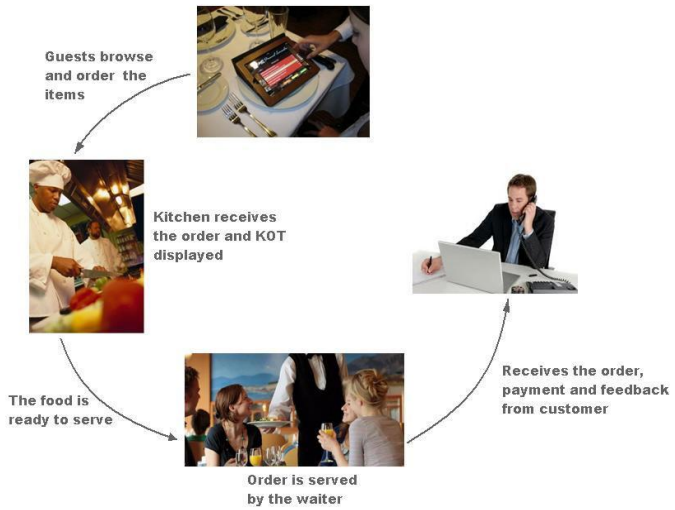
1] Transactions are atomic, consistent, isolated, and durable (ACID) even after system crashes and power failures.
2] Zero-configuration - no setup or administration needed.
3] Implements most of SQL
4] A complete database is stored in a single cross-platform disk file.
5] Supports terabyte-sized databases and gigabyte-sized strings and blobs.
6] Small code footprint: less than 350KiB fully configured or less than 300KiB with optional features omitted.
7] Faster than popular client/server database engines for most common operations.
8] Simple, easy to use API.
9] Written in ANSI-C. Bindings for dozens of other languages available separately.
10] Well-commented source code with 100% branch test coverage.
11] Available as a single ANSI-C source-code file that you can easily drop into another project.
12] Self-contained: no external dependencies.
13] Cross-platform: Unix (Linux, Mac OS-X, Android, iOS) and Windows (Win32, WinCE, WinRT) are supported out of the box.
14] Easy to port to other systems.
15] Sources are in the public domain. Use for any purpose.

Suggested Uses For SQLite:

Application File Format: Rather than using fopen() to write XML or some proprietary format into disk files used by your application, use an SQLite database instead. You'll avoid having to write and troubleshoot a parser, your data will be more easily accessible and cross-platform, and your updates will be transactional. Database For Gadgets: SQLite is popular choice for the database engine in cellphones, PDAs, MP3 players, settop boxes, and other electronic gadgets. SQLite has a small code footprint, makes efficient use of memory, disk space, and disk bandwidth, is highly reliable, and requires no maintenance from a Database Administrator. Website Database: Because it requires no configuration and stores information in ordinary disk files, SQLite is a popular choice as the database to back small to medium-sized websites. Stand-in For An Enterprise RDBMS: SQLite is often used as a surrogate for an enterprise RDBMS for demonstration purposes or for testing. SQLite is fast and requires no setup, which takes a lot of the hassle out of testing and which makes demos perky and easy to launch

IV. SYSTEM ARCHITECTURE

Fig. 4.1: System flow



System architecture of project could be described as:

When the customer enters the restaurant, he would surf on the tablet to order his menu. He could also surf quickly if he has already decided upon what to order. He would click the item he wants to order and after he is sure he wants each item in the list, he would click confirm. The confirmed order would be displayed on the display screen in the kitchen. After the chef has completed preparing the item, it would be notified to the customer. After the customer has completed eating the food, bill would be directly displayed on his tablet as well as managers system.

V. Technical Specification

The technologies which are used to implement the system are:
1) Android version 2.2.3 (Smart Phone) and Android version 2.2 – 4.0 for Tablets is required.
2) Java SE 6 Programming Language is used to develop the software.
3) Eclipse Indigo is used as a Rapid Application Development Tool (RAD) or as an Integrated Development Environment (IDE) for coding the software.
4) JSP/SERVLET is used for Remote Database Access from the main system of the restaurant.
5) SQLite is a light weight Database which is going to be used for database access from handheld device or the tablet.

5.1 Advantage

- Wastage of paper is avoided as our implementation is working just on tablet and does not need any paper work. e.g.-For taking the order, we are not using papers. Also, our menu card would be digitized.
- A customer going into restaurant does not has to wait for the waiters to take the order. As soon as he occupies a seat, he would order whatever he needs.
- As soon as the order is ready, it would be notified to the customer. So, there would not be any issue of late delivery in spite of the food being ready.

5.2 Disadvantage

- Tablets would cost us more as they are more costly the simple paper. Hence, it would cost us more. Though it would be a onetime investment, it would certainly be more costly.
- If we compare our system with traditional paper based system, more maintenance would be needed. Some technical assistance would also be needed.

5.3 Application

We are going to implement our system in restaurants to ease the management of the Restaurant and also give a technical touch which would help atomize the working of Restaurant.

VI. Conclusion

The proposed system would attract customers and also adds to the efficiency of maintaining the restaurant's ordering and billing sections.

Acknowledgment

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Effect of Soil Type and Irrigation Pattern on Seedlings Growth (Jatropha Curcas)

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Abstract- **S**urvival of seedlings in arid and semi arid zones plantation is strongly affected by water availability and soil type. Many tree species are able to adapt themselves in response to soil moisture content. Three soil types, sand, clay and mixture soil (according to the main classification of soil in Sudan) with different irrigation pattern (2, 4and 7 days intervals) were used for propagation of Jatropha Curcas for three months. The experiment was conducted at the nursery and was laid out in a randomized complete design with three replications and the growth characteristics were measured on monthly bases The aim of this study was to evaluate the effect of irrigation pattern and soil types in seedlings growth characteristics of Jatropha Curcas. Seeds used in the experiment were collected from South Kordofan. The results obtained supported earlier findings by some workers. Shoot length showed no significant different over all treatment. The control (normal nursery practices) was the best for all seedling characteristics measurements. These. results may suggest that for successfully a forestation programs, low nursery cost in raising Jatropha seedling recommended to irrigation that practices in the nursery, Since there are no significant differences of juvenile growth performances between the three soil type recommended that caution shouldn't done when Jatrohpa plant

I. INTRODUCTION

Few years ago there was limited attention in Jatropha .It is tropic species from the family *Euphobiaceae* which is a wild green thorny shrub that farmers in India, Africa and Central America would often plant to fence in livestock like goats or make soap. Recently wide attention has been oriented in the cultivation of the species for using its oil as a diesel fuel, primarily as it is drought resistant and can be cultivated on marginal land, without competing crop food production (Heller1996; Grimm 1996; RF 1998). Now *Jatropha curcas* has gained much attention (Grimm1996; Heller 1996; Henning 2000a; Pratt *et al*. 2002) of different research organizations, governments, public and international developmental agencies and industries in the tropics and subtropics; Jatropha is unique among renewable energy sources in terms of the number of potential benefits that can be obtained from its widespread cultivation. Its cultivation requires simple technology and relatively modest funds investment. The seeds of. *Jatropha curcas* are a good source of oil; seed contains 35-40 %viscous oil known as 'curcas oil'. Petroleum based fuel, which can be used as a diesel alternative (Foid et al 1996, Mandpe 2005). Although the seed meal, after extraction of oil, is rich in protein, it is toxic to rats, mice and ruminants and therefore cannot be used as an animal feed but after treatment, the seeds or seed cake can be used as an animal feed (Makkar *et al*. 2001). Its nitrogen - rich seed cake is a good soil fertilizer (RF, 1998; Makkar *et al*. 2001, FACT Foundation, 2006). Traditionally the seeds are used for medical treatments and soap production (Duke 1983; Henning 2002). Several cases of J. curcas nut poisoning in humans after accidental eating of the seeds have been reported with symptoms of giddiness, vomiting and diarrhea and in the extreme condition even death has been recorded **(**Becker&Makkar, 1998**)**

Jatropha grows on well-drained soils with good aeration and is well adapted to marginal soils with low nutrient content (FACT Foundation, 2006). It grow under a wide range of rainfall regimes (200 mm to over 1500 mm per annum) It is not sensitive to day length (ICRAF,2003). In many African countries, it is grown as a live fence and can be used to recover eroded areas (Heller 1996; Joker and Jepsen 2003, FACT Foundation, 2006).J*atropha* can reach 6 meters or more (Heller 1996; Makkar *et al*., 2001). Its leaves and stems are toxic to animals (FACT Foundation, 2006) the plant itself is very strong and can be an excellent applicant for reclaiming eroded zones,

In Sudan, *Jatropha curcas* is widely spread. It is found in Khartoum State, Kassala State, and Kordfan States in the west. It is mentioned as a local plant in some books describing the plants of the Sudan these days the, governments is beginning to introduce the species to be planted at a large scale for bio fuel production purposes. This will need a very large amount of seedling and seeds; the adequate seedling needs to be found in order to increase seed yield and oil content and more healthy plant by reducing the usual length traditionally adopt. As Henning (2000c) has noticed that pre-cultivation of *Jatropha* seedlings in poly-ethylene bags is more suitable and helps to speed up the appropriate of a plantation by at least 3 months.

The aim of the present study on Seed germination and growth parameters of the seedling will help in distributing the plant according to the classification of soil types and amount of rainfall in the Sudan. Providing information about the different growth parameters during juvenile stage will add an important dimension to the earlier ones about the establishment and development of this important plant species.

Research objective was to analyze the effects of different re-establishment techniques through water interval and soils type on seedling growth and survivorship during the nursery stage, to investigate the growth characteristic of the plant species at the nursery stage for successful establishment and development.

Specific objectives include: Determination of growth parameters such as height, width, crown and root /shoot ratios, in different soil types and different irrigation patterns, and verifying the style of growth performance accordingly

II. MATERIAL AND METHODS

The experiment was conducted at the Forestry nursery Center at AD Duwem city White Nile state during the period 26/3/2012-26/6/2012 Seeds used in the experiment were collected from South Kordfan. Three soil types (according to the main classification of soil in Sudan) were used for growing tree seeds at nursery with three irrigation patterns. The main soils are sand, clay, and sandy clay soil (50-50%). Irrigation patterns vary from 2, 4, and 7 days interval. The first irrigation started at the same time for all replications, and then followed by mentioned intervals. Polythene bags of 25cm width and 30cm length used for growing the seeds. the experiment was laid out in a randomized complete design with three replications. Seedlings were raised in the nursery for a period of three months and growth variables were measured on monthly bases including seedling height, crown height, and relation between them. Records were taken daily for estimating germination rate

Data analysis

All results were statistically analyzed with the statistical package MSTATC program. Analysis of variance (ANOVA) and difference among treatments means were determined by Duncan, Multiple Range test at (P = 0.05) level.

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III. RESULTS

Table (1a) Growth characteristics of 1st month Jatropha seedling of the three soil types

Growth characteristic	Soil Types		
	Sand	Clay	Mixture
Growth at root color(cm)	5.5111a	5.600a	5.1778a
No .of leaves/seedling	5.4444a	5.6667a	5.2222a
Total length(cm)	32.8889a	28.9556b	30.2222ab
Shoot length(cm)	19.0667a	17.8778a	17.3111a
Root length(cm)	13.778a	10.5111a	12.322a
Crown length(cm)	8.2222a	7.6667a	6.5556a
Root/Shoot Ratio	.37a	.41a	.36a

Similar letters in the same raw are not significantly different using Duncan' Multiple Range test(p=0.05)level

Result in table (1a,b) showed that most growth characteristic is not significantly different. The clay soil shows the lowest total length

Table (1b) Growth characteristics of 1st month Jatropha

Growth characteristic	Irrigation interval		
	2days	4days	7days
Growth at root color(cm)	5.5111a	5.40000a	5.3778a
No .of leaves/seedling	5.3333a	5.5556	5.4444
Total length(cm)	3o.3556a	30.2222a	31.08889a
Shoot length(cm)	18.0222a	17.9222a	18.3111a
Root length(cm)	11.2222a	12.5556a	12.8333a
Crown length(cm)	6.3333a	7.8889a	8.2222a
Root/Shoot Ratio	.37b	.46a	0.39ab

seedling of the three irrigation interval

Table (2) Growth characteristics of 1st month Jatropha seedling of the three irrigation interval

Growth characteristic	Irrigation interval		
	2days	4days	7days
Growth at root color(cm)	5.5111a	5.4000a	5.3778a
No .of leaves/seedling	5.3333a	5.5556a	5.4444a
Total length(cm)	3o.3556a	30.2222a	31.08889a
Shoot length(cm(	18.0222a	17.9222a	18.3111a
Root length(cm)	11.2222a	12.5556a	12.8333a
Crown length(cm)	6.3333a	7.8889a	8.2222a
Root/Shoot Ratio	.37b	.46a	.39ab

Similar letters in the same raw are not significantly different using Duncan' Multiple Range test

Result in table (2) show that the growth characteristic is not significantly different

Table (3) Growth characteristics of 2nd month Jatropha seedling of the three soil types

Growth characteristic	Soil Types		
	Sand	Clay	Mixture
Growth at root color(cm)	11.2111a	10.8778.a	11.0778a
No .of leaves/seedling	11.222a	10.8889a	11.333a
Total length(cm)	48.5556b	48.1556.b	52.6889a
Shoot length(cm)	18.6446a	18.4889a	18.778a
Root length(cm)	30.5111ab	29.1778b	33.5556a
Crown length(cm)	13.1111a	13.0000a	12.7444a
Root/Shoot Ratio	.69a	.69a	.67a

Similar letters in the same raw are not significantly different using Duncan' Multiple Range test

Result in table (3) show that the most growth characteristic are not significantly different. The clay soil shows the lowest root length and the mixture soil shows the highest total length

Table (4)Growth characteristics of 2nd month gatropha seedling of the three Irrigation interval

Growth characteristic	Irrigation interval		
	2days	4days	7days
Growth at root color(cm)	11.778a	11.3444a	11.5000a
No .of leaves/seedling	11.0000a	11.1111a	11.3333a
Total length(cm)	48.5333b	49.4889ab	51.3778a
Shoot length(cm)	18.6444a.	18.5111a	18.7556a
Root length(cm)	29.6667a	31.5556a	32.0222a
Crown length(cm)	13.4111a	12.8889a	12.5556a
Root/Shoot Ratio	.71a	.67a	.67a

Similar letters in the same raw are not significantly different using Duncan' Multiple Range test

Result in table (4) show that the most growth characteristic is not significantly different however 2days irrigation interval has the lowest total length

Table (5) Growth characteristics of 3rd month old Jatropha cuscas seedling s of the three soil types

Growth characteristic	Soil Types		
	Sand	Clay	Mixture
Growth at root color(cm)	13.58a	13.49a	12.91a
No .of leaves/seedling	12.56a	11.56a	12.56a
Total length(cm)	65.96a	67.56a	67.38a
Shoot length(cm)	19.73a	20.06a	19.93a
Root length(cm)	17.28ab	15.28b	19.00a
Crown length(cm)	46.22a	47.49a	47.49a
Root/Shoot Ratio	.85a	.85b	.93a

Similar letters in the same raw are not significantly different using Duncan' Multiple Range test

Result in table (5) show that the most growth characteristic is not significantly different. The clay soil shows the lowest root length

Table (6)Growth characteristics of 3rd month Jatropha seedling of the three Irrigation intervals

Growth characteristic	Irrigation interval		
	2days	4days	7days
Growth at root color(cm)	13.42a	12.74a	13.81a
No .of leaves/seedling	12.56a	11.56a	12.56a
Total length(cm)	64.78b	63.69b	72.42a
Shoot length(cm)	19.72a	19.91a	20.09a
Root length(cm)	17.17a	17.17a	17.22a
Crown length(cm)	45.04b	43.78b	52.38a
Root/Shoot Ratio	.95a	.95a	.87a

Similar letters in the same raw are not significantly different using Duncan' Multiple Range test

Result in table (6) show that the most oh growth characteristic are not significantly different however 7days irrigation has the highest total length, and crown length

Table (7) Growth characteristics of 3rd month Jatropha seedling of the three Irrigation intervals interact with three soil type

treatment Parameters/	2days			4days			7days		
	sand	clay	mixture	sand	clay	mixture	sand	Clay	mixture
diameter	14.87a	13.73ab	12.47d	12.87acd	12.75cd	12,63cd	13,80abc	14.00ab	13.63abc
No of leaves	11.67bcd	10.67d	12.00bcd	12.33abc	11.00cd	12.67ab	13.67a	13.00ab	13.00ab
Total length	65.27c	64.87c	64.20cd	61.93d	63.2cd	65.93c	70.67b	74.60a	72.a0b
Shoot length	19.60a	19.83a	19.73a	19.40a	20.40a	19.93a	20.20a	19.93a	20.13a
Crown length	45.67c	45.00cd	44.47cd	42.53d	42.8d	46.00c	50.47b	54.67a	52.00b
Root length	17.17bc	16.00bcd	18.44ab	16.67bc	14.17d	20.67a	18.00bc	15.67cd	18.00bc
Germination	62.00a	56.33b	41.67c	35.33d	52.00b	54.00b	52.33b	54.67b	43.33c
Root/shootratio	.87b	.80c	.93b	.85b	.69f	1.03a	.89b	.78d	.89b

Similar letters in the same raw is not significantly different using Duncan 'Multiple Range test

Result in table (7) show that the most of growth characteristic different the 2days irrigation interval with sand soil has the highest germination mean, and the biggest diameter at root color, shoot length show not significantly different over all treatment However4days irrigation with clay soil has the lowest root and length, crown length and 7days irrigation with sand has highest number of leaves. The result also indicate that 7day irrigation clay soil has the highest crown length

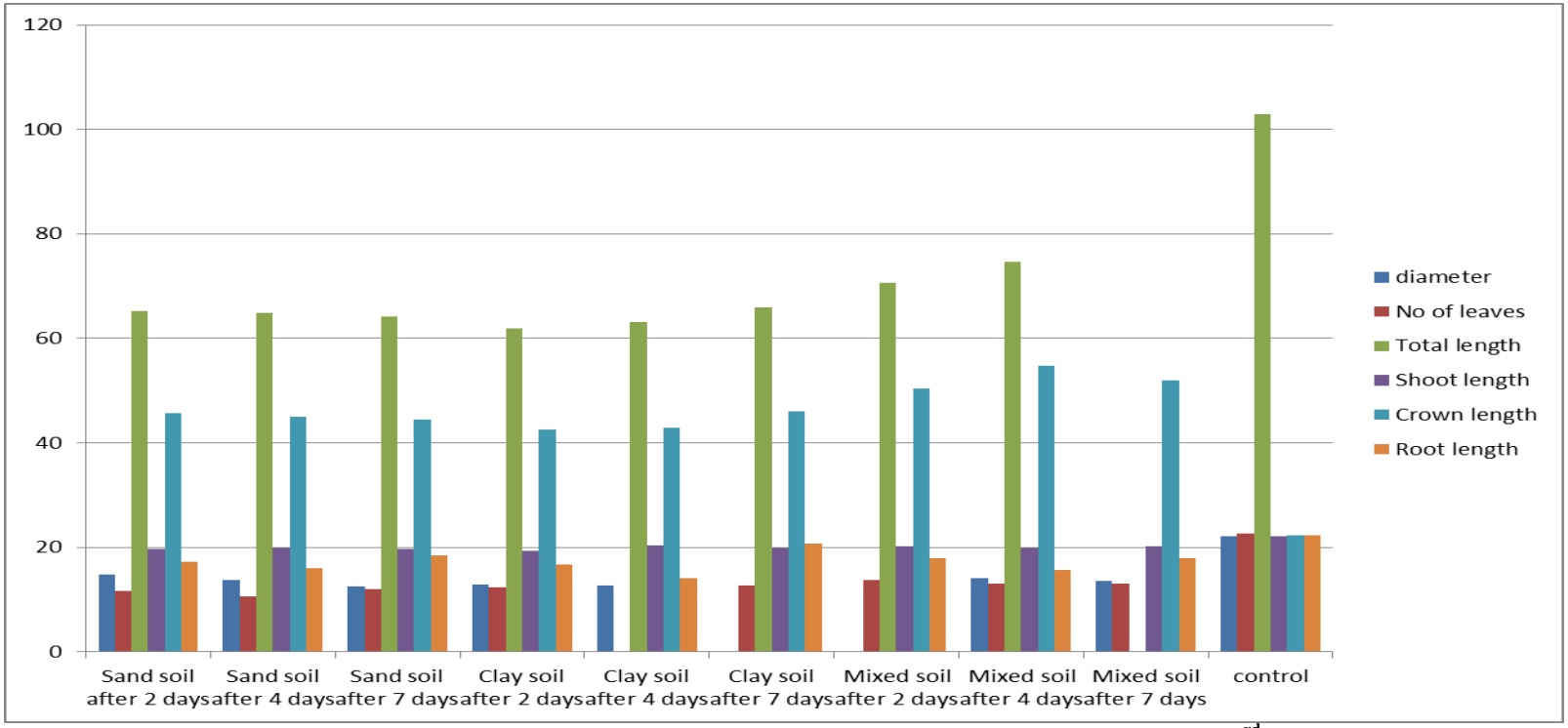


Figure (1) Growth parameter for of three soil type via three irrigation interval with the control of 3rd month growth

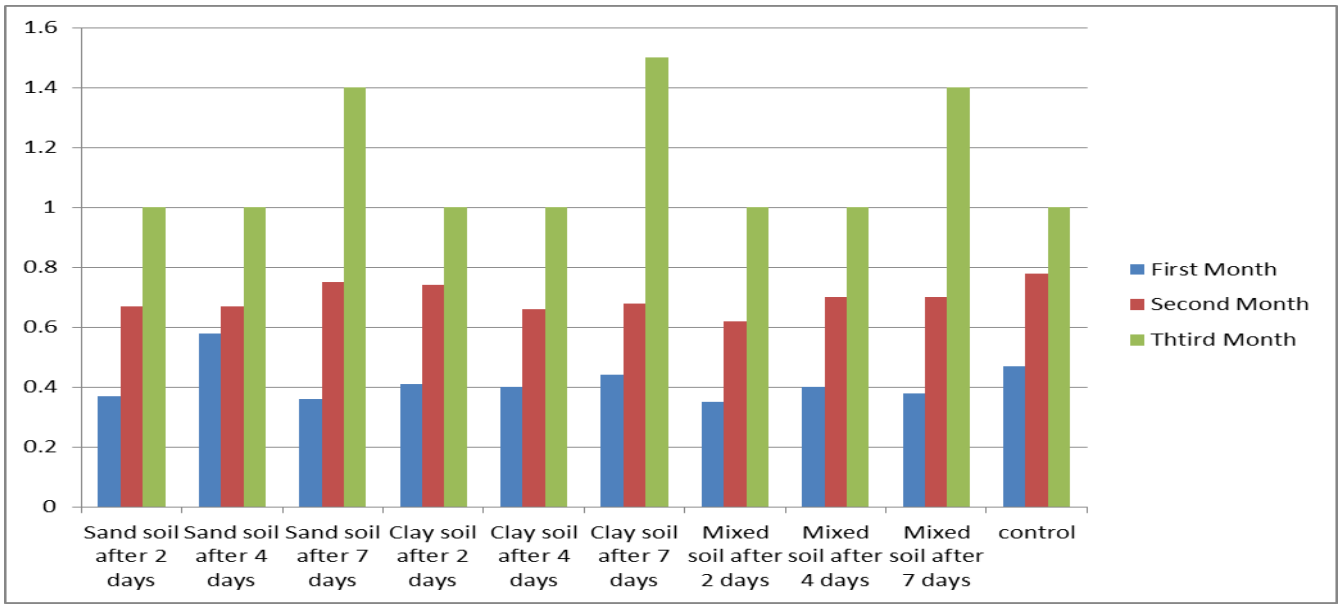
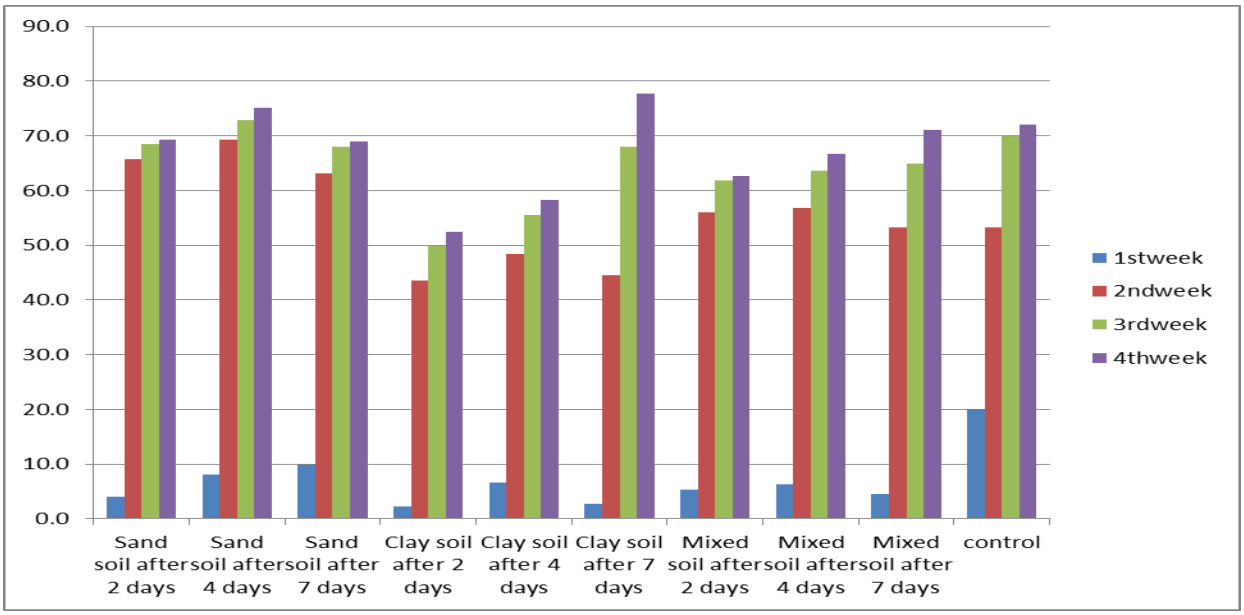


Figure (2) Figure Ratio of Root length/Shoot length for three month of three soil type via three irrigation interval with the control



Figure(3)germination percentage for 3rdmonth of three soil type via three irrigation interval with the control

Growth parameter show the best record of all parameter for control as shown in figure (1)but 4days irrigation with mixture soil show the best root/shoot ration and germination percentage for 2nd,3rd.and4th month as show in figure (2)and(3) respectively

IV. Discussion

Table (7) shows mean seedling characteristic by water interval interact with soil type for Jatrohpa in third month of growth .Seeding growth was evaluated in term of measurement of 8 growth parameters. The result showed significant differences for all growth characteristic of the nine treatment .From the result obtained it is clear that shoot in all treatment grow in the same habits The lowest root length attained by 4days/clay This may be explained in the light of findings obtained by Ahmed (in Ibrahim 1988)for Acacia subspecies that root length is critical importance at early stage as deeper primary root system provide contact to more reliable source of water in short time than available to shallow root system . From figure(2) the ratio of root length/shoot length increase by month this is important under difficult condition particularly low soil moisture the ratio of root length to shoot length is considered as an important factor in survival of plants. The ratio is believed to be large for species grow in dry region this some as quoted by Abbatt (1984). Ahmed (1982) claimed that seedling of arid land tree species are characterized by that their roots grow taller than shoots even if they are watered. And 2day irrigation interval for sand soil Have highest diameter at root color and germination mean. Highest number of leaves record by 7days irrigation with sand the produced relatively higher number of leaves per seedling .This is an indication of adaptation species to drier condition compared to the other species Algunaid (2004)

V. Variation in growth

From table (1,2,5,6)Sand soil showed good growth for three month while clay showed better growth in the first month and showed lowest growth for the root in the 3rd months. For the root growth among the three soil it was found not significant during initial 1st month while these insignificant were evened down for three soil by the 3rd months. Variation in growth in the three soil of special interest to the tree breeder as it makes selection among soil available as the three soil investigated in the study not differ significantly from each other this is when using shoot height as only criteria for evaluating growth performance and productivity this is mainly due to the important of early height growth in determining the success of seedling establishment .Seedling that will grow rapidly could have practical advantages, damage from animal for instance , would be lessened because the terminal part of the tree would sooner be unreachable to browsing . Fast growing seedling can compete better with other vegetation Demster (1972) argued that outstanding seedling can maintain superior growth and rates for considerable periods.

VI. Conclusion

The conclusions drawn from this study may be summarized as follows:

- The results support earlier findings obtained by some workers.
- Findings from interaction between irrigation interval and soil type pre-germination 2days irrigation interval with sand soil has the highest germination mean, and the biggest diameter at root color,
- Shoot length show no significantly different over all treatment
- The control (normal nursery practices) the best for all seedling parameter measurement
- 4days irrigation with clay soil has the lowest root and crown length and 7days irrigation with sand has highest number of leaves. On the other hand 7day irrigation clay soil has the highest crown length

VII. Recommendation

1. Since there are no significant difference of juvenile growth performances between the three soil type(according to the main classification of soil in Sudan) recommended that caution shouldn't be done when Jatrohpa plant
2. For successfully a forestation programs, less nursery cost in raising Jatropha seedling recommended to irrigation that practices in the nursery
3. The result of this study is important for utilization successful and establishment of the species in the nursery and later in the field.
4. Further studies on seed from different provenances of the country

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Assessing the Effects of Packaging Materials on the Moisture Content and Surface Finish of Packaged Furniture Products

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Abstract- The study investigated the effects of packaging materials on moisture content and surface finish of furniture products. Unfinished, semi-finished, and finished furniture parts were packaged for two, four and six weeks using five different packaging materials. Four hundred and fifty (450) strips were prepared from Iroko (Millicia excelsia). Five different packaging materials were selected for the three levels of furniture parts after which a change in moisture contents and colour were determined. The mean moisture content (%) changes determined on unfinished furniture parts with plastic, cardboard, metallic, plastic plus cardboard (PC) and plastic plus cardboard and metallic (PCM) showed a rise in MC with plastic and cardboard showing highest(2.26%) and lowest(0.37%) respectively while in semi-finished furniture parts, PC and metallic recorded the highest and least of 1.72% and 0.65% respectively. However, cardboard recorded 0.12% upon the initial moisture content whiles metallic and cardboard packaging recorded highest and the least rise in MC of 2.21% and 0.14% respectively for finished furniture parts. There was insignificant difference in the surface finish. It was concluded that cardboard is the superior packaging material with the lowest MC increase throughout the three time levels.

Index Terms- Packaging; furniture; plastic; cardboard; metallic materials.

I. INTRODUCTION

1.1 The furniture industry in Ghana

In recent times, there has been a rapid development in the processing of furniture in Ghana as a result of increase development in technology. A typical example is the American technology for processing small diameter logs from both natural forests and plantations with wood mizer mills (Okai**,** 2002) that is gaining popularity in the country. In Ashanti and Brong Ahafo Regions, wood mizer mills have been purchased and installed in most companies. These companies stand the chance of turning lumber into finished products such as furniture for domestic and export markets.

In addition, the use of computer-aided design of furniture is catching up with most furniture producing companies in the country. According to the International Tropical Timber Organization (ITTO) there are about 600 small to medium scale furniture producers in the country with the bulk in Kumasi (ITTO, 2004). Most of these companies received their training from the numerous Technical Institutions, Polytechnics, Intermediate Technology Transfer Units and the informal apprenticeship training centres in the country.

Regrettably, there has been a considerable decline in the export of timber products including furniture. In the 1970s there was a sharp decline in the output of furniture exports which was attributed to the general poor state of the economy in the 1970s and early 1980s, obsolescence and continuous breakdown of logging and processing equipment, poor packaging, the over-valued exchange rate, inadequate transport and poor state of rail and harbour facilities, (Baidoe,1987).

Since 1989, the export performance of furniture industry has not been encouraging. Furniture export statistics for 1990 indicates that Ghana earned USD 37,288,400.00 showing 1.86% increase on the 1989 furniture export earning of USD 3,554,270.00 (Appiah, 1990).

Despite a 27.96% drop in volume in 1991 (from 2754 m^3 in 1990 to 1984 m^3 in 1991) furniture export in 1999 increased to about 2720 m^3 which earned the country $7,143,810.00. The international furniture trade is valued at $67 billion and is expected to grow in future because of the encouraging retail market prospects from the USA and Europe, the two largest furniture markets (Ratnasingam, 1998). Unfortunately, the export volume of furniture in 2000 was decreased as against the past performance. The export volume dropped by 8.13% from 2720m^3 in 1999 to 2499m^3 in 2000.

Data available from the Timber Industry Development Division of the Forestry Commission (TIDD) indicates that contracts for a volume of 2,475 pieces of furniture parts were processed and approved for export during the fourth quarter of 2009. There was a sharp increase in furniture parts export as a result of two major parts shipments for the Togolese market undertaken by Portal Limited, a Takoradi based timber firm. The TIDD report indicates that Mim Scanstyle Ltd, once the major furniture parts exporter, submitted almost no contracts for export approval during the same period (TIDD, 2009).

The poor export performance of the furniture industry since 1989 to the present shows that the industry is still facing serious problems. Among the numerous problems identified was difficulty in packaging the furniture. Currently, the few companies that export furniture do so in their knocked-down form. Even with this, there are many problems associated with moisture changes, which lead to discolouration, bruising and staining of the furniture products. Other problems like scratches and sometimes breakages are also real.

Furniture is inherently predisposed to damage during transit and delivery. Damaged goods are unacceptable to

consumers and the producer must therefore either scrap the items or undertake costly and time-consuming repairs which leads to high looses to the industry.

Traditionally, the primary function of packaging is to contain and protect the product (Kotler, 1984). Of late, attracting attention, describing the product and making the sale also form part of the functions of packaging. Improving packaging is often limited by a perceived increase in packaging material purchase costs.

The purpose of this study was to investigate the effects of five packaging materials on the moisture content and the surface finish of the packaged furniture and to make recommendations on the most appropriate materials and method for packaging furniture.

Information gathered from interviews and questionnaires administered to the furniture companies that export furniture revealed that they are faced with problems associated with packaging such as moisture changes and discolouration of the surface finish. In this respect, the study when undertaken successfully will help to achieve this ultimate goal by way of prescribing improved methods and materials for packaging furniture by the industries.

The research will also lead to the reduction of the number of damages experienced by furniture exporters in order to reduce the amount of compensations paid on defected goods and thereby maximizing their profit. Moreover, the study will help bring about sustainable utilization of our limited forest resources.

1.2 Furniture produced in Ghana

Page (1973) states that the bulk of furniture produced by all sector in the import-substituting sector of the industry consists of wooden and upholstered furnishings, designed for domestic and commercial use. According to Prempeh (1993), furniture is fully assembled, finished and upholstered before sale on the local market; for the large firms in the export market like Scanstyle Mim Limited produces knocked down furniture and the range consist of garden chairs of all kinds, indoor chairs and occasional furniture. Knocked down furniture is furniture that may easily be folded or taken apart and packed flat for transport. It is also called demountable or packaged furniture (Gloag, 1969).

Knocked down processing allows for better-packed and less breakage in transit. It also promotes compact storage in the warehouse, making handling very easy and increasing the strength of the joint. Unfortunately, the nature of knocked down processing allows firms to avoid the use of heavily protected locally produced inputs such as primary paints, lacquers and fabrics

1.3 Packaging of furniture

Human beings have always protected food and drink in containers, using skins, leaves, and gourds, and then baskets, pottery, and, as early as 1500 BC (Covell, 2002). According to Willinston (1988), proprietary products such as panels and furniture parts for the shoulder trade were also packed in individual cartons. The trend of packaging and branding will continue in the foreseeable future.

Packaging is all the activities of designing and producing the container or wrapper for a product (Kotler, 1984). It can also be described as the technology used to contain, protect, and preserve products throughout their distribution, storage and handling, and at the same time to identify them, provide instructions for their use, and promote them (Covell,2002).

Packaging may include the product's primary package or container, secondary package and the tertiary or shipping package. Recently, labelling and printed information appearing on the package is also part of packaging. Packaging must maintain the condition of its contents. It must exclude all the undesirable conditions to prevent it from becoming unfit for use during the period designated as its shelf life. The pack must also prevent the product leaking, especially if there is a corrosive or poisonous chemical inside. The pack must identify the contents and their amount by print and pictures, and, if necessary, provide instructions on use, as well as any hazard warnings. The latter is essential when the pack contains pharmaceuticals or chemicals, either domestic or industrial.

1.3 Materials for packaging furniture

A wide range of materials ranging from "tamperproof" to "waterproof" devices is in use as packaging materials. Common ones in use are cellophane (plastic film), polyethylene, plywood, solid wood, shredded paper, wood wool, cellulose wadding, corrugated cardboard, foam, polystyrene (Styrofoam), natural and synthetic papers, plastic and metallic containers. Some of these materials act as accessories to the main pack.

Sawn wood or lumber has a wide range of different end uses. The various end uses are from construction to repairs, packaging, furniture, mining, shipping, sleeper etc (Amoako, 1993). Wood intended for furniture production must be kiln dried to a very low moisture content of about 8 to 12 % (Rietz 1957). A few of the more important reasons according to him are that seasoning reduces gross weight and thereby reducing shipping and handling costs, imparts dimensional stability, increases most strength properties, increases fastener holding power and thereby joint strength, increases electrical resistance, improves paintability and glueability, and finally, improves the thermal properties of wood. In addition to these advantages, drying wood below the fiber saturation point renders it impervious to biological degradation so long as it is not re-wetted. Attack by wood destroying fungi, in particular, is prevented.

Most of the small and medium scale furniture manufacturing industries all over the country use air seasoned lumber. The green lumber is neatly stacked with stickers separating one layer from another. In some cases, they are left at the mercy of the weather. Some companies even use processed timber with high moisture content and in fact, some finished products in the form of furniture, doors and windows are being processed from the green lumber. However, green timber may contain a sizeable amount of water. Seasoning is a costly and time consuming process and would not be employed unless there were valid reasons why it is required. So far, only the large scale furniture companies like SCANSTYLE Mim Ltd and DUPAUL employ the use of kiln in drying the timber for furniture construction

Cardboard is one of the widely used packaging materials in the world. According to Pröyry (2003), the exportation of paper and paperboard from the North America and Europe to the other parts of the world is high and it will continue to increase till the year 2015. Unfortunately paper based (board) packaging

components are subject to deterioration and reduced performance through ingress of moisture. For this reason, board components are ideally stored in heated and temperature stable areas away from warehouse doors which remain open for significant periods of time. Corrugated and board sheet material should be stored flat if kept for prolonged periods (FIET, 2001).

Plastic packaging materials come in the form of cellophane, foam, polystyrene and polyethylene. These can be thermoplastic or thermosetting. The thermoplastic such as the polyethylene is the most common. They are described as linear polymers, straight chained -though may branch occasionally and crystalline. This is why they soften if the plastic is heated (Metcalfe *et al*, 1966). But some like the polystyrene are amorphous which makes it not having a sharp melting point. According to Ashby and Jones (1998), polyethylene and the polystyrene are the common plastics in use in the packaging industry. Plastics have a mechanical property of being less stiff, lower density, lower strength and with less hardness. These properties may deteriorate rapidly with quite small increase in temperature (Bolton 1988). According to FIET (2001), plastic packaging materials and components are less affected by ambient environment, temperature and humidity fluctuations, but they should be dry before use.

The basic metallic materials used in the packaging industry are aluminium and steel. Metallic container is advantageous in exportation of furniture as it is locked against pilfering and sealed against the weather, usual packing requirements are relaxed, and the freight is billed as a volume shipment. Interchange of material is expedited, and containers can be used for storage; some terminals are fitted with electrical outlets for maintaining refrigerated containers. Damage claims on container cargo have been found to be much lower, and pilfering has been almost eliminated (Covell, 2002).

Plywood is normally used as packaging material because its behavior can be predicted with reasonable accuracy by calculations used in structural theory and normal engineering practices. The performance of plywood is similar to that of solid wood of the same species generally, but it is subjected to the following modifications: The permissible working stress parallel to the grain is often higher in plywood than solid wood of the same species; transit load in both parallel and perpendicular directions; has greater dimensional stability and rigidity than solid wood and it is advantageous in terms of strength to weight ratio and resistance to split and impact.

1.4 The Hazards of Distribution

The key to selecting packaging, which protects product, is to understand the hazards within the distribution environment. Ideally packaging should match the hazards inherent in handling and distribution of the product such that it is adequately protected, whilst avoiding excessive use of packaging which increases purchase and distribution costs. The deep understanding of the hazards allows informed selection of packaging materials and packaging formats, reducing the use of unnecessary packaging and reducing product damage rates. Distribution hazards are typically categorized as shock by drop or impact and vibration.

1.5 Forms of damage

Breakage is a common form of damage to furniture legs, cornices, doors, glass panels, and chair arms. Breakage is mostly caused by impacts either by shock or by drop; if products are consistently damaged in this way during transport and storage, more cushioning should be applied to the vulnerable component or alternative product delivery options considered.

Bruising is also the result of impact, probably a less severe impact than would cause component breakage. Bruising is generally the result of inadequate cushioned coverage, particularly along the edges, around the corners, and over the flat panel surfaces of the furniture unit.

Scratching occurs when a packaged piece of furniture comes into contact with a sharp edge or point. This may be a protruding bolt inside a vehicle, a door handle in the warehouse or factory, or the handle of another piece of furniture, which is inadequately protected. Scratching is also caused within knocked-down furniture by dust and dirt, loose fittings, screws, handles or other components which are allowed to move about within the pack during transport and handling. Scratching can also happen when a package is being opened by customers or their agents. Corrugated cases are often sealed by pressure-sensitive tape, and anyone trying to open these cases is tempted to use a knife or other sharp blade to slit the tape, unaware that a vulnerable finished surface lies just beneath. A piece of stout board underneath the taped joint can help.

Abrasion is a particular problem in packages containing knocked-down furniture. It occurs whenever finished timber comes into contact with an unsuitable packaging material or with another abrasive component within the same package. It may also be caused by sudden shocks, vibrations, effects of changes in moisture content and temperature.

1.6 Discolouration

Natural wood surfaces, both solid and veneers, are liable to change colour when exposed to natural light. The stronger the light, the greater the colour change. (Kellogg and Meyer, 1982).This may not be a problem if the whole unit or panel is exposed, because when the article is in use it will age naturally. However, if only part of the unit is exposed, the colour variations between the exposed and unexposed portions may prove unacceptable to the purchaser, despite the fact that in time the colour differences will disappear.

Climatic conditions have detrimental effects on the surface finish of the packaged furniture products. Moisture, dehydration and temperature damage can take a variety of forms. High temperatures can cause softening of adhesives (Kollman and Côte, 1984) and, of greater importance to the packer, softening of lacquers. It is easy for lacquers to mark if they become soft whilst in contact with even the least abrasive packaging material, and harder packaging materials, such as single-faced or double-faced corrugated fibreboard, may cause severe marking.

Furniture makers have little control over the environment in which their products will be transported or stored, so it is important to make the package as weather resistant as possible. Enclosing the article in a plastic (polyethylene) bag can help. A silica gel packet of appropriate size may be placed in the bag to absorb excess moisture and prevent the formation of condensation if high humidity is likely to be a problem.

II. METHODOLOGY

The materials which were used for the study were plastic sheets for waterproofing, corrugated paperboard for shock and vibration proof metallic container for tamper proofing, electronic moisture meter, samples of *Millicia excelsia* (Iroko), sellotape and a marker. The electronic moisturemeter (capacitance type) was used to measure the moisture content of the wood samples.
Kiln dried samples of *Millicia excelsia* (Iroko) were selected for the study. Sellotape was also used to seal the plastic and cardboard packs. The waterproof marker was used to mark all selected samples to ensure easy identification.

The study was carried out at two different sites – Scanstyle Mim Company Limited and Bibiani Logs and Lumber Company (BLLC). Mim was selected because it is one of the biggest mills in the country and also the leading exporter of furniture. BLLC was also selected because of its large concession in *Millicia excelsia* (Iroko) predominant forest district of the country. It also has a modern computerised kiln.

Feasibility studies carried out at the mills also indicated that availability of logs for processing was not a problem and the companies had enough selected species for the study. The two companies were also ready to release information and samples for the study. The Faculty of Renewable Natural Resources' Workshop of Kwame Nkrumah University of Science and technology, Kumasi was used for the execution of the project.

Millicia excelsia (Iroko) was selected for the study because the initial survey of the company revealed that contracts for immediate processing demanded this species and was the leading species in the exportation of furniture in the country. In addition, the logs were available in sufficient quantities to ensure continuous production and it is in abundance in the local mills in the community to ensure easy access to the material.

Kiln dried samples of *Millicia excelsia* (Iroko) were planed and sawn into dimensions of 20 × 20 × 100 mm strips. Four hundred and fifty (450) strips were prepared and sanded. These were divided into three groups: unfinished, semi-finished and finished furniture parts of one hundred and fifty (150) strips each.

The unfinished furniture was made up of smoothly sanded strips without any protective coating. Sanding sealer was applied to the semi-finished furniture and the finished furniture received both sanding sealer and lacquer. After preparation, specimens were marked using waterproof ink to identify them.

Five different packaging materials were selected based on their use by the furniture companies and their availability. These were plastic sheet (P), cardboard (C), metallic container (M), a combination of plastic and cardboard (PC) and a combination of plastic, cardboard and metallic container (PCM).

These materials were thoroughly checked for any punchers and defects. They were later cut to size to suit the volume of samples to be packaged.

After preparing the specimens, the initial moisture content was taken with the capacitance type moisturemeter at room temperature (25°C). The average prevailed temperature and relative humidity at the workshop during the period of the experiment were 27.9°C and 73.6% respectively. Every specimen was tested and photographed with a digital camera to ascertain the initial colour, after which they were immediately packaged.

Each group was made up ten (10) replicates for each packaging method. For instance, a pack containing ten specimens (replicates) of unfinished furniture parts was packaged in plastic sheet. This was repeated for other packages.

The set-ups were then stored for periods of two (2), four (4) and six (6) weeks under the then prevailing weather conditions. These three time levels were selected based on the length of time that products usually remain in the pack. A product for local consumption stays in the pack for a maximum of two weeks before they are unwrapped. The four week time level was also based on the maximum period it takes for shipping a product to the international market. Finally, a product upon reaching their destinations takes a maximum of six weeks before they are unwrapped.

After each period, the specimens were unwrapped and the moisture content quickly taken to determine the moisture content differences. Photographs were also taken at each period with the same digital camera and used to compare with the initial colour to determine any change in colour of the surface finish based on colour constancy.

The means and standard deviations of the data were determined. Descriptive analysis was used to discuss all the results and a comparative analysis was used to describe the differences in colour of the surface finish.

The experimental design used was split plot in complete randomized design (CRD). The data was subjected to the analysis of variance (ANOVA); and F-test at 1% and 5% significance were used to find out if there were significant differences in mean moisture changes in the five different packaging materials, time and the interaction between packaging materials and time.

Regression analysis was also carried out to determine the correlation between the time of packaging and the change in moisture content of furniture parts packaged using the various packaging materials.

Colour chart was also used to aid in the visual comparison of colours of the initial and final surface finish of the furniture parts.

III. RESULTS

3.1 Effect of packaging materials on the moisture content of furniture parts

The results of the effect of packaging materials on the moisture content(MC%) of furniture parts from Iroko (*Millicia excelsia*) within the three different time levels of 2, 4 and 6 weeks were obtained as shown with the example below:

Table 1: Effect of plastic material on the moisture content of an unfinished furniture parts

Duration (Weeks)		Replications										Mean	SD	
		1	2	3	4	5	6	7	8	9	10			
2	A	8.8	10.3	10.4	9.8	9.5	8.7	9.8	9.5	9.2	10.1			
	B	10.1	12.4	11.4	11.8	11.8	10	10.7	10.9	11	11.1			
	C	1.3	2.1	1	2	2.3	1.3	0.9	1.4	1.8	1	1.15	0.504315	
4	A	9.5	10.1	9.9	10.3	10.1	9.8	9.8	10.1	9.5	9.4			
	B	10.8	11.1	10.7	11.1	11.1	11.1	10.9	11.1	10.8	10.7			
	C	1.3	1	0.8	0.8	1	1.3	1.1	1	1.3	1.3	1.09	0.202485	
6	A	9.1	9.2	9.8	9.2	9.8	9.4	9.4	9.2	9.2	9.1			
	B	10	10.7	10.6	10.1	10.3	10.3	10.5	10.3	10.4	10.8			
	C	0.9	1.5	0.8	0.9	0.5	0.9	1.1	1.1	1.2	1.7	1.06	0.347051	

A=MC(%) after packaging; B=MC(%) before packaging; C=change in MC(%); Positive values=rise in MC(%)

The results obtained from the effects of the five different packaging materials on the moisture content of unfinished furniture parts from Iroko (*Milicia excelsia*) indicates that all samples recorded higher MC as compare to MC before packaging except metallic(3.33%) and PCM (13.33%) packaging that recorded a drop in moisture content. The standard deviation (SD) also showed a high variability in the MC of specimens that went through all the five packaging materials.

It was also realised that a higher mean MC was recorded for two weeks packaging with plastic (1.51%) and metallic(1.53%) packaging whiles cardboard, PC and PCM recorded higher moisture contents at the six weeks packaging (0.98%,1.79% and1.14% respectively). Plastic and metallic packaging recorded the least mean moisture contents of 1.06% and 0.74% respectively in the week six while cardboard, PC and PCM recorded least mean MCs of 0.84%, 1.13% and 0.72% respectively in the fourth week.

In the case of semi-finished furniture parts, a higher mean MC was recorded for the samples that underwent two and four weeks packaging with plastic (1.88%) and cardboard (1.45%) materials respectively whiles metallic, PC and PCM recorded higher MC at the six weeks packaging (2.21%, 1.86% and 1.57% respectively). of cardboard, metallic and PC packaging respectively that recorded a drop in MC. The SD also showed a high variability in the MC of specimens that went through all the five packaging materials. plastic and cardboard packaging recorded the least mean MC of 0.88% and 0.14% respectively in the six weeks while PC recorded least mean MC of 0.65% and metallic and PMC, 0.65% and 1.09% respectively in the fourth week.

The results obtained from the effects of the five different packaging material on the MC of finished furniture parts also showed a higher mean MC recorded for the samples that underwent two weeks packaging with plastic (2.26%) and cardboard (0.37%), metallic (2.02%) and PC (1.9%) materials whiles PCM recorded higher mean MC of 2.07 in the sixth week. However, plastic and PC packaging recorded the least mean MC of 0.88% and 0.14% respectively in the six week, while cardboard, metallic packaging also recorded least MC of 0.12% and 0.65%respectively in the week four, and PMC recorded least MC of 1.16% in week two.

IV. Discussions

4.1 Effect of packaging materials on moisture content of unfinished furniture parts.

In the experiment conducted in this study, the plastic packaging showed a significant increase in moisture content of the unfinished furniture parts with time as shown in Figure 1. The general practice in many of the industries is to wrap their unfinished furniture parts with plastic sheets with the view of protecting them from moisture absorption in the course of transit or at the warehouse. Unfortunately, pressure build-ups in the plastic wrapped cases, and can easily cause the furniture to absorb the little moisture in the enclosed air.

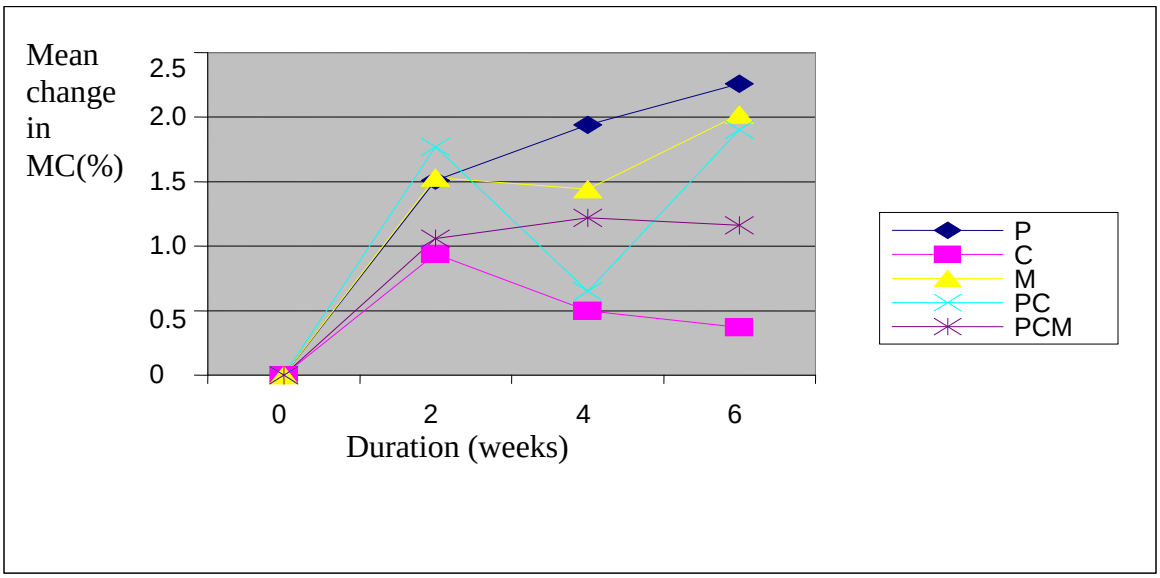

Figure 5.1 Effect of packaging materials on moisture content of unfinished furniture

Kollman and Côte (1984) emphasised that, wood is likely to absorb or desorbs moisture depending on the moisture content of the wood until equilibrium is attained. It was also noticed that, after the wood has taken up the moisture in the surrounding air after the first period (two weeks), it started releasing it. This depicted a slight decrease in moisture contents by 0.42% in week four with a further drop of about 0.03% in week six.

The cardboard packaging also recorded an increased (0.94%) in moisture content after the second week and dropped by 0.1% in week four and rose again in week six by 0.14%. Metallic packaging also showed similar trend of a rise and a continual drop after week two while Plastic plus cardboard packaging and PCM also showed a clear trend of rise and fall manner in moisture content with the three different time levels studied. This trend is presumed to have been caused by fluctuating temperature that caused the fluctuating humidity to affect the cardboard and in effect affecting the moisture content.

4.2 Effect of packaging materials on moisture content of semi-finished furniture parts

Effect of packaging materials on the moisture content of semi-finished furniture parts revealed general increase in the first two weeks. The highest and lowest moisture contents increase were 1.7% and 0.5% for plastic and cardboard respectively as shown in figure 2. The trend is due to the inability of the moisture to escape from the plastic pack, whiles in the case of the cardboard packaging, it was assumed to have absorbed some of the moisture, which was to have been taken by the furniture parts.

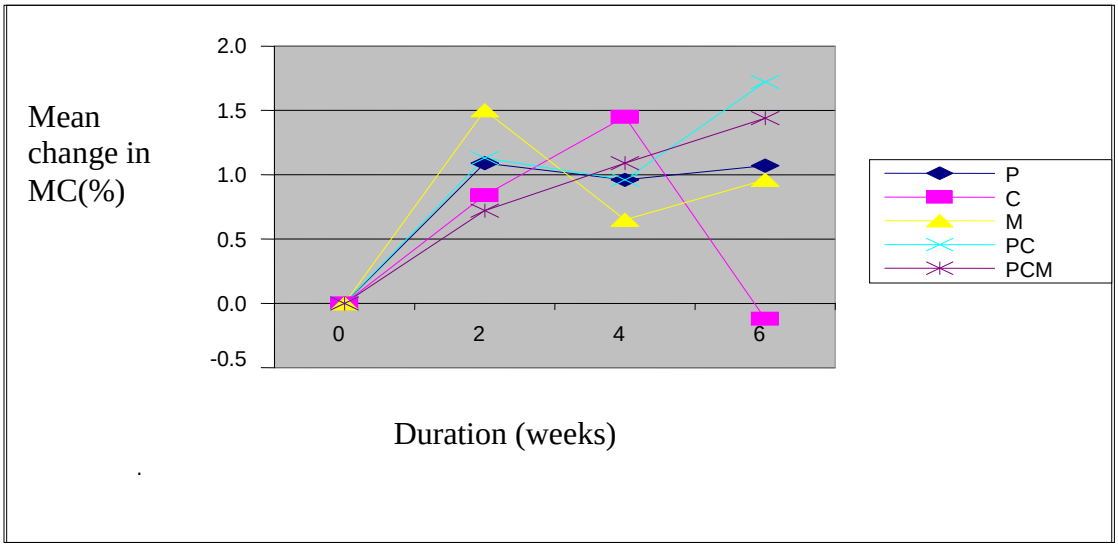

Figure 2. Effects of packaging materials on moisture content of semi-finished furniture

The fourth week duration revealed a different case, where all the materials that recorded high increase in MC rather gave low moisture content and vice-versa. Moisture content of furniture parts with plastic packaging dropped by 0.92%. This may be due to the fluctuating temperature and humidity during the fourth week period. However, it could also be seen that the change in temperature and humidity was too low to effect much change in the moisture contents of the packaged furniture parts.

Cardboard packaging recorded a lower increase in moisture content as compared to metallic, which gave a higher MC increase of 2.21%. Apart from plastic, that also showed a decrease in the four-week period, all the other three packaging methods with metallic parts recorded an increase in the four-week period as shown in figure 2. This can be attributed to the fact that the accumulated air in the corrugated inner layer of the cardboard tend to absorb the moisture in the pack since air is a good conductor of heat.

4.3 Effect of packaging materials on moisture content of finished furniture parts

The test conducted on the finished furniture parts to determine the effect revealed that plastic packaging recorded higher MC after the two weeks period and started dropping until it recoded a mean MC below the initial moisture content of the samples by the sixth week.

Cardboard that recorded low mean MC throughout all the three time levels also recorded a drop in MC as against the initial MC of the samples. This indicates that finished furniture parts that are to be packaged for a period of about six weeks must be packaged with cardboard. The metallic packaging also kept on dropping from 2% to 0.7% as indicated in figure 3.

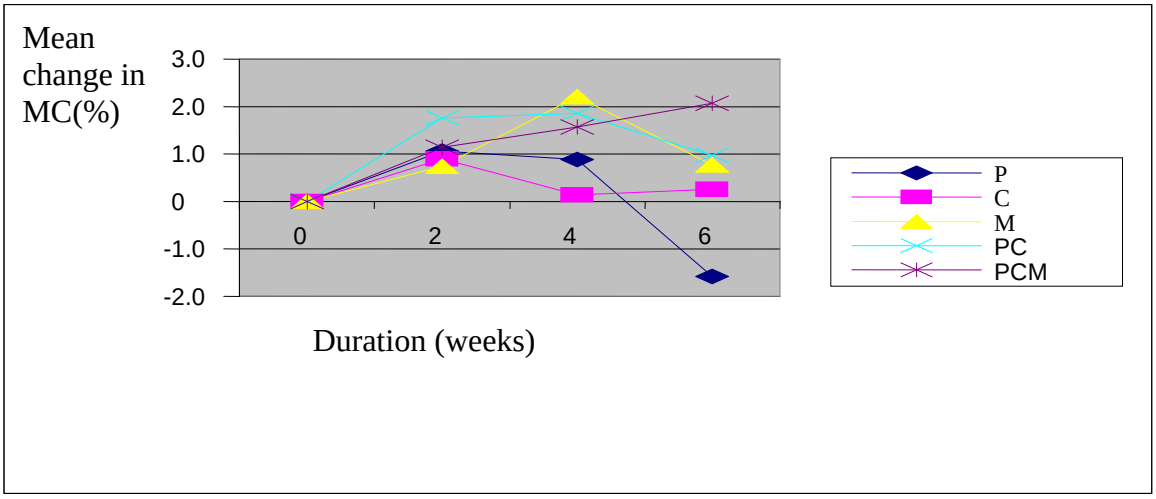

Figure 3. Figure 5.3 Effects of packaging materials on moisture content of finished furniture.

In the case of the plastic and cardboard packaging, the moisture contents of the different packaging materials were consistently decreasing after the two-week period as indicated in the figure 3, but the drop was minimal (0.2%) in week four, and subsequently dropped heavily by 1% in week six. This trend of change can be due to the continual heat generated in the packs because of the finish applied on the parts, which are highly volatile.

4.4 Discolouration of the surface finish of furniture parts.

A visual comparative analysis conducted with the photographs taken before and after packaging revealed that, there were no significant change in the colour with the samples packaged for the period of two weeks. However, 10% of the unfinished specimens packaged with plastic slightly darkened in colour from 14 to 12 on the colour chart. This could be attributed to the rise in MC (1.4%) of the furniture parts after packaging since moisture, dehydration and temperature has damaging effect on packaged furniture. The four-week duration also showed no difference in colour but in the case of the finished furniture parts, about three out of ten specimens were slightly faded from 14 to 16. This may also be due to the internal heat generated in the pack for two weeks as indicated in the work of Kollman and Côte, (1984)

V. Conclusion and Recommendations

This study on the effects of packaging materials on the moisture content (MC) and surface finish of packaged furniture was necessitated by the need to prevent the number of packaging problems associated with moisture changes that confront furniture manufacturers and exporters in Ghana. The study looked at the effect of five different packaging on the moisture content with time. The test was also concurrently undertaken on unfinished, semi-finished and finished furniture parts. It was found generally that cardboard packaging recorded the least change in MC throughout the test. However, plastic and metallic packaging recorded higher moisture contents in most of the test conducted.

In the case of the unfinished furniture, it was realized that cardboard and the combination of plastic and cardboard recorded the least and highest MC changes respectively for two weeks packaging. The four weeks packaging revealed the combination of plastic, cardboard and metallic as the method with least MC changes whiles metallic packaging recorded the highest.

However, metallic recorded the lowest while plastic and cardboard packaging recorded the highest MC change.

Semi-finished furniture parts packaging also proved that, cardboard recorded the lowest and highest moisture content respectively for the two weeks packaging. Metallic and cardboard-recorded least and highest moisture content respectively for four weeks while cardboard and metallic recorded lowest and highest MC changes respectively in the sixth week.

The finished furniture parts' test further proved cardboard's superiority as the packaging material with the lowest MC increase throughout the three time levels. That is, plastic for two weeks, plastic plus cardboard for four weeks and plastic plus cardboard and metallic for six weeks.

In the case of the effect of the packaging materials on the surface finish of the furniture parts, it was found generally that there was no clear-cut difference in the colours of the furniture parts before and after packaging. Some recorded slight differences but it was too slight that it could not be quantified.

It was noted that not every packaging situation should call for the same packaging material. The selection of a packaging material must be based on most likely events that are "normally" encountered. A packaging material to protect against every event would be extremely expensive if not practicable.

In view of the observations made during the study, the following recommendations are made:

1. For better understanding of the test, adequate samples of different ecological zones must be tested to confirm the absorption rate of the wood with the five different packaging materials.
2. Products to be packaged for six weeks are to be packaged in metallic container. However, for the problem of moisture ingression of the furniture parts due to high humidity on sea, they must be packaged with the plastic and cardboard and metallic combination as being done by some companies.
3. Furniture parts to be packaged for less than six weeks must be packaged with cardboard but in the case of high humidity, the plastic plus cardboard and metallic can be used with an absorbent material enclosed to absorb any form of moisture likely to build-up on the surfaces to change the colour of the wood.
4. Wood finishes add to the beauty and protect the product, but no finish is indestructible, therefore where plastics are to be employed on finished furniture, gaps should be created between the furniture surfaces and the plastic sheet since their ingredients can react to soften the finish to smudges and streaks when in contact with hard surfaces.

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Distribution of Silica in Different Density Fractions of Kusmunda Coals, Korba Coal Field

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Abstract - The huge amount of coal ash generated in our power plants can serve as important source for extraction of important ash constituents. Kusmunda area in Korba Coal field is an important source for power coals and therefore a bulk coal sample from the area has been studied in details as a test case for distribution of mineral matter in coal ash in different possible beneficiated and constituted dispatch material to power plants taking silica as a major and marker ash constituent. While, silica as percent of ash from samples under study remains within close limits, its retention in higher density is found much higher than in corresponding lower density.

Index Terms- Ash constituents, Silica, Density fraction

I. INTRODUCTION

Our main concern in utilization of coal for power industries has mostly been directed towards extraction of its fuel value while ash, the by-product of combustion of coal in our giant boilers, has so far been treated as hazardous waste of power industry despite its enormous mineral value for strategic and general industries. Almost the entire inorganic family of elements of periodic table is found to be present in the coal ash; some of these elements are present in major proportions while some in minor to trace levels in different combination states in coal ash. These constitute the inorganic value of coal ash and considering the enormous amount of ash production in power industry, coal ash can be considered as a mega potential source for these elements.

In view of the fact that silica constitutes nearly 60 per cent of the bulk of ash and at say 40 per cent ash in coal level, it represents 24 parts by weight of coal put in boiler, the present studies have been confined to the distribution of silica, as we try to fractionate a bulk coal sample from Kusmunda area, Korba Coal field, in the state of Chhattisgarh, into different size and density fractions, a normal physical method for obtaining suitable quality coals for a particular industry.

II. EXPERIMENTAL PARTICULARS

A bulk coal sample has been drawn from the Kusmunda area, Korba Coal field and brought to the Central Institute of Mining and Fuel Research CIMFR laboratory, Bilaspur for processing and preparation of desired laboratory samples for characterization and detailed studies.

The bulk sample after weighing was passed through 100 mm screen crushing any plus 100 mm material manually and without using much violence. The bulk now is put on screens of 25mm, 13mm, 6mm, 3mm, and 0.5mm apertures. Thus different size fractions 100-25mm, 25-13mm, 13-6mm, 6-3mm, 3-0.5mm and -0.5mm were obtained. Individual size fractions were weighed and aliquot representative samples were separated by thorough mixing and coning and quartering. These aliquots of individual size fractions were weighed and used for density separation using float and sink test. For this purpose, liquids of different densities of 1.3, 1.4, 1.5, 1.6, 1.7, 1.8, and 1.9 g/cc. were prepared by mixing benzene, carbon tetrachloride and bromoform in desired proportions.

During actual of gradually experiment, solutions increasing densities were taken in different containers in a row and a particular coal sample, say 100-25 mm material, was put in first container containing 1.3 density liquid. The material floating in the liquid was carefully separated and labeled as sample of < 1.3 density. The sink material was taken out and dried in air and then put in solution of next higher density, say 1.4. The float of this density was labeled as 1.3-1.4 density material. The sink was again put in next higher density liquids i.e. 1.5. The process was continued till we separated the floats and sinks in 1.9 density solution. The floats and sinks of different densities thus separated are dried in air and weighed individually. The process was repeated for aliquots of different size fractions separated from the bulk coal sample.

Floats and sinks of all size fractions were weighed individually and crushed to prepare 72 mesh size coal samples as per BIS: 436, (Part 1/ Section 1) - 1964 and IS: 436, (Part 2/ Section 2) - 1976.

A bulk coal sample representing properties and qualities of original bulk sample collected from the mine was also prepared by mixing powdered (72 mesh) individual size fractions in weight proportions (as obtained on screening of the bulk coal sample).

The 72 mesh ground samples of individual size fractions, their respective floats and sinks and the constituted bulk sample formed the material for characterization and ash analysis in the laboratory.

About 15-20 g of individual sample(s) is taken and uniformly distributed in a silica dish. The dish with the sample is now introduced in a ventilated muffle furnace at room temperature and the temperature of the furnace is gradually raised to 850^0C and maintained them till complete combustion. The ash is cooled to room temperature and finely ground using an agate mortar. Thus finely ground and uniformly mixed ash is again heated for an hour or so at 815^oC in muffle furnace to burn any unburnt carbon residue in the ash.
Ash for individual floats and sinks of different densities of 100-25mm, 25-13mm, 13-6mm, 6-3mm, 3-0.5mm size and the bulk coal samples have been obtained as above and subjected to detailed ash analysis following BIS: 1355-1984.
The distribution of Silica as a marker and major constituent of coal ash in different density fractions of individual size fractions has been followed as below. Though the major proportion of silica found in coal ash, is never present in coal as such (free silica), the results have been projected in terms of silica on coal for ease of appreciation.

III. RESULTS AND DISCUSSION

The results of distribution of silica in different size fractions have been summarized in Table-1.

Table: 1 Distribution of Silica (SiO_2) in different size fractions

Size Fraction(in mm)	Wt. (in Tonnes)	Silica(in Tonnes)	Silica as % of mass of size consist	Silica as Part of Total Silica in bulk coal
100-25	54.9	14.24	25.93	61.48%
25-13	8.8	1.97	22.43	8.51%
13-6	13.1	2.69	20.56	11.61%
6-3	6.2	1.03	16.57	4.45%
3-0.5	10.1	1.66	16.46	7.16%
-0.5	6.9	1.57	22.72	6.78%
Bulk Coal	100.0	23.16	23.16	100%

From Table 1 it has seen that 100 tonnes of the bulk coal contains 23.16 tonnes silica in its ash. Of this 14.24 tonnes silica is present in 100-25mm, 1.97 tonnes silica in 25-13mm, 2.69 tonnes in 13-6mm, 1.66 tonnes in 3-0.5mm and 1.57 tonnes silica is present in -0.5mm size material.
The results of distribution of silica in different density fractions separated from the 100-25 mm size fraction material obtained from the bulk coal sample have been summarized in the table below.

Table- (2a) SiO_2 in 100-25mm Screen Fraction

Density range	Wt %	Dry ash%	SiO_2%	SiO_2 as % of dry coal	Fraction of SiO_2 in 100-25mm coal	Fraction of SiO_2 in bulk coal	Fraction % of total SiO_2 in bulk coal
<1.3	4.4	9.1	61.16	5.56	0.24	0.132	0.6
1.3-1.4	20.7	17.8	61.60	11.0	2.27	1.246	5.4
1.4-1.5	25.1	28.6	63.52	18.08	4.56	2.503	10.8
1.5-1.6	15.6	39	64.44	25.13	3.92	2.152	9.3
1.6-1.7	8.1	46.7	64.27	30.01	2.43	1.334	5.8
1.7-1.8	6.8	53.8	65.00	35.00	2.38	1.307	5.6
1.8-1.9	5.9	60.8	65.22	39.65	2.34	1.285	5.5
>1.9	13.4	89.8	64.83	58.20	7.80	4.282	18.5
Overall	100	40.4	64.20	25.93	25.93	14.24	61.5

The 100-25 mm size fraction representing 54.9 per cent mass of the bulk Kusmunda coal sample is found to have 14.24 tonnes of SiO_2 in its coal ash; this is 61.5 % of the total 23.16 tonnes of SiO_2 in coal ash from the original bulk coal sample. This is further distributed in different densimetric fractions between <1.3 density and >1.9 density. In <1.3 density material, SiO_2 concentration is found to be the lowest, a mere 0.13 tonne and increases up to 2.5 tonnes in 1.4-1.5 density material. Silica is practically found constant at around 1.3 tonnes in fractions between 1.6-1.9 densities. The >1.9 density material has retained 4.28 tonnes of silica in its ash. Thus 1.4-1.6 density material retains 20.1% silica, > 1.9 density material retains 18.5 % while 28.2 % of silica of original bulk coal sample is found distributed in <1.3, 1.3-1.4, and 1.6-1.9 density fractions

The distribution of silica in 25-13 mm size fraction has been presented in table below.

Table- (2b) SiO_2 in 25-13mm screen fraction

Density range	Wt %	Dry ash%	SiO_2%	SiO_2 as % of dry coal	Fraction of SiO_2 in 25-13mm coal	Fraction of SiO_2 in bulk coal	Fraction % of total SiO_2 in bulk coal
<1.3	17.6	8.8	60.24	5.30	0.93	0.082	0.4
1.3-1.4	21.6	19.1	60.77	11.61	2.51	0.221	1.0
1.4-1.5	17.6	28.4	62.28	17.69	3.11	0.274	1.2
1.5-1.6	13.9	38.2	63.13	24.12	3.35	0.295	1.3
1.6-1.7	6.8	47.6	64.42	30.66	2.09	0.184	0.8
1.7-1.8	4.9	54	65.32	35.27	1.73	0.152	0.7
1.8-1.9	3.9	59	65.56	38.68	1.51	0.133	0.6
>1.9	13.7	81.2	64.60	52.52	7.20	0.634	2.7
Overall	100	35.3	63.54	22.43	22.43	1.98	8.5

The 25-13mm size fraction represents 8.8 per cent of the mass of the bulk sample and is found to contain 1.98 tonnes of total 23.16 tonnes of SiO_2 contributing 8.5% to the total SiO_2 of the original bulk coal sample. On density separation, highest proportion 0.634 tonne is found retained in > 1.9 density material while the < 1.3 density fraction has a mere 0.082 tonnes of silica in its ash. The 1.3-1.6 density material is found to retain 0.79 tonnes of silica and the remaining 0.469 tonnes of silica are found distributed in 1.6-1.9 density material. In terms of total silica in the bulk coal sample, 3.5 % of it is found in 1.3-1.6 density fraction,0.21 % in 1.6-1.9 density fractions, 0.4 % in <1.3 density material and the remaining 2.7 % is found retained in the >1.9 density material

Silica distribution in different density fractions of 13-6 mm size fraction of the bulk Kusmunda coal sample has been presented in the table below.

Table- (2c) SiO_2 in 13-6mm screen fraction

Density range	Wt %	Dry ash%	SiO_2%	SiO_2 as % of dry coal	Fraction of SiO_2 in 13-6 mm coal	Fraction of SiO_2 in bulk coal	Fraction % of total SiO_2 in bulk coal
<1.3	24.3	6.5	58.66	3.81	0.93	0.122	0.5
1.3-1.4	21.3	16.7	60.38	10.09	2.15	0.282	1.2
1.4-1.5	13	26.9	62.97	16.94	2.20	0.288	1.2
1.5-1.6	13.3	36.5	63.30	23.10	3.07	0.402	1.7
1.6-1.7	6.6	45.9	64.17	29.45	1.94	0.254	1.1
1.7-1.8	2.6	51.2	63.86	32.70	0.85	0.111	0.5
1.8-1.9	3.6	56.3	65.97	37.19	1.34	0.176	0.8
>1.9	15.3	81.8	64.57	52.81	8.08	1.058	4.6
Overall	100	32.4	63.47	20.56	20.56	2.69	11.6

The 13-6mm size fraction representing 13.1 per cent of the bulk Kusmunda coal sample contributes 2.69 tonnes or 11.6% silica of total silica of the bulk coal. Silica concentration is found maximum in >1.9 density material amounting to 1.058 tonnes constituting 4.6 % of total silica of the bulk coal sample. The 1.5-1.6 density fraction retains 0.40 tonne representing 1.7% of total silica. The material up to 1.5 densities retains 0.69 tonnes of silica forming 2.9 % of silica of original coal. The remaining 0.94 tonnes is found distributed in 1.6-1.9 density material representing 2.4 % of total silica of the bulk coal sample.

Silica distribution in 6-3 mm size fraction has been given in table below.

Table- (2d) SiO_2 in 6-3mm screen fraction

Density range	Wt %	Dry ash%	SiO_2%	SiO_2 as % of dry coal	Fraction of SiO_2 in 6-3mm coal	Fraction of SiO_2 in bulk coal	Fraction % of total SiO_2 in bulk coal
<1.3	33.1	3.7	58.5	2.16	0.72	0.045	0.2
1.3-1.4	19.2	12.6	58.9	7.42	1.42	0.088	0.4
1.4-1.5	13.8	25	61.04	15.26	2.11	0.131	0.6
1.5-1.6	8.8	35.2	62.85	22.12	1.95	0.121	0.5
1.6-1.7	4.9	42.8	63.42	27.14	1.33	0.082	0.4
1.7-1.8	2.4	46.8	65.02	30.43	0.73	0.045	0.2
1.8-1.9	2.8	51.4	64.45	33.13	0.93	0.058	0.2

>1.9	15.0	77	63.67	49.03	7.35	0.456	2.0
Overall	100	26.4	62.80	16.54	16.54	1.03	4.4

The 6-3mm size fraction from Kusmunda coals representing 6.2 per cent of the bulk, contributes 4.4% or 1.03 tonnes of 23.16 tonnes of the total silica in bulk coal. Silica is found more concentrated in 1.4-1.6 density fraction representing 0.252 tonnes or 1.1 % of total silica of bulk sample. Silica in density fraction up to 1.4 densities is 0.6 % and between 1.6-1.9, it would be 0.185 tonnes or 0.8 % of total silica in bulk coal sample. The >1.9 density material retains 0.456 tonnes or 2 per cent of total silica present in the bulk coal.

Silica distribution in 3-0.5 mm finer coals is presented below.

Table- (2e) SiO_2 in 3-0.5mm screen fraction

Density range	**Wt %**	**Dry ash%**	**SiO_2%**	**SiO_2 as % of dry coal**	**Fraction of SiO_2 in 3-0.5mm coal**	**Fraction of SiO_2 in bulk coal**	**Fraction % of total SiO_2 in bulk coal**
<1.3	2.2	3	56.18	1.69	0.37	0.037	0.2
1.3-1.4	35.8	9.5	58.89	5.59	2.00	0.202	0.9
1.4-1.5	9.9	23.9	60.12	14.37	1.42	0.143	0.6
1.5-1.6	6.8	33.9	62.11	21.06	1.43	0.144	0.6
1.6-1.7	3.3	39.4	62.56	24.65	0.81	0.082	0.4
1.7-1.8	2.5	44.8	62.99	28.22	0.71	0.072	0.3
1.8-1.9	2.1	48.2	63.44	30.58	0.64	0.065	0.3
>1.9	17.4	81.8	64.12	52.41	9.12	0.921	4.0
Overall	100	26.3	62.57	16.50	16.50	1.67	7.2

The 3-0.5mm size material forming 10.1 per cent of the bulk coal sample from Kusmunda area, contributes 1.67 tonnes or 7.3% of total SiO_2 present in bulk coal. The >1.9 density material analyses 0.921 tonnes of silica in its ash representing 4% of total silica of bulk coal. The <1.3 density fraction retains 0.037 tonnes , a mere 0.2% of total silica of bulk sample while 1.3-1.6 density fraction is found to retain 2.1% of total silica. Density fractions between 1.6-1.9 densities retain 0.219 tonnes of silica forming 1 % of total silica in bulk coal.

It is thus seen that in coal sample from Kusmunda area, major proportion of total silica is retained in 100-25 mm fraction. The contribution of other size fractions in retention of silica is in the order 13-6mm fraction > 25-13 mm fraction > 3-0.5 mm fraction > -0.5mm material > 6-3mm size fraction. Again in different size fractions, the maximum silica is found to be retained in material > 1.9 density.

IV. BENEFICIATION OF DIFFERENT SIZE FRACTIONS AT 1.6 DENSITY

The coal beneficiation studies are usually aimed at obtaining a suitable density for separation of it two or more products of the exercise to sustain criterion of ash and gainful utilization of coal resources. Based on our studies of separation of individual size fractions of Korba coals in liquids of densities between 1.3 and 1.9, we have selected 1.6 and 1.8 densities of separation of individual size fractions for presentation of yield and quality of separated floats and sinks and distribution of silica in them.

The results of the exercise have been compiled as below**:**

Table: (3a) Distribution of Silica in different size fractions at 1.6 density liquid

Size fractions	**Floats at 1.6 density**			**Sinks at 1.6 density**		
	Yield	**Ash%**	**Floats(Silica)**	**Yield**	**Ash%**	**Sinks(Silica)**
100-25 mm	36.12	26.40	6.03	18.78	63.40	8.21
25- 13mm	6.20	22.61	0.86	2.6	65.70	1.09
13-6mm	9.42	18.76	1.09	3.68	64.30	1.60
6-3mm	4.64	13.60	0.38	1.56	62.10	0.65
3-0.5mm	7.54	11.70	0.52	2.56	65.00	1.14

When individual size fractions obtained from the bulk coal sample from Kusmunda area, Korba CF are put in a liquid of 1.6 density, the yield of floats would be 36.12, 6.20, 9.42, 4.64 and 4.14 tonnes with 26.40, 22.61, 18.76,13.6 and 11.7 per cent ash for 100-25 mm, 25-13 mm, 13-6 mm, 6-3 mm and 3-0.5 mm size fractions respectively. The corresponding sinks would be 18.78, 2.6, 3.68, 1.56 and 2.56 tonnes with 63.4, 65.7, 64.3, 62.1 and 65.0 per cent ash. This would also leave an unwashed 6.9 tonnes of x 0.5 mm fines containing 1.57 tonnes of silica in its ash.

A 100 tonne bulk Kusmunda coal sample with 23.16 tonnes of silica in its ash retains. Of this 14.24 tonnes of silica is present in ash of 100-25mm size fraction. After beneficiation treatment at 1.6 density 6.03 tonnes of silica is retained in floats while 8.21 tonnes of silica goes with the sinks.

The 25-13mm and 13-6mm size fractions, respectively retain 0.86 tonnes and 1.09 tonnes of silica in the floats while 1.09 and 1.60 tonnes of silica goes with the sinks .In 6-3mm and 3-0.5mm sizes, 0.38 and 0.52 tonnes of silica remains in the floats while 0.65 and 1.14 tonnes goes along with the sinks of these size consists respectively.

Thus sinks at 1.6 density still have sufficient fuel value and have potential for their utilization in power generation with or without requiring specialized high ash combustors, the ash generated in these boilers will be found more concentrated in terms of silica and so also for other ash constituents.

In quest of extracting higher quantities of power coals while leaving behind as little quantity as possible of true rejects devoid of practically any meaningful caloric content.

V. BENEFICIATION OF DIFFERENT SIZE FRACTIONS AT 1.8 DENSITY

The floats and sinks at 1.8 density level can be investigated in terms of yield of floats and distribution of silica. The results of the exercise have been as per below.

Individual size fractions separated from the bulk coal sample were put in a solution of 1.8 density prepared by mixing appropriate quantities of carbon tetra chloride and bromoform liquids.

The separated floats and sinks were air dried and weighed and laboratory samples were prepared for characterization and ash analysis.

The results of the exercise have been compiled on the next page.

Table: (3b) Distribution of Silica in different size fractions at 1.8 density liquid

Size Fraction	At 1.8 density					
	Yield	Ash%	Floats(Silica)	Yield	Ash%	Sinks(Silica)
100-25mm	44.3	30.70	8.67	10.6	73.80	5.57
25-13mm	7.25	26.55	1.21	1.55	76.00	0.76
13-6mm	10.62	22.00	1.46	2.48	72.50	1.23
6-3mm	5.10	16.31	0.51	1.10	69.40	0.52
3-0.5mm	8.13	13.86	0.67	1.97	71.90	0.99

It is pertinent to note that ash in sinks of all the size fractions at 1.8 density analyse pretty high ash and can be considered as true dirts. However, they still have some fuel value which can be extracted in fluidized bed combustors for power generation.

The yield of floats for 100-25 mm size fraction would be 44.3 tonnes with 30.7 per cent ash. For 25-13 mm material, the yield of floats at 1.80 sp gr would be 7.25 tonnes with 26.55per cent ash. The floats of 13-6 mm fraction would be 10.62 tonnes with 22.0 per cent ash while the floats for 6-3 mm and 3-0.5 mm material are found to be 5.10 and 8.13 tonnes with 16.31 and 13.86 per cent ash respectively. The yield of sinks would be 10.6 tonnes with 73.80 per cent ash for 100-25mm, 1.55 tonnes with 76.00 per cent ash for 25-13mm, 2.48 tonnes with 72.50 per cent ash for 13-6mm, 1.10 tonnes with 69.40 per cent ash for 6-3 mm and 1.97 tonnes with 71.90 per cent ash for 3-0.5mm size fractions leaving 6.9 tonnes of unwashed 0.5 mm fines.

The amount of silica retained in floats of these different size fractions is found to be 8.67 tonnes for 100-25 mm, 1.21 tonnes for 25-13 mm, 1.46 tonnes for 13-6 mm, 0.51 tonnes for 6-3 mm and 0.67 tonnes for 3-0.5 mm size consists. The unwashed x0.5 mm material retains 1.57 tonnes of silica.

Sinks for these size fractions at 1.80 density has been found to be 5.57, 0.76, 1.23, 0.52 and 0.99 tonnes respectively.

Thus if 93.10 tonnes of 100-0.5 mm material is put in a liquid of 1.8 sp.gravity, the yield of floats would be 75.40 tonnes retaining 12.52 tonnes of silica while the material sinking in the liquid would be 17.70 tonnes with 9.07 tonnes of silica.

VI. BENEFICIATION OF DIFFERENT CUMULATIVE SIZES AT 1.6 DENSITY

The coals excavated in a mine is generally fractioned in two or three broad size fractions suggested on the basis of detailed screen analysis data generated in a laboratory study. The exercise enables coal producers to optimize their coal supplies while suiting the requirements of the consumer industries. Therefore, ash generated in a particular plant and therefore, quantities of different ash constituents would depend upon the quality and quantity of combination fractions included or excluded from the coal supplies. Again coal beneficiation may be pursued for such constituted samples to improve their quality leaving one or more size fraction(s) untreated and mixed in the product for final dispatch.

Accordingly studies were made to ascertain the distribution of ash constituents in different possible combinations in coal supplies. The results of beneficiation of the possible combinations at 1.6 and 1.8 density of separation have been summarized below:

Table: (4a) Distribution of silica in different cumulative sizes at 1.6 density liquid

Size fractions	At 1.6 density					
	Yield	Ash%	Floats(Silica)	Yield	Ash%	Sinks(Silica)
+25mm	36.12	26.40	6.03	18.78	63.30	8.21
+13mm	42.32	25.90	6.89	21.38	63.60	9.32
+6mm	51.74	25.40	7.98	25.06	63.70	10.92
+3mm	56.38	24.30	8.36	26.62	63.60	11.57
+0.5mm	63.92	22.80	8.88	29.18	63.70	12.73

On density separation of broad size constituted samples at 1.6 sp. gravity, the yield of floats increases while the ash per cent decreases as we include more and more finer sizes .The yield of floats of +25mm has been found to be 36.12 tonnes with 26.40 per cent ash, For + 13 mm material the yield would be 42.32 tonnes with 25.90 per cent ash, for +6mm, the yield of floats would be 51.74 tonnes with 25.40 per cent ash while the yield of floats for +3 and +0.5 mm constituted samples would be 56.38 and 63.92 tonnes with 24.30 and 22.80 per cent ash respectively.

This also increases proportion of ash constituents in the floats and sinks on inclusion of more fine sizes in the constituted samples. Thus silica retained in the floats ash increases from 6.03 tonnes to 8.88 tonnes as we move from +25 mm material to +0.5 mm constituted sample. Silica for the sinks of these samples also follows the same trend as it increases from 8.21 tonnes to 12.71 tonnes on moving from +25 mm to +0.5 mm constituted sample. The corresponding finer size material (-25mm,-13 mm, -6 mm, -3 mm and -0.5 mm) accordingly would have lesser retention of these constituents in their ash.

VII. BENEFICIATION OF DIFFERENT CUMULATIVE SIZES AT 1.8 DENSITY

On repeating the exercise of separation of floats and sinks from constituted samples at 1.8 density, the results of distribution of silica obtained have been compiled below.

Table: (4b) Distribution of silica in different cumulative sizes at 1.8 density liquid

Size fractions	At 1.8 density					
	Yield	Ash%	Floats(Silica)	Yield	Ash%	Sinks(Silica)
+25mm	44.30	30.70	8.67	10.60	73.80	5.57
+13mm	51.55	30.10	9.88	12.05	74.10	6.33
+6mm	62.17	28.70	12.12	14.63	73.80	6.78
+3mm	67.27	27.80	12.63	15.73	73.50	7.31
+0.5mm	75.40	26.30	13.30	17.70	73.30	8.29

The amount of floats and sinks separated from constituted broad size fractions of Kusmunda coal sample at 1.8 density increases from + 25 mm material to +0.5 mm coals. The ash per cent of floats decreases as we include more and more finer size material while the ash per cent of corresponding sinks are higher than floats but do not follow a regular order. The ash per cent of sinks of +13mm size is highest in 1.8 density and lowest in +0.5mm size. In floats, silica increases from 8.67 tonnes to 13.30 tonnes as we move from +25 mm floats to +0.5 mm floats. Similar trend is maintained in sinks too and the silica increases from 5.57 tonnes to 8.29 tonnes as we move from +25 mm sinks to +0.5 mm sinks.

The Indian coals, due to their drift origin have, been responsible for generation of extremely large quantities of ash in major power plants of the country. Since coal ash is known to contain a number of major and minor ash constituents of commercial significance the huge tonnage of ash generated in our plants can be treated as source material for these constituents. In the light of this the present studies were under taken on a bulk coal sample from Kusmunda mine, Korba Coalfield in Korba District of Chhattisgarh**.**

To evaluate the movement and distribution pattern of silica in different density fractions individual size fractions from the bulk coal sample were subjected to float and sink tests using organic liquid mixtures of different densities between 1.3 and 1.9. Aliquots of separated samples representing density fractions of individual size consists of the original bulk sample were characterized and studied for silica content in their ash.

It is seen that there is an increase in silica content as we move to higher density fractions. Thus the100-25mm size coal which has14.24 tonnes SiO_2 in its coal ash amounting to 61.5 per cent of the total 23.16 tonnes of SiO_2 in coal ash from the original bulk coal sample. In <1.3 density material, SiO_2 retained is 0.13 tonne and the amount increases up to 2.5 tonnes in 1.4-1.5 density material. It is practically found constant at around 1.3 tonnes in fractions between 1.6-1.9 densities. The >1.9 density material has 4.28 tonnes of silica in its ash. Thus 1.4-1.6 density material retains 20.1 per cent silica, > 1.9 material retains 18.5 per cent while 28.2 per cent of silica of original bulk coal sample is found distributed in <1.3, 1.3-1.4, and 1.6-1.9 density fractions.

The 25-13mm size fraction contains 1.98 tonnes of of SiO_2 constituting 8.5 per cent of total SiO_2 of the original bulk coal sample. On density separation, silica is found concentrating more in higher density fractions, It is 0.634 tonnes in >1.9 density material while <1.3 density fraction has a mere 0.082 tonnes of silica in its ash. Similarly the 1.3-1.6 density material is found to retain 0.79 tonnes of silica and the remaining 0.416 tonnes of silica are found distributed in 1.6-1.9 density material. In terms of total silica in the bulk coal

sample, 3.5 per cent of it is found in 1.3-1.6 density fraction,0.21 per cent in 1.6-1.9 density fractions, 0.4 per cent in <1.3 density material and the remaining 2.7 per cent is found retained in the >1.9 density material

The 13-6mm size fraction of Kusmunda coals contributes 2.69 tonnes or 11.6 per cent silica of total silica in bulk coal. Silica concentration is found maximum in >1.9 density material amounting to 1.058 tonnes constituting 4.6 per cent of total silica of the bulk coal sample. The 1.5-1.6 density fraction retains 0.40 tonne which would be 1.7 per cent of total silica The material up to 1.5 density retains 0.69 tonnes of silica forming 2.9 per cent of silica of original coal. The remaining 0.54 tonnes is found distributed in 1.6-1.9 density material representing 2.4 per cent of total silica of the bulk coal sample.

The 6-3mm size fraction from Kusmunda coals contributes 4.4 per cent or 1.03 tonnes of 23.16 tonnes of the total silica in bulk coal. Silica is found more concentrated in 1.4-1.6 density fraction representing 0.252 tonnes or 1.1 per cent of total silica of bulk sample. Silica in density fraction up to 1.4 densities is 0.6 per cent and between 1.6-1.9, it would be 0.185 tonnes or 0.8 per cent of total silica in bulk coal sample. The >1.9 density material retains 0.456 tonnes or 2 per cent of total silica present in the bulk coal.

The 3-0.5mm size material from Kusmunda coals contributes 1.67 tonnes or 7.3 per cent of total SiO_2 present in bulk coal. Silica is found more concentrated in >1.9 density material, 0.921 tonnes, representing 4 per cent of total silica of bulk coal. The <1.3 density fraction retains 0.037 tonnes, a mere 0.2 per cent of total silica of bulk sample while 1.3-1.6 density fraction retains 2.1 per cent of total silica. Density fractions between 1.6-1.9 densities retain 0.219 tonnes of silica forming 1 per cent of total silica in bulk coal

Based on the results of separation of individual size fractions of Korba coals in liquids of densities between <1.3 and >1.9, 1.6 and 1.8 densities have been selected for separation of individual size fractions for presentation of yield and quality of separated floats and sinks and distribution of ash constituents in them.

When this sample is fractionated in different size fractions and put in a liquid of 1.6 density, the yield of floats would be 36.12 tonnes of 100-25 mm material, 6.2 tonnes of 25-13 mm fraction, 9.42 tonnes of 13-6 mm fraction, 4.64 tonnes of 6-3 mm material and 7.54 tonnes of 3-0.5 mm material, leaving 6.9 tonnes of unwashed 0.5 mm fines.

A 100 tonne bulk Kusmunda coal sample has 23.16 tonnes of silica in its ash. Out of this 14.24 tonnes of silica is present in ash of 100-25mm size fraction. After beneficiation treatment at 1.6 density 6.03 tonnes of silica is retained in floats while 8.21 tonnes of silica remains in sinks. In 25-13mm and 13-6mm size fractions, respectively 0.86 tonnes and 1.09 tonnes of silica is retained in the floats while 1.11 and 1.60 tonnes of silica goes with the sinks of these size fractions. In 6-3mm and 3-0.5mm sizes, 0.38 and 0.52 tonnes of silica is present in floats while 0.65 and 1.13 tonnes is retained in sinks of these size consists respectively.

When this sample is fractionated in different size fractions and put in a liquid of 1.8 density, the yield of floats would be 44.3 tonnes of 100-25 mm material, 7.59 tonnes of 25-13 mm fraction, 11.10 tonnes of 13-6 mm fraction, 5.27 tonnes of 6-3 mm material and 8.34 tonnes of 3-0.5 mm material, leaving 6.9 tonnes of unwashed 0.5 mm fines.

The floats of 100-25mm material at 1.80 density will have 8.67 tonnes of silica leaving 5.57 tonnes in the sinks. The floats of 25-13mm size consists will have 1.21 tonnes of silica while the sinks would carry 0.76 tonnes, the floats of 13-6mm, 6-3mm and 3-0.5mm material will have 1.46, 0.51 and 0.67 tonnes of silica in their ash respectively while corresponding sinks would carry 1.23, 0.52 and 0.98 tonnes of silica respectively.

Again coal beneficiation may be pursued for special constituted samples to improve their quality leaving one or more size fraction(s) untreated and mixed in the product for final dispatch. Accordingly studies were made to ascertain the distribution of silica in different possible combinations in coal supplies.

On density separation of broad size constituted samples at 1.6 and 1.8 sp. gravities, the yield of floats and sinks increases as we include more and more finer sizes. This also increases proportion of silica in the floats and sinks leaving the corresponding untreated finer size consists poorer in these constituents.

The amount of floats separated from Kusmunda coals from different constituted coal samples at 1.6 sp. gravity shows a continuous increase from 36.12 per cent of the size consist for +25 mm material to 63.92 per cent for +0.5 mm coals. Since finer coals have better separation of dirt and coaly matter, and hence the lighter coal particles get preferentially included in the floats, there is an increase in the yield of sinks from +25mm i.e. 18.78 per cent to 29.18 per cent for +0.5 mm constituted material.

Silica increases from 6.03 tonnes of the total silica in bulk coal sample in floats of + 25 mm sample to 8.88 tonnes in floats of + 0.5 mm material and in sinks increases from 8.21tonnes of total silica in bulk coal for + 25 mm material to 12.73 tonnes of total silica in +0.5 mm sinks at 1.6 sp. gravity.

The amount of floats separated from Kusmunda coals at1.8 density increases from 44.3 per cent of the + 25 mm material to 75.4 per cent of the +0.5 mm coals. The same trend is followed weights of sinks from 10.6 per cent for + 25 mm coals to 17.7 per cent for the sinks of +0.5mm.

Silica in floats as per cent of total silica in bulk coal increases from 8.67 tonnes for + 25 mm material to 13.30 tonnes for floats of +0.5 mm coals. Similar increase in concentrations of these constituents is found as we move from sinks of +25mm coals to sinks of +0.5 mm coals from 5.57 tonnes to 8.31 tonnes.

VIII. CONCLUSION

It is seen that there is an increase in silica content as we move to higher density fractions. The100-25mm size coal which has14.24 tonnes SiO_2 in its coal ash amounting to 61.5 per cent of the total 23.16 tonnes of SiO_2 in coal ash from the original bulk coal sample. When this sample is fractionated in different size fractions and put in a liquid of 1.6 density, the yield of floats would be 36.12 tonnes of 100-25 mm material with 6.03 tonnes silica leaving 8.21 tonnes in sinks. After getting fractionated in different size fractions and

put in a liquid of 1.8 density, the yield of floats would be 44.3 tonnes of 100-25 mm material with 8.67 tonnes silica leaving 5.57 tonnes in the sinks. Thus for silica separation beneficiation of 100-25mm size fraction at 1.8 density can be suggested.

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Fast and Slow Modes on Dust Ion Acoustic Solitary Waves in A Warm Plasma

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***Abstract*-** Using Korteweg-de Vries (KdV) equation, solitons are investigated in a dusty plasma consisting of positive ion, negatively charged dust grain and Boltzmann distributed electron. Both compressive and rarefactive solitons are found to exist in the dusty plasma for fast and slow modes. Greater Z_d (Z_d is the number of elementary charges residing on the dust grain) is found to yield higher amplitude solitons. Further, it is observed that higher temperature ratio gives higher amplitude in case of slow mode and gives smaller amplitude in case of fast mode.

1. INTRODUCTION

Dusty or Complex plasmas are ordinary plasmas with embedded solid particles. The particles can be made of either dielectric or conducting materials and can have any shape. Dusty plasmas are present in interstellar clouds, circumstellar clouds, planetary rings, comets, cometary tails, asteroid zones, earth's atmosphere and magnetosphere [1- 11]. Dusty plasma can be produced in laboratories by Modified Q-machine, dc discharges, rf discharges etc [1]. Dusty plasma plays an important role in plasma crystals [12], coating and etching of thin films [13] etc. Rao *et al.* [14] have first reported the existence of the dust acoustic waves for low frequency in unmagnetized dust plasma and verified in a laboratory experiment by Barken *et al.* [15]. The linear and nonlinear features of both dust-acoustic wave and dust ion-acoustic wave have also reported in experimental and theoretical observations [16 - 19]. Roychoudhury and Mukharjee [20] have reported that the finite dust temperature restricts the region for the existence of nonlinear solitary waves. Shukla and Silin [21] have reported about the existence of dust ion acoustic waves for higher frequency. Nakamura and Sharma [22] have shown that the presence of negatively charged dust in plasma decreases the velocity of soliton and increases its width for a given hight of the peak. Ghosh [23] has investigated the role of negative ions in dusty plasma with variable dust charge. Tiwari and Mishra [24] have investigated the dust ion acoustic solitons in the plasma consisting positively or negatively charged dust and discussed the conditions in parameter space to obtain compressive and rarefactive dressed solitons. EI-Wakil *et al.* [25] have investigated the dust ion acoustic waves propagation in an inhomogeneous dusty plasma considering positive and negative dust grains. Mowafy *et al.* [26] have investigated the effect of dust charge fluctuation on the propagation of dust ion acoustic waves in inhomogeneous mesospheric dusty plasma. Das and Chatterjee [27] have investigated the formation of large amplitude double layers in a dusty plasma constituting warm dust grains, non-thermal electrons and two temperature isothermal ions using Sagdeev's pseudopotential technique. Baluku *et al.* [28] have investigated the dust ion acoustic solitons in an unmagnetized dusty plasma consisting cold dust particles, adiabatic fluid ions and electrons satisfying kappa distribution. Attia *et al.* [29] have investigated the higher order effects of positive and negative dust charge fluctuation on the propagation of dust ion acoustic waves in a weakly inhomogeneous, weakly coupled, collision less and unmagnetized mesospheric dust plasma consisting of four components dusty plasma. Barman and Talukdar [30] have investigated the propagation of ion-acoustic waves in a warm dusty plasma with electron inertia. Tiwari *et al.* [31] have investigated characteristics of ion acoustic soliton in dusty plasma, including the dynamics of heavily charged massive dust grains using Sagdeev potential method. Alinejad [32] has studied the properties of arbitrary amplitude dust ion acoustic solitary waves in a dusty plasma containing warm adiabatic ions, electrons following flat-trapped velocity distribution and arbitrarily (positively or negatively) charged mobile dust through pseudo-potential method. Recently Maitra [33] has investigated the dust acoustic solitary waves in a magnetized dusty plasma considering the effect of dust size and dust charge variations.
Very recently Malik *et al.* [34] have studied the soliton propagation in a moving electron-positron pair plasma having negatively charged dust grain including finite temperatures of the electrons and positrons. They have obtained two types of modes propagating in opposite directions.

In this paper, we consider a three component unmagnetized dusty plasma system consisting of positive ion, negatively charged dust grain and Boltzmann distributed electron. Our main objective is to study the effect of temperature on the amplitude and width of dust ion acoustic solitary waves in an unmagnetized plasma for fast and slow modes. The paper is organized as: In Sec 1, introduction, in Sec 2, the basic set of equations governing the plasma model and derivation of the KdV equation, in Sec 3, solitary wave solution, in Sec 4, discussion and finally, conclusions in Sec 5.

2. BASIC EQUATION AND DERIVATION OF KDV EQUATION

We consider a one dimensional, unmagnetized collision less dusty plasma with positive ion, negatively charged dust grain and Boltzmann distributed electron. The dynamics of dust ion acoustic waves is governed by the following equations:

$$\frac{\partial n_i}{\partial t} + \frac{\partial}{\partial x}\left(n_i v_i\right) = 0 \qquad (1)$$

$$\frac{\partial v_i}{\partial t} + v_i \frac{\partial v_i}{\partial x} + \frac{\alpha}{n_i}\frac{\partial p_i}{\partial x} + Z_d \frac{\partial \phi}{\partial x} = 0 \quad (2)$$

$$\left(\frac{\partial}{\partial t} + v_i \frac{\partial}{\partial x}\right) p_i + 3p_i \frac{\partial v_i}{\partial x} = 0 \quad (3)$$

$$\frac{\partial n_d}{\partial t} + \frac{\partial}{\partial x}\left(n_d v_d\right) = 0 \quad (4)$$

$$\left(\frac{\partial}{\partial t} + v_d \frac{\partial}{\partial x}\right) v_d + \frac{1}{Q'}\left(\frac{\alpha}{n_d}\frac{\partial p_d}{\partial x} - Z_d \frac{\partial \phi}{\partial x}\right) = 0 \quad (5)$$

$$\left(\frac{\partial}{\partial t} + v_d \frac{\partial}{\partial x}\right) p_d + 3p_d \frac{\partial v_d}{\partial x} = 0 \quad (6)$$

$$n_e = e^{\phi} \quad (7)$$

Again for charge imbalances these equations are to be combined by the Poisson equation

$$\frac{\partial^2 \phi}{\partial x^2} = n_e - n_i + Z_d n_d \quad (8)$$

where, *i*, e and d stand for positive ion, electron and negatively charged dust grain respectively, $Q' = \frac{m_d}{m_i}$ (= dust grain to ion mass ratio) , $\alpha = \frac{T_i}{T_e}$ (= ion to electron temperature with $T_i = T_d$, T_d is the dust temperature) and Z_d is the number of elementary charges residing on the dust grain.

We have normalized densities n_i, n_e and n_d by the unperturbed densities n_{e0}, pressures p_i, p_d by the characteristic ion pressure $k_b n_{e0} T_i$, time t by the inverse of the characteristic ion plasma frequency i.e., $\omega_{pi}^{-1} = \left(\frac{m_i}{4\pi n_{e0} e^2}\right)^{\frac{1}{2}}$, distance x by the electron Debye length $\lambda_{De} = \left(\frac{k_b T_e}{4\pi n_{e0} e^2}\right)^{\frac{1}{2}}$, velocities by the ion-acoustic speed $C_s = \left(\frac{k_b T_e}{m_i}\right)^{\frac{1}{2}}$, and the potential ϕ by $\frac{k_b T_e}{e}$; k_b is the Boltzmann constant.

To derive the KdV equation from the set (1) – (8), we use the stretched variables

$$\xi = \varepsilon^{\frac{1}{2}}\left(x - Ut\right), \quad \tau = \varepsilon^{\frac{3}{2}} t \quad (9)$$

with phase velocity *U* of the wave and using the following expansions of the flow variables in terms of the smallness parameter ε:

$$\begin{aligned} n_i &= n_{i0} + \varepsilon n_{i1} + \varepsilon^2 n_{i2} + \ldots, \\ n_d &= n_{d0} + \varepsilon n_{d1} + \varepsilon^2 n_{d2} + \ldots, \\ v_i &= \varepsilon v_{i1} + \varepsilon^2 v_{i2} + \ldots, \\ v_d &= \varepsilon v_{d1} + \varepsilon^2 v_{d2} + \ldots, \end{aligned} \quad (10)$$

$$\phi = \varepsilon\phi_1 + \varepsilon^2 \phi_2 + \ldots\ldots$$

With the use of of the transformation (9) and the expansion (10) in the normalized set of equations (1) – (8), we get the expression for the phase velocity and the KdV equation as follows-

$$\frac{r Z_{dd}^2}{\left(Q'U^2 - 3\alpha\right)\left(1 - rZ_d\right)} + \frac{Z_d}{\left(U^2 - 3\alpha\right)\left(1 - rZ_d\right)} = 1, \; r = \frac{n_{d0}}{n_{i0}} \quad (11)$$

This gives $U^2 = \frac{-b \pm \sqrt{b^2 - 4ac}}{2a}$

Where $a = \left(1 - rZ_d\right)$,

$b = -\left\{3\alpha\left(1 + Q'\right)\left(1 - rZ_d\right) + \left(Z_d Q' + rZ_d^2\right)\right\}$

$c = 9\alpha^2\left(1 - rZ_d\right) + 3\alpha\left(Z_d + rZ_d^2\right)$

and $\frac{\partial \phi_1}{\partial \tau} + p\phi_1 \frac{\partial \phi_1}{\partial \xi} + q\frac{\partial^3 \phi_1}{\partial \xi^3} = 0 \quad (12)$

where $p = \frac{A}{2B}$ and $q = \frac{1}{2B}$ with

$$A = \frac{3n_{i0}Z_d^2\left(U^2 + \alpha\right)}{\left(U^2 - 3\alpha\right)^3} - \frac{3n_{d0}Z_d^3\left(Q'U^2 + \alpha\right)}{\left(Q'U^2 - 3\alpha\right)^3} - 1$$

and $B = \frac{n_{d0}Q'Z_d^2 U}{\left(Q'U^2 - 3\alpha\right)^2} + \frac{n_{i0}Z_d U}{\left(U^2 - 3\alpha\right)^2}$

From the expressions of A and B, we observe that the nonlinear dust ion-acoustic solitons, in this model of plasma, exist when $U^2 \neq 3\alpha$, $\frac{3\alpha}{Q'}$ subject to the condition $Z_d < \frac{1}{r}$.

3. SOLITARY WAVE SOLUTION

To find a stationary solution of the KdV equation (12), we use the transformation $\chi = \eta - V\tau$. Using the boundary conditions $\phi_1 = \frac{\partial \phi_1}{\partial \eta} = \frac{\partial^2 \phi_1}{\partial \eta^2} = 0$ as $|\eta| \to \infty$, the KdV equation (12) can be integrated to give $\phi_1 = \frac{3V}{p}\sec h^2\left(\frac{1}{2}\sqrt{\frac{V}{q}}\chi\right)$, where *V* is the velocity with which the solitary waves travel to the right.

Thus, the wave amplitude of the soliton is given by $\phi_0 = \frac{3V}{p}$ and the corresponding width by $\Delta = 2\sqrt{\frac{q}{V}}$.

4. DISCUSSION

In presence of ion temperature in this model of plasma, both fast and slow compressive and rarefactive solitons are found to exist. The ion temperature is responsible for generating the fast and slow modes. For the fast mode, both compressive and rarefactive solitons are exist, but for slow mode only rarefactive solitons of small amplitude are found to exist. It is to be mentioned that the compressive fast ion-acoustic solitons exist

only for smaller values of $Z_d(>1)$ [Fig.1 (a)] and its amplitude increase as Z_d increases. Furthermore, they are found to exist only in upper regime of $Z_d(>1)$. On the other hand, rarefactive fast ion-acoustic solitions exist [Fig. 1(a)] only in the upper regime of Z_d for fixed $V=0.05$, $r=0.05$ and $Q'=10$ for different values of $\alpha=0.05(1),0.15(2),0.25(3)$. However, the character of the fast compressive solitons changes to fast rarefactive solitons after certain Z_d^* characterizing an uncountable region. The corresponding widths [Fig.1 (b)] of the fast (compressive and rarefactive) ion-acoustic solitons nonlinearly decrease with Z_d with negligible difference for $\alpha=0.05(1),0.15(2),0.25(3)$ for fixed $V=0.05$, $r=0.05$ and $Q'=10$. To the contrary, the amplitudes [Fig. 2(a)] of slow rarefacetive solitons increases parabolically (for higher values of α) with Z_d for fixed $V=0.05$, $r=0.05$ and $Q'=10$ for different values of $\alpha=0.05(1),0.15(2),0.25(3)$. But the corresponding widths [Fig. 2(b)] of the slow rarefactive solitons decreases parabolically (for higher values of α) with Z_d for the same set of values of the parameters. The amplitudes [Fig. 3(a)] of fast compressive soliton decrease linearly but rapidly with α for fixed $V=0.20$, $r=0.05$ and $Q'=10$ for different values of $Z_d=6,7,8$. The corresponding widths [Fig. 3(b)] of the fast compressive solitons are decrease linearly but slowly with α. On the other hand, the amplitudes [Fig. 4(a)] of slow rarefacetive solitons increase (numerically) uniformly with α for fixed $V=0.20$, $r=0.05$ and $Q'=10$ for different values of $Z_d=6,7,8$. But the corresponding widths [Fig. 4(b)] of the slow rarefactive solitons increase uniformly with α for fixed $V=0.20$, $r=0.05$ and $Q'=10$ for different values of $Z_d=6,7,8$. The amplitudes [Fig. 5(a)] of fast compressive solitons decreases sharply in the lower regime of $Q'(\le 4)$ and then decreases very slowly with Q' fixed $V=0.05$, $r=0.05$ and $Z_d=8$ for different values of $\alpha=0.10,0.15,0.20,0.25$. The corresponding widths [Fig. 5(b)] of the fast compressive solitons nonlinearly decrease with Q' fixed $V=0.05$, $r=0.05$ and $Z_d=8$ for different values of $\alpha=0.10,0.15,0.20,0.25$. On the other hand, the amplitudes [Fig. 6(a)] of slow rarefacetive solitons increases (numerically) nonlinearly with Q' for fixed $V=0.05$, $r=0.05$ and $Z_d=8$ for different values of $\alpha=0.10(1),0.15(2),0.20(3),0.25(4)$. But the corresponding widths [Fig. 6(b)] of the slow rarefactive solitons increases sharply in the lower regime of $Q'(\le 4)$ and then decreases very slowly for fixed $V=0.05$, $r=0.05$ and $Z_d=8$ for different values of $\alpha=0.10(1),0.15(2),0.20(3),0.25(4)$.

5. CONCLUSION

In this investigation, we have studied the effect of ion temperature and dust temperature in a three-component dusty plasma using reductive perturbation method. Two modes namely slow and fast are observed corresponding to different phase velocities. Both compressive and rarefactive solitons are found to exist in the dusty plasma for fast and slow modes. Greater Z_d is found to yield higher amplitude solitons. Further, it is observed that higher temperature ratio gives higher amplitude in case of slow mode and gives smaller amplitude in case of fast mode.

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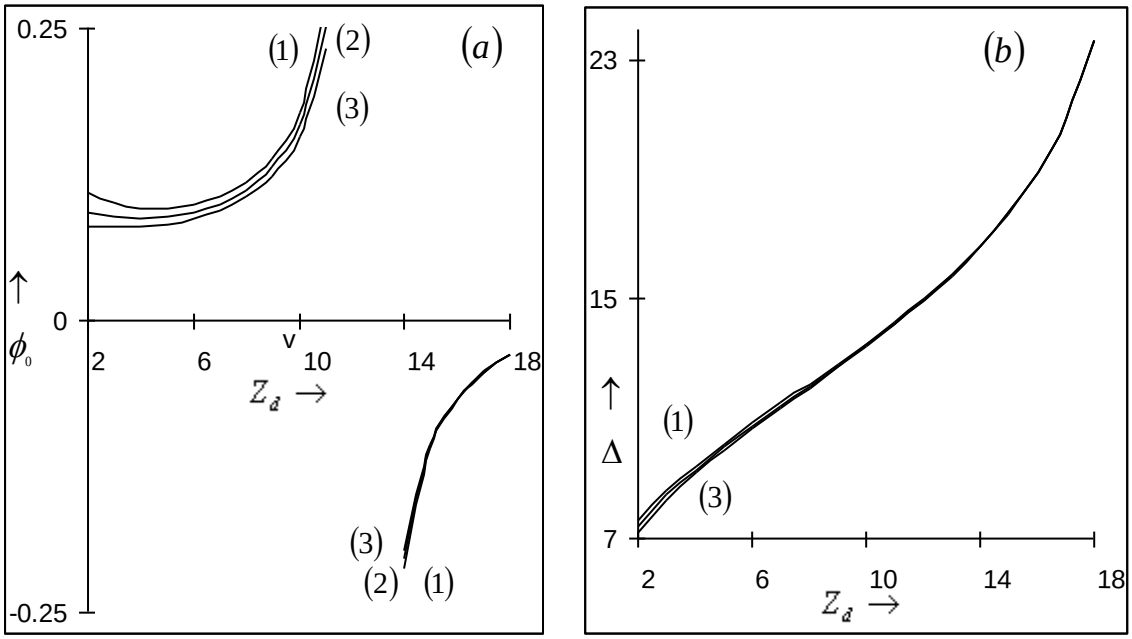

Figure1. Amplitudes (a) and widths (b) of fast compressive and rarefactive dust ion-acoustic solitons versus Z_d for fixed $V = 0.05$, $r = 0.05$ and $Q' = 10$ for different values of $\alpha = 0.05(1), 0.15(2), 0.25(3)$.

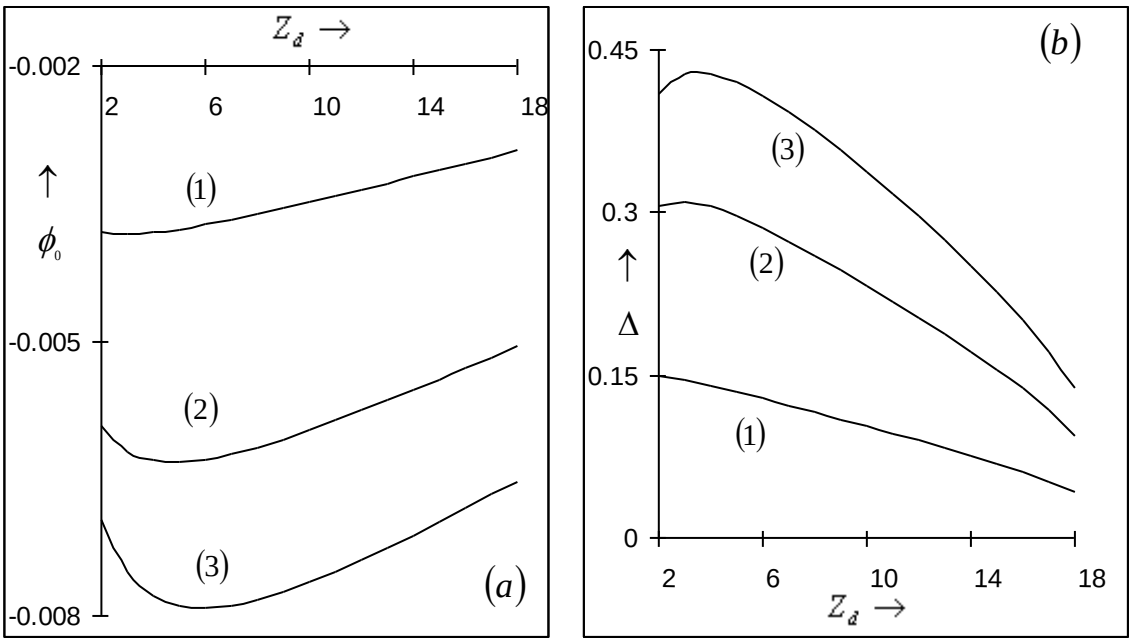

Figure 2. Amplitudes (a) and widths (b) of slow rarefactive dust ion-acoustic solitons versus Z_d for fixed $V = 0.05$, $r = 0.05$ and $Q' = 10$ for different values of $\alpha = 0.05(1), 0.15(2), 0.25(3)$.

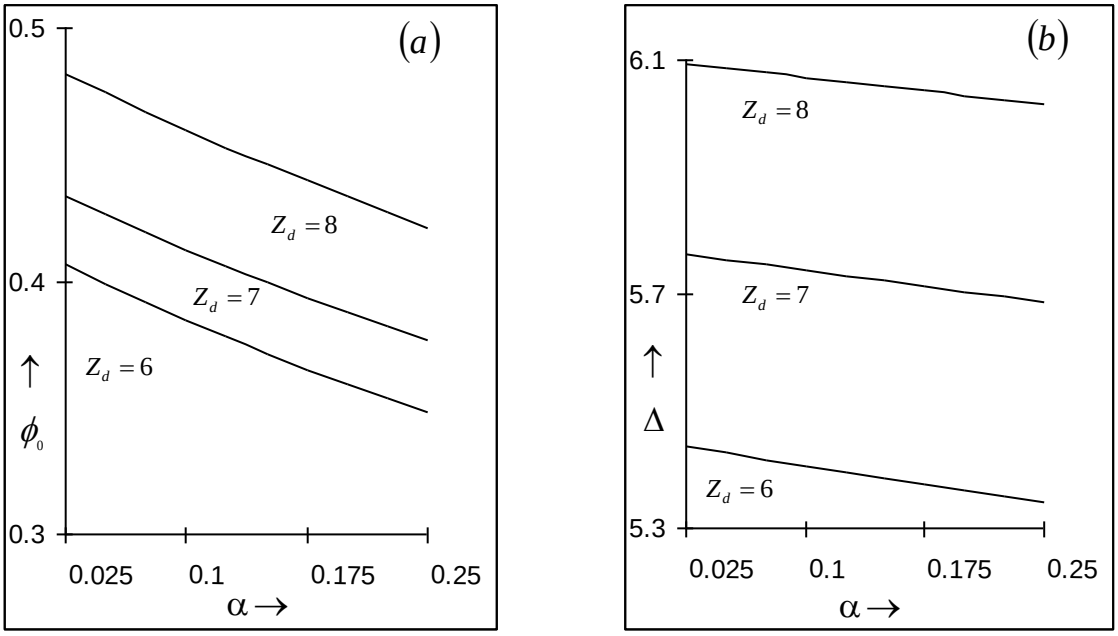

Figure 3. Amplitudes (a) and widths (b) of fast compressive dust ion-acoustic solitons versus α for fixed $V = 0.20$, $r = 0.05$ and $Q' = 10$ for different values of $Z_d = 6,7,8$.

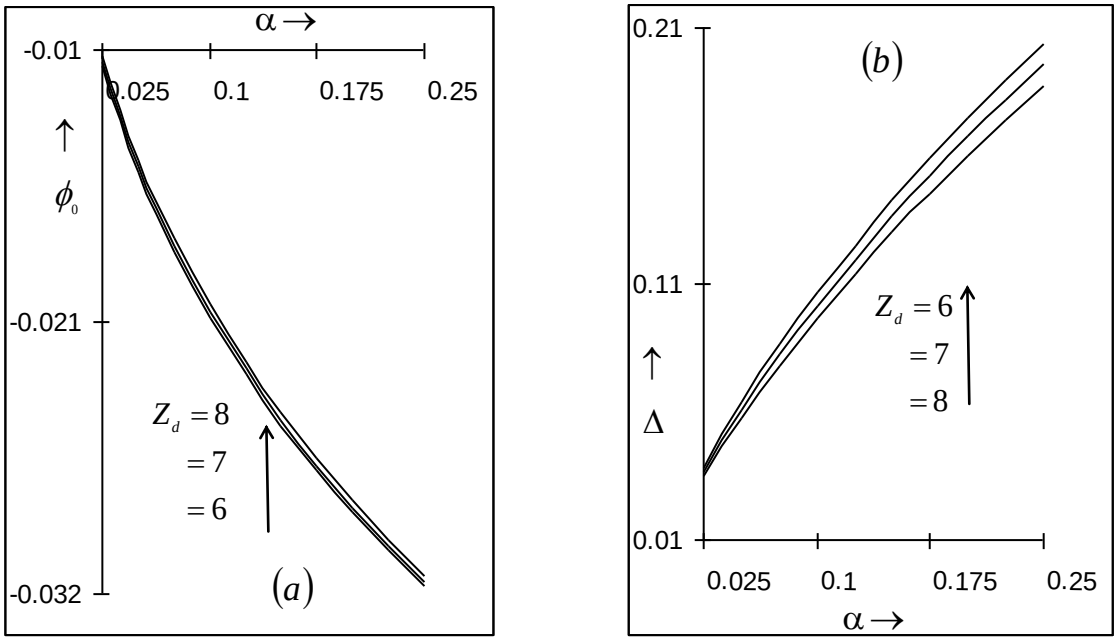

Figure 4. Amplitudes (a) and widths (b) of slow rarefactive dust ion-acoustic solitons versus α for fixed $V = 0.20$, $r = 0.05$ and $Q' = 10$ for different values of $Z_d = 6,7,8$.

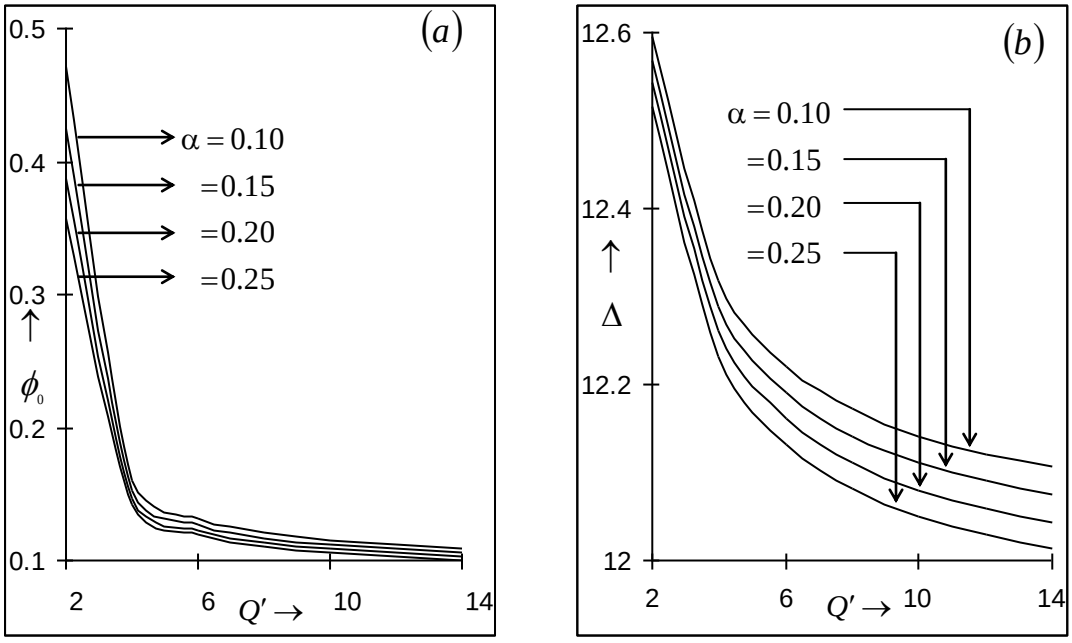

Figure 5. Amplitudes (a) and widths (b) of fast compressive dust ion-acoustic solitons versus Q' for fixed $V = 0.05$, $r = 0.05$ and $Z_d = 8$ for different values of $\alpha = 0.10, 0.15, 0.20, 0.25$.

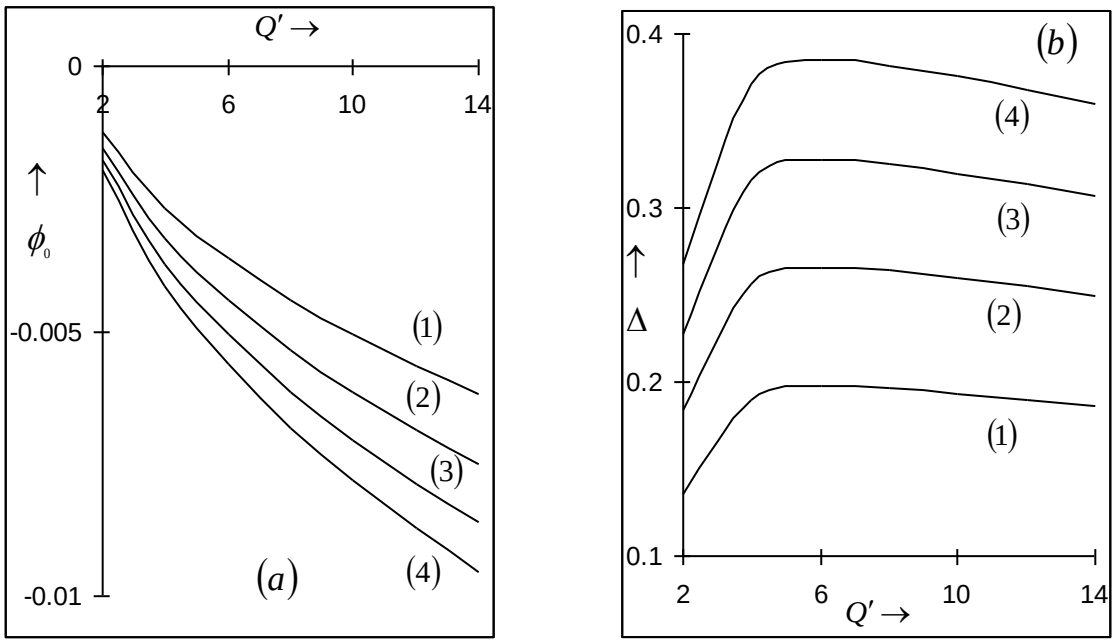

Figure 6. Amplitudes (a) and widths (b) of slow rarefactive dust ion-acoustic solitons versus Q' for fixed $V = 0.05$, $r = 0.05$ and $Z_d = 8$ for different values of $\alpha = 0.10(1), 0.15(2), 0.20(3), 0.25(4)$.

Diminutive-Tenure Vigor Harmonizing Accompanied by Escalating Breeze Vigor Stratum

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***Abstract*-** Escalating stratums of breeze vigor, which is variable, difficult to predict accurately, and escalating connected via power electronic converters, are changing how electricity grids are planned, designed, and operated. . Systems with high breeze penetration are also experiencing dramatic changes to the operating regimes of conventional resources to areas with large loads. Optimal diminutive-tenure (minutes to day ahead) vigor harmonizing for systems with high breeze penetration, which is the focus here, requires high-quality breeze forecasts and advanced scheduling methodologies. These advances from the traditional scheduling approach include: dynamic reserve targets, higher resolution scheduling periods, more frequent scheduling, and the use of stochastic optimization techniques. Here, some of the possible evolutions in optimal diminutive-tenure vigor harmonizing to better deal with breeze vigor uncertainty are investigated. The focus is mainly on managing reserves through changes in scheduling, in particular market structure (more regular and higher resolution scheduling), reserve procurement (dynamic as opposed to static), and improved operational planning (stochastic as opposed to detenureinistic). Infrastructure changes including flexible plant, increased demand side participation, more interconnection, transmission, larger harmonizing areas, and critically improved forecasting can also be significant and are dealt with in the discussion. The evolutions are tightly coupled, their impact is system-dependent and so no "best" set is identifiable but experience of system operators will be critical to future developments.

***Index Terms*-** Vigor harmonizing, market design, power system operations, reserve allocation, scheduling, unit commitment, breeze power.

I. INTRODUCTION

Escalating stratums of breeze vigor, which is variable, difficult to predict accurately, and escalating connected via power electronic converters, are changing how electricity grids are planned, designed, and operated [1]. For example, the spatially distributed, asynchronous nature of breeze vigor is driving upgrades in the transmission system, with deployment of high voltage direct current transmission (HVDC) becoming escalating popular to connect areas with good breeze resources to areas with large loads. Systems with high breeze penetration are also experiencing dramatic changes to the operating regimes of conventional generators, which must now operate more flexibly in order to accommodate variable breeze power. The displacement of conventional generation also impacts power system dynamics as the voltage support and frequency response previously supplied by these units are also displaced [2], [3]. The increased variability and uncertainty that comes with increased breeze vigor penetrations exists across multiple time scales and makes vigor harmonizing more challenging. Long tenure vigor harmonizing is complicated by the fact that the capacity value of breeze for a given system can vary significantly from year to year [4]. Optimal diminutive-tenure (minutes to day ahead) vigor harmonizing for systems with high breeze penetration, which is the focus here, requires high-quality breeze forecasts and advanced scheduling methodologies.The performance of these approaches is heavily influenced by infrastructural and portfolio changes in the power system. In particular, a more flexible portfolio, more demand side participation, increased interconnection, transmission, larger harmonizing areas, and improved breeze forecasting [5]. The remainder of the paper is arranged as follows: Section II briefly summarizes how diminutive-tenure vigor harmonizing is currently achieved through the scheduling process and how large-scale breeze vigor penetration may impact this process. Section III describes advancements to the traditional scheduling methodology that are being implemented in industry and/or proposed in the literature. Section IV discusses longer tenure infrastructural developments in the power system that will impact diminutive-tenure vigor harmonizing with escalating stratums of breeze vigor. Section V concludes

II. DIMINUTIVE-TENURE VIGOR HARMONIZING AND BREEZE VIGOR

The primary objective of optimal diminutive-tenure vigor harmonizing is to minimize costs while maintaining the balance between supply and demand at, or above, a desired reliability stratum. The problem can be studied by modeling unit commitment (UC), which shorts the commitment schedule of units, in combination with economic dispatch (ED), which detenureines the dispatch stratum of

those units in real time. UC tools commit units, typically day-ahead, based on the demand forecast and requirement for reserves and are subject to both unit constraints (e.g., minimum generation) and system constraints (e.g., transmission capacity). Reserves, with various activation times, ensure sufficient generation is available to meet forecast errors, contingencies, and variations over diminutives time resolutions than the resolution of the UC and dispatch (typically one hour down to 5 min). Therefore, committed units need to be able to manage primary, secondary, and tertiary frequency control as well as meet the ramp requirements over all time frames. As breeze vigor increases, the most impacted reserve categories are regulating reserves and load following reserves together with supplemental/replacement reserves (see [6] for discussion on reserve tenureinology)." If necessary, the system operator may recommit units intraday to allow for significant changes in demand or contingencies. Intraday markets perform a similar function where they exist. Demand follows daily, weekly, and seasonal patterns and as such demand forecasts are relatively accurate. Consequently, UC optimization approaches have traditionally been detenureinistic, with uncertainty in demand and power generation being accounted for by provision of reserves. Breeze power forecasts by contrast are relatively inaccurate, particularly in the day-ahead time-scale, as error increases strongly with time horizon. This can be seen in Fig. 1 which illustrates breeze power forecast error at various time horizons on the 2020 Dutch system. This study used an atmospheric model to generate breeze speed forecasts. In the diminutive-tenure (1–6 hours ahead), information from online breeze or breeze power measurements have to be used in addition to the numerical weather prediction model data to reach a good performance [9].. At low penetrations of breeze power, additional reserves can be scheduled to cover the additional uncertainty due to breeze power. However, as the breeze power penetration grows, it becomes escalating inefficient to rely on

TABLE I
Evolutions For Short-Term Energy Balancing With Increasing Wind Energy Penetrations

		Scheduling rate	
Explanation		Once per day	More regular scheduling
Dynamic assets procurement	A assets requirement that is based on dynamic forecast error estimates at different time horizons	Breeze power increases tertiary assets significantly, but the impact will be more limited when the forecast uncertainty is accounted for dynamically.	The combined impact of more regular UC and dynamic assets procurement would help to keep tertiary assets requirement relatively low most of the time.
Stochastic UC	Optimization of UC decisions over several scenarios for possible outcomes of breeze and demand	Improves the reliability and yields more optimal UC.	Reduces tertiary assets procurement and improves UC optimally further.
Scheduling resolution	Scheduling period is shorted eg from hourly to five minutes	Ramp within the scheduling period will be smaller, which reduces regulating assets. Scheduling accuracy will be improved.	

existing methods for reserve quantification and scheduling. Section III explores evolutions to scheduling that are being studied and in some cases applied in industry

III. Scheduling Evolutions

Table I summarizes the evolutions in the scheduling methodology that are currently being deployed and/or proposed for diminutive-tenure vigor harmonizing with high stratums of breeze vigor. Different methods, which can account for the uncertainty of breeze power output, are

Presented in the first column, while the top row categorizes these methods in tenures of when they are undertaken, i.e., once per day or more regularly. The different methods can be complimentary. For example, more regular and higher resolution commitment and dispatch can be done in place of, or as part of dynamic reserve procurement. In reality, combinations of these different strategies will be employed.

A. Scheduling Frequency

A more frequent UC, ED, and reserve procurement achieves two things: portion of the procured reserves can be released later and less expensive reserves can be used more often. Increased frequency enables the use of more up-to-date forecasts and real system information. By using updated information, the reserves carried on the system can be reduced as the operating period gets closer, as illustrated by Fig. 2. In general, repeating UC and reserve procurement in the intraday would still require that a 24-hour or longer UC is carried out to accommodate slower starting units and to ensure availability of capacity; however, these schedules should then be updated whenever new information is available. In addition, this approach allows commitment decisions for quicker starting units to be made closer to real time, delaying commitment decisions until more accurate forecasts are available. In effect, fewer units need to be scheduled for startup, which reduces the procurement costs. The rationale for more frequent scheduling was proposed by Schlueter *et al.* in has not been cited in recent literature

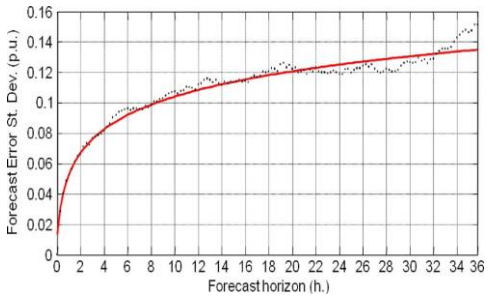

Fig. 1: Normalized standard deviation of breeze power forecast error for 12 GW installed capacity versus forecast horizon.

Tuohy *et al.* [13] show that escalating the frequency of commitment from 6 hours to 3 hours can bring tangible benefits in tenures of cost and reliability in the Irish system; however, modeling limitations prevented any benefits of decreasing the planning period further from being quantified. Similarly, [14] demonstrates benefits when moving from day-ahead to 3-hour ahead gate closure in the UC. More regular UC and ED may also cause some additional costs. Operational costs for some power plants may increase due to diminutive preparation time. This increases the importance of accurate modeling of certain unit constraints, for example, startup times of units, which may be longer than the time between commitments [15]. While research demonstrates benefits for more regular scheduling, in power exchanges the liquidity of the intraday market.

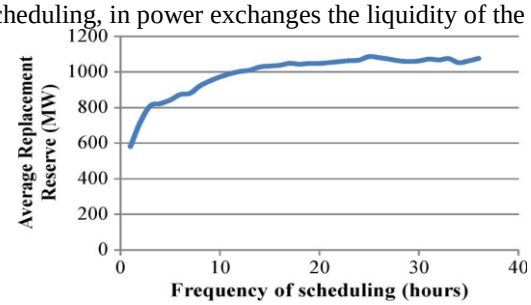

Fig. 2. Example of the trade-off between the reserve requirements and the frequency of commitment [11]. Replacement reserve is similar to the tertiary reserve defined here

has been low—at least in Europe [16]. This hinders the realization of possible benefits from more regular scheduling. One reason is that generators may expect higher profits in the harmonizing market and, therefore, do not bid intraday [17]. They may also be hindered by bilateral contracts. Hence, intraday has been an expensive method to balance forecast errors. This leads to self-harmonizing, which is suboptimal, or to the use of harmonizing markets, which is on average more expensive due to the diminutives response time. Therefore, the results from these models may overestimate the benefits of more regular scheduling

B. Dynamic Reserve Procurement

Meteorological conditions govern the probable range of breeze power output and breeze power forecast errors also tend to vary with these conditions [19]. As a simple example, if the predicted breeze power output is low, downward error cannot be large. Therefore, a static reserve stratum is not appropriate. Rather, dynamic reserve constraints which are functions of the breeze forecast error and/or the diminutive-tenure variability of breeze power output should be implemented, where the reserve requirement is based on the present stratum of breeze power output, and the expected uncertainty and diminutive-tenure variability of breeze. Taking dynamic reserve allocation as a starting point, the influence of breeze power on different operating reserve categories has been detailed in [20]. In situations with very high stratums of breeze generation where the regulating power plants are being displaced, breeze power plants need to provide the regulation. The alternative is that breeze power plants will have to be curtailed in order to accommodate the minimum generation stratums of the regulating power plant.

C. Scheduling Resolution

Power systems with a significant amount of breeze power could benefit from higher resolution scheduling (e.g., 5 min instead of one hour). This has been recently implemented in several power systems [31] and in many cases breeze power has been at least a partial motivator. Ramps within the diminutives dispatch interval will be smaller, which enables a reduction of regulation reserves acting within the scheduling interval [32].

IV. CONCLUSION

Diminutive-tenure vigor harmonizing to manage the variability and uncertainty of breeze power is evolving. Scheduling evolutions including scheduling frequency, dynamic reserve procurement, higher scheduling resolution, and stochastic UC are being proposed

and some are being implemented. Frequent scheduling takes advantage of new data closer to real-time and helps to reduce exposure to uncertainty. With more frequent scheduling, the procured reserves can be released later and less expensive reserves can be used more often. Dynamically scheduling reserves reduces the quantity of reserve procurement. Scheduling at higher resolution can reduce the need for reserve, while stochastic scheduling produces solutions which may inherently carry required reserves and are robust against forecast uncertainty. Each of these scheduling evolutions impact on how system operations and decision making can be organized to better manage reserve requirements. Infrastructure developments including increased system flexibility, increased demand side management, interconnection, transmission, larger harmonizing areas, and improved breeze forecasting will also improve diminutive-tenure vigor harmonizing performance. The scheduling evolutions discussed here are tightly coupled and complimentary to the infrastructure developments, and the overall best solution is system dependent and will be detenureined by further research and experience.

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Histopathological Alterations Induced by Lindane (Gamma-Hexachlorocyclohexane) in a Minor Carp, Aspidoparia Morar Inhabiting Jammu Waters.

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Abstract- The objective of the present study was to identify the degree of damage to the histological architecture of haemopoietic tissues viz., liver, anterior kidney and spleen in *A. morar* exposed to various sublethal concentrations (10%, 20% and 30% of LC_{50} value) of lindane for a period of 60 days. The exposed liver had distention of sinusoids, necrosis, vacuolation and degenerative process in the cellular architecture. Anterior kidneys manifested necrosis, tubular vacuolation, followed by total damage of the tissue. Spleen on the other hand showed deposition of haemosiderin pigments, necrosis, vacuolation followed by proliferation of melanomacrphage centres. Total degeneration of the splenic tissue was observed by the end of the experiment. The results of this histological analysis of various fish tissues indicate a direct correlation between insecticide exposure and histopathological disorders observed.

Index Terms- Lindane, Histological disorders, Haemopoietic tissues, Aspidoparia morar.

I. Introduction

Persistent polluting chemicals are substances which have long life mainly due to their chemical stability. DDT, dieldrin, lindane, endosulfan are well known examples of chemicals which can be found in various compartments of the aquatic environment as well as in living organisms (Bro-Ramussen, 1996). Among different such insecticides lindane is an organochloride insecticide which has been used on a wide variety of crops as well as in public health to control insect borne diseases (with fungicides) and as seed treatment (Ortiz *et al*., 2001).

Aquatic organisms quickly accumulate and store persisting organochlorine insecticide lindane. Fish are very susceptible to bioaccumulation in the tissues as they take up lindane residues from the water through the gills and skin (Gopal *et al*., 1982). Histopathology has been used as a sublethal test for evaluating toxic effects of water pollutants in fish (EIFAC, 1983). Histopathological alterations can also be utilized as tools to get a clear idea as to what extent an organism is affected at tissue or cellular level. Exposure of fish, *A. morar* to various sublethal resulted in various histological alterations in the haemopoietic tissues viz., liver, anterior kidney and spleen. Being haemopoietic, these organs are of vital importance in fish.

II. Materials and Methods

1.1 Sampling site

The fish, *A. morar* (Ham.) for present study were netted from Nikowal region of River Tawi. The area is located towards the southern part of Jammu about 8-12 kms. From main R.S. Pura Tehsil in Jammu district of Jammu and Kashmir.

1.2 Histological analysis

The test fishes were dissected open in ringer solution and their liver, anterior kidney and spleen were fixed in Bouin's fixative.

After the termination of each experiment, the test fishes after collection of their blood were dissected open in ringer saline solution and their liver, anterior kidney and spleen were fixed in Bouin's fixative. For thymus, the lower jaw and snout was removed and the rest of the head was fixed. Later the head was cut into smaller sections.

After post fixation treatment and routine dehydration and clearing, these tissues were embedded in histowax of 54-56ºC. 5-7 µm thick section of these were cut on microtome and stained using haematoxylin eosin stain.

III. Results

Liver histopathology: The normal histology of liver of fish, *A. morar* showed the presence of a) **hepatocytes** (Fig. 1) cells involved in the synthesis of proteins, cholesterol, bile salts and phospholipids. b) **sinusoids** (Fig. 1) (the blood vessels similar to capillaries but with discontinuous epithelium) and c) **melanomacrophage centers (MMCs)** (Fig. 1) (pigment containing cells and are normally located in the stroma of haemopoietic tissue of the liver).

During present studies, *A. morar* after exposure to sublethal concentrations of lindane exhibited many alterations in liver tissue, the chief being i) distention or widening of blood sinusoids observed after 15^{th} day of the experiment (Fig. 2), ii) necrosis and vacuolation (during 35^{th} day of the experiment) (Fig. 3), iii) degenerative process was observed in cellular architecture during 50^{th} day of the experiment (Fig. 4) iii) total degeneration of the tissue by the end of 60^{th} day of experimental period (Fig. 5).

Kidney histopathology: Microscopic examination of head kidney of control group showed the presence of a) **haemopoietic tissue** (blood forming tissue) (Fig. 6) and

b) **renal tubules** (excretory in function) (Fig. 6).

Compared to control group, cellular structure of kidney exposed to lindane manifested changes viz., i) necrosis of kidney tissue which was observed after 20th day of the experiment (Fig. 7), ii) tubular vacuolation by 30th day of the experiment (Fig. 8), iii) Signs of tubular degeneration observed after 40 days of the experiment (Fig. 9) iv) total damage of the cellular make up of kidney structure by the end of experimental period of 60 days (Fig. 10).

Spleen histopathology: Microscopic examination of spleen from the control fish showed the presence of i) red pulp (comprising of erythrocytes) (Fig. 11) ii) white pulp (comprising of leucocytes) (Fig. 11) and iii) melanomacrophage centres (Fig. 11) (which are the important component of reticuloendothelial system and act as main repository for iron containing compounds (Agius, 1979).

Upon treatment with lindane presently normal histology of *A. morar* was observed to get disrupted resulting in various histopathological effects which include i) deposition of haemosiderin pigments observed after 15th day of the experiment (Fig. 12), ii) necrosis and vacuolation which became more marked during 35th day of the experimental period (Fig. 13), iii) proliferation of MMCs by 45 days of the experiment (Fig. 14) and iv) inflammation and total degeneration of the splenic tissue by the end of the experiment (Fig. 15).

IV. Discussion

The histopathological alterations in the liver tissue under the influence of lindane toxicity resulted in distention or widening of blood sinusoids which evidently means that haemopoiesis being one of the synthetic metabolic processes get hampered by insecticide intoxication.

Necrosis of the hepatocytes which implies cell death, is an advanced and usually irreversible stage of degeneration characterised by dead hepatocytes (Pal, 2006). Distention of blood sinusoids evidently means that haemopoiesis being one of the synthetic metabolic processes possibly get hampered by the insecticide intoxication. In this context findings of Gupta (2008),Vinodhini and Narayanan (2009), Pathan *et al*., (2010) who also reported severe necrosis of hepatocytes and widening of blood sinusoid to hamper the haematopoietic machinery of the fish lends a strong support for the present viewpoint.

Random distribution of vacuoles in the hepatocytes of affected fishes as observed presently indicate an imbalance between the rate of synthesis of substance in the parenchymal cells and their release into the circulation. In tune to present findings Kabir and Begum (1978), Shastry and Sharma (1979), Gingerich (1982), Vinodhini and Narayanan (2009), Oliva *et al*. (2010) and Velmurugan *et al*. (2007) too, observed vacuolation of the hepatocytes in their fishes exposed to different xenobiotics. This, all of them stressed, certainly affect the release of substances in the general circulation.

Since liver is the site of various metabolic activities its the total degeneration which was observed after 60 days of the experimental duration completely disrupt the metabolic processes of the liver (including haemopoiesis) and hence result in the malfunctioning of the liver tissue.

The various histopathological changes observed in the kidney tissue are suggestive that kidney are malfunctioning. This simply is indicative of the fact that haemopoietic machinery possibly gets affected under the influence of lindane toxicity and hence may result in inhibition of further release of erythrocytes in the general circulation.

In this context, the observations of Bucher and Hofer (1993) and Ranzani-Paiva *et al*. (1997) can be quoted who too behold that histopathological alterations in kidneys are the most plausible causatives for decline in haematological parameters on exposure to various xenobiotics. As an important organ of immunity response elaboration (Zapata and Cooper, 1990), pathological changes induced in kidney tissue by lindane toxicity present author states, by affecting defense system, can definitely disturb homeostasis and health of fishes.

Lindane induced deposition of haemosiderin pigments in the splenic tissue which is one of the breakdown products of haemoglobin from senescent erythrocytes as reported by Zapata and Cooper (1990). According to Hibiya (1982) deposition of haemosiderin causes disease haemosiderosis when because of increase in the rate of destruction of erythrocytes in spleen there is resultant decline in the number of mature erythrocytes in the circulating blood. Deriving support from these findings deposition of haemosiderin pigments observed during present studies, it appears may represent the diseased condition called haemosiderosis.

Vacuolation of the splenic cells is indication of disruption of the synthetic machinery of this organ and definitely the normal mechanism of release of substances form the tissue into the general circulation.

Proliferation of MMCs in the spleen as observed presently finds association with either normal ageing or to prolonged starvation or infectious diseases (Reviewed by Couillard *et al*., 1999). Lindane intoxication, it appear, induce a sort of stress condition which may then cause the proliferation of MMCs. MMCs proliferation are known to affect the erythrocyte synthesis and can block haemopoiesis or erythropoeisis in fish (Gupta, 2008). All the above alterations lead to the total damage of the tissue thereby hampering its normal functioning.

V. Conclusion

All the histopathological observations indicated that exposure to sublethal concentrations of lindane caused structural effects in the liver, anterior kidney and splenic tissues of *A. morar*. Liver, anterior kidney and spleen histopathological alterations, such as those observed in this study may result in severe physiological problems ultimately leading to the death of the fish. In conclusion, the findings of the present histological investigations demonstrate a direct correlation between pesticide exposure and histopathological disorders observed in several tissues.

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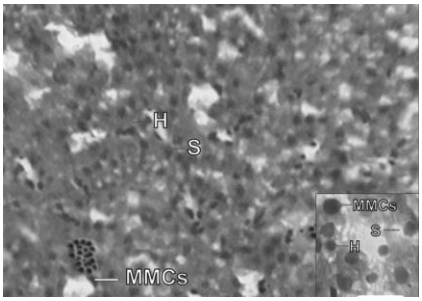

Fig. 1 Microphotograph of liver from control fish showing Hepatocytes (H), Sinusoids (S) and melanomacrophage centres (H&E×1000)

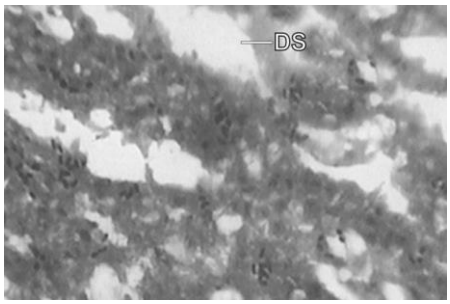

Fig. 2 Microphotograph of liver tissue from lindane reated fish showing distended blood sinusoids (DS) after 15th day of the experiment (H&E×1000)

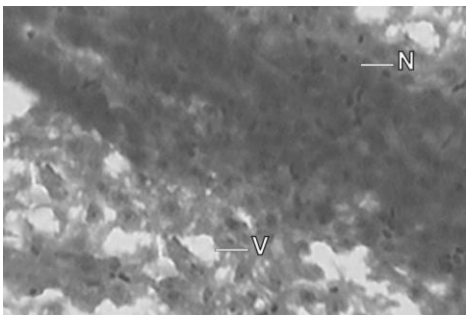

Fig. 3 Microphotograph of liver tissue from lindane treated fish showing Necrosis (N) and Vacuolation (V) after 3th day of the experiment (H&E×1000)

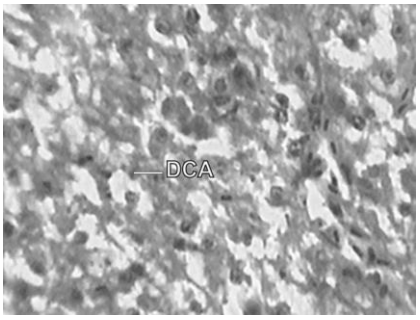

Fig. 4 Microphotograph of liver tissue from lindane treated fish showing Degenerative cellular architecture during 50th day of the experiment (H&E×1000)

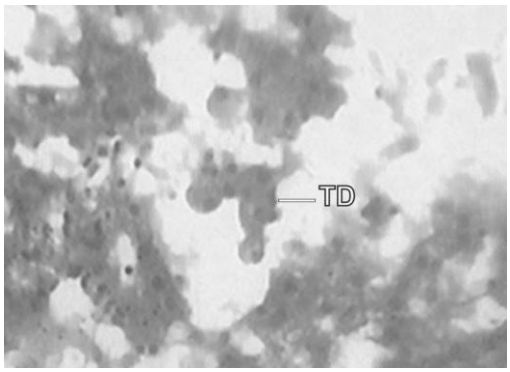

Fig. 5 Microphotograph of liver tissue from lindane treated fish showing Total degeneration of the tissue after 60th day of the experiment (H&E×1000)

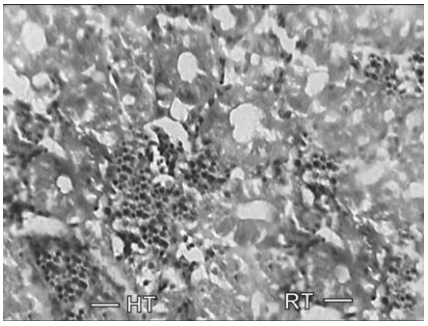

Fig. 6 Microphotograph of kidney tissue from control showing haemopoietic tissue (HT) and renal tubules (RT) (H&E×1000)

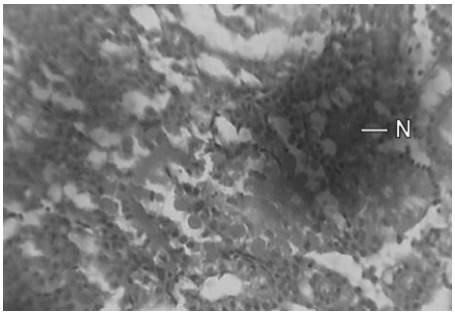

Fig. 7 Microphotograph of kidney tissue from lindane treated fish showing Necrosis (N) after 20th day of the experiment (H&E×1000)

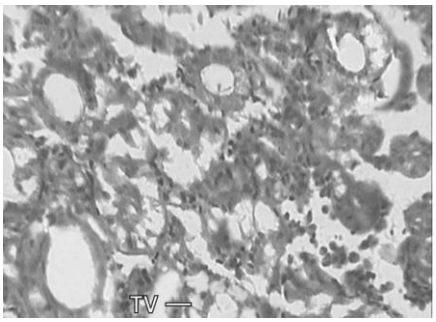

Fig. 8 Microphotograph of kidney tissue from lindane treated fish showing Tubular vacuolation (TV) after 30th day of the experiment (H&E×1000)

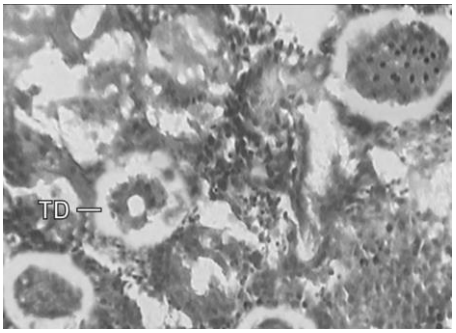

Fig. 9 Microphotograph of kidney tissue from lindane treated fish showing Tubular degeneration (TD) after 40th day of the experiment (H&E×1000)

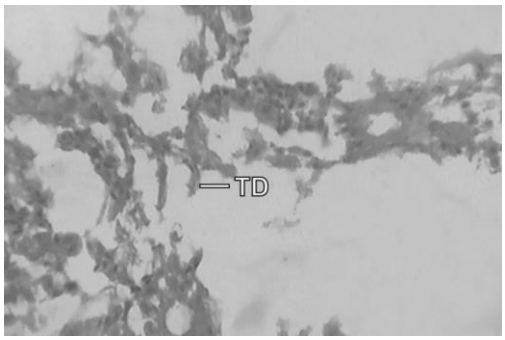

Fig. 10 Microphotograph of kidney tissue from lindane treated fish showing Total degeneration (TD) of the tissue after 60th day of the experiment (H&E×1000)

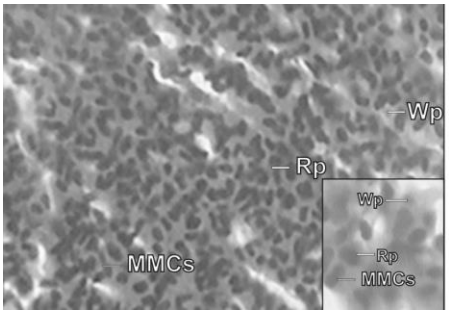

Fig. 11 Microphotograph of splenic tissue from control showing Red pulp (Rp), white pulp (Wp) and melanomacrophage centres (MMCs) (H&E×1000)

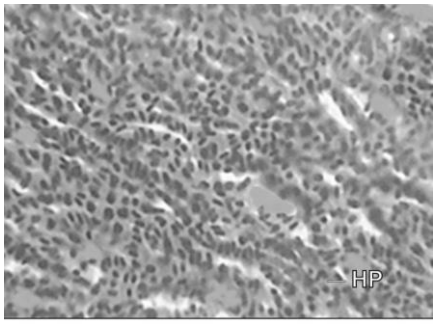

Fig. 12 Microphotograph of splenic tissue from lindane treated fish showing deposition of Haemosiderin pigments (HP) after 15th day of the experiment (H&E×1000)

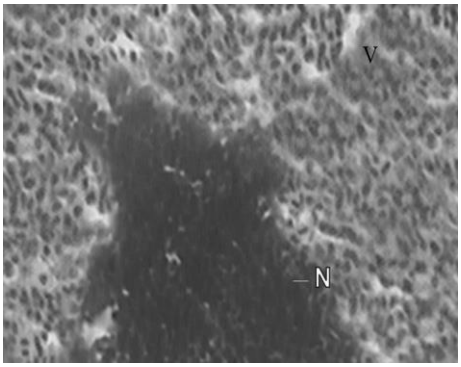

Fig. 13 Microphotograph of splenic tissue from lindane treated fish showing Necrosis (N) and Vacuolation after 35th day of the experiment (H&E×1000)

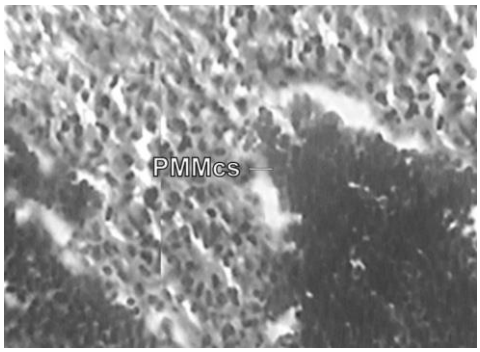

Fig. 14 Microphotograph of splenic tissue from lindane treated fish showing Proliferation of MMCs after 45th day of the experiment (H&E×1000)

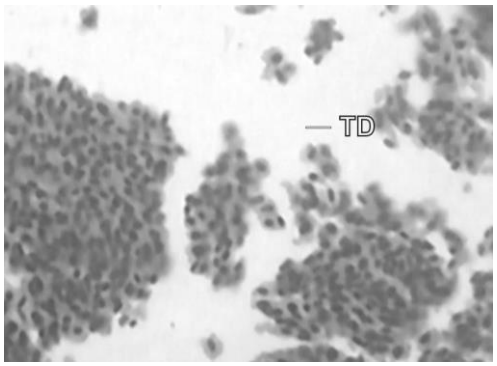

Fig. 15 Microphotograph of splenic tissue from lindane treated fish showing Total degeneration of the tissue after 60th day of the experiment (H&E×1000)

Code Bloat Problem in Genetic Programming

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***Abstract*-** The concept of "bloat" in Genetic Programming is a well-established phenomenon characterized by variable-length genomes gradually increasing in size during evolution [1]. Bloat hampers the efficiency and ability of genetic programming for solving problems. A range of explanations have been proposed for the problem of bloat, including destructive crossover and mutation operators, selection pressure and individual representation. Different methods to avoid bloat and to control bloat have been proposed by researchers. This paper proposes a theoretical analysis of code bloating problem and the discussion on the work already done by various authors to handle bloat in genetic programming.

Index Terms- Bloat, Double Tournament, Elitism, Genetic Programming, Spatial Structure.

I. INTRODUCTION

During the evolution of solutions using Genetic Programming (GP) there is generally an increase in average tree size and depth without a corresponding increase in fitness—a phenomenon commonly referred to as bloat [1]. Bloat is a well-known phenomenon in Genetic Programming. An individual program in GP could be of any size. Such flexibility in representation provides more freedom in searching solutions, but at the same time it causes the bloat problem, individuals growing unnecessarily large. Apparently big individuals are computationally more expensive to evaluate during evolution. If they are final solutions, then their execution time in applications would increase accordingly. That would not be desirable for situations where speed is a requirement such as in real-time systems. Furthermore bloat makes these evolved programs even more difficult to comprehend [17].

Code bloating presents a serious problem in scaling GP to larger and more difficult problems. First, bloat consumes computing resources, making the search process slower and slower, and eventually forcing it to stop when all available resources have been exhausted. Second, bloated candidate solutions are often more difficult to modify in meaningful ways, hampering the ability of GP to breed and discover better solutions. Third, bloating can slow the grade-assessment process. In a very real sense, bloating makes genetic programming a race against time, to find the best solution possible before bloat puts an effective stop to the search [29].

Three main methods for controlling bloat are commonly proposed: set an upper bound to the complexity of individuals in the population; introduce an explicit fitness penalty (parsimony measure) that biases against larger individuals [10]; and apply genetic operators designed to target redundant code or the bias against offspring size increases [12]. Many authors have shown a number of theoretical advances in understanding bloat [3], [9]–[12]. Poli [9] reduced bloat by a stochastic approach to setting the fitness of above average-sized individuals to zero. Stringer and Wu [10], [11] showed that a shrinking effect on genome length occurred for a chunking GA once the population had essentially converged and selection had become random. Skinner et al. [12] provided a theoretical argument for this observed tendency of variable length genomes to shrink when selection is not considered (i.e., under the process of genetic drift). The paper implied that the presence of a lower absorbing boundary (a genome size that once reached cannot be reduced further), combined with no upper bound, results in a reduction of the average size of a population under drift. Although this theoretical model contributes to an understanding of individual size dynamics, the concept has not at present formed the basis for new bloat control methods. Related to these results other research [13] has shown that for large, discrete programs, fitness convergence of the population is possible and has been used to explain sub quadratic growth of program size.

In this paper, we have analyzed and presented the problem of code bloat in GP, its types and variants and the effective measures taken by various authors to prevent or control bloat. The rest of the paper is organized as follows: section II contains the description of the code bloat problem, section III discuss the work done by various authors to prevent bloat and section IV presents different methods present in the literature to avoid code bloating in GP.

II. CODE BLOAT IN GENETIC PROGRAMMING

"In a very real sense, bloating makes genetic programming a race against time, to find the best solution possible before bloat puts an effective stop to the search". While bloat is well-defined and can be identified, there are currently no consensual explanations on why it occurs. Authors have presented different explanations regarding bloat. Three popular theories can be found in the literature to explain it [36]:

– The introns theory states that bloat acts as a protective mechanism in order to avoid the destructive effects of operator's once relevant solutions have been found. Introns are pieces of code that have no influence on the fitness: either sub-programs that are never executed, or sub-programs which have no effect;

– The fitness causes bloat theory relies on the assumption that there is a greater probability to find a bigger program with the same behavior (i.e. semantically equivalent) than to find a shorter one. Thus, once a good solution is found, programs naturally

tend to grow because of fitness pressure. This theory states that code bloat is operator-independent and may happen for any variable length representation-based algorithm. As a consequence, code bloat is not to be limited to population-based stochastic algorithm (such as GP), but may be extended to many algorithms using variable length representation.

– The removal bias theory states that removing longer sub-programs is more dangerous to do than removing shorter ones (because of possible destructive consequence), so there is a natural bias that benefit to the preservation of longer programs. While it is now considered that each of these theories somewhat capture part of the problem there has not been any definitive global explanation of the bloat phenomenon. At the same time, no definitive practical solution has been proposed that would avoid the drawbacks of bloat (i.e. increasing evaluation time of largetrees) while maintaining the good performances of GP on difficult problems. Some common solutions rely either on specific operators e.g. size-fair crossover or different fair mutation on some parsimony-based penalization of the fitness or on abrupt limitation of the program size such as the one originally used by Koza. Also, some multi-objective approaches have been proposed. Some other more particular solutions have been proposed but are not widely used yet.

Authors have distinguished bloat into two main types: structural bloat and functional bloat [32]:

A. Structural Bloat

The structural bloat is defined as the code bloat that necessarily takes place when no optimal solution can be approximated by a set of programs with bounded length. In such a situation, optimal solutions of increasing accuracy will also exhibit an increasing complexity (larger programs), as larger and larger code will be generated in order to better approximate the target function.

B. Functional Bloat

Another form of bloat is the functional bloat, which takes place when program length keeps on growing even though an optimal solution (of known complexity) does lie in the search space. Most of the works cited earlier are in fact concerned with functional bloat which is the most surprising, and the most disappointing kind of bloat. There are various levels of functional bloat: cases where the length of programs found by GP runs to infinity as the number of test cases run to infinity whereas a bounded-length solution exists, and also cases where large programs are found with high probability by GP where as a small program is optimal.

III. Previous Work Done

Liu, Cai, Ying, and Le in [29] stated that GP using a size or depth limit (LGP) is a common approach to battle bloat, but LGP is not ideal in size control and searching efficiency. In their paper, besides extended the concept of bloated individual in LGP, and the concept of Candidate Crossover Points Set is presented. A new variants of LGP, named RLGP, which adds some restrictions in genetic operations (crossover, swap, and mutation), is proposed. RLGP introduces Candidate Crossover Points Set (CCPS) into crossover operations. Finally, in even 3, 4, and 5-parity problem, strongly positive results are reported regarding both size control and searching efficiency.

Thomas Helmuth, Lee Spector and Brian Martin [30], introduced a new node selection method that selects nodes based on a tournament, from which the largest participating sub-tree is selected. Size-based tournaments differentiate between internal nodes of different sizes, whereas Koza 90/10 treats all internal nodes equally. This method of size-based tournaments improves performance on three standard test problems with no increases in code bloat as compared to unbiased and Koza 90/10 selection methods.

Whigham [1], has presented an implicit model of bloat control based on a spatially structured population with local elitism; referred to as SS+E. Regular spatial structures (such as a ring or torus) maintain diversity and slow bloat by effectively reducing the population size. In addition, elitism reduces the growth of introns, especially once the population has largely converged and cannot easily find fitness improvements. Previous panmictic models with elitism found that this resulted in crossover largely becoming a copying operator, resulting in convergence to non optimal solutions. Most bloat control methods tradeoff controlling size and fitness, however SS+E appear to balance this tradeoff without compromising overall fitness.

Langdon and Poli [2] have described a way to control bloat using a fix size or depth limit (LGP) in which the bloat is controlled by applying the limit to the allowed individual size or depth simply. Individuals exceeding the limits are removed from the population. Because individual size or depth is calculated easily during evaluation, this approach only requires relatively little additional computation.

Stringer [4], has handled bloat by explicitly setting an upper bound on the depth of evolved trees or by incorporating a parsimony pressure that adjusts the fitness of individuals by a tradeoff between performance and size.

Bleuler, Brack, Thiele, and Zitzler [15] proposed a nonparametric method, Double Tournament, this method is similar to a multi objective approach to bloat, however the objectives of fitness and size are treated separately. Hence, there are two tournaments: one based on parsimony, which produces an initial set of winners, and a subsequent tournament that selects a subset of those winners based on fitness.

Sara Silva and Ernesto Costa [31] presented two important variations on a recently successful bloat control technique, Dynamic Maximum Tree Depth, intended at further improving the results and extending the idea to non tree-based GP. Dynamic Maximum Tree Depth introduces a dynamic limit on the depth of the trees allowed into the population, initially set with a low value but increased whenever needed to accommodate a new best individual that would otherwise break the limit. The first variation to this idea is the Heavy Dynamic Limit that, unlike the original one, may fall again to a lower value after it has been raised, in case the new best individual allows it. The second variation is the Dynamic Size Limit, where size is the number of

nodes, instead and regardless of depth. The variations were tested in two problems, Symbolic Regression and Parity, and the results show that the heavy limit performs generally better than the original technique, but the dynamic limit on size fails in the Parity problem. The possible reasons for success and failure are discussed.

Soule and Foster [32], introduced the concept of removal bias, arguing that neutral branches of code (i.e., introns) are likely to be small, however their replacement with crossover does not have this restriction. Hence, the children produced from neutral crossover events are likely on average to increase in size. In this paper they described that the initial population are likely to be small but introns grows on increasing in size after crossover operation and the size of the individual can be very large so to restrict the size of the individual two forms of nondestructive crossover (NDC) were presented: a child would replace a parent if it was at least as fit as the parent, or in the strict version the child had to exceed the parent's fitness. These methods were tested with a maze navigation problem and a parity problem, with both examples showing a reduction in bloat and an improvement in convergence to fit solutions. However, since crossover is often destructive, strict elitism can reduce the effectiveness of crossover as a search mechanism, especially once the population has begun to converge.

A review and comparison of the most common methods is given in Luke and Panait [18], while discussion on the causes of bloat maybe found in Soule and Heckendorn and the Field Guide to Genetic Programming [25].These previous methods generally take into account individual size to control bloat. However, a number of researchers have also considered methods that do not explicitly consider the size of individuals and therefore bloat reduction results as a side effect.

Over the years a range of methods have been introduced to manage bloat [1]-[33] There are two major approaches to dealing with GP tree bloat. First, by improving breeding, selection, and tree-generation, GP can be made to search more efficiently to find better individuals before bloat sets in. Second, various techniques can help GP put off bloat as long as possible, lengthening the search interval [4]. Following things are considered under these two approaches while handling bloat:

- treating fitness and size as a multiobjective optimization;
- using disassortative mating based on two species (one selected on fitness, the other on fitness and size);
- explicitly reducing the fitness of above average-sized individuals (referred to as the Tarpeian method);
- eliminating programs where the parent and child fitness are similar using a modified tournament selection operator that uses either fitness, depth or an ordered combination of both for selection;
- placing a form of resource constraint on the population so that larger individuals are discouraged;
- using a waiting room for individual entry into a population, with time proportional to size;
- biasing selection for removal from a population based on size;
- explicitly simplifying individuals after each generation;
- dynamically extending an initially low maximum tree depth only when a child is produced that is fitter than the best individual and larger then this size limit;
- viewing size as a resource constraint, that can only be extended by fitness improvements; and
- applying specific genetic operators to reduce the size of large individuals or maintain the size of children to parents.

IV. Methods to Avoid Bloat

Different methods have been proposed in the literature to avoid bloat. Some of them are described as follows:

A. Using a fix size or depth limit (LGP)

The most common way to avoid bloat is to limit the size (number of nodes) and depth (height of the tree) of the individual [2]. The limitations on size and depth can be considered during initialization of population by using suitable algorithm for tree generation. Individuals exceeding the limits are removed from the population. Because individual size or depth is calculated easily during evaluation, this approach only requires relatively little additional computation. This method efficiently restricted individual from bloating.

B. Parsimony Pressure

Parsimony Pressure is not only another popular bloat control technique in GP, but also has been used in a wide variety of arbitrary-length representations tended to get out of control. It is the second most common method for controlling bloat. This is done by adding a tree size penalty as an additional criterion in the individual's fitness assessment. Individual having larger tree size is allotted lower fitness. Many researchers till date have used parsimony pressure as a method to avoid bloat while using GP for problem solving. Such usage is divided into two broad categories: parametric parsimony pressure, where size is directly defined as a numerical factor in fitness, and pareto parsimony pressure, where size is considered as a separate objective in a pareto-optimization procedure.

C. Local Elitism Method

While performing crossover or mutation operation, new individual or child are generated. The child can only replace the parent in new generation if its fitness is better than or equal to the parent fitness otherwise it will be retained with some probability [34]. In this way better individuals can be taken to the new generation.

D. Using modified genetic operators

The genetic operators, crossover and mutation can be modified to avoid the problem of bloat. For example in FEDS crossover (Fitness, Elitism, Depth limit & Size), the concept of fitness, elitism, depth limit and tree size for generating the next generation individuals through crossover operation is proposed to handle the problem of code bloating in GP [35]. In [37], authors

have used point mutation operation to avoid increase in tree size during mutation.

V. CONCLUSION

In this paper, we have done theoretical study of an important issue in Genetic Programming known as code bloating. Bloating hampers the performance of program structures designed using GP. A lot of research has been done to find the actual cause of bloat and methods to avoid and reduce the problem of bloat. For this various strategies including modification in genetic operators (crossover and mutation) have been presented.

Our contribution in the paper can be summarized as follows:

1) We have presented the theoretical concept of code bloating in GP.
2) We have identified different types of bloat and their root causes.
3) We have discussed about different methods to avoid and control bloat in GP.

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An Overview of Advance Microcontroller Bus Architecture Relate on APB Bridge

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Abstract — The Advanced Microcontroller Bus Architeture (AMBA) is a widely used interconnection standard for System on Chip (SoC) design. An AMBA-based microcontroller typically consists of a high-performance system backbone bus (AMBA AHB or AMBA ASB), able to sustain the external memory bandwidth, on which the CPU, on-chip memory and other Direct Memory Access (DMA) devices reside. This bus provides a high-bandwidth interface between the elements that are involved in the majority of transfers. This paper present three distinct buses and their comparison. By considering merits of APB , AMBA can be design by using HDL.

Index Terms — AMBA, AHB, ASB, APB, Difference of buses

I. INTRODUCTION

Today in the era of modern technology micro-electronics play a very vital role in every aspects of life of an individual, increasing use for micro-electronics equipment increases the demand for manufacturing its components and its availability [4].Embedded system designers have a choice of using a share or point-to-point bus in their designs [2]. Typically, an embedded design will have a general purpose processor, cache, SDRAM, DMA port, and Bridge port to a slower I/O bus, such as the Advanced Micro controller Bus Architecture (AMBA) Advanced Peripheral Bus (APB). In addition, there might be a port to a DSP processor, or hardware accelerator, common with the increased use of video in many applications. As chip-level device geometries become smaller and smaller, more and more functionality can be added without the concomitant [2] increase in power and cost per die as seen in prior generations.

The Advanced Microcontroller Bus Architecture (AMBA) was introduced by ARM Ltd 1996 and is widely used as the on-chip bus in system on chip (SoC) designs. AMBA is a registered trademark of ARM Ltd. The first AMBA buses were Advanced System Bus (ASB) and Advanced Peripheral Bus (APB). In its 2nd version, AMBA 2, ARM added AMBA High-performance Bus (AHB) that is a single clock-edge protocol. In 2003, ARM introduced [2,4] the 3rd generation, AMBA 3, including AXI to reach even higher performance interconnect and the Advanced Trace Bus (ATB) as part of the Core Sight on-chip debug and trace solution.
These protocols are today the de-facto standard for 32-bit embedded processors because they are well documented and can be used without royalties AMBA's target is to help designer of embedded system to meet challenges like design for low power consumption, to facilitate the right-first-time development of Embedded Microcontroller Products with one or more CPUs or signal processors, to be technology-independent and to encourage modular system [4]. To minimize the silicon infrastructure required supporting efficient on-chip and off-chip communication for both operation and manufacturing test [1].

This paper discusses the architecture of AMBA in the section II, section III deals with the various bus methods and their comparison is discuss in section IV. Finally section V and VI gives proposed work and conclude the paper.

II. ARCHITECTURE OF AMBA BASED SIMPLE MICROCONTROLLER

An AMBA-based microcontroller typically consists of a high-performance system backbone bus (AMBA AHB or AMBA ASB), able to sustain the external memory bandwidth, on which the CPU, on-chip memory and other Direct Memory Access (DMA) devices reside. This bus provides a high-bandwidth interface between the elements that are involved in the majority of transfers[3]. Fig1 shows AMBA based Simple Microcontroller. Also located on the high performance bus is a bridge to the lower bandwidth APB, where most of the peripheral devices in the system are located. AMBA APB provides the basic peripheral macro cell communications infrastructure as a secondary bus from the higher bandwidth pipelined main system bus [1]. Such peripherals typically:

(i) Have interfaces which are memory-mapped registers
(ii) Have no high-bandwidth interfaces
(iii) Are accessed under programmed control.

The AMBA specification [2] has become a de-facto standard for the semiconductor industry, it has been adopted by more than 95% of ARM's partners and a number of IP providers. The specification has been successfully implemented in several ASIC designs. Since the AMBA interface is processor and technology independent, it enhances the reusability of peripheral and system components across a wide range of applications.

The AMBA specification [1,3] has been derived to satisfy the following four key requirements.

(i) To facilitate the right-first-time development of Embedded Microcontroller Products with one or more CPUs or signal processors.
(ii) To be technology-independent and ensure that highly reusable peripheral and system macro cells can be migrated across a diverse range of IC processes and be appropriate for full-custom, standard cell and gate array technologies.
(iii) To encourage modular system design to improve processor independence, providing a development road-map for advanced cached CPU cores and the development of peripheral libraries.
(iv)To minimize the silicon infrastructure required supporting efficient on-chip and off-chip communication for both operation and manufacturing test.

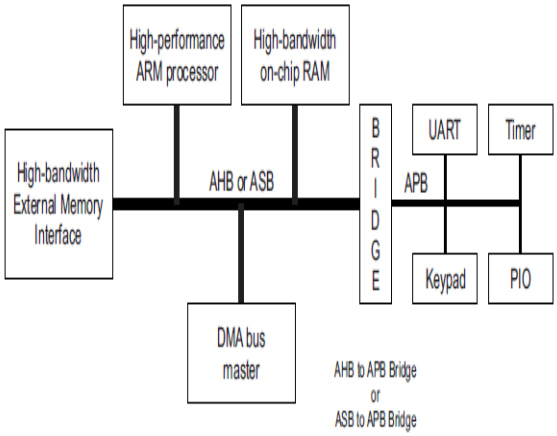

Fig. 1. AMBA based Simple Microcontroller

III. DIFFERENT AMBA BUSES

The Advanced Microcontroller Bus Architecture (AMBA) is ARM's no-cost, open specification[1,3,4], which defines an on-chip communications standard for designing high performance Embedded Microcontrollers. Three distinct buses are defined within the AMBA specification:

(A) The Advanced High-performance Bus (AHB)
(B) The Advanced System Bus (ASB)
(C) The Advanced Peripheral Bus (APB).

(A) The Advanced High-performance Bus (AHB): AHB is a new generation of AMBA bus which is intended to address the requirements of high-performance synthesizable designs. It is a high-performance system bus that supports multiple bus masters and provides high-bandwidth operation. Bridging between this higher level of bus and the current ASB/APB can be done efficiently to ensure that any existing designs can be easily integrated. An AMBA AHB design may contain one or more bus masters, typically a system would contain at least the processor and test interface. However, it would also be common for a Direct Memory Access (DMA) or Digital Signal Processor (DSP) to be included as bus masters.

The external memory interface, APB bridge and any internal memory are the most common AHB slaves. Any other peripheral in the system could also be included as an AHB slave. However, low-bandwidth peripherals typically reside on the APB.

(B) The Advanced System Bus (ASB): ASB is the first generation of AMBA system bus. A typical AMBA ASB system may contain one or more bus masters. For example, at least the processor and test interface. However, it would also be common for a Direct Memory Access (DMA) or Digital Signal Processor (DSP) to be included as bus masters.

The external memory interface, APB bridge and any internal memory are the most common ASB slaves. Any other peripheral in the system could also be included as an ASB slave. However, low-bandwidth peripherals typically reside on the APB.

(C) The Advanced Peripheral Bus (APB): The APB is part of the AMBA hierarchy of buses and is optimized for minimal power consumption and reduced interface complexity. The AMBA APB appears as a local secondary bus that is encapsulated as a single AHB or ASB slave device. APB provides a low-power extension to the system bus which builds on AHB or ASB signals directly. The APB bridge appears as a slave module which handles the bus handshake and control signal retiming on behalf of the local peripheral bus. By defining the APB interface from the starting point of the system bus, the benefits of the system diagnostics and test methodology can be exploited.

The AMBA APB should be used to interface to any peripherals which are low bandwidth and do not require the high performance of a pipelined bus interface. The latest revision of the APB[6,7] is specified so that all signal transitions are only related to the rising edge of the clock. This improvement ensures the APB peripherals can be integrated easily into any design flow. These changes to the APB also make it simpler to interface it to the new AHB. An AMBA APB implementation typically contains a single APB bridge which is required to convert AHB or ASB transfers into a suitable format for the slave devices on the APB. The bridge provides latching of all address, data and control signals, as well as providing a second level of decoding to generate slave select signals for the APB peripherals.

IV. COMPARISON OF BUSES

Following table shows the comparison among the AMBA buses:

AHB	ASB	APB
High performance	High performance	Low power
Pipelined operation	Pipelined operation	Latched address and control
Multiple bus masters	Multiple bus masters	Simple interface
It consists of master, slave, arbiter decoder	It consists of master, slave, arbiter decoder	It consist of APB bridge and slave

Table 1: Comparison of buses

V. PROPOSED WORK

As discuss above, APB is good choice for implementing Advanced Microcontroller Bus Architecture by using HDL.

VI. Conclusion

Implementation of proposed work i.e. AMBA APB prvides the basic peripheral macro cell communications infrastructure as a secondary bus from the higher bandwidth pipelined main system bus .Such peripherals typically:

(i) Have interfaces which are memory-mapped registers
(ii) Have no high-bandwidth interfaces
(iii) Are accessed under programmed control.

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Influence of Injection Timing on Performance and Emission Characteristics of Simarouba Biodiesel Engine

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Abstract- Efforts are being made throughout the World to reduce the consumption of liquid petroleum fuels wherever is possible. Biodiesel is recently gaining prominence as a substitute for petroleum based diesel mainly due to environmental considerations and depletion of vital resources like petroleum and coal. According to Indian scenario, the demand for petroleum diesel is increasing day by day hence there is a need to find out an appropriate solution. This study investigates influence of injection timing of 20% blend simarouba biodiesel on performance and emission characteristics. The effect of varying injection timing was evaluated in terms of thermal efficiency, specific fuel consumption, carbon monoxide, hydrocarbons and oxides of nitrogen. By retarding injection timing brake thermal efficiency can be improved of S20.

Index Terms- injection timing, simarouba, biodiesel

INTRODUCTION

The recent survey on the world energy consumption highlights that a major portion of the total energy consumed is derived from the combustion of fossil fuels. Unfortunately, the reserves of fossil fuels, specially the liquid fuels are not unlimited and may exhaust, if not utilized economically, within few decades. Recently, there has been a growing concern about the increasing air pollution caused by the combustion of petro diesel. In addition, depleting resources of conventional fuels has caused an increase in its price. Biodiesel is a renewable fuel which is produced from vegetable oil or animal fat through a chemical process and can be used as either direct substitute, extender or as an additive to fossil diesel fuel in compression ignition engines. The most promising feature of biodiesel is that it can be utilized in existing design of diesel engine with no or very little modifications. It has a proven performance for air pollution reduction. Biodiesel is typically produced through the reaction of vegetable oils or animal fat with methanol or ethanol in the presence of catalyst to yield glycerol as major by product[1] (biodiesel chemically called methyl or ethyl ester). However, the price of biodiesel is presently more as compared to petro diesel [2]. Higher cost of biodiesel is primarily due to the raw material cost [3].One non edible biodiesel feedstock is simarouba glauca.

It is commonly accepted that there is some advancement of injection time when biodiesel is used in place of diesel because of its bulk density. The higher bulk density and viscosity transfers the pressure wave through fuel pipe lines faster and an earlier needle lift will lead to advanced injection. Due to the difference in cetane number, it is often suggested that injection timing be retarded to attain more complete combustion of vegetable oil based fuels [4]. Late injection of fuel into the combustion chamber helps in reducing the NOx emission of a diesel engine [5]. Evaluation of Simarouba esters indicates its superiority over many other vegetable oils in terms of engine performance, emissions, ease of use and availability. Simarouba glauca belongs to family simarubaceae, commonly known as “The Paradise Tree” or “King Oil Seed Tree”, is a versatile multipurpose evergreen tree having a height of 7-15 m with tap root system.

In India, it is mainly observed in Andhra Pradesh, Karnataka and Tamil Nadu etc. It can adapt a wide range of temperature, has the potentiality to produce 2000-2500 kg seed/ha/year ; can grow well in marginal lands/wastelands with degraded soils and therefore considered as a major forest tree. All parts of Simarouba are useful in some way or the other. The plant is also known for its medicinal properties.

Simarouba seeds

An effort is made in this study to evaluate the effect of varying the injection timing on the combustion, performance and emissions of a 5.2 kW engine fuelled with simarouba methyl ester of this oil (S20) for establishing the appropriate injection timing. The aim was to establish the modifications required in small, constant speed, direct injection diesel engines ,so that these can be made to run on Simarouba biodiesel (S20) with better performance and at the same time improve the emissions.

IDENTIFY, RESEARCH AND COLLECT IDEA

Biodiesel is produced by the transesterification of vegetable oils with alcohols to produce esters. [6]. Bio-diesel has become more attractive recently because of its environmental benefits and the fact that it is made from renewable resources [7]. Biodiesel has demonstrated a number of promising characteristics, including reduction of exhaust emissions [8]. Although there are many ways and procedures to convert vegetable oil into a Diesel like fuel, the trans-esterification process was found to be the most viable oil modification process [9]. Trans-esterification is the process of using an alcohol (e.g. methanol, ethanol or butanol), in the presence of a catalyst, such as sodium hydroxide or potassium hydroxide, to break the molecule of the raw renewable oil chemically into methyl or ethyl esters of the renewable oil, with glycerol as a byproduct. Glycerol is the major value-added byproduct produced from oil and fat from transesterification reactions performed during biodiesel manufacturing processes.

For a diesel engine, fuel injection timing is a major parameter that affects the performance and exhaust emissions. The state of air into which the fuel injected changes as the injection timing is varied, and thus ignition delay will vary. If injection starts earlier, the initial air temperature and pressure will be lower, so that the ignition delay will increase. If injection starts later (when piston is closer to TDC), the temperature and pressure will be slightly higher, a decrease in ignition delay will proceed. Hence, injection timing variation has a strong effect on the engine performance and exhaust emissions, especially on the brake specific fuel consumption (Bsfc), brake thermal efficiency (BTE) and NOx emissions, because of changing maximum pressure and temperature in the cylinder [10,11].

STUDIES AND FINDINGS

Properties of diesel and simarouba biodiesel

Property comparison of Diesel and Simarouba bio diesel are shown in table 1

SL.NO.	Characteristics	Diesel	Simarouba 100%	Simarouba 20%(S20)
1	Calorific value (KJ/Kg)	43000	39800	42360
2	Viscosity at 40^0C	2.6-5	4.8	3.4
3	Cetane number	50	51	51
4	Flash point (^{0}C)	55	165	70
5	Specific gravity	0.84	0.867	0.8454

EXPERIMENTAL SET UP, PROCEDURE AND OBSERVATION

The experiment aims at determining appropriate proportions of biodiesel & diesel for which higher efficiency is obtainable. Hence .experiments are carried out at constant speed, comparing the performance of compression ignition engine operated on blends of diesel. The S20 blend is checked under loads 20%,40%,60% and 80% with injection timing 15.1^0,20.5^0 and 25.5^0BTDC by constant injection pressure 200 bar and compression ratio 17.5. The samples are prepared by using the 1000 ml measuring jar and a 10 ml graduated test tube.

Fig.1 shows the schematic diagram of the complete experimental setup for determining the effects of waste cooking oil as bio diesel fuel additives on the performance and emission characteristics of compression ignition engine. It consists of a single cylinder four stroke water cooled compression ignition engine connected to an eddy current dynamometer. It is provided with temperature sensors for the measurement of jacket water, calorimeter water, and calorimeter exhaust gas inlet and outlet temperature. It is also provided with pressure sensors for the measurement of combustion gas pressure and fuel injection pressure. An encoder is fixed for crank angle record. The signals from these sensors are interfaced with a computer to an engine indicator to display P-Ɵ, P-V and fuel injection pressure versus crank angle plots. The provision is also made for the measurement of volumetric fuel flow. The built in program in the system calculates indicated power, brake power, thermal efficiency, volumetric efficiency and heat balance. The software package is fully configurable and averaged P-Ɵ diagram, P-V plot and liquid fuel injection pressure diagram can be obtained for various operating conditions.

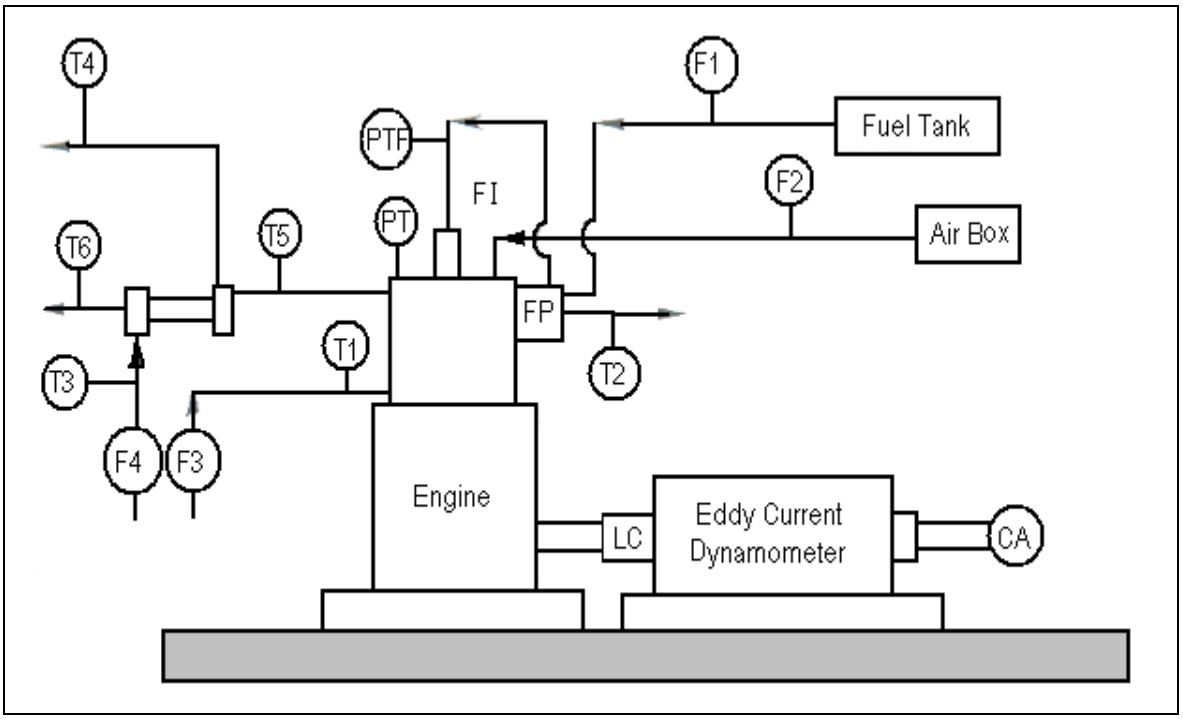


Fig. 3.1 Schematic Diagram of the Experimental Set-up

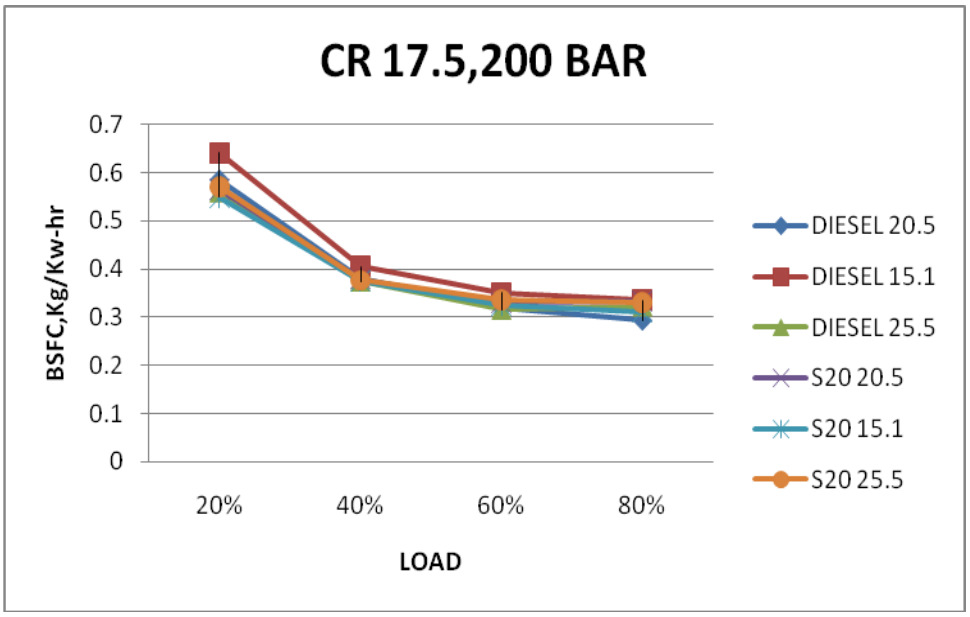
PT Combustion Chamber Pressure Sensor
PTF Fuel Injection Pressure Sensor
FI Fuel Injector
FP Fuel Pump
T1 Jacket Water Inlet Temperature
T2 Jacket Water Outlet Temperature
T3 Inlet Water Temperature at Calorimeter
T4 Outlet Water Temperature at Calorimeter
T5 Exhaust Gas Temperature before Calorimeter
T6 Exhaust Gas Temperature after Calorimeter
F1 Liquid fuel flow rate
F2 Air Flow Rate
F3 Jacket water flow rate
F4 Calorimeter water flow rate
LC Load Cell
CA Crank Angle Encoder
EGC Exhaust Gas Calorimeter

ENGINE SPECIFICATIONS

SL.NO	Engine parameters	specification
1	Engine type	TV1(Kirloskar ,four stroke)
2	Rated power	5.2 KW at 1500 rpm
3	Bore	87.5 mm
4	Stroke	110 mm
5	Cubic capacity	661 cc
6	Compression ratio	17.5:1
7	Injection pressure	200 bar
8	Injection timing	20.5^0 BTDC

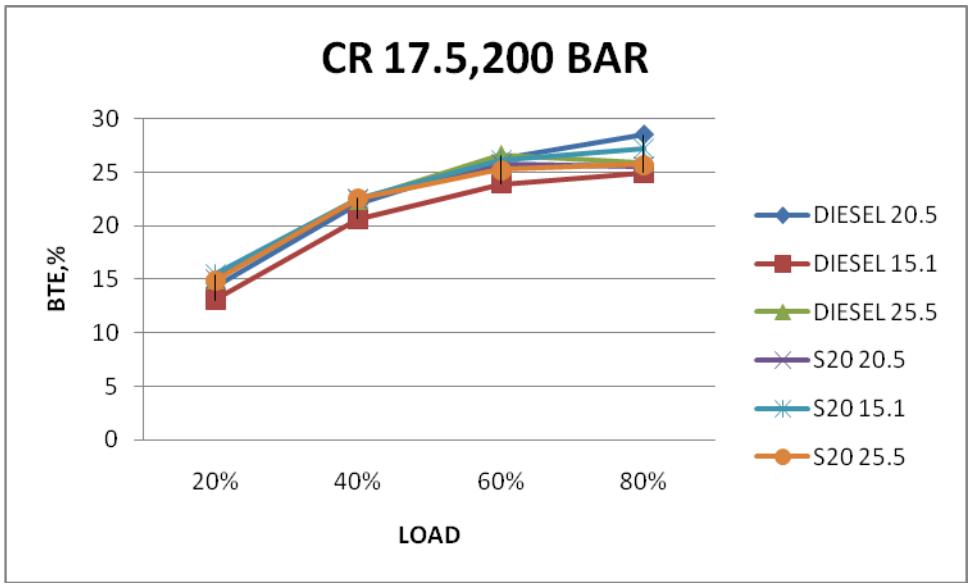
RESULTS AND DISCUSSIONS

Brake specific fuel consumption



While retarded injection timing BSFC is more in diesel comparative to S20 biodiesel.

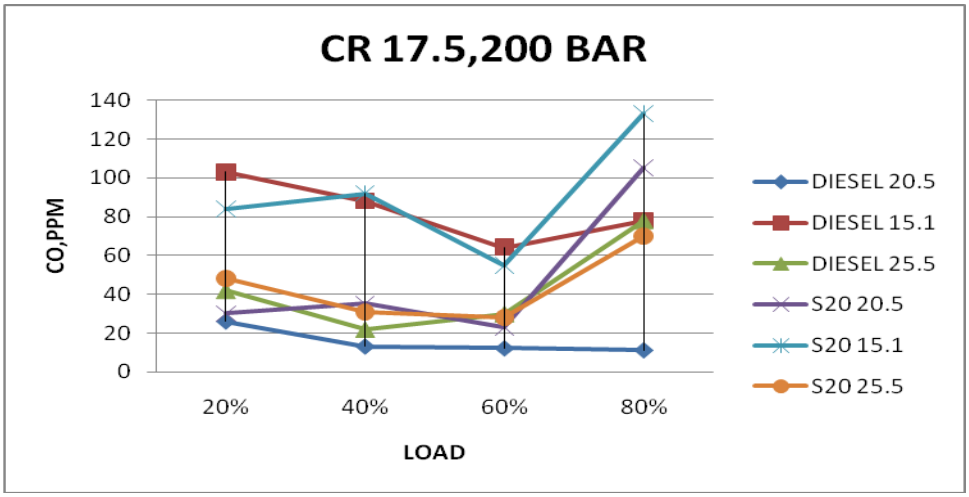
Brake thermal efficiency(BTE)



Less brake thermal efficiency is obtained while retarded injection timing for diesel. For retading injection timing S20 give better brake thermal effiency.

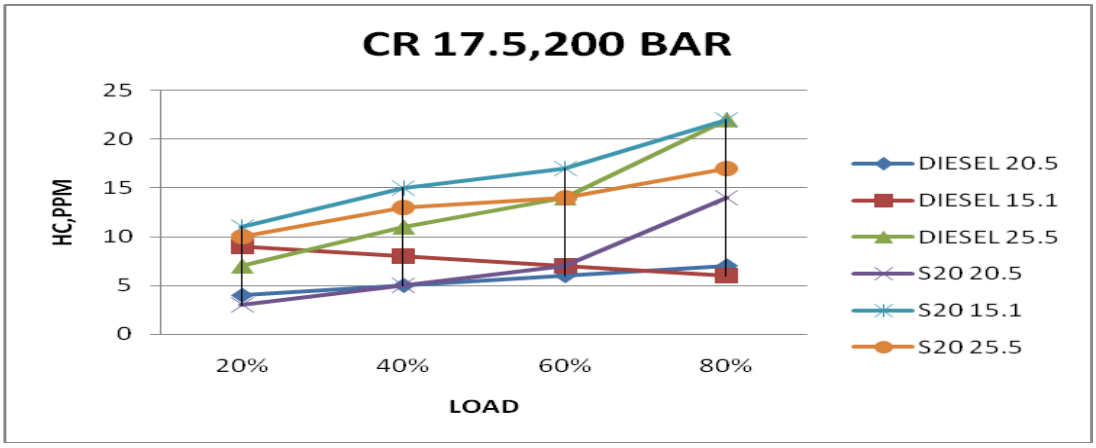
Carbon monoxide(CO)

Very less carbon monoxide emission for diesel while at standard injection timing 20.5^0.

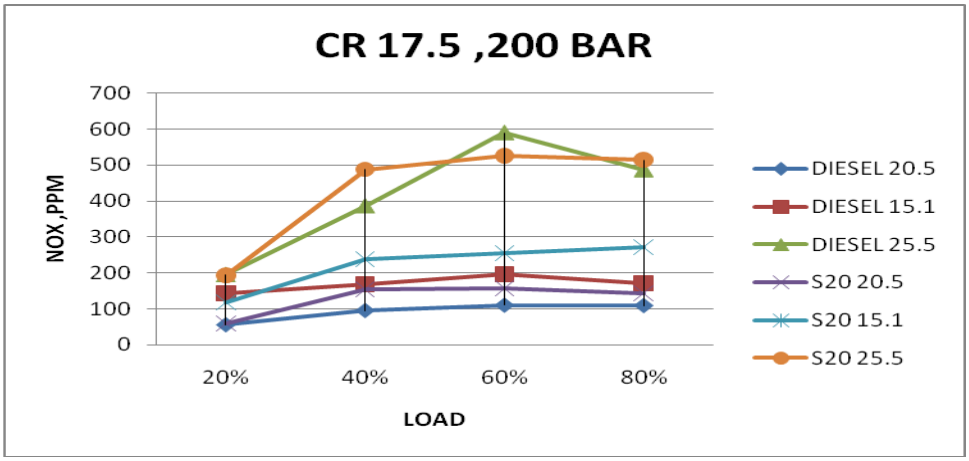


Hydrocarbons(HC)

Retarding injection timing causes more hydrocarbon emission.



NOX EMISSION



APPENDIX

BSFC-BRAKE SPECIFIC FUEL CONSUMPTION

BTE -BRAKE THERMAL EFFICIENCY

BTDC-BEFORE TOP DEAD CENTRE

CO- CARBON MONOXIDE

HC- HYDROCARBONS

NOX- OXIDES OF NOTROGEN

PPM- PARTS PER MILLENNIUM

S20-SIMAROUBA 20% BLEND DIESEL

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Bandwidth Enhanced in Wave Rectangular Dielectric Resonator Antenna (WRDRA)

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Abstract- WLAN requires low-cost and compact antennas with sufficient bandwidth. Chip antennas are particularly attractive as they are not substrate dependent and are solderable with standard reflow processes. Therefore, low-cost substrates can be used for the circuit board with no major consequence on radiation. Among possible antenna solutions dielectric resonator antennas (DRAs) offer an assembly that is surface mount technology (SMT) process compatible and their manufacturing can be based on standard processes. They also exhibit excellent properties regarding bandwidth and quasi omni directional radiation. When compactness is a main issue, conducting plates can be placed at the electric walls of the fundamental mode of the DRA to reduce its dimensions. However, this also decreases the bandwidth. Wider bandwidth structures have been proposed but they remain larger compared to a wavelength. In this paper, we propose a new DRA topology with both reduced size and large bandwidth. It has been designed for the low band of Hiperlan2 and IEEE802.11a standards [5.15–5.35] GHz. The proposed structure has been optimised using 3D EM simulations and is compared to a more conventional compact design.

A novel technique for producing enhanced band width in micro and mm wave region of spectrum is presented. A new design of compact & broadband wave dielectric resonator antenna is proposed using co-axially probe feed technique. Two different WDRA are designed and their characteristic behaviors are compared. Finally, parametric study of Second Antenna has been done. With the proper design the resonant behavior of the antenna is found, over which the leaky wave DRA produces extended bandwidth. Numerous designs for the WRDRA are simulated and bandwidths exceeding 20% are achieved.

Index Terms- Wave Rectangular Dielectric Resonator Antenna (WRDRA), Resonant, Broad band

I. INTRODUCTION

Recently, interest in dielectric resonator antennas has increased because of their attractive features such as small size, high radiation efficiency (98%), wide bandwidth, and high power capability for radar applications and base stations. The dielectric resonator antenna is made from high dielectric constant materials and mounted on a ground plane or on a grounded dielectric substrate of lower permittivity.

Design curves will be provided for the circular disc and hemisphere dielectric resonators. Use of these models with other geometries is discussed. Different excitation mechanisms are demonstrates such as the probe, slot, image line and waveguides. Applications of dielectric resonators in arrays are provided with discussion on the mutual coupling level and the wide scanning capabilities of the dielectric resonator antenna array. The array bandwidth limit is discussed based on the element size and the spacing between the array elements. Techniques for broadband applications are discussed. Some of the techniques are based on the material properties and some depends on the DRA shape. Several examples are provided. Some elements would provide a matching bandwidth over 40% with reflection coefficients better than –10dB for 50 Ohms ports. Finally, Techniques for size reduction of the DRA are presented to demonstrate the flexibility of the DRA to satisfy the required small size for some applications. The technique will result in small size and keeping wide bandwidth. The applications of the DRA for spatial power combiners are presented. The DRAs are placed in an oversized hard horn to provide uniform field distribution. Recent developments of the dielectric resonators as a multifunction device will be also provided. In this application we will show the use of the same DR as an antenna with low quality factor and as a resonator with high Q-factor.

Modern communication systems require wide bandwidth to support the demand of high data rate transfer for various multimedia applications. To fulfill this requirement, most wireless mobile systems have to be operated at the millimeter wave frequencies [1]-[2]. For ease of space allocation, it is highly desirable to have small size, low profile equipment. Hence, the antennas for modern wireless communication system should be low in profile and efficient in high frequencies.

Dielectric resonator antennas (DRA) have been the interest of research and investigation due to its highly desirable characteristics such as small size, light weight, highly efficient in microwave and mm wave spectrum. The most popular shape studied for practical antennas applications have been the cylindrical dielectric resonator antennas, rectangular dielectric resonator antennas, spherical dielectric resonator antennas and many more different structure are reported. The stacked DRA has also been tested [3]-[7] with a resulting increase in bandwidth that is much wider than the bandwidth of the micro strip antennas.

The dielectric resonator antennas based on NRD guides have few publications. The technique of NRD guided antenna was proposed by Yoneyama and Nishida [8]-[10] . Although, it is classified as open dielectric waveguide, it has attractive feature of no radiation [11]-[13]. However, introducing suitable perturbation to the NRD guide structure can produce leaky waves that propagate away from the dielectric slab to the open ends.

This mechanism makes the NRD guide working as a leaky wave antenna.

Several techniques have been proposed to generate leaky waves from NRD guide, such as foreshortened sides of parallel metal plate's technique, Asymmetric air gape technique, Trapezoidal dielectric slab technique and many more.

A NRD guide itself is a transmission line and so it is non-radiative. In this research paper a novel dielectric antenna based on NRD guide model has been presented. A wave rectangular dielectric resonator antenna (WRDRA) has been designed. WRDRA is excited by coaxial probe feed mechanism. The WDRA is parametrically studied and different approaches are presented to achieve an extended bandwidth nearly20% at -10dB. The study also shows the dual resonance behavior WRDRA at frequency value of 22.14GHz and 24.97GHz. The dependence of band width on the various parameters and the geometries of the system show that higher band width with desired radiation characteristics can be achieved with such dielectric resonator antenna based on NRD guides. Therefore, it is necessary to extend extensive research and study on this topic, because it can provide an alternative device to achieve wider band width characteristics.

II. Antenna Design

First proposed design uses a substrate of relative permittivity of 2.4 and dimension 210mmX 152mmX 0.6mm. The upper surface of the substrate has finite conductivity layer. This has been done to minimize the back-lobe radiation phenomenon. The rectangular dielectric resonator of relative permittivity 8.2 is used, having dimension of 148mmX 6mmX 5.2mm. The top layer surface of the dielectric resonator is perturbed by embedding a strip of thickness 0.8mm and length equal to that of the dielectric material. The co-axial probe feed mechanism is used for the excitation of WDRA. In Fig. 1(a), the proposed antenna is presented in 3D view. The proposed First Antenna is shown in fig. 1(b). In the second antenna design the perturbation is increased by just modifying the embedded dielectric material strip of First Antenna design, by removing material of thickness 0.1mm from between the upper face. The antenna design is shown in fig. 1(c). This is termed as Second Antenna designed. The parametric study of the Second Antenna is done by varying the probe penetration length, l into the dielectric material of the resonator. The variation of the length is done from -0.6mm to 5.4mm in step size of 0.6mm. The antenna designed for parametric study, showing probe length l, is shown in fig. 1(d).

III. Simulations

The designed antenna is simulated on Ansoft HFSSv10 simulation software. The simulation of LWRDRA has been done in three stages. At first, the First Antenna was simulated and results were recorded. In second stage, the Second Antenna was simulated and the results which were obtained were compared with that of the First Antenna. At last, the parametric variation of the project variable, l was done. The result obtained was studied in detail to know the relation between probe positions, probe length and the height of the dielectric material.

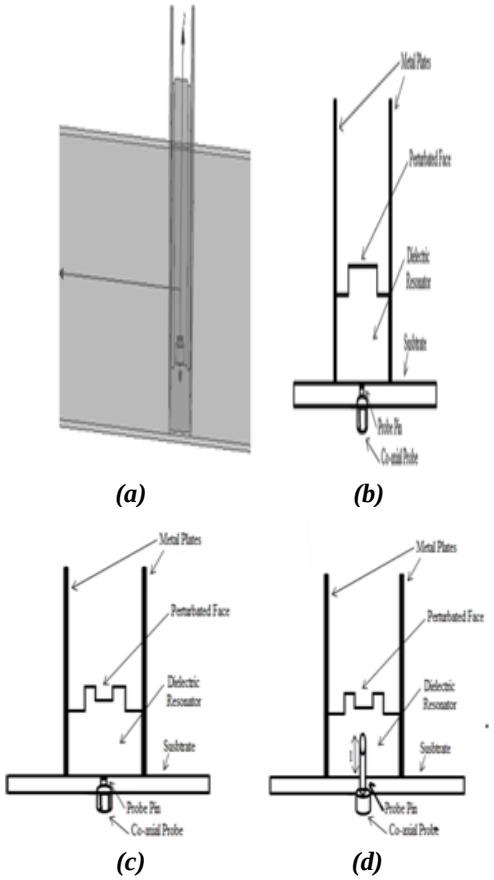

Figure 1(a) WRDRA Designed & Simulated on HFSS 3D view, (b) First Antenna Designed, (c) Second Antenna Design and (d) Second Antenna Design, with probe pin, l set as project variable for parametric study of Antenna

A. Simulation Result of First Antenna

The S_{11}Vs Frequency plot of first antenna is shown in Fig. 2. It can be seen that antenna is well matched at 9.89GHz having return loss of -23.01dB. It has a bandwidth (-10dB) of 800MHz which corresponds to nearly 8.2% in the frequency range of 9.3GHz-10.1GHz.

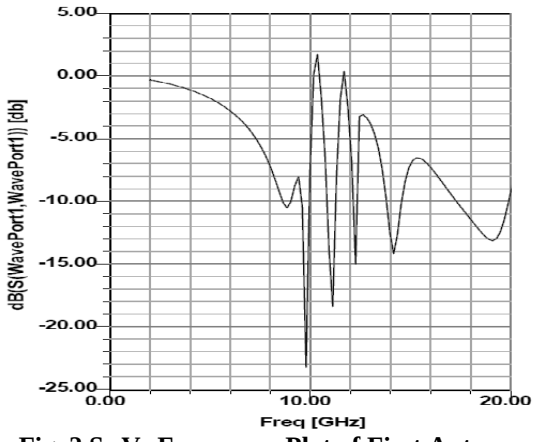

Fig. 2 S_{11}Vs Frequency Plot of First Antenna

B. Simulation Result of Second Antenna

The S_{11}Vs Frequency plot of second antenna is shown in fig. 3. This graph shows that the second antenna is well matched at frequency of 24.25GHz and having return loss of -28.59dB. It has -10dB bandwidth of 16.54% in the frequency range of 22.34GHz-26.37GHz. This shows the 100% increase in the bandwidth as compared to the first antenna design.

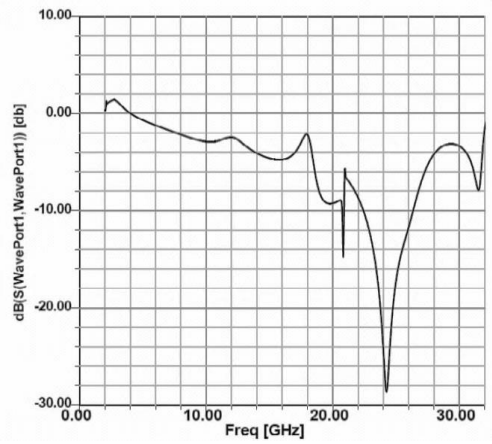

Fig. 3 S_{11}Vs Frequency Plot of Second Antenna

C. Simulation Result of Parametric Study of Second Antenna by Varying the Probe Length, l Penetration into the Dielectric Resonator of WRDRA

As discussed in Section II, the parametric variation in the probe length, l was done by setting, l as project variable. The simulated results are shown in Fig. 4(a), 4(b), 4(c). During the variation of probe length, l, for l = -0.6mm to 0.6mm, it is observed that the there is decrease in bandwidth as well as increase in return loss. The matching frequency gets shifted to higher value. The frequency range over which the bandwidth is calculated gets shifted to higher value (shown in Fig. 4(a)). For l = 1.2 and l=1.8, there is increase in bandwidth as well as resonant frequency. For l = 2.4mm to 3.6mm, the resonant frequency tends to decrease and bandwidth along with return loss starts to increase. At l = 4.2mm, we observed a dramatic decrease in the return loss to -31.60dB, the resonant frequency decrease to a value of 18.63GHz at -10db, calculated bandwidth is 8.4%. At l = 4.8mm, dual resonance behavior of the antenna is observed. The WRDRA resonates in the frequency range of 21.62GHz-25.03GHz. The bandwidth obtained at this frequency range is nearly 20%. The tabulated result of the parametric variation of probe length, l has been presented in Table I.

Fig. 5 shows the S_{11} Vs Frequency plot of Second Antenna at resonant condition. It is found that as l = 4.8mm, the WRDRA acts as dual resonant leaky wave antenna. The resonant frequencies of WRDRA are 22.14GHz and 24.97GHz. Calculated bandwidth for the design at dual resonance is found to be 14.6% (-10dB) with overall bandwidth of the WRDRA at -10dB bandwidth is nearly 20%.

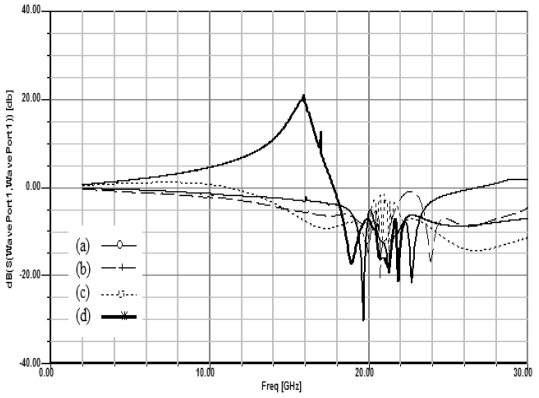

Fig. 4(a) S_{11}Vs Frequency Plot of Second Antenna graph showing parametric variation of Probe length, l inside the WDRA(a) l = -0.6mm, (b) l = 0mm (c) l = 0.6mm (d) = 1.2mm

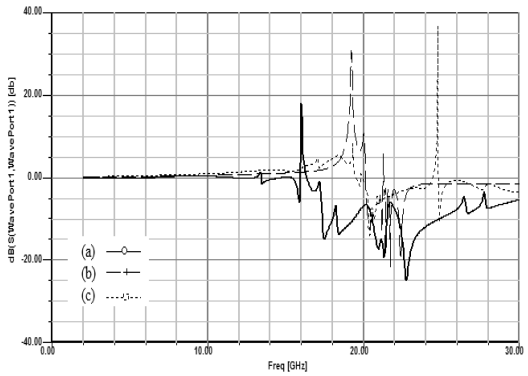

Fig. 4(b) S_{11}Vs Frequency Plot of Second Antenna graph showing parametric variation of Probe length, l inside the WDRA (a) l = 1.8mm, (b) l = 2.4mm (c) l = 3.0mm

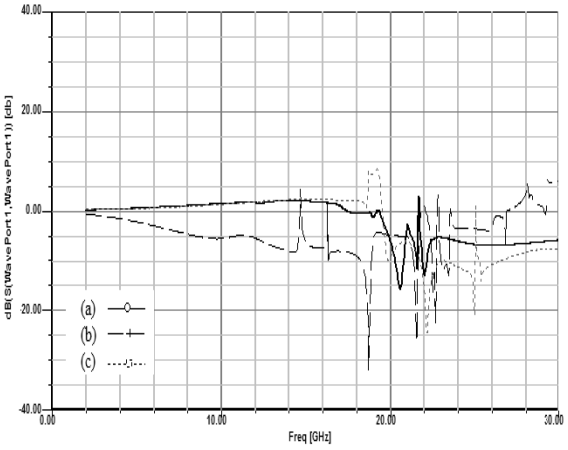

Fig. 4(c) S_{11}Vs Frequency Plot of Second Antenna graph showing parametric variation of Probe length, l inside the LWDRA (a) l = 3.6mm, (b) l = 4.2mm (c) l = 4.8mm

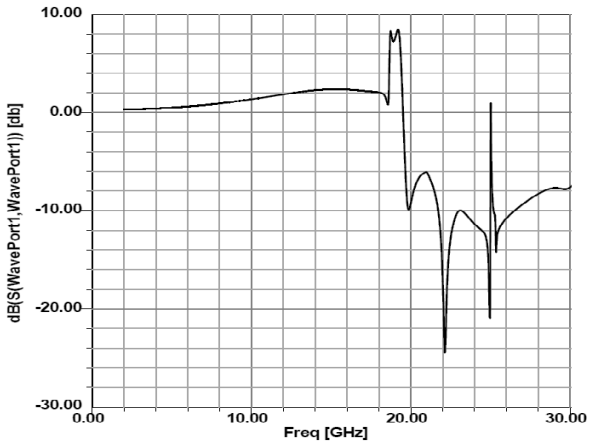

Fig. 5 S_{11}Vs Frequency Plot of Second Antenna at l = 4.8mm showing the Dual Resonance Behavior of WRDRA at -10dB

TABLE I VARIATION OF PROBE LENGTH INSIDE WRDRA

S. No.	Probe Length (mm)	Low Freq. (GHz)	High Freq. (GHz)	Reso-nant Freq. (GHz)	Band-width (%)	Return Loss (-dB)
1	-0.6	19.24	19.97	19.66	3.72	30.58
2	0.0	20.58	21.31	20.74	3.48	20.55
3	0.6	21.05	21.47	21.26	1.97	15.63
4	1.2	21.62	22.14	21.82	2.37	21.41
5	1.8	22.29	24.82	22.76	10.7	25.10
6	2.4	22.08	22.60	22.85	2.32	20.18
7	3.0	20.14	20.84	20.33	3.41	14.12
8	3.6	20.18	20.89	20.85	3.45	15.72
9	4.2	18.21	18.93	18.63	8.40	31.60
10	4.8	21.62	26.53	22.14	20.3	24.52

IV. CONCLUSION

A new, comprehensive dual resonance WRDRA has been designed. It is found that the First Antenna which has less perturbation on dielectric resonator's upper surface has bandwidth of 8.2% (-10dB) and good matching at frequency 9.89GHz but as the perturbation is increased as seen in Second Antenna design, the bandwidth gets increased to a new value of 16.54% (-10dB) and good matching of the system occurs at 24.25GHz.Thus, it is found that as the perturbation of dielectric resonator of WDRA increases, the bandwidth of the system along with matching frequency gets shifted to some higher value. In First Antenna perturbation of upper surface of dielectric resonator was half as compared to that of Second Antenna, and the numerical results obtained shows that values of bandwidth and the resonant frequency of WRDRA depends upon the perturbation of the surface. Analysis suggests the relation of direct proportionality between perturbation of dielectric resonator surface and the bandwidth and resonant frequency of the WRDRA. Further 16% increase in the bandwidth is obtained by increasing the probe penetration into the dielectric material. Results obtained by the parametric study of probe penetration length inside the dielectric resonator material of the antenna demonstrate that the dual resonance behavior of the WRDRA is obtained when the antenna is co-axially feed and the position of excitation is at the 3/4^{rth} distance from the center of the resonator and the penetration length is equal to 0.8 times that of the height of the rectangular dielectric resonator. By applying the composite technique the extended band width can be produced.

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Analysis & Trouble Shooting of the Contaminations, Arised in North Indian Zones during Plant Tissue Culture

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Abstract- In the present investigation Jatropha and pea plant leafs was used, to germinated the plant by artificial condition through plant tissue culture method. We have taken three methods Method A, B & C to germinate callus, seed, bud and stem. 30 Sample were taken to analyze.
In Method A 60% samples were contaminated (fungus and bacteria) and 20% was shown growth.
In Method B 66% was contaminated by fungus and 19% were contaminated by bacteria and only 15% are shown growth.
In Method C 85% samples were contaminated by fungus and bacteria, only 15 % samples was show growth.
The study shows most of all contamination was due to Fungus, a very less contamination were due to bacteria. Among three method used with different Sterilization process. Method A was best suited for Callus Culture, Method B for Seed Culture & Method C for Stem & Bud Culture.

Index Terms- Analysis of contamination occurs in plant tissue culture, Biotechnology, Graph for Stem Culture, callus culture, seed culture, bud culture.

I. INTRODUCTION

Biotechnology is name given to the methods and techniques that involve the use of living organisms like bacteria, yeast, plant cells etc or their parts or products as tools (for example, genes and enzymes). They are used in a number of fields: food processing, agriculture, pharmaceutics, and medicine, among others. Plant tissue culture can be defined as culture of plant seeds, organs, explants, tissues, cells, or protoplasts on nutrient media under sterile conditions.

Plant tissue culture, also referred to as cell, in vitro, axenic, or sterile culture, is an important tool in both basic and applied studies, as well as in commercial application *(Adams, R.P. and Adams, J. E. (eds.) 1992, Conservation of Plant Genes, DNA Banking and In Vitro Technology. Academic Press, New York.)*

Plant tissue culture is the aseptic culture of cells, tissues, organs and their components under defined physical and chemical conditions in vitro. The theoretical basis for plant tissue culture was proposed by Gottlieb Haberlandt in his address to the German Academy of Science in 1902

on his experiments on the culture of single cells *(Bhojwani, S.S. (eds.) 1990, Plant Tissue Culture: Applications and Limitations. Elsevier, Amsterdam)*

He opined that, to my knowledge, no systematically organized attempts to culture isolated vegetative cells from higher plants have been made. Yet the results of such culture experiments should give some interesting insight to the properties and potentialities that the cell, as an elementary organism, possesses. Moreover, it would provide information about the interrelationships and complementary influences to which cells within a multicellular whole organism are exposed (from the English translation, *(Lewin, B. 2004 Genes VIII, Oxford University Press, New York)*

He experimented with isolated photosynthetic leaf cells and other functionally differented cells and was unsuccessful, but nevertheless he predicted that one could successfully cultivate artificial embryos from vegetative cells. He, thus, clearly established the concept of totipotency, and further indicated that the technique of cultivating isolated plant cells in nutrient solution permits the investigation of important problems from a new experimental approach. On the basis of that 1902 address and his pioneering experimentation before and later, Haberlandt is justifiably recognized as the father of plant tissue culture. Other studies led to the culture of isolated root tips *(Old, R.W. and Primrose, S.B. 1990.) (Singer, M. and Berg, P. 1991. Genes and Genomes, University Science Books, Mill Valley, California and Blackwell Science Publishers)*

II. MATERIALS AND METHODS

Reagents used:
Extran (5% v/v) / tween 20.
0.1 % (w/v) Mercuric chloride solution
0.7% (w/v) Agar

Growth regulators:N^6-benzyladenine (BA)/ kinetin (Kn) /indole-3-acetic acid (IAA)/indole 3- butyric acid (IBA) / α naphthalene acetic acid (NAA).
70 % Ethanol
1% Sodium hypochlorite

METHOD A - Young shoots and callus of jatropha plant as a source tissue were washed, first under running tap water followed by a detergent Extran (5%v/v) / tween 20 for 5 min. After thorough wash in water, source tissue was surface sterilized using 0.1 % (w/v) mercuric chloride solution for 7-9 min. After repeated washing with sterile water (3 times or 15 min) node segment were cut into appropriate size (1.0-1.5 cm) and cultured on sterile nutrient medium

METHOD B -Capsules were washed with 2–3 drops of Tween 20 and surface-disinfected in 0.1 % mercuric chloride solution for 2 min followed by 70 % ethanol for 30 s and surface flaming. Finally, the capsules were rinsed five times with sterile distilled water and dried. All subsequent work was carried out aseptically in laminar flow cabinets. The capsules were then dissected longitudinally with a surgical blade and extracted seeds were spread thinly over the surface of liquid culture medium contained in 25 × 150 mm glass test tubes (20 test tubes per capsule) each containing 10 ml of medium

METHOD C- Take explants part of plant and wash with 2 drops of tween 20 around 15 minute and then washed the explants with sterile distilled water around 2 times and then dip the explants part in NaClO 10 % concentration and dip around 15 min and then washed up to 2 times and then planted on MS media

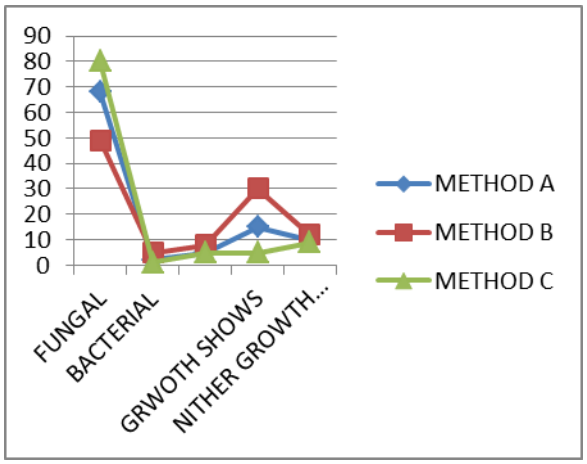

Fig 2: Graphical representation of contamination and growth in seed culture by all the three methods (method A, B and C)

III. RESULTS AND DISCUSSION

Graph for Callus Culture

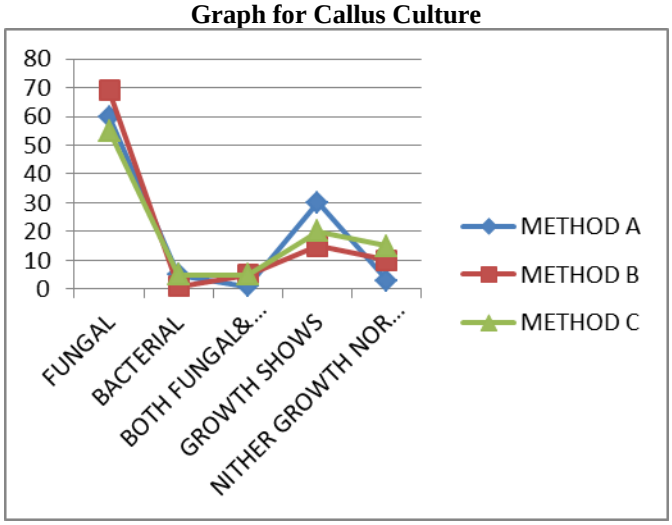

Fig 1: Graphical representation of contamination and growth in callus culture by all the three methods (method A, B and C)

Graph for Bud culture

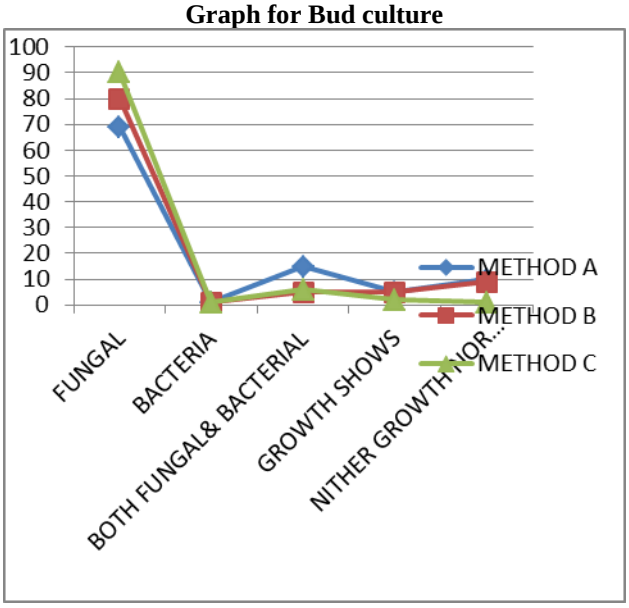

Fig 3: Graphical representation of contamination and growth in Bud culture by all the three methods (method A, B and C)

Graph for Stem Culture

Graph for Seed Culture

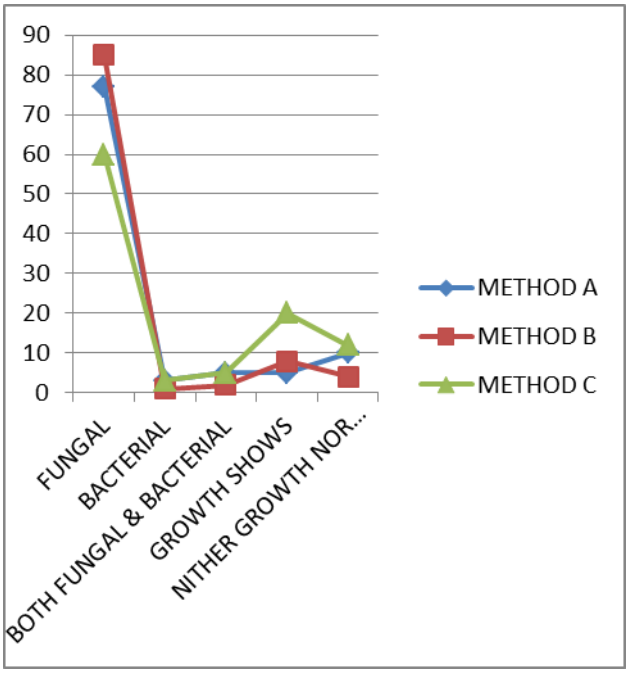

Fig 4: Graphical representation of contamination and growth in Stem culture by all the three methods (method A, B and C)

GRAPHICAL REPRESANTATION OF ALL THE CULTURES AND THEIR RESPECTIVE CONTAMINATION.

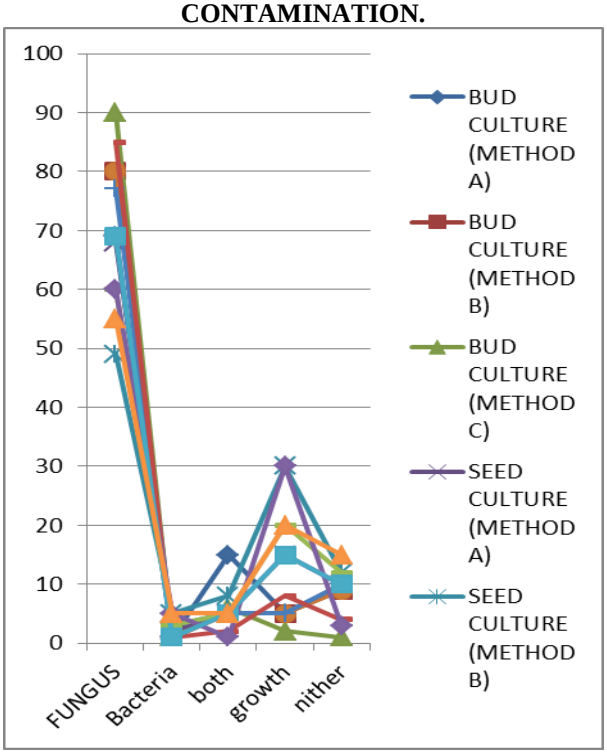

Fig 6: Graphical representation of contamination and growth in Callus culture, Stem culture, Bud culture and Seed culture by all the three methods (method A, B and C)

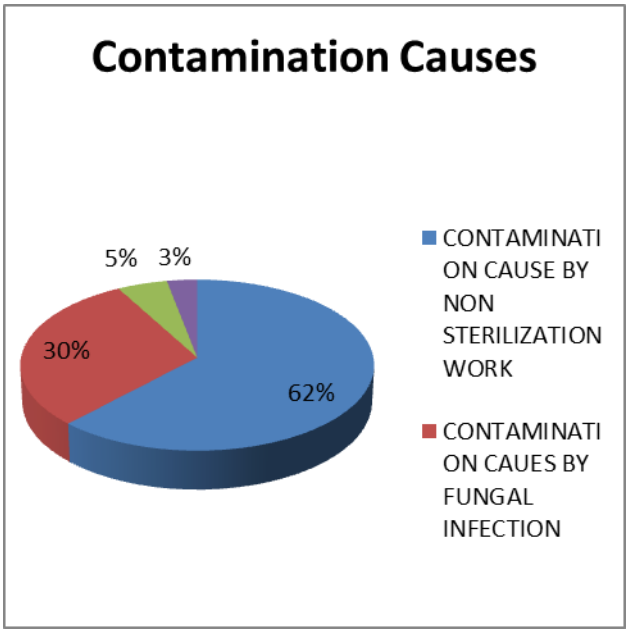

Fig 5: Pie chart depicting the percentage of contamination arising because of different reasons.

Above pie chart shows few other causes of contamination which we faced during our study. From this it is clear that in plant tissue culture field mostly contamination occurs when sterilization conditions are not maintained properly. Around 30 samples of callus cultures were taken, out of which 22 samples were clearly contaminated due to non sterilization conditions. And 30% samples were mostly contaminated by fungal infection, and 5% samples are mostly contaminated by lack of nutrition in MS media due to which other growth inhibits the growth of samples. And 3 % samples were mostly contaminated or died due to overheating of the samples in white light.

IV. CONCLUSION

The ability to establish and grow plant cell, organ, and tissue cultures has been widely exploited for basic and applied research, and for the commercial production of plants (micro-propagation). Regardless of whether the application is for research or commerce, it is essential that the cultures be established in vitro free of biological contamination and be maintained as aseptic cultures during growth and storage successful micro propagation.

Contamination in plant tissue culture could have very serious consequences on the result. These contaminations could originate from any of the sources as discussed above. Besides fungal and bacterial contamination, non sterilization work plays a major role in causing contamination in plant tissue culture work. In this study we focused mainly upon the percentage of fungal contamination, bacterial contamination, contamination caused by non sterilization work and contamination caused by loss of nutrition.

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Assured Data Transfer under Auditing in Distributed Circumstances

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Abstract- Distributed considerations are the major acquiring epitome. The domain abbreviates monetary value related on computation. The service provided on scattered location to its users on demand across the cyberspace. The data and other resources used by the user are stored in the open environment. The circumstance issues more on data security and user fear on missing bound on their data. To enrich security on data, the security mechanisms are implemented, though the data integrity is unnoticed to user. To overcome the problem and achieve data integrity the method of auditing is established through Third Party Auditing (TPA). In addition to auditing the sensitive data in uploading over the dispersed area are protected by DES encryption algorithm.

Index Terms- Open environment, data security, data integrity, auditing, TPA, DES encryption algorithm.

I. INTRODUCTION

Cloud computing is a type of computing that relies on *sharing computing resources* rather than having local hosts or personal devices to handle applications. In cloud computing, the word cloud (also phrased as "the cloud") is used as a metaphor for "*the Internet*," so the phrase *cloud computing* means "a type of Internet-based computing," where different services -- such as servers, storage and applications -- are delivered to an organization's computers and devices through the Internet. Cloud computing is an on-demand service that is obtaining mass appeal in corporate data centers. The cloud enables the data center to operate like the Internet and computing resources to be accessed and shared as virtual resources in a secure and scalable manner. Cloud computing opens up a new world of opportunities for businesses, but mixed in with these opportunities are numerous security challenges that need to be considered and addressed prior to committing to a cloud computing strategy. Cloud computing security (simply referred "cloud security") is a sub-domain of information technology security. In order to protect the cloud data and applications, a set of security policies, methodologies and control technologies are implemented in the associated cloud security infrastructure. To confidently leverage cloud solutions, cloud security is needed.

Major issues are compliance and access control related. Security concerns associated with cloud computing fall into two categories: security issues faced by cloud providers and those faced by customers. This is the main reason why security in the cloud is a shared responsibility: both the provider and the customer must ensure that proper measures are taken in order to protect the client's data and to ensure that the infrastructure is secure. Cloud providers and their clients can negotiate terms around liability, intellectual property and end-of-service when signing the Service Level Agreement. Cloud computing security challenges fall into two broad categories:

- Data Protection: Securing the data both at rest and in transit
- User Authentication: Limiting access to data and monitoring who accesses the data

Implementing a cloud computing strategy means placing critical data in the hands of a third party, so ensuring the data remains secure both at rest (data residing on storage media) as well as when in transit is of paramount importance. Data needs to be encrypted at all times, with clearly defined roles when it comes to who will be managing the encryption keys. In most cases, the only way to truly ensure confidentiality of encrypted data that resides on a cloud provider's storage servers is for the client to own and manage the data encryption keys.

Data resting in the cloud needs to be accessible only by those authorized to do so, making it critical to both restrict and monitor who will be accessing the data through the cloud. In order to ensure the integrity of user authentication, user need to be able to view data access logs and audit trails to verify that only authorized users are accessing the data. These access logs and audit trails additionally need to be secured and maintained for as long as the user needs or legal purposes require.

However, security concerns become relevant as we now outsource the storage of possibly sensitive data to third parties. The security issues are considered, a secure overlay cloud storage system –FADE (File Assured Deletion) provides fine-grained access control and assured deletion for outsourced data on the cloud [2]. The stored data in the cloud is accessed through various techniques, the relevant work is done with role based access control in cloud secure the users data [3].

The data leaving in third parties hand is managed in secure form, to ensure the security authentication mechanism is implemented. The security in the cloud database is attained with hybrid encryption method which concerned about small sized data [4]. The cloud service providers, issues and services are categorized in study [1] which entitles the base level of cloud utility.

The analysis on encryption algorithm details the use of algorithm over cloud based on confidentiality, integrity and availability in best approach manner. The confidentiality on data on the cloud is ensured up with RSA algorithm in earlier [5]. Two –factor authentication technique fulfilled the data integrity measure on the cloud. The fulfilment is achieved through Diffie-

Hellman key exchange algorithm [6]. To monitor the usage of encryption algorithms in data security the new cloud environment is designed with java in concern including RSA and MD5 for resource allocation controlled by client and cloud admin [7].

To be effective, distributed data security depends on more than simply applying appropriate data security procedures and countermeasures. This paper guarantees the data security and data accountability on the open environment with beneficial supportive methodologies. Data uploading with encryption is highly beneficial in cloud data security. The TPA automatically logs the usage of data on the cloud with access control policies and authenticated logging mechanism.

II. Problem Statement

Data security is a vital issue in the cloud computing environments. In cloud, the data can be physically located anywhere in any data centre across the distributed network. The cloud nature issues more with user authentication, data integrity and confidentiality. The data hosted in the cloud is completely under the third party control to ensure the data usage in the cloud, the data on the cloud environment is accounted with Third Party Auditing. The Distributed Information Accountability (DIA) framework completes the progress of data accountability over cloud. The sensitive data is achieved

III. Data Storage

Cloud-based web hosting and locally-stored server hosting are the two major data storage available. Each of these has benefits, though it's better to be turning to using cloud-based web hosting. While attacks could happen on a server that's on the Internet, business owners without much office space to house web servers are finding the trade-off satisfactory. The benefits of using cloud-based hosting for access to the desktop/service/files anywhere we need them is a key selling point for many. It could also be known as the back-up place of data. The stored data under third party control must be guaranteed in correctness and availability. The major issue is to effectively detect any unauthorized data modification. To address this problem regarding security the data to be stored upon cloud must be encrypted and authenticated. The issue is recovered up with erasure code [6] which safeguards from Byzantine failure but the technique occupies more memory.

IV. Data ACCESS

Accessing data from the outsourced environment should be easy to get into access. The environment should also provide a safe data access. The access is controlled in different categories such as Role Based Access Control [2] and file based access [1]. Access rights are monitored with authentication and authorization techniques in better way to attain security.

V. Security Techniques

To enrich security in the cloud, the various security techniques are available. The data over public network can be protected by the method of encryption. Encryption converts that data by any encryption algorithms using the "key" in scrambled form. Only user having access to the key can decrypt the encrypted data.

Encryption algorithm plays a big role in providing data security against malicious attacks. Encryption algorithm can be categorized into symmetric key (private) and asymmetric key (public) key.

In symmetric key encryption or secret key encryption, only one key is used to encrypt and decrypt data. In asymmetric key encryption, two keys are used: public and private keys. Public key is used for encryption and private key is used for decryption.

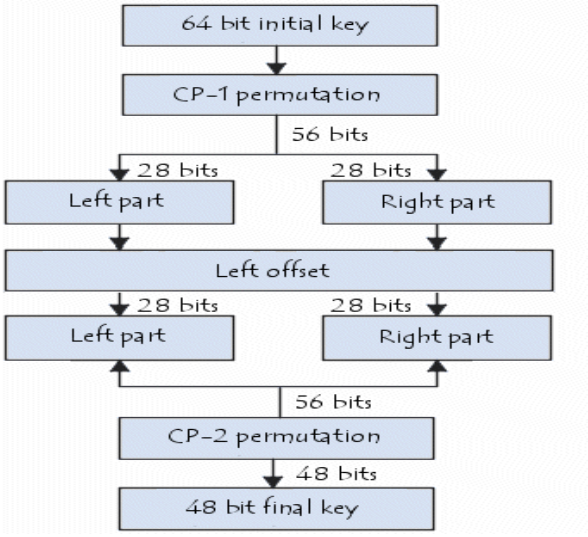
The data in the cloud allowed being more secure with encryption algorithms and to valid the original data hashing technique is used. The algorithms analyzed are RSA, DES, and AES.

RSA

RSA is a commonly used public key cryptography algorithm. The first, and still most commonly used asymmetric algorithm RSA is named for the mathematicians Rivest, Shamir, and Adleman. RSA today used in hundreds of software products and can be used for key exchange, digital signature, or encryption of small blocks of data. RSA uses a variable size encryption block and a variable size key. The key pair is derived from a very large number, n-that is the product of two prime numbers chosen according to special rules. Since it was introduced in 1977, RSA has been widely used for establishing secure communication channels and for authentication the identity of service provider over insecure communication medium. In the authentication scheme, the server implements public key authentication with client by signing a unique message from the client with its private key, thus creating what is called a digital signature. The signature is then returned to the client, which verifies it using the server's known public key.

DES

Data Encryption Standard (DES) is a widely-used method of data encryption using a private (secret) key. DES originated at IBM in 1977 and was adopted by the U.S. Department of Defence. It is specified in the ANSI X3.92 and X3.106 standards and in the Federal FIPS 46 and 81 standards. There are 72 quadrillion or more possible encryption keys that can be used. For each given message, the key is chosen at random from among this enormous number of keys. Both the sender and the receiver must know and use the same private key. DES applies a 56-bit key to each 64-bit block of data. The process can run in several modes and involves 16 rounds or operations.



AES

The Advanced Encryption Standard (AES) is a National Institute of Standards and Technology specification for the encryption of electronic data. AES is a new cryptographic algorithm that can be used to protect electronic data. Specifically, AES is an iterative, symmetric-key block cipher that can use keys of 128, 192, and 256 bits, and encrypts and decrypts data in blocks of 128 bits (16 bytes). Unlike public-key ciphers, which use a pair of keys, symmetric-key ciphers use the same key to encrypt and decrypt data. Encrypted data returned by block ciphers have the same number of bits that the input data had. Iterative ciphers use a loop structure that repeatedly performs permutations and substitutions of the input data.

VI. Auditing on Data

Data uploaded in the cloud environment is utilized only whenever required. Until then the data on the cloud must be secure. In the same time the user of the data in need of monitoring the access over the sensitive information which get shared on the third party environment. The access rights on the data in the cloud are provided by the owner. Though the access rights are set there is essential to get notice on time of access and to maintain log. Audit on accessing over the data is maintained by TPA.

VII. Implementation

To ensure the data security the proposed framework includes an idea of auditing the data along allows uploading files in public and private forms and the authentication mechanism is managed to monitor the logs.

DATA UPLOADING

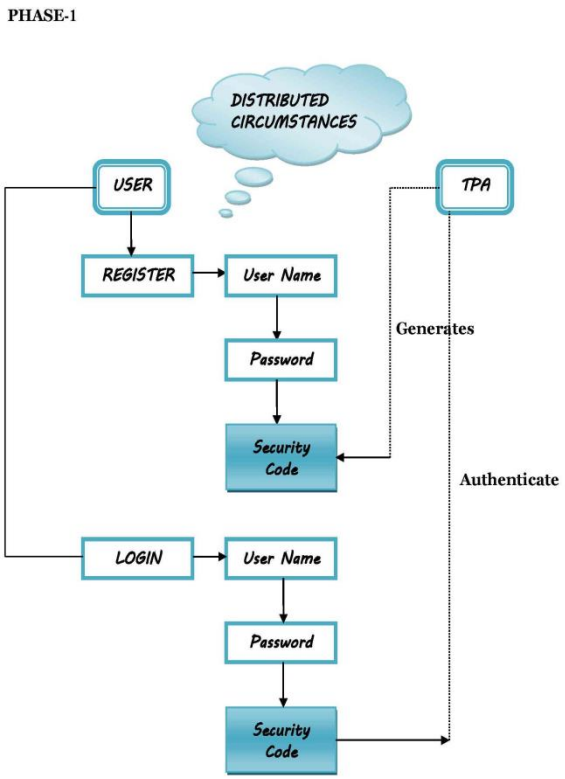
The uploading data in the cloud is being in two different sets. The data which is more sensitive is authorized to encrypt before adding into the cloud circumstances which is meant as Private data. The private data is encrypted using DES algorithm which the key generated is shared to download the data. The DES algorithm aid high-level security, efficient and economical manner. The public data is uploaded as such in the same format without any modification.

ACCESS MECHANISM

The access rights on the data are provided by the data owner. The access rights on data sharing comprise view mode and modify mode. The view mode is which allow the members who get shared with data only possible to view the contents unable to download or modify. The modify mode is download enable mode where the member with data access can download and modify the data as specified by the owner.

DATA ACCOUNTING

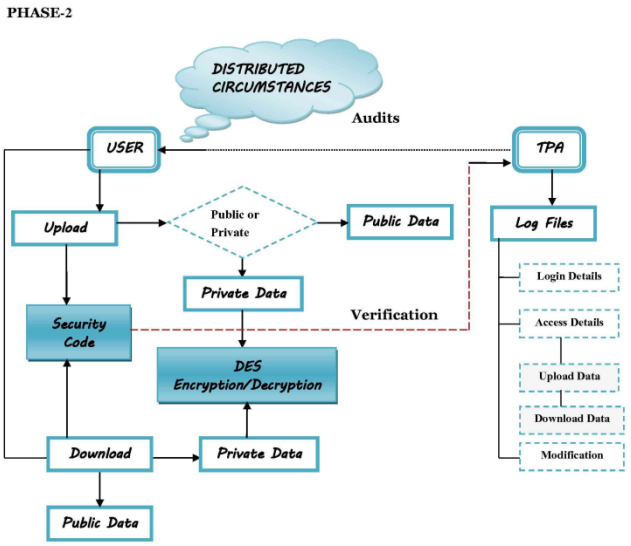
The data accountability is maintained with the log file created in every access on the data. The generated log files are the monitor display at the owner side. The mechanism maintenance is attained by the Third Party Audit (TPA). The TPA has rights only on authenticating the access. The log file includes the contents of data access time, unauthorized access, modification details.



DATA DOWNLOAD

The owner of the data is only supposed to download the data uploaded in the private mode. Every access on data is reported by TPA. The access record consists of data being shared or rest on the cloud circumstances. The private data download includes simple steps as on data upload. The encrypted data is decrypted using the key generated while uploading the content. As per DES encryption logic, pair of key generated, the key left

for decrypt is to download. This decrypt on data validates the integrity of the original data. This ensures data security.



VIII. CONCLUSION

The Distributed circumstances provide an enormous facility in taking challenges over the data in the concept of whenever and wherever required. The problem with data security issues in cloud data storage and data transfer. The DIA framework attains idea on security on data in rest as well as in transit. The data storage is secured with DES algorithm, access rights are authenticated with logging mechanism and the data integrity is achieved through accountability method. Provision of security to the users' data on the cloud will definitely encourage the data owner to outsource the data and utilize the service beneficially.

Future Enhancement

The proposed framework achieves security in designed form. When coordinating different sets of operating system along with every distributed circumstances, the data integrity in need to verify through auditing mechanism. To resolve the speed of data in uploading and downloading, AES encryption algorithm will be implemented as per analysis on computability with the framework.

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Path Coefficient and Correlation response for Yield Attributes in Pigeon Pea (Cajanas cajan L.)

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Abstract- Correlation and path coefficient analysis were completed with 84 Pigeon pea genotypes to find out association among characters and to measure the direct and indirect contribution of ten characters on grain yield. The experiment was laid out in a Randomized Complete Block Design with three replications. The genotypic correlation studies indicated that grain yield per plant exhibited stable positive association with all traits except days of pod initiation ((-0.2191). While the phenotypic correlations revealed that days to pod initiation (-0.0125), pod length (-0.0177) and100 seed yield (-0.0475) were negatively correlated with grain yield per plant. The direct effects of phenotypic path correlation coefficient analysis revealed that the grain yield per plant had positive and significant correlation with number of seed per pod (0.0789) and it was followed by number of seeds per plant (0.7493) days to 50% of flowering (0.1812) plant physical maturity (0.0876), plant height (0.091) number of seed per pod (0.0789) & number of seeds per plant (0.7493) were positive and the rest effects of few characters were negative like days to pod initiation (-0.0406), primary (-0.0305), secondary braches (-0.0409), and number of pods per plant (-0.252). Path analysis revealed high positive and direct influence of number of seed per pod with seed yield per plant followed by number of seed per pod and number of pods per plant. Moreover, it was noticed that high indirect contribution was via grain yield per pod with most of the yield components. Hence, these two parameters (number of seed per pod and number seed per plant) should be given more consideration while deciding about selection criteria of genotypes for drought resistance.

Index Terms- Pigeon pea, Yield attributes Path coefficient analysis

I. INTRODUCTION

Pigeon pea (*Cajanus cajan* L.) commonly known as gram is the fifth most important food legume crop in the world after soybean, groundnut, dry beans and peas. This crop occupies an essential place in our daily diet as very high-quality source of protein. It is mainly cultivated for its dry seeds and green vegetables in dry areas of the tropics and subtropics.The major pigeon pea producing areas in the world are India, Eastern Africa, Central and South America, the Caribbean and West Indies. India with a total area of 2, 6 million hectares and an average yield of 719 kg/ha (Sharma and Jodha, (1982) produces nearly 92% of the world's entire pigeon pea crop. To increase its yield potential, several genetic improvement methods have been employed. Seed yield being the most important and polygenically controlled complex character, is also governed by many physiological changes within the plant and influenced by many environmental factors when cultivated, hence it is not an efficient character for selection. Interrelationship among direct and indirect effect of component characters of yield is important in predicting the correlated response to direct selection and in the detection of traits with much effects as markers. The present study was undertaken to elucidate the association between yield and its attributes in pigeon pea, over drought condition.

II. MATERIAL AND METHODS

The present investigation comprised 84 genetically diverse true breeding genotypes of Pigeon pea procured from ICRISAT along with a local check for yield and other its components. The genotypes are WC-25,WC-41,WC-2,WC-32, WC-1, WC-34, WC-5, WC-20, PPE-45-2, WC-9, NALLAKANDI, WC-42, WC-39, ICP-7035, ICP-7066, WC-44, 4985-10, PRG-158, 4985-11, ICP-10531, 4985-7, WC-30, ICP-20062, 4979-2, ICP-6364, ICP-77303, ICP-7044, WC-6, WC-3, WC-16, ICP-8863, WC-24, WC-10, ICP-20036, WC-13, ICP-7532, WC-26, WC-31, WC-29, WC-11, ICP-7529, WC-8, 4985-4, WC35, ICP-13198, ICP-6682, ICP-98008, ICP-85060, WC-45, WC-21, WC-14, LRG-41, WC -17, WC-7, ASHA, WC-19, ICP-87089, WC-37, ICP-2711, PRG-100, 4985-1, 332WR, WC-18, ICP-7068, ICP-8634, ICP-909, ICP-7061, WC-36, ENT-11, WC-43, ICP-97253, ICP-6628, LAXMI, WRG-79, 4978-5, ICP-85063, WC-15, LRG-37, TTB-7, BRG-2, LRG-30, 87089, ICP-332, 4985-4. The experiment was laid out in a Randomized Complete Block Design with three replications. The following data was collected. They are days to 50% flowering, days to pod initiation, days to physical maturity, plant height at harvest (meter), Number of effective primary, secondary branches, Number of pods per plant, Number of seed per plant, pod length, Number of seeds per pod , 100 seed weight and total yield.

III. RESULTS

The simple genotypic and phenotypic correlation for the association among the characters studied for the 84 genotypes were shown in Table 1&2. The genotypic correlation of Total grain yield per plant was found to be positively correlated with days to 50% flowering (0.3862), days to physical maturity (0.4367), plant height (0.5774), and number of primary (0.1367) and secondary branches (0.0838), Number of pods per plant

(0.8053), Number of seed per plant (0.8032), pod length (0.0345), Number of seed per pod (0.0345)and 100 seed yield (0.1636). The days to pod initiation (-0.2191) was negatively correlated with Total plant yield. The phenotypic correlation of Total grain yield per plant was found to be positively correlated with days to 50% flowering (0.2214), days to physical maturity (0.2476), plant height (0.2676), and number of primary (0.0045) and secondary branches (0.0502), Number of pods per plant (0.5641), Number of seed per plant (0.5636) and Number of seed per pod (0.106) The days to pod initiation (-0.0125), pod length (-0.0177) and100 seed yield (-0.0475) was negatively correlated with Total plant yield. The phenotypic and genotypic correlation of days to 50% flowering was positively correlated with days to pod initiation, days to physical maturity, plant height, number of primary and secondary branches, Number of pods per plant, Number of seeds per plant, pod length, and Total plant yield, but negatively correlated with Number of seeds per pod and 100 seed weight. The genetic correlation of days to pod initiation was positively correlated with days to 50 % of flowering, days to maturity, primary branches, Number of pods per plant and Number of seed per plant, but negatively correlated with plant height, secondary branches, Number of seeds per pod, 100 seed weight and Total plant yield. The phenotypic correlation of days to pod initiation was positively correlated with all traits except days to maturity, Number of seeds per pod and Total plant yield. The phenotypic and genotypic correlation of days to maturity was negatively correlated with secondary branches; Number of seeds per pod, pod length and 100 seed weight, reaming all traits was positively correlated. The genetic correlation of plant height, primary and secondary branches were positively correlated with all traits, but negatively correlated with 100 seed weight. The phenotypic correlation of plant height was negatively correlated with pod length and 100 seed weight, also primary branches of phenotypic correlation was negatively correlated with Number of pods per plant and Number of seed per plant. The phenotypic and genotypic correlation of Number of pods per plant and Number of seed per plant were negatively correlated with pod length in case of genetic correlation Number of seeds per pod and 100 seed weight also negatively correlated.

The results of genotypic and phenotypic correlation coefficients were partitioned into direct and indirect effects through various yield contributing characters (Table 3). The direct effects of days to 50% of flowering (**0.1812**) plant physical maturity (**0.0876**), plant height (**0.091**) number of seed per pod (**0.0789**) & number of seeds per plant (**0.7493)** were positive and the rest effects of few characters were negative like days to pod initiation (**-0.0406**), primary (**-0.0305**), secondary braches (**-0.0409**), and number of pods per plant (**-0.252**). The highest direct effect was exhibited by number of seed per pod (**0.0789)** and it was followed by number of seeds per plant (**0.7493)**.

IV. Discussion

In the present investigation, the phenotypic correlation coefficient indicating the strong inherent association for the various traits studied pointing out the possibility of effective phenotypic selection. Total grain yield per plant per plant exhibited stable positive association with number of seed per pod followed by number of pods per plant, plant height, days to maturity and 50 percent of flowering at phenotypic levels. Thus, it can be inferred that selection based on any one of these traits either alone or in combination, would result in identifying high yielding genotypes as reported by Sharma *et al.* (1989) and Singh *et al.* (1990). When interrelationship of different characters were monitored in individual as well as combined over seasons, it was observed that number of seed per pod exhibited positive and significant association with number of pods per plant and secondary branches through indirect contribution. The plant height was positively correlated with grain yield. This was supported by Rahman et al., (1995), Spanner et al., (1996), Kumar and Kumar (1997) Gautan et al., (1999), similarly accordance with our results. Plant height positively correlated with Total plant yield. The high value of genotypic and phenotypic variation suggest that there is good scope for yield improvement through selection for pods/plant, seeds/plant and yield/plant. These findings are in agreement with other reports. (Kumar and Dubey et al., 1996).

V. Conclusion

Number of seed per pod and number of seed per plant had the maximum contribution in determining grain yield in pigeon pea. It was also observed that high indirect contribution was also exhibited via Number of seed per pod by most of the yield components and hence these two traits may be given more emphasis while selecting high yielding pigeon pea genotypes for drought condition.

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Table 1:Genotypic correlation

S.No.		50% days of Flowering	Days of pod initiation	Days to maturity	Plant Height (cm)	primary branches	secondary branches	No. of pods/ plant	No. of seeds/plant	Pod length (cm)	No. of seeds/pod	100 seed weight (gm)	Total yield
1	50% days of Flowering	1	1.3674	0.348	0.1879	0.1016	0.0049	0.3466	0.3435	0.1454	-0.0629	-0.197	0.3862
2	Days of pod initiation		1	0.0106	-0.2677	1.5742	-0.2113	0.0333	0.0297	-0.5248	-0.2558	-0.5464	-0.2191
3	Days to maturity			1	0.3024	0.0302	-0.2079	0.2993	0.2961	-0.2887	-0.0404	-0.0332	0.4367
4	Plant Height				1	0.3102	0.529	0.534	0.5375	0.5001	0.2499	-0.2983	0.5774
5	primary branches					1	0.4127	0.1151	0.1225	0.0091	0.4692	-0.01	0.1367
6	secondary branches						1	0.3403	0.3523	0.0872	0.483	-0.0408	0.0838
7	No. of pods/ plant							1	0.9999	-0.313	-0.0675	-0.2034	0.8053
8	No. of seeds/plant								1	-0.3199	-0.0602	-0.2024	0.8032
9	Pod length									1	0.7582	-0.3913	0.0345
10	No. of seeds/pod										1	0.2056	0.1636
11	100 seed weight											1	0.0215
12	Total plant yield yield												1

Table 2: Phenotypic correlation

S.No.		50% days of Flowering	Days of pod initiation	Days to maturity	Plant Height	primary branches	secondary branches	No. of pods/ plant	No. of seeds/plant	Pod length	No. of seeds/pod	100 seed weight	Total plant yield
1	50% days of Flowering	1	0.1179	0.344	0.1456	0.0739	0.0022	0.207	0.2044	0.0311	-0.0271	-0.0408	0.2214
2	Days of pod initiation		1	-0.0036	0.0079	0.1603	0.0102	0.0431	0.0432	0.0079	-0.0383	0.0169	-0.0125
3	Days to maturity			1	0.2307	0.0294	-0.13	0.1899	0.1873	-0.0655	-0.0106	-0.0034	0.2476
4	Plant Height				1	0.1701	0.2957	0.2848	0.2862	-0.0205	0.0661	-0.0107	0.2676

5	primary branches					1	0.158	-0.0108	-0.0079	0.0225	0.1107	0.0622	0.0045
6	secondary branches						1	0.1449	0.1485	0.0867	0.092	0.0373	0.0502
7	No. of pods/ plant							1	0.9995	-0.0539	0.052	0.0077	0.5641
8	No. of seeds/plant								1	-0.0541	0.052	0.0075	0.5636
9	Pod length									1	0.0929	-0.0291	-0.0177
10	No. of seeds/pod										1	-0.0371	0.106
11	100 seed weight											1	-0.0475
12	Total plant yield												1

Table 3. The Phenotypic Path correlation Coefficient of direct and indirect effects for different characters on yield and its components in Pigeon pea.

	50% days of Flowering	Days of pod initiation	Days to maturity	Plant Height (cm)	Primary branches	secondary branches	No. of pods/ plant	No. of seeds/plant	No. of seeds/pod	100 seed weight (gm)
50% days of Flowering	**0.1812**	-0.0048	0.0301	0.0133	-0.0023	-0.0001	-0.0523	0.1532	-0.0021	0.0011
Days of pod initiation	0.0214	**-0.0406**	-0.0003	0.0007	-0.0049	-0.0004	-0.0109	0.0324	-0.003	-0.0004
Days to maturity	0.0623	0.0001	**0.0876**	0.021	-0.0009	0.0053	-0.048	0.1404	-0.0008	0.0001
Plant Height	0.0264	-0.0003	0.0202	**0.091**	-0.0052	-0.0121	-0.072	0.2145	0.2052	0.0003
primary branches	0.0134	-0.0065	0.0026	0.0155	**-0.0305**	-0.0065	0.0027	-0.0059	0.0087	-0.0016
secondary branches	0.0004	-0.0004	-0.0114	0.0269	-0.0048	**-0.0409**	-0.0366	0.1113	0.0073	-0.001
No. of pods/ plant	0.0375	-0.0017	0.0166	0.0259	0.0003	-0.0059	**-0.2527**	0.7489	0.0041	-0.0002
No. of seeds/plant	0.0371	-0.0018	0.0164	0.0261	0.0002	-0.0061	-0.2525	**0.7493**	0.0041	-0.0002
No. of seeds/pod	-0.0049	0.0016	-0.0009	0.006	-0.0034	0.0038	0.0131	0.0389	**0.0789**	0.001
100 seed weight	-0.0074	-0.0007	-0.0003	-0.001	-0.0019	-0.0015	-0.002	0.0056	-0.0029	**-0.026**

A Novel Approach towards Building Automation through Dali-WSN Integration

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Abstract- The building automation systems (BAS) include the monitoring and control of building services and networks. Since different manufacturers usually deal with one aspect of building automation such as heating ventilation and air conditioning, lighting control, different kinds of alarms, etc, the Final building automation system has different subsystems which are finally taken to an integrated building management system. The development of a prototype to be used in a wireless sensor network (WSN) which also integrates Digital addressable lighting interface (DALI) protocol can provide a solution for centralization of building automation services. Since DALI is a well-established standard and it has been adopted by major electronic ballasts' suppliers it is very easy to find DALI compliant devices. Despite it is designed for lighting control, DALI has also been adapted to other applications, such as motor or fan controllers, proximity alarms, etc. Adapting the standard to a WSN allows integrating DALI devices as a part of the WSN, expanding the traditional DALI bus and removing wires, which results in a reduction of installation costs. For more efficient energy utilization of wireless sensor network, the IEEE 802.15.4 standard is modified to a deployable energy efficient 802.15.4 MAC protocol (DEEP).

Index Terms- Building Automation System, Wireless Sensor Network, Digital addressable lighting interface, IEEE 802.15.4, Deployable energy efficient 802.15.4 MAC protocol.

I. INTRODUCTION

The building automation system (BAS) had gained great amount of attention in recent years. BAS are implemented to monitor the parameters defining living and working conditions within a building, and control its electrical equipment. BAS were initially developed to control heating, ventilation and air conditioning (HVAC) systems. Through time we have gone through several kinds of controllers, e.g. pneumatics, analog circuits, microprocessors, etc. Even though other home systems like lighting, security should also use automation, they are usually installed in a different system than HVAC. This division of the two subsystems increases the end consumer cost due to additional investment in communication hardware and software for integrating HVAC and lighting in a single control point.

Different manufacturers are concerned about only one aspect of building automation and hence the final building automation system has different subsystems which are finally taken to an integrated building management system. The need of a centralized monitoring control centre makes necessary the integration of all BA applications. However, the new and exciting opportunities to increase the connectivity of devices within the building for the purpose of building automation remains largely unexploited.The WSN-DALI integration can provide a solution for centralization of building automation services.The digital addressing capability of the DALI is being used for this purpose.

II. MATERIALS AND METHOD

1. DALI

A digitally addressable lighting interface (DALI) network consists of a controller and one or more lighting devices that have DALI interfaces. The controller can monitor and control each light by means of a bi-directional data exchange. The DALI protocol permits devices to be individually addressed and it also incorporates Group and Scene broadcast messages to simultaneously address multiple devices. Each lighting device is assigned a unique static address in the numeric range 0 to 63, making possible up to 64 devices in a standalone system. Alternatively, DALI can be used as a subsystem via DALI gateways to address more than 64 devices. Data is transferred between controller and devices by means of an asynchronous, half-duplex, serial protocol over a two-wire differential bus, with a fixed data transfer rate of 1200 bit/s.

DALI requires a single pair of wires to form the bus for communication to all devices on a single DALI network. The network can be arranged in a bus or star topology, or a combination of these. The DALI System is not classified as separated extra low voltage and therefore may be run next to the mains cables or within a multi-core cable that includes mains power. The DALI data is transmitted using Manchester encoding and has a high signal to noise ratio which enables reliable communications in the presence of a large amount of electrical noise.

Earlier generations of DALI devices stored configuration data in EEPROM, which was problematic due to the limited number of write cycles supported by EEPROMs. In current generations of DALI devices, RAM is used in preference to EEPROM during normal operation, which significantly reduces the number of EEPROM writes and thus extends their lifetimes. This use of RAM, however, is patented and therefore mandates payment of a license fee. DALI is a dedicated lighting control interface. It uses differential pair with fixed baud of 1200bps. Each DALI bus can address up to 64 devices. And when configured as a subsystem of the whole building management system, more devices can be

controlled. DALI is based on Digital Signal Interface (DSI), which is also a lighting control interface.

2. WIRELESS SENSOR NETWORK

A wireless sensor network (WSN) consist of autonomous sensors that are distributed to monitor physical or environmental conditions such as temperature, sound, pressure, etc. and to cooperatively pass their data through the network to a main location. The more modern networks are bi-directional, also enabling control of sensor activity. The development of wireless sensor networks was motivated by military applications such as battlefield surveillance; today such networks are used in many industrial and consumer applications, such as industrial process monitoring and control, machine health monitoring, and so on.

The WSN is built of nodes – from a few to several hundreds or even thousands, where each node is connected to one or sometimes several sensors. Each such sensor network node has typically several parts: a radio transceiver with an internal antenna or connection to an external antenna, a microcontroller, an electronic circuit for interfacing with the sensors and an energy source, usually a battery or an embedded form of energy harvesting.

A sensor node might vary in size from that of a shoebox down to the size of a grain of dust, although functioning "motes" of genuine microscopic dimensions have yet to be created. The cost of sensor nodes is similarly variable, ranging from a few to hundreds of dollars, depending on the complexity of the individual sensor nodes. Size and cost constraints on sensor nodes result in corresponding constraints on resources such as energy, memory, computational speed and communications bandwidth. The topology of the WSNs can vary from a simple star network to an advanced multi-hop wireless mesh network. The propagation technique between the hops of the network can be routing or flooding.

To reduce communication costs, some algorithms that remove or reduce nodes redundant sensor information and avoiding forwarding data is of no use. As nodes can inspect the data they forward, they can measure averages or directionality of readings from other nodes. For example, in sensing and monitoring applications, it is generally the case that neighboring sensor nodes monitoring an environmental feature typically register similar values. This kind of data redundancy due to the spatial correlation between sensor observations inspires the techniques for network data aggregation and mining.

Several standards are currently either ratified or under development by organizations including WAVE2M for wireless sensor networks. There are a number of standardization bodies in the field of WSNs. The IEEE focuses on the physical and MAC layers; the Internet Engineering Task Force works on layers 3 and above. In addition to these, bodies such as the International Society of Automation provide vertical solutions, covering all protocol layers. Finally, there are also several non-standard, proprietary mechanisms and specifications. Standards are used far less in WSNs than in other computing systems which make most systems incapable of direct communication between different systems. However predominant standards commonly used in WSN communications include wireless HART , IEEE1451, ZigBee / 802.15.4, ZigBee IP and 6LoWPAN. Here we are utilizing 802.15.4 WSN.

3. TOOL FOR MODEL-BASED DESIGN AND SIMULATION

Simulink provides a graphical user interface (GUI) for building models as block diagrams, allowing to draw models as would with pencil and paper. Simulink also includes a comprehensive block library of sinks, sources, linear and nonlinear components, and connectors. The interactive graphical environment simplifies the modeling process, eliminating the need to formulate differential and difference equations in a language or program. Models are hierarchical, so can build models using both top-down and bottom-up approaches.

After defining a model, you can simulate its dynamic behavior using a choice of mathematical integration methods, either from the Simulink menus or by entering commands in the MATLAB Command Window. The menus are convenient for interactive work, while the command line is useful for running a batch of simulations. Using scopes and other display blocks, see the simulation results while the simulation runs, then change parameters and see what happens. The simulation results can be put in the MATLAB workspace for post processing and visualization.

III. SYSTEM MODELLING

Despite being designed for lighting control, DALI can also adapted to other applications, such as motor or fan controllers, proximity alarms, etc. This possibility is being utilized for extending the integration of DALI and wireless sensor network to a centralized building automation system. The systems such as heating ventilation and air-conditioning (HVAC), security systems are hence included along with lighting system to extend the capability of DALI. The lighting control is implemented using human sensors and an led light indicator. The heating ventilation and air conditioning (HVAC) system is simplified and implemented using temperature sensors, AC motor control system and power displays for fan control and a humidity sensor, DC heater and power display for heater control. The efficient usage of WSN is being enhanced with deployable energy efficient MAC protocol (DEEP). DALI is based upon the master-slave principle. The implementation of a DALI master controller is using an IEEE 802.15.4-based wireless sensor network (WSN). The Nodes which compose the wireless sensor network (WSN) have a microcontroller unit (MCU) and an IEEE 802.15.4-compliant transceiver. The DALI address generation takesplace here. The slave section is composed of the DALI address comparison and device control section. The automation of the devices with in our concerned system is being done here. The full system is being modeled with this concept.

The implementation of deployable energy efficient MAC protocol is based on request, acknowledge concept. This concept is implemented in heater system which represents high power devices which are controlled using request and acknowledge

system. The system is mainly implemented in two section, the DALI master control section where digital address generation and transmission takesplace and a slave section where the DALI address comparison and device control section lies.

1. DALI-WSN MASTER CONTROL SECTION

The DALI master control section consists of wireless sensor network (WSN). Nodes which compose the wireless sensor network (WSN) have a microcontroller unit (MCU) and an IEEE 802.15.4 compliant transceiver.

For lighting control, we make use of the human sensors in three rooms, when there is human entry indicated in the human sensor of the concerned room, the light in the concerned room is on. For fan control, we make use of temperature sensors in two rooms, when there is indication of high temperature, the fan runs at high speed else at low speed. The security system, we make use of an intrusion sensor, when there is an indication of an intruder, the alarm indicated by an led light is on. The heater system make use of humidity sensor and works upon DEEP protocol, ie, upon the request acknowledgement principle.

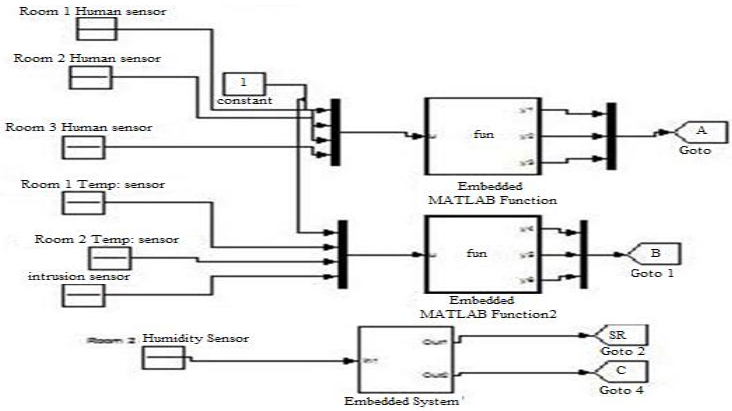

Figure 1: DALI-WSN master control section

2. DALI-WSN SLAVE SECTION

The DALI slave section composes of the address comparison and device control mechanisms. The controls of the devices are done according to the digital address being send by the master section. If the correct digital address of the device is being received, the particular device is being activated henceforth.

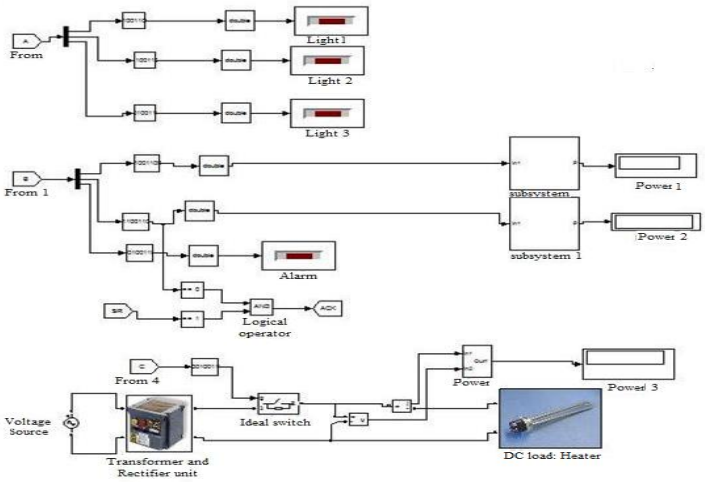

Figure 2: DALI-WSN slave section

For the lighting control, if the correct digital address of the lights of any three rooms is being received, then the light for the corresponding rooms are kept on, otherwise the lights are always at the off condition. Hence it works in such a way that whenever the presence of human in room is detected, the digital address of that particular room is send by the wireless transceiver and upon the reception of the digital address by the slave section, the light is kept on. For the alarm system too same procedure is being done. As such upon the indication of an intruder in a room by the intrusion sensor, the alarm in the respective room is on. This is upon the reception of the particular digital address of the alarm device send by the wireless transceiver.

IV. RESULTS AND DISCUSSION

DALI, when alone used for lighting control in building automation can be represented as shown in fig.5.1.1. When the input to the human sensor of a particular room is high, then the light in that room will be on, else it will be off. In Fig. 3, the human sensor inputs of room1 and room3 is high and that of room 2 is low, hence the lights in room1 and room3 is on and that of room2 is off. Whenever the human entry in a particular room is detected then the lights in that particular room are kept on.

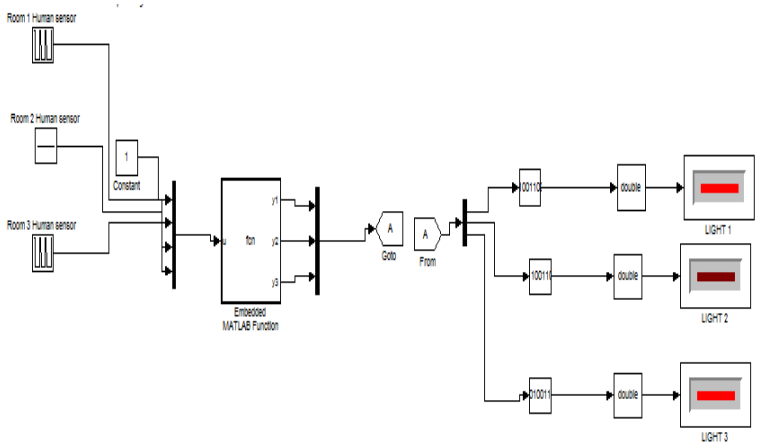

Figure 3: Simulation diagram of building lighting automation

Although DALI was designed for lighting purposes, the up and down commands or direct levels can be also used for setting a fan speed , detection of intruder in a room etc .They can be implemented as shown in Fig. 4. The room1 and room2 is having fan controllers and an intrusion sensor is included to add the features of security system in buildings. The inputs are provided to the sensors so that room1 temperature sensor senses a low temperature, room2 temperature sensor senses a high temperature, and the intrusion sensor senses the presence of an intruder. Hence we obtain the simulation result as for room1 fan runs at low speed consuming less power, room2 fan runs at high speed consuming more power, and intrusion alarm is on.

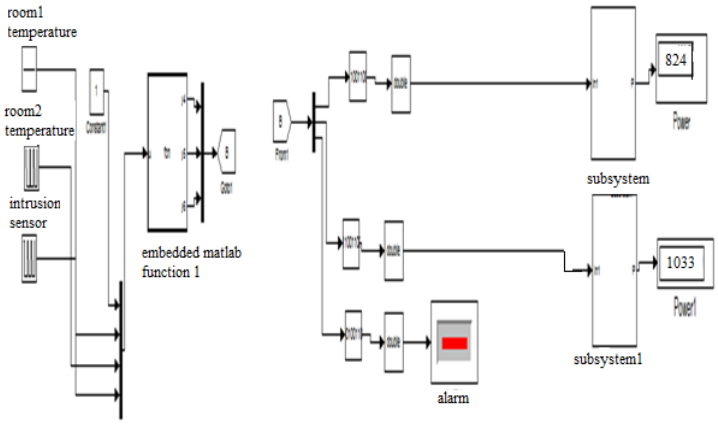

Figure 4: Simulation diagram of fan controller and security system in BAS

For high power consuming devices such as room heater systems, the DEEP protocol can be implemented for the purpose of conformation of the efficiency of the wireless sensors. This is implemented in Fig. 5. Even though the room2 humidity sensor senses a high humidity, there is again made a conformation with the room2 temperature sensor. Since the temperature sensor senses high itself, the heater is kept off and room2 fan still runs at high speed consuming high power. The room1 temperature is sensed low; hence it runs at low speed consuming less power. The intrusion sensor senses a low and hence intrusion alarm is off.

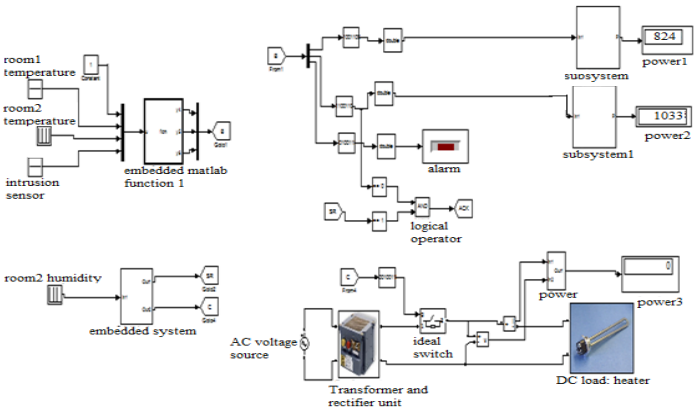

Figure 5: Simulation diagram of BAS implementing DEEP

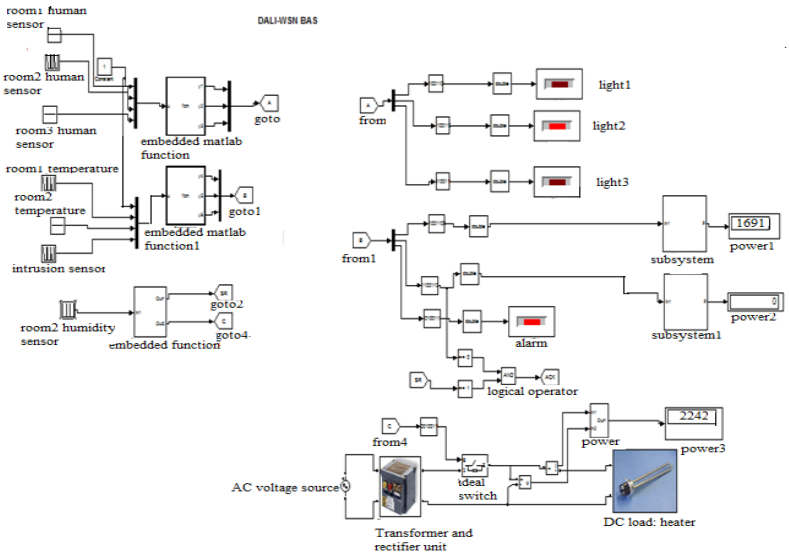

Figure 6: Simulation diagram of final BAS with DALI-WSN integration

The final building automation takes into account all the above systems and integrates them together into a single building automation system. It is implemented as shown in Fig.6. The room1 and room3 human sensors are kept low, and room2 human sensor at high, hence the room2 light is only on. The intrusion sensor senses high and so it is on. The temperature sensor for room1 is high and for room 2 is sensed low. The humidity sensor for room2 senses high. Incorporating them, the room1 fan runs at high speed consuming more power, room2 fan is off and the room2 heater is kept on. Hence the full building automation can be controlled through the integration of DALI with wireless sensor network (WSN).

V. CONCLUSION

DALI was initially introduced for lighting control alone for the purpose of building automation. But the capability that it can also be adapted to other applications, such as motor or fan controllers, proximity alarms, etc is being utilized here. Beneath that expanding the traditional DALI bus and removing wires, results in a reduction of installation costs. Adapting a standard wireless sensor network allows integrating DALI devices as a part of the WSN .Hence opted the IEEE 802.15.4wireless sensor network.

DALI is based upon the master-slave principle. Implementing a DALI master controller is using an IEEE 802.15.4-based wireless sensor network. Nodes which compose the wireless sensor network have a microcontroller unit (MCU) and IEEE 802.15.4-compliant transceiver. DALI-WSN integration is done through the digital addressing of devices. The analysis of the concept of WSN-DALI integration is done with the powerful simulation modeling capabilities of Matlab /Simulink. The different subsystem functions are connected through dedicated Matlab functional blocks through codings. The utility of digital addressing capability of DALI makes the integration of blocks

together and for the purpose of centralized building automation. Direct comparison with digital address of the devices enables the centralized integration of building automation. Besides the lighting control, the heating ventilation and air-conditioning (HVAC) and security systems are also being included for the purpose of centralized building automation.

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Effect of nano-ZnO in lowering yellowing of aliphatic amine-cured DGEBA-based epoxy coatings on UV exposure

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Abstract- Epoxy resins as binders are best known in protective coating field for their excellent corrosion resistance. However, the exterior durability of these resins is severely affected by their poor weathering resistance. The exposure of epoxies to outdoor environment causes chemical reactions within them, which leads to deterioration of their useful properties. A few hours of UV-exposure leads to chalking and discoloration of epoxies, caused as a result of photo-degradation. The present work aims to study the extent to which the color change and yellowing of epoxy caused due to weathering of an aliphatic amine namely, Diethylenetriamine (DETA) cured diglycidyl-ether of bisphenol-A (DGEBA) based epoxy system, when exposed to accelerated weathering conditions, could be lowered by means of a UV-blocking additive, namely nano Zinc oxide (ZnO). A titanium dioxide (TiO_2) based white coating with DGEBA epoxy and DETA was formulated and applied on mild steel (MS) panels for the weathering study. The coated panels were exposed in an UVB-weatherometer. Nano zinc oxide particles were synthesized by chemical route and characterized for size, shape and structure. At 2% (by weight) of total coating formulation, nano-ZnO was found to be an effective UV-stabilizer and lowered the degradation in the epoxy-hardener cured system on UVB exposure. The enhanced resistance to color change and yellowing was indicated by lowered color change (dE) and yellowness index values (YI) for the coatings on UVB exposure.
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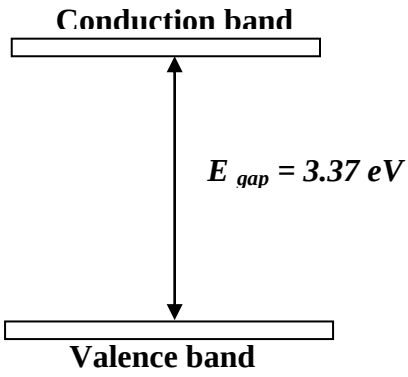
Index Terms- Epoxy, nano ZnO, yellowing, UV exposure

I. INTRODUCTION

Among the different classes of polymers, epoxy resins are known for their high range of attainable properties and versatility. Depending on the chemical structure of curing agents, and curing conditions, it is possible to attain: toughness, chemical resistance, flexibility, high strength and hardness, good heat and electrical resistance. However, inherently epoxies have poor weathering resistance owing to the presence of an aromatic moiety in their polymeric backbone structure. It is this aromatic moiety which absorbs at 300 nm and degrades in the presence of UV light and humidity. This consequently results in discoloration, yellowing and chalking [1, 2].

The best approach to stabilization with respect to weathering degradation in polymers would be to not allow radiation to be absorbed by vulnerable groups and these preventive techniques comprises of screening and absorption. The effectiveness of the stabilizers against weathering depends on solubility, ability to stabilize in different polymer matrix, the distribution in matrix, evaporation loss during processing and use. There are both organic and inorganic UV blockers. The emerging awareness of toxicity of organic UV-stabilizers like HALS and other organic-based UV-scavengers have initiated the need for, making the best use of alternate inorganic metallic oxides like zinc oxide and titania as UV-stabilizers. These oxides basically protect the polymer matrix from degradation by acting as effective UV-blockers.

Zinc oxide (ZnO) ultrafine particles are well known as UV blocking materials, making them feasible to be widely used in polymers, fabrics, woods and cosmetic materials [3]. With the accelerated development of nano-particle technology, nano ZnO has the ability to offer UV protection to coatings and underlying substrates while also being transparent in the visible spectrum. Also, dispersing nano-particles, instead of larger particles, allows a coating formulator to increase the interfacial content in the coating matrix significantly which in turn imparts highly specialized functionality to surface coatings [4, 5]. Nano ZnO particles thus have become an option as formulating materials for UV protection. Also, since they are inorganic and particulate, they have added advantages of being stable and non-migratory within an applied coating, offering better effectiveness and a longer service life. A distinguishing feature of the nano ZnO containing composite coating is a drop in transmission around 370 nm, which corresponds to the absorption of UV radiation by nano ZnO particles (E_{gap} = 368 nm) and is indicative of the UV protecting ability of the coating [6, 7].



Band gap in Zinc oxide

The present work studies the effect of nano-sized zinc oxide particles on lowering the color change and yellowing of an aliphatic amine cured epoxy system on UVB exposure at different loading concentrations. As epoxy coatings deteriorate rapidly and loss their aesthetic properties with no time under UV exposure, the only aim of this study was to observe, to what extent, a simple TiO_2 based, white DGEBA epoxy cured with an aliphatic amine (DETA) can withstand yellowing or color change with and without nano ZnO under UV (B) light.

II. EXPERIMENTAL

Materials

Diglycidyl ether of bisphenol A (DGEBA) epoxy with EEW of 185, DER 331 was obtained from DOW Chemicals. An amine-based hardener namely; Diethylenetriamine (DETA) was used for curing of epoxy. Zinc sulfate heptahydrate ($ZnSO_4 \cdot 7H_2O$), aqueous ammonia solution (NH_4OH), ammonium bicarbonate (NH_4HCO_3) and anhydrous alcohol were obtained from Sigma-Aldrich for the synthesis of nano ZnO particles. BYK additives were used for the formulation of the TiO_2 based epoxy coating.

Synthesis of nano-zinc oxide

The first stage involved synthesis of the precursor, zinc carbonate hydroxide (ZCH) using 1:10:1 ratio (volume ratio) of aqueous ammonia (7 mol/L): $ZnSO_4 \cdot 7H_2O$ solution (0.5 mol/L): NH_4HCO_3 (3 mol/L). Aqueous ammonia followed by NH_4HCO_3 solution, both, were drop-wise added to a continuously stirred solution of $ZnSO_4 \cdot 7H_2O$ at room temperature. The reaction mixture was then heated up to 60^0C and stirred for 30 min. The ZCH precipitate was then filtered and washed with deionized water to ensure complete removal of sulphate ions (SO_4^{2-}). The ZCH precursor was dried and calcined at 400^0C for two hours to obtain nano sized zinc oxide particles.

The synthesized nano-powders were investigated by Transmission Electron Microscopy [TEM, Philips CM200 electron microscope], Scanning Electron Microscopy (SEM, Model no.S3400, Hitachi) and X-ray diffraction [XRD, X'Pert Pro Philips] for its shape, size and crystallinity. FTIR [JASCO FTIR 6100] studies of nano-powders were carried out to study the structure of the oxides.

Preparation of epoxy coated samples

The mild steel (MS) samples to be coated were de-greased, cleaned and then roughened mechanically with abrasive paper (emery paper grade no.: 100). A TiO_2 based white coating was formulated with 35% pigment concentration, 50% epoxy resin concentration and rest with the appropriate concentrations of additives and xylene as solvent. The nano ZnO modified epoxy coatings were prepared by addition of firstly the micron sized TiO_2 pigment along with the required amount of nano ZnO at different loading levels, maintaining the 35% pigment concentration in the coating followed by the additives. Nano zinc oxide was added to the epoxy resin using ultra-sonnication at 1%, 2% and 5% (by weight) of the total coating formulation. The only purpose to take TiO_2 as the pigment was to formulate a white coating and hence evaluate discoloration/ yellowness of the coating on weathering. The pigment concentration was always maintained at 35% with and without nano ZnO while formulating the coatings. The coating formulation for different loading of nano ZnO in the coating is as tabulated as follows

Component	Neat Epoxy (%)	1% nano ZnO + Epoxy (%)	2% nano ZnO + Epoxy (%)	5% nano ZnO + Epoxy (%)
TiO_2 pigment	35.0	34	33	30
Nano ZnO	0	1	2	5
Resin	50.0	50.0	50.0	50.0
BYK 530	2.0	2.0	2.0	2.0
BYK 333	1.0	1.0	1.0	1.0
BYK 9076	1.5	1.5	1.5	1.5
BYK 320	0.5	0.5	0.5	0.5
Xylene	10	10	10	10

As mentioned in the table above, the resin content was kept constant throughout in the coatings, with different nano ZnO concentrations. The resulting blend was ultrasonnicated for 20 minutes and the shear time via ultrasonication was kept constant for all loadings of nano ZnO. After mixing TiO_2 pigment, nano ZnO and additives, at the end, hardener and solvent were mixed to obtain the nano ZnO modified epoxy coatings. The formulated coating was applied using brush on the surface treated MS panels so as to achieve uniformly coated panels with good finish and then allowed to hard cure.

Weathering Test

The coated MS panels were subjected to accelerated weathering in a UV weatherometer [QUV Weatherometer, Q-Lab Products & Services] equipped with UVB-313 nm lamps. The test cycle in UVB-weatherometer comprised of 4 hours UVB simulation at 60^0C followed by 4 h of condensation (UVB-lights off during condensation) at 50^0C as per accordance of ASTM G-154. The epoxy coated MS panels were exposed to two test cycles, i.e, 16 hours in the weatherometer.

The coatings exposed in UVB weatherometer were characterized for color change (dE) and yellowness index (YI) using a spectrometer (BYK-Gardener Spectrometer) equipped with Color-Lab Quality Control software. The data were reported on L, a, b scales and overall color difference was given in the following equation [1]:

$$\mathbf{dE = [\Delta L^2 + \Delta a^2 + \Delta b^2]^{1/2}}$$

where; $\Delta L = L_2 - L_1$, $\Delta a = a_2 - a_1$, and $\Delta b = b_2 - b_1$

Respective numbers 1 and 2 are denoted to samples before and after the exposure test.

The yellowness index was measured as per ASTM D 1925, formulated as:

$$\mathbf{YI = 100\ (1.28\ X_{CIE} - 1.06\ Z_{CIE}) / Y_{CIE}}$$

As per ASTM D 1925, the conditions for measurement are as follows: Illuminant: C, Standard Observer function: 2^0, denoted as C/2^0. X, Y and Z are the tristimulus values and 1.28 and 1.06 are the coefficients as per C/2^0 conditions.

The surface morphology of the coatings was studied using SEM microscopy [SEM, Model no.S3400, Hitachi] to look for the distribution of nano zinc oxide at different loading levels in the coatings.

III. RESULTS

Characterization of nano-ZnO

FTIR analysis

The FTIR spectrum for synthesized zinc oxide is shown in figure.1. The characteristic Zn – O peaks were observed at around 468 and 480 cm^{-1}, confirming the zinc oxide structure [6, 8]

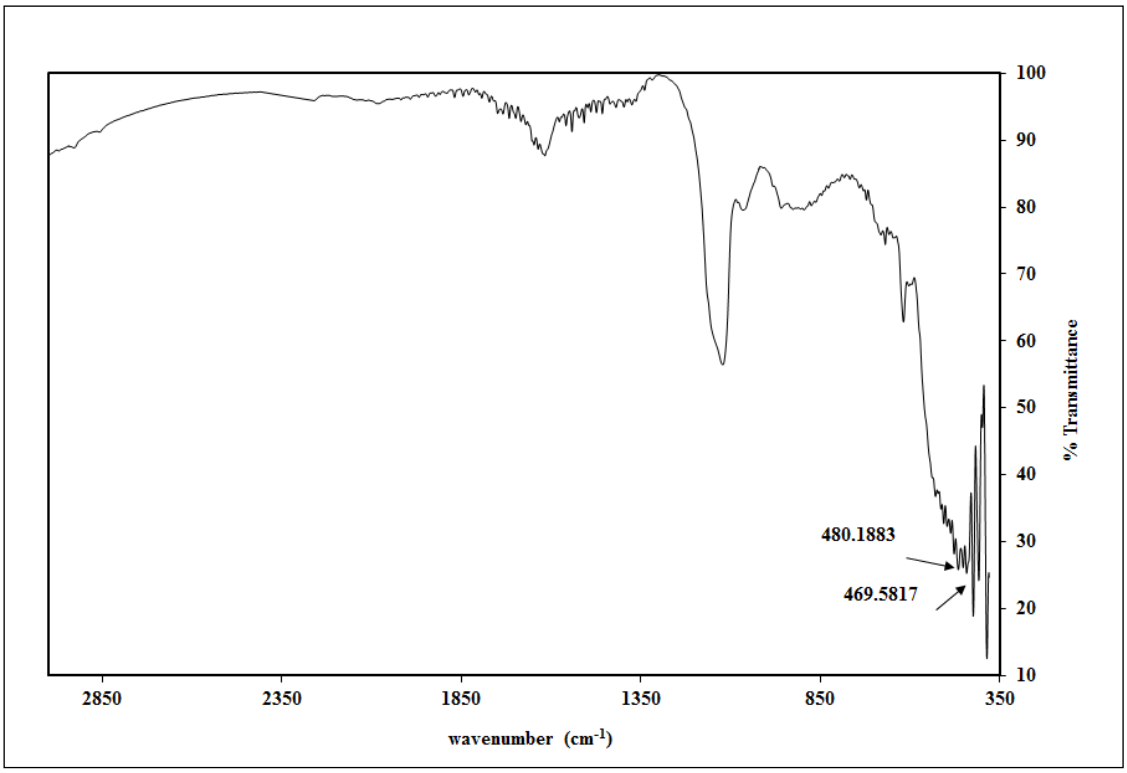


Figure.1. FTIR spectrum of nano zinc oxide

XRD analysis

The XRD spectrum for the synthesized zinc oxide particles is shown in figure.2. The XRD analysis identified well indexed diffraction peaks in good agreement with those of a wurtzite hexagonal structure of ZnO (ICSD Reference code 01-075-0576) [9-11].

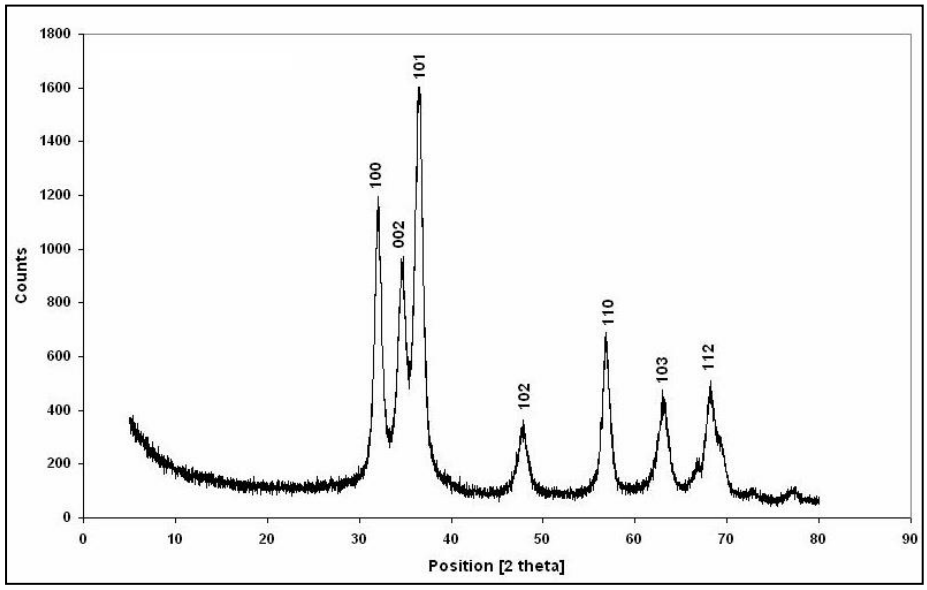


Figure.2. XRD spectrum of nano ZnO

SEM & TEM analysis

The SEM micrographs showed the zinc oxide particles having flake-like morphology (figure.3). TEM results confirmed that the flake-like particles were composed of small spherical particles with diameter ranging from 7 to 20 nm (figure.4.a). The nano size of synthesized zinc oxide particles was thereby confirmed from TEM analysis. The lattice fringes of synthesized zinc oxide were clearly

identified in the HR-TEM image as shown in figure.4.b. The inter-planar spacing i.e. d-spacing was obtained as 0.29 nm which closely matches to the d-spacing of 100 plane (0.281nm) of zinc oxide [12-14].

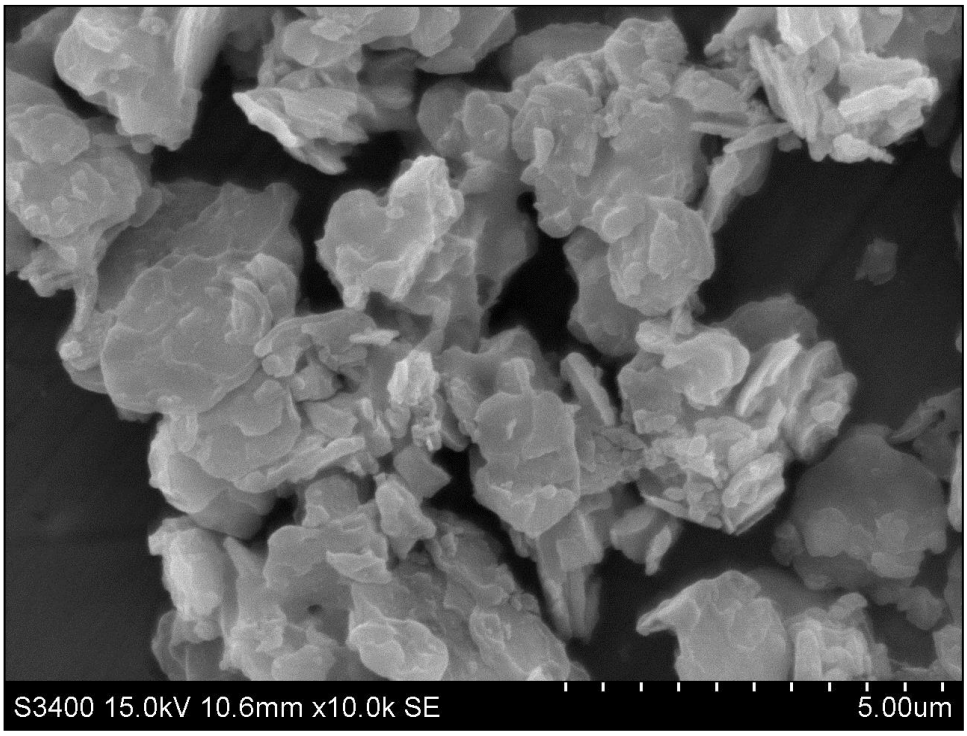


Figure.3. SEM micrograph of nano ZnO

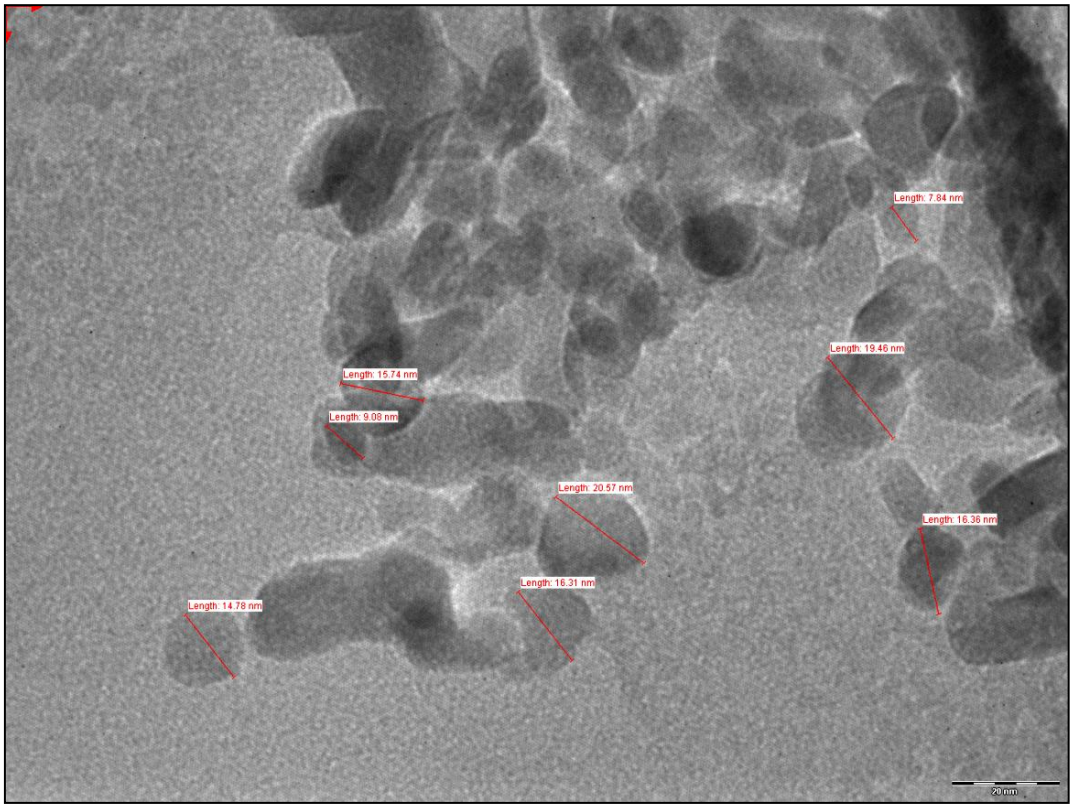


Figure.4. (a) TEM image of nano ZnO

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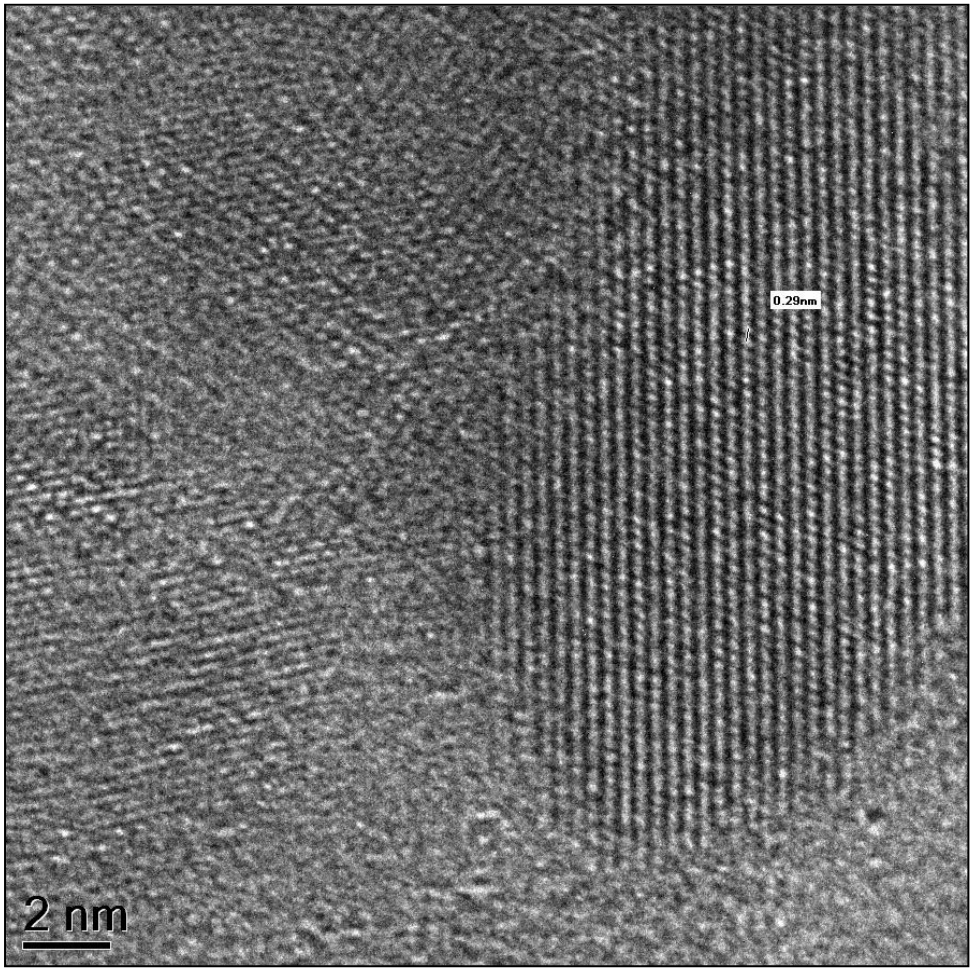


Figure.4. (b) HR-TEM image of nano ZnO

Characterization of epoxy coated panels

Color change (dE) and Yellowness Index (YI) measurements of UVB-weathered epoxy coated panels

The epoxy coating formulated with and without nano-ZnO, was applied on 0.7 mm thick MS panels by brush. The coating thickness was determined using a thickness measuring gauge [Elcometer DFT Gauge] and the coating thickness for the coating systems were found to be in the range of 85-100 microns. The dE and YI results were evaluated after a mere 16 hours of UV (B) exposure. i.e, two cycles of UV (B) weathering were each cycle of 8 hours comprises of 4 hour UV (B) light followed by 4 hour condensation. The dE and d [YI] values for the DETA cured epoxy coatings for the neat and nano ZnO modified systems are as shown in figure.5.a-5.b. The discoloration observed in epoxies on weathering is attributed to the formation of a quinone- methide structure and has been confirmed in many cited studies by both infrared and UV-Vis analysis [15]. The nano ZnO incorporated coatings show lowered color change and yellowing than the neat epoxy system, however, the minimal color change and yellowing on UVB exposure was exhibited by the system incorporated with 2% nano-ZnO comparative to 1% and 5% loading levels. The color change lowered by 10 units and yellowing lowered by eight units at 2% loading level of nano ZnO in the coating.

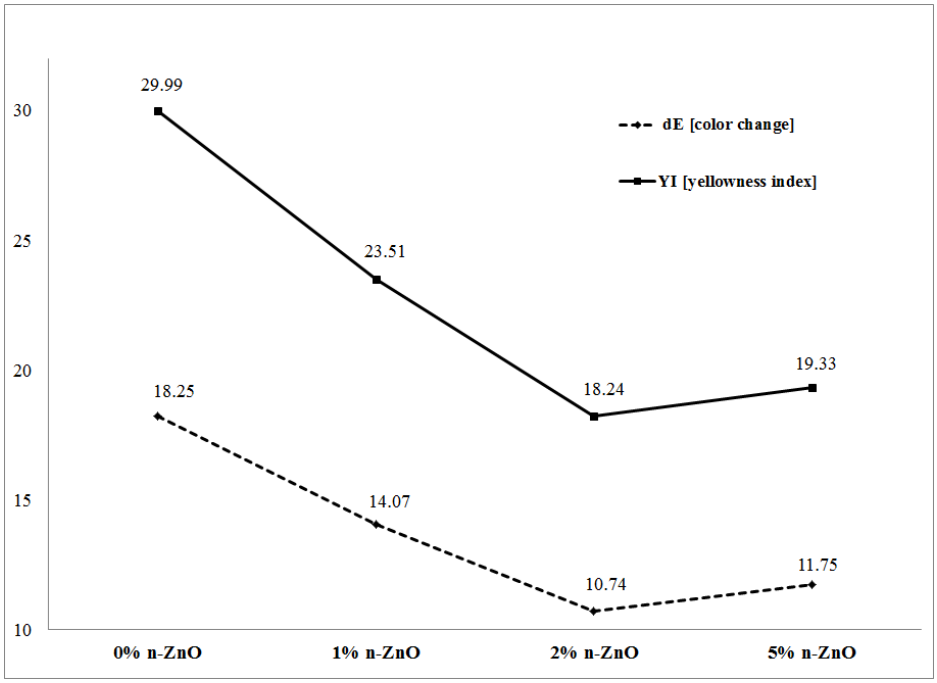


Figure.5.a. Color change (dE) and Yellowness index (YI) plot after 2 UVB test cycles (16 hours)

The extent of weathering resistance, offered by the nano ZnO modified epoxy coating systems were determined by evaluating the percentage reductions in color change (dE) and yellowing (YI) during the exposure. The percentage reduction values were calculated as follows: Suppose after 2 UVB cycles in weatherometer,
Neat Epoxy = A
Modified Epoxy = B
Where A and B are dE (or YI value) after the weathering cycle, then;
% Reduction in dE (or YI) = [(A-B) / A] x 100
The percentage reduction values in figure. 5.b are relative to the dE and YI value for the neat epoxy coating without nano ZnO i.e, 0% nano ZnO and hence don't have the column bar for 0%.

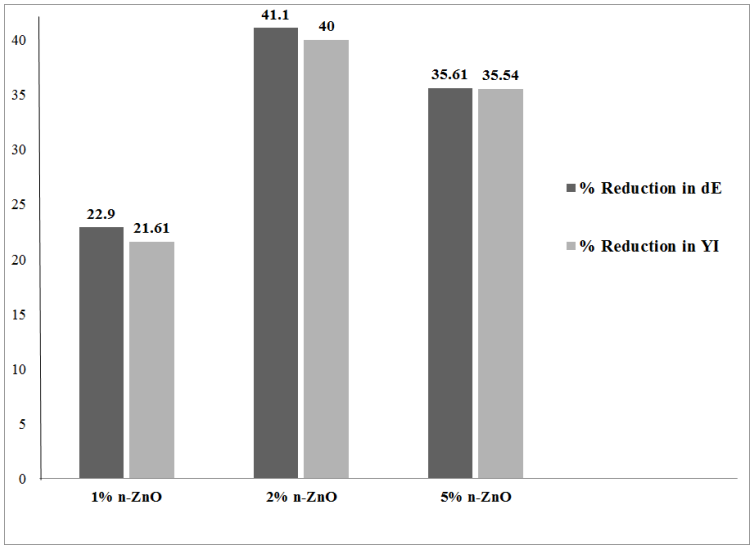


Figure.5.b. Color change [dE] and Yellowness index [YI] percentage reduction plot after 2 UVB test cycles (16 hours)
Table.1 Color change [dE] and Yellowness index [YI] results after 2 UVB test cycles (16 hours)

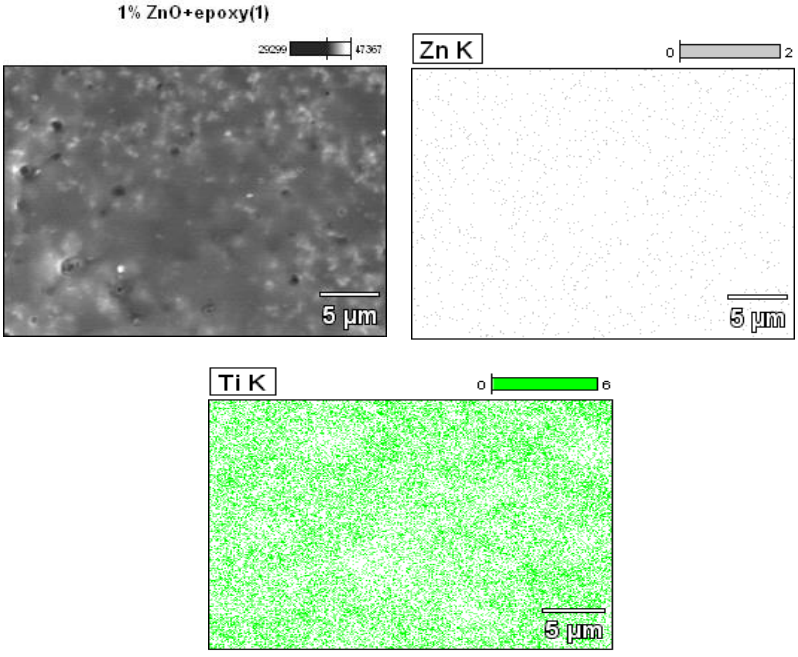
System	dE	YI	% Reduction in dE	% Reduction in YI
Neat Epoxy	18.25	29.99		
1% nano ZnO + epoxy	14.07	23.51	22.9%	21.61%
2% nano ZnO + epoxy	10.74	18.24	41.1%	40%
5% nano ZnO + epoxy	11.75	19.33	35.61%	35.54%

Both the lower (1%) and higher concentration (5%) concentration of nano zinc oxide does not provide efficient weathering resistance compared to the 2% loading level. As observed from the spectrophotometric results, the dE and YI values for 2% and 5% loading of nano ZnO in the epoxy coating showed a difference of only a single unit. The undesired performance at higher concentration of nano ZnO is attributed to the agglomeration of nano ZnO particles in the coating matrix, while at 1% loading level; the concentration is just not enough to impart appreciable weathering resistance to the coating. At 2% loading level of nano-ZnO, color change and yellowing was lowered by 40% compared to the neat epoxy system and thus proved to be the optimized loading level in the coating for effective weathering resistance.

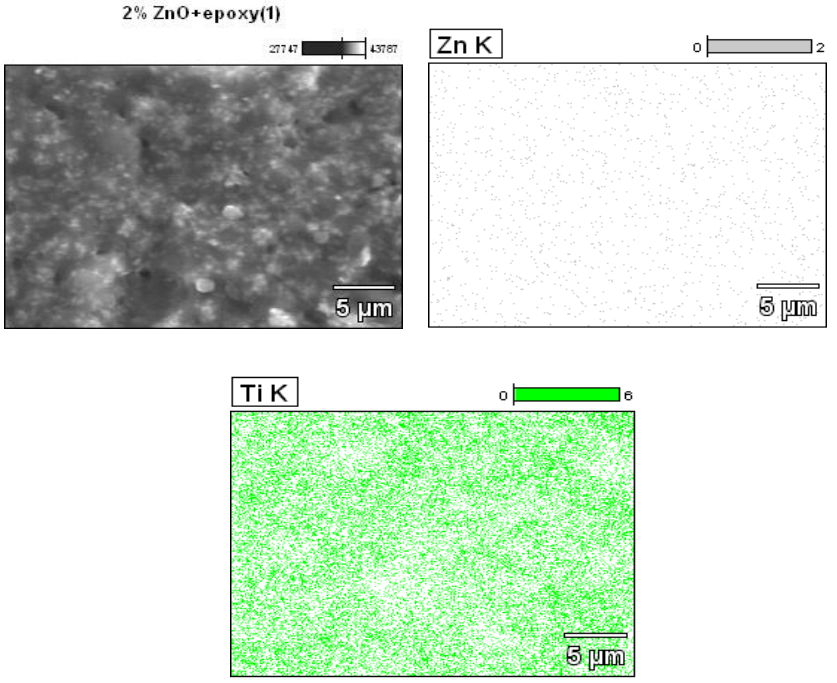
The enhanced weathering resistance for the nano-ZnO incorporated coating is due to UV-absorption property of nano ZnO owing to its large band gap energy. Zinc oxide has a band-gap at around 3.37 eV corresponding to 376 nm and thus it absorbs light that matches or exceeds this band gap energy. UV-range of solar spectrum lies within this range and thus UV light gets absorbed by zinc oxide particles. The large band gap energy inherent of ZnO and the size of particles in nano-dimensions together, effectively imparts resistance to yellowing to the epoxy system under study at the optimized loading level of 2% (by weight of epoxy resin).

SEM and EDX analysis

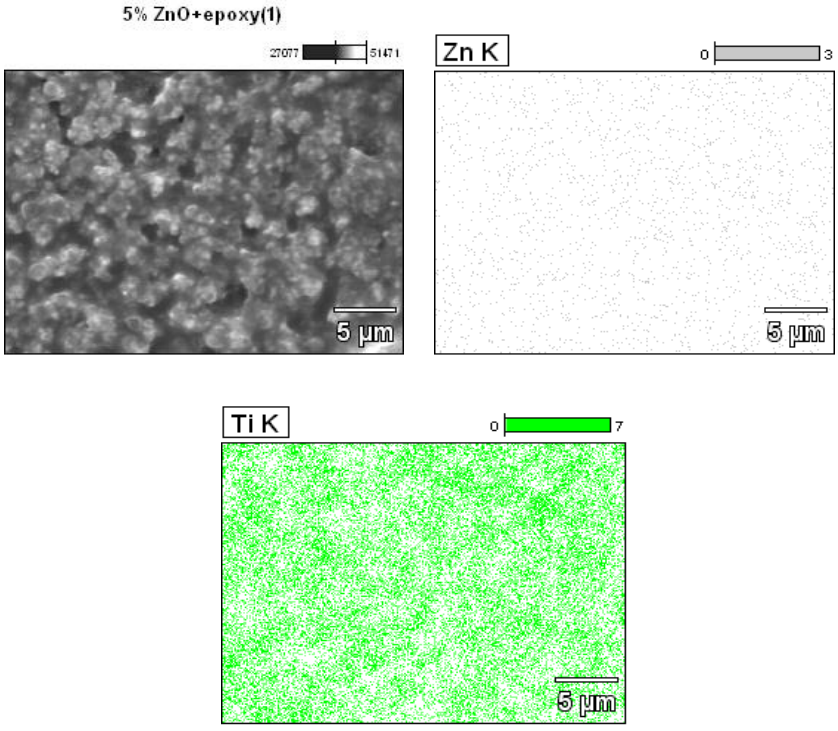
The SEM micrographs and EDX mapping images of the epoxy coating at 1%, 2% and 5% loading levels is as shown in figure.6.a -6.c



(a)



(b)



(c)

Figure.6. SEM micrographs and EDX mapping images of epoxy coating incorporated with (a) 1% nano ZnO (b) 2% nano ZnO (c) 5% nano ZnO

The elemental analysis of Zn (due to ZnO) and Ti (due to TiO_2) was carried by EDAX technique and results are shown in table.2.

Table.2. EDX elemental analysis

System	Elements (Weight %)	
	Ti	Zn
1% n-ZnO + Epoxy	58.89	1.98
2% n-ZnO + Epoxy	54.86	3.87
5% n-ZnO + Epoxy	52.88	7.46

At 1% and 2% (by weight) loading levels of nano-ZnO the distribution is fairly uniform with no presence of agglomerates. However at 5% loading level, the nano particles get agglomerated leading to a very non-uniform inhomogeneous distribution in the epoxy matrix. The agglomeration sites are more pre-dominant in 5% nano ZnO coating followed by the epoxy coating with 2% nano ZnO and least in 1% nano ZnO incorporated epoxy coating. And as mentioned in the coating formulation table, TiO_2 content is lesser in 5% nano ZnO incorporated coating than in 2% and 1% nano ZnO incorporated coating. So, the agglomeration phenomenon as observed in the SEM micrographs is being asserted to be associated with nano ZnO content in the coating matrix and not the TiO_2 pigment.
The agglomerates at 5% loading, in the SEM image can be seen almost projected out from binder matrix. Agglomeration of nano-particles is un-desirable as it makes nano-particles lose their high surface area and thus the desired functional properties [5]. The agglomerated nano ZnO would also protect the material underneath the agglomerate. However, the result of agglomeration is that the nano ZnO would not be evenly distributed and so there will be regions in the coating that would be starved of zinc oxide and not protected, by zinc oxide in the coating.

With increase in nano ZnO concentration in the coating, the distribution of nano ZnO varied and SEM study was carried out to observe the same. The SEM images showed that the agglomerates formed at 5% loading of nano ZnO, were not binded or wetted by the resin effectively resulting in inhomogeneous distribution. We inferred from the SEM images that with the same resin content available in the coatings with 1%, 2% and 5% nano ZnO, as the agglomeration tendency of the nano ZnO increased from 1% to 5% loading, the binding or wetting of nano ZnO by resin was poor. It can be visibly observed that at 5% loading, the agglomerates projecting out of the coating matrix due to insufficient binding with the resin. Consequently, the epoxy coating with 5% nano ZnO showed higher dE and YI value by a unit compared to the epoxy coting with optimum loading of 2% nano ZnO.

IV. CONCLUSION

1. NANO ZNO SYNTHESIZED WITH FLAKE-LIKE MORPHOLOGY, LOWERED DISCOLORATION AND YELLOWING BY 40 % AT AN OPTIMIZED LOADING LEVEL OF 2% (BY WEIGHT OF EPOXY RESIN) IN THE EPOXY COATING SYSTEM UNDER STUDY.

2. LOWER CONCENTRATION OR A HIGHER CONCENTRATION THAN THE OPTIMUM LOADING OF 2% RESULTED IN LOWERING THE RESISTANCE TO DISCOLORATION AND YELLOWING.

3. THE FLAKE-LIKE MORPHOLOGY AND NANO-SIZE DIMENSIONS, ALONG WITH THE OPTIMUM LOADING CONCENTRATION OF ZNO, HELPED IN EFFECTIVE LOWERING OF DISCOLORATION AND YELLOWING OF THE EPOXY COATING ON ACCELERATED UVB EXPOSURE.

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Study on Machining Parameters of TiB_2 Reinforced Aluminum 6063 Composites

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Abstract- In this experimental study, three TiB_2/Al metal matrix composites (MMCs) with 40 μm in mean size were produced using a melt stirring squeeze casting route. Test was carried out to find the parameters influencing surface roughness (R_a) and material removal rate (MRR) on machining the TiB_2 reinforced Al-6063 composite materials. The orthogonal array, the signal-to-noise ratio, and analysis of variance were employed to study the performance characteristics in turning operations of 5 and 10 wt. % TiB_2 particles reinforced aluminum (Al-6063) metal matrix composites. Taguchi method was used to find the optimal cutting factors for surface roughness (Ra) and material removal rate (MRR). The factors considered were weight percentage of TiB_2, cutting speed, depth of cut and feed rate. The plan for the experiments and analysis was based on the Taguchi L_{27} orthogonal array with three cutting factors and two carbide tools of K_{10} and K_{20} type. The optimal parametric combination for K_{10} and K_{20} carbide insert was found to be feed, speed, depth of cut and percentage of TiB_2 reinforcement. The analysis of variance (ANOVA) results show that feed the most significant process parameter on surface roughness followed by speed/TiB_2 reinforcement for both K_{10} and K_{20} carbide insert tools. From MRR data also, it is clear that the speed and feed are the most significant parameters followed by the TiB_2 filler loading/depth of cut for the two carbide tools. The confirmatory test for the optimal condition yields low R_a value and high MRR value for the K_{20} type carbide tool.

Index Terms- TiB_2 reinforced 6063 aluminum composites, Taguchi Method, Surface Roughness, Material Removal Rate, K_{10} and K_{20} type Carbide Tool.

I. INTRODUCTION

In order to meet the demands of engineering applications, rigid ceramic fillers were added to the metal matrix such as aluminum and its alloys. Through compound effect, the strength, stiffness and wear resistance of metal matrix composites (MMCs) were greatly enhanced compared to unreinforced alloys. Besides, MMCs also have some outstanding properties like low density, high modulus, high thermal conductivity and low thermal expansion which make them find increasing applications in automobile, aerospace, electronics and medical industries [1]. There are manufacturing techniques with which it is possible to produce high quality MMC components to near-net shape. Unfortunately, for reasons as component design and dimensional tolerance requirements, the need for machining can not be completely eliminated [2]. However, MMCs are difficult to be machined to a good surface quality due to their low plasticity, non-uniformity and high abrasive nature of the ceramic reinforcement, which causes rapid tool wear rate and excessive machining induced defects [3-5].

In-situ composites are multiphase materials where the reinforcing phase is synthesized within the matrix during composite fabrication. Recently, In-situ techniques have been developed to fabricate aluminum-based metal matrix composites [6-8], which can lead to better adhesion at the interface and hence better mechanical properties. Literature review of aluminum-based composites show that this system has been widely investigated both in terms of understanding the mechanism of formation and property evaluation of composites [9-16]; the main dispersoids being SiC and Al_2O_3. Although investigations with varying volume fraction of dispersoids have been done, meaningful results have been obtained when the dispersoid percentage has been restricted to 15. As far as ceramic reinforced aluminum matrix composites are concerned, the machined surface possesses various flaws and damages. Generally, machined under conventional methods, four types of Surface/ subsurface damages may be produced when machining powder-formed MMCs: cracked SiC particles, delaminated matrix–reinforcement interface, pulled out/missing particles and work-hardened matrix [14]. Crack that initiated from either a fractured SiC particle or a delaminated matrix–reinforcement interface was also occasionally seen. And for the cast-MMCs, two additional types of subsurface defects were identified, i.e. numerous cracked SiC particles due to casting and collapsed or halved voids due to machining. Using polycrystalline diamond (PCD) tools and under high speed turning, Gallab and Sklad[3] investigated that the machined surface revealed grooves, voids around the SiC particles, pulled-out SiC particles, and fractured or crushed SiC particles. They found that the plastic deformation due to machining extended to about 60–100 μm below the machined surface.

Mechanical and physical properties are greatly influenced by the condition of the work piece surface. Hence, the effect of machining on the type of surface defects, as well as on their distribution, has a major impact on the performance of the machined component. Such machine-induced defects could be a concern when using MMCs in a critical or precise application.

Again the size of the dispersoids is dependent on/or restricted to the processing route adopted. There are different routes to synthesize Al–TiB_2 composites, but in-situ approach is gaining importance due to simplicity of its fabrication. Among the reinforcements, TiB_2 has emerged as a promising candidate for Al-based composites. This is due to the fact that TiB_2 is stiff, hard and more importantly, does not react with aluminum to form reaction product at the interface of reinforcement and matrix. TiB_2 is a refractory compound that exhibits outstanding features such as high melting point (2790°C), high hardness (86 HRA or 960 HV) and high modulus characteristics. Its resistance to plastic deformation even at high temperatures portrays it to be a good potential reinforcing candidate in an aluminum matrix.

5 and 10 wt. % TiB_2 particles reinforced aluminum (Al-6063) metal matrix composites produced by using master alloys of Al-Ti & B by stir casting process to obtain the material for the experiment [6-8,17]. Few researchers have dedicated themselves to the study of diamond turning of SiC/Al composites [18-21]. Unfortunately, most researchers limited their studies to relating the effect of the machining cutting parameters and reinforcement to the work piece surface roughness. To the best of the authors' knowledge, no research has yet comprehensively demonstrated the type of the surface/subsurface defects and the factors that affect them when ultra-precision machining SiC/Al composites. Surface roughness has become the most significant technical requirement and it is an index of product quality. In order to improve the tribological properties, fatigue strength, corrosion resistance and aesthetic appeal of the product, a reasonably good surface finish is desired. Nowadays, the manufacturing industries specially are focusing their attention on dimensional accuracy and surface finish. In order to obtain optimal cutting parameters to achieve the best possible surface finish, manufacturing industries have resorted to the use of handbook based information and operators' experience. This traditional practice leads to improper surface finish and decrease in the productivity due to sub-optimal use of machining capability. This causes high manufacturing cost and low product quality [18, 22-24]. In addition to the surface finish quality, the material removal rate (MRR) is also an important characteristic in turning operation and high MRR is always desirable [22, 24]. Several researchers carried out experiments on machining of MMCs.

Channakesava rao et al. [25] have experimented with different cutting tools and reported that, the crater wear is not substantial in K_{10} tools and is having superior wear resistance and produce continuous chips. Hoecheng et al. [26] have studied the effect of speed, feed, depth of cut, rake angle and cutting fluid on the chip form, forces, wear and surface roughness. Yuan et al. [27] have studied the effect of percentage volume reinforcement, cutting angle, feed rate and speed on the surface integrity in ultra precision diamond turning of MMCs.

The main concerns when machining MMCs is the extremely high tool wear due to the abrasive action of the ceramic fibers or particles. Therefore, materials of very high resistance to abrasive wear are often recommended. The HSS tools are inadequate, cemented carbide tools are preferred for rough machining and PCD tools for finish machining operations [28]. The cost of PCD tools increase the cost of production so it is necessary to carryout basic machinability studies, in order to find cutting conditions using carbide tools, which can result in high productivity at low cost. Most of the studies on Al/SiC-MMCs composite machining show that minimizing the surface roughness has been very difficult and is to be controlled. Taguchi emphasizes on the fact that quality provides robustness and immune to the uncontrollable factors in the manufacturing state. This approach helps to reduce the large number of experimental trials when the number of process parameters increases. [29-31]. In the present work, a procedure has been introduced to assess and optimize the chosen factors to attain minimum surface roughness and maximum material removal rate by incorporating response table, response graph and analysis of variance (ANOVA) technique.

II Experimental Details

A Materials and method of fabrication

Aluminium 6063 alloy was selected as the base line material as it possesses good formability, weldability, machinability and corrosion resistance, with medium strength compared to other grades of aluminium alloys. Its nominal chemical composition is shown in Table 1. The commercially available Al-6063 matrix alloy and master alloy are melted in an electric resistance furnace. The percentage weight of Al-Ti &B % was varied from 0-10 wt % in steps of 5 wt%. The mixture of matrix alloy and master alloy were melted in an electric resistance furnace at a temperature of 800^0C and allowed to stand for a duration of about 30 min to get melts. The melt was degassed using commercially available chlorine based tablets (Hexachloroethane) to remove the entrapped gases before stirring the melt using stirrer to get in-situ composites of TiB_2 in Al-6063 alloy. The melt is poured into the preheated metallic moulds. The different wt% composition (0, 5 and 10) of Al6063-TiB_2 composites rods are prepared of size Ø 22 mm x120 mm length.

Table 1. Chemical composition of base alloy Al-6063.

Element	Mg	Si	Fe	Cu	Mn	Zn	Ti	Cr	Al
Wt.%	0.45 -0.9	0.2-0.6	0.35	0.1	0.1	0.1	0.1	0.1	Balance

B Microscopy, Density and Hardness Measurement

In order to know the dispersion of TiB_2 in Al-6063, the samples were analyzed by scanning electron microscope (SEM). The density of the composites was obtained by the Archimedes's principle of weighing small pieces cut from the composite disc first in air and then in water. Then, theoretical density of composite and its alloy was calculated from the chemical analysis data. The porosity of the composites was also determined. The hardness of the composites and matrix alloy were measured after polishing to a 3 mm finish. The magnification of the images was 500x. Hardness of all samples was measured by Brinell hardness tester and mean of at least five readings was taken to represent the sample.

C Cutting conditions

The experiments were carried with four factors at three levels each as shown in Table 2. The factorial design used is a standard L_{27} (3^{13}) orthogonal array. This orthogonal array is chosen due to its capability to check the interactions among factors. The turning trials were carried out on the CNC turning center (MITSUBISHI-EZ Motion NC E60) in dry machining condition shown in Figure 1. The insert used is of K_{10} type carbide tool (SANDVIK: DNMG 3005 TOOL: $K_5 - K_{20}$) and K_{20} type carbide tool (SANDVIK: DNMG 3215 TOOL: $K_{15} - K_{35}$) shown in Figure 2 (a) and 2(b) respectively. The material used is 0, 5 and 10wt % TiB_2 reinforced Al-6063 composite material cut into the of size Φ20 mm x 60 mm length as shown in Figure 3.

The surface roughness (R_a) of the machined surface was measured using HANDYSURF E-35A instrument shown in Figure 4 and the MRR is calculated using the following equation [18].

$$\text{MRR } (mm^3/min) = \frac{[\text{Intial weight of the work(gms)} - \text{Final Weight of the work(gms)}]}{\text{Density } (gms/mm^3) \times \text{Machining time(min)}}$$

Figure 1: CNC turning center. (MITSUBISHI-EZ Motion NC E60)

Figure 2: (a) SANDVIK: DNMG 3005 Tool (K_{10})

(b)SANDVIK: DNMG 3215 Tool (K_{20})

Figure 3: Composite Material 0, 5 & 10 wt%TiB_2 Reinforced with Al-6063.

Table 2: Factors (process parameters) and Levels Used in the Experiments

Process Factors	Levels		
	1	**2**	**3**
A-Cutting Speed (m/min)	37.69	75.39	113.07
B-Feed rate (mm/min)	0.05	0.10	0.15
C- Depth of Cut (mm)	0.25	0.50	0.75
D- Material (Al6063 + wt%TiB_2)	0	5	10

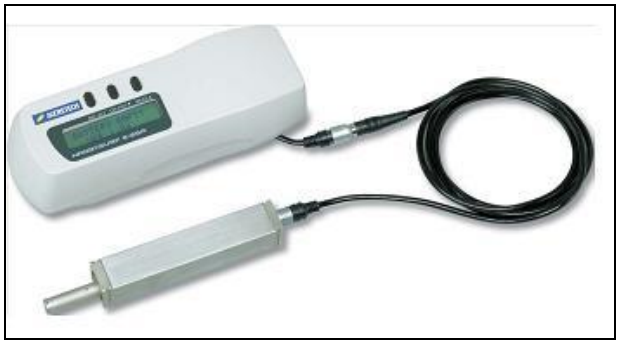


Figure 4. HANDYSURF E-35A, Surface Roughness Measuring Instrument.

III Results and Discussion

A Microstructure, Density and Hardness of TiB2/Al-6063 Composites

The properties of the MMCs depend not only on the matrix, particle, and the volume fraction, but also on distribution of reinforcing particles and interface bonding between the particle and matrix. In practical way, to achieve a homogenous distribution is difficult. The photomicrographs of the aluminum composite reinforced with 5 and10 wt. % of TiB_2 are shown in Fig. 5(a) and (b) respectively. The particles, with the average particle size of 25 µm, mainly formed in the surface showed a character of homogenous distribution.

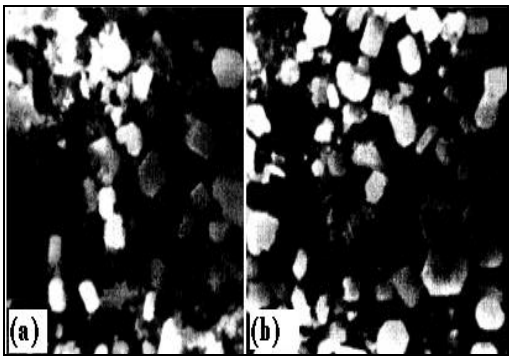


Figure 5. Photomicrographs of Al-6063 composites (a) 5 wt% TiB_2 and (b) 10 wt% TiB_2.

The variations of density and hardness of the composites are shown in Figure 6. The density and hardness of the MMCs increased more or less linearly with the weight fraction of particles in the alloy matrix due to the increasing ceramic phase of the matrix alloy. A significant increase in both density and hardness was found in 10 wt. % TiB_2 into aluminum composite. The increase in density indicates that particle breakage may not have any significant influence on the composites. It is believed to achieve an improvement of the bonding between the particle and matrix. The porosities of

composites were evaluated from the difference between the expected and the observed density of each sample. The variations of porosity level in these composites are 1.3 and 1.5 % for 5 and 10 wt. % TiB_2 in aluminum composites respectively. The porosity level increased, since the contact surface area was increased.

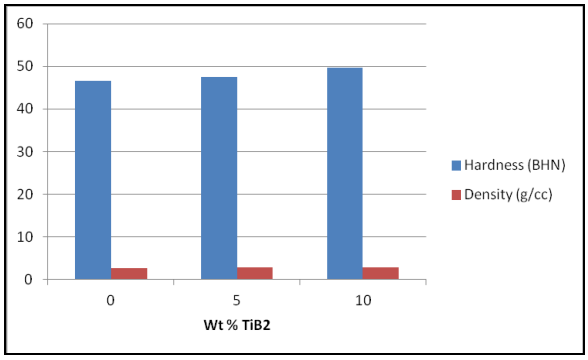


Figure 6. Density and hardness of TiB_2/Al MMCs.

B Analysis of Variance of Surface Roughness and Material Removal Rate

Assessing the factors and its effects on surface roughness (R_a) and material removal rate (MMR) of TiB_2/ Al-MMC machining process has been carried out through response table, response graph and analysis of variance (ANOVA) technique. The influence of machining parameters on R_a and MMR has been performed using response table. Response tables are used to simplify the calculations needed to analyze the experimental data. The complete response table for a three level, four factors using L_{27} orthogonal array is shown in Table 3.

C Surface Roughness

The response data recorded in Table 3, for surface roughness subjected to ANOVA for finding the significant factors at above 85% confidence levels and the result of ANOVA for these response parameters are presented in the Tables 4, 5 and 6 respectively for both type of carbide tool. For MRR subjected to ANOVA for finding the significant factors, at above 95% confidence levels and the result of ANOVA for these response parameters are presented in the Tables 7, 8and 9 respectively for both type of tool. Average S/N ration for each level of experiment and the different values of S/N ration between maximum and minimum are shown in Table 4 for surface roughness. The feed rate and cutting speed are two factors that have highest different values of 8.34 and 4.75 respectively for K_{10} type tool & the feed rate and TiB_2 reinforcement are two factors that have highest different values of 6.12 and 3.61 respectively for K_{20} type tool. Based on the Taguchi prediction that the bigger different values of S/N ratio will give more effect or more significant. Increase in the feed rate will increase the surface roughness significantly.

The results of analysis of variance for surface roughness are shown in Table 5. DF (degree of freedom), SS (sum of squares), MS (mean of squares), F (variance ratio), P (Significant factor) and percentage contribution of each level [30-34].

Table 5. Shows that the feed rate and the cutting speed have more influence on the surface roughness value. The significant factor (P) values for both are 0.000 and 0.005 respectively for K_{10} tool. In statistical analysis of Taguchi method, the smallest P value gives more significant effect on responded surface roughness parameters. The values for the feed rate and the cutting speed are significant. The contributions are 60.29% and 17.29% respectively for K_{10} type tool. The type of work material contributes about 8.18%, where as the contribution of the depth of cut and the interactions are insignificant.

Table 6. provides values for K_{20} tool, the feed rate(0.024) and TiB_2 reinforcement o (0.158) are significant followed by the speed of about 0.238. The values for the feed rate and reinforced material are significant. Their contributions are 38.52% and 13.22% respectively for K_{20} type tool. The cutting speed contributes about 9.53%, where as the contribution of interaction, speed and material is significant and contribute about 11.20%, the depth of cut and the other interactions are insignificant.

The most significant factor, which affects the surface roughness measured in turning the 5 and 10wt % TiB_2 reinforced with Al-6063 composite material, is the feed rate for both type of tool and therefore the surface roughness can be controlled with a suitable feed rate value. Previous researchers suggest the similar results. They claimed that the surface roughness strongly depends on the feed rate followed by the cutting speed or the reinforced material.[32-34].

Table 3. Factor Settings, Surface Roughness Data and MRR Data.

Exp No	Cutting Speed (m/min)	Feed (mm/min)	Depth of cut (mm)	wt% TiB_2	For K_{10} Tool		For K_{20} Tool	
					Material Removal Rate (mm^3)/min)	Surface Roughness (R_a) µm	Material Removal Rate (mm^3)/min)	Surface Roughness (R_a) µm
1	37.69	0.05	0.25	0	0.136	0.40	0.152	0.40
2	37.69	0.05	0.50	5	0.380	0.67	0.276	0.73
3	37.69	0.05	0.75	10	0.236	0.83	0.099	1.90
4	37.69	0.10	0.25	5	0.360	1.03	0.282	1.47
5	37.69	0.10	0.50	10	0.321	1.37	0.719	1.73
6	37.69	0.10	0.75	0	0.470	0.67	0.590	0.57
7	37.69	0.15	0.25	10	0.395	1.70	0.325	1.70
8	37.69	0.15	0.50	0	0.683	1.97	0.724	1.87

9	37.69	0.15	0.75	5	0.909	2.13	1.066	1.95
10	75.39	0.05	0.25	5	0.302	1.87	0.339	0.60
11	75.39	0.05	0.50	10	0.180	0.90	0.553	0.90
12	75.39	0.05	0.75	0	0.599	1.10	0.793	1.00
13	75.39	0.10	0.25	10	0.488	1.23	0.367	1.10
14	75.39	0.10	0.50	0	2.269	1.20	1.653	0.87
15	75.39	0.10	0.75	5	1.522	1.13	1.616	1.20
16	75.39	0.15	0.25	0	2.559	2.33	0.925	1.73
17	75.39	0.15	0.50	5	2.016	3.10	2.678	2.32
18	75.39	0.15	0.75	10	1.209	1.97	1.396	1.57
19	113.07	0.05	0.25	10	0.312	0.53	0.504	0.63
20	113.07	0.05	0.5	0	0.585	0.47	0.717	0.60
21	113.07	0.05	0.75	5	0.791	0.63	0.727	0.87
22	113.07	0.10	0.25	0	1.856	0.57	1.037	0.67
23	113.07	0.10	0.50	5	2.316	1.23	2.397	1.67
24	113.07	0.10	0.75	10	1.775	0.80	1.925	0.83
25	113.07	0.15	0.25	5	2.005	2.07	1.525	1.50
26	113.07	0.15	0.50	10	2.150	1.43	2.865	1.37
27	113.07	0.15	0.75	0	3.022	1.22	3.032	0.70

Surface roughness plays an important role in many areas and is a factor of great importance in the evaluation of machining accuracy. Although many factors affect the surface condition of a machined part, machining parameters such as cutting speed, feed rate and depth of cut have a significant influence on the surface roughness for a given machine tool and work piece set-up.

The studies on machining characteristics of TiB_2/6063Al MMCs acquire more importance due to the presence of abrasive phase in metal matrix. The presence of TiB_2 in metal matrix is reported to increase the hardness, tensile strength and heat resistance. The rate of change in these properties is dependent on the weight fraction of TiB_2 added to the matrix alloy. The cutting speed plays an important role in deciding the surface roughness. At high cutting speeds, the surface roughness decreases. At low speeds, the built-up edge (BUE) is formed and also the chip fracture readily producing the rough surface [35]. As the speed increases, the BUE vanishes, chip fracture decreases, and hence the roughness decreases.

The increase in depth of cut results in high normal pressure and seizure on the rake face and promotes the BUE formation. Hence, the surface roughness increases along with increase in depth of cut. The increase in feed rate, increases the surface roughness linearly upto 0.15 mm/rev. At feed rates between 0.15 and 0.05 mm/rev, the BUE forms readily and is accomplished by feed marks resulting in increased roughness.

The results shown prove that the surface roughness of TiB_2/Al-6063 composite is highly influenced by the feed rate, cutting speed and % weight fraction of TiB_2 particles in the work piece. The depth of cut also plays a significant role on composite machining process in deciding the surface roughness.

Table 4. Average for S/N Ratio and Main Effect of Surface Roughness.

Process parameters	Design ation	For K_{10} Tool					For K_{20} Tool				
		Levels			Max-Min	Rank	Levels			Max-Min	Rank
		1	2	3			1	2	3		
Cutting Speed (m/min)	A	-0.39	-3.65	- 1.10	4.75	2	-1.55	-1.32	0.79	2.34	3
Feed rate (mm/min)	B	2.64	0.11	-5.69	8.34	1	2.24	-0.42	-3.89	6.12	1
Depth of cut (mm)	C	-0.82	-1.55	-0.55	0.99	4	0.31	-1.71	-0.67	2.03	4
Wt% TiB2 material	D	0.61	-2.64	-0.90	3.25	3	1.65	-1.96	-1.76	3.61	2

Table 5. ANOVA Analysis of S/N Ratio for Surface Finish of K_{10} Tool

Process parameters	Designation	DF	SS	MS	F	P	SS (%)
Cutting Speed (m/min)	A	2	2.01	1.01	14.00	0.005	17.29
Feed rate (mm/min)	B	2	7.02	3.51	48.80	0.000	60.29
Depth of cut (mm)	C	2	0.20	0.10	1.39	0.319	01.70

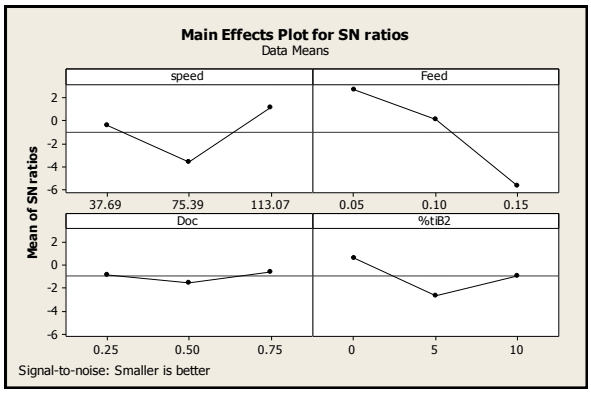
%TiB2 material	D	2	0.95	0.48	6.63	0.030	08.18
Speed*Feed	A*B	4	0.35	0.09	1.21	0.396	02.99
Speed * DOC	A*C	4	0.27	0.07	0.94	0.500	02.32
Speed * %TiB2	A*D	4	0.40	0.10	1.40	0.338	03.46
Error		6	0.43	0.07			03.77
Total		26	11.64				100.00

R-Sq = 96.29%

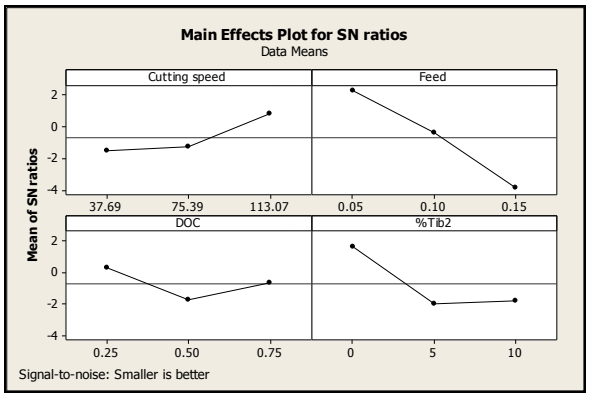
Table 6. ANOVA Analysis of S/N Ratio for Surface Finish of K_{20} Tool

Process Parameters	Designation	DF	SS	MS	F	P	SS (%)
Cutting Speed (m/min)	A	2	0.71	0.36	1.84	0.238	09.53
Feed rate (mm/min)	B	2	2.87	1.43	7.43	0.024	38.52
Depth of cut (mm)	C	2	0.29	0.15	0.76	0.509	03.89
wt%TiB2 material	D	2	0.98	0.49	2.55	0.158	13.22
Speed*Feed	A*B	4	0.41	0.10	0.53	0.722	05.50
Speed * DOC	A*C	4	0.19	0.05	0.25	0.900	02.68
Speed * %TiB2	A*D	4	0.84	0.21	1.08	0.443	11.20
Error		6	1.16	0.19			15.55
Total		26	7.45				100.0

R-Sq=84.45%



(a) K_{10} tool



(b) K_{20} tool

Figure 7(a) and (b) Main Effect Plots for S/N Ratio of process parameters on Surface Roughness.

Figure 7(a) shows that the S/N ratio of surface finish for K_{10} tool, the B1 (low level of feed) is at the maximum value with 2.64 of S/N ratio, decrease dramatically to B2 (0.11) and then to B3 (-5.69). S/N ratio for B2 and B3 decreased, due to increase in the feed rate. The parameter A1 (low level speed) is at the minimum value at -0.39 and it increased to a value of 1.10 as the speed increases. The material composition is at the maximum value of 0.61 and decreased to -2.64 as the inclusion of TiB_2 reinforcement increases [36].

The signal to noise ratio are shown in Figure 7(b) for the process factors on surface roughness for K_{20} tool. The value B1 (low level of feed) is at the maximum value with 2.24, decrease dramatically to B2 (-0.42) and then to B3 (-3.89). S/N ratio for B2 and B3 decreased, due to increase in the feed rate. The parameter D1 (low level of material reinforcement) is at the maximum value at 1.65 and it decreased to a value of -1.76 as the feed increases. The cutting speed is at the minimum value of -1.55 and increased to 0.79[37]

From the above S/N ratio results for feed variation values, the K_{20} tool provides low values compared to K_{10} tool. From the analysis, the optimal machining parameters for the TiB_2/Al MMCs machining process are achieved for the maximum surface finish. The optimal conditions arrived are:(i) Cutting speed (113.07 m/min),(ii) % weight fraction of TiB_2 (0),(iii) Depth of cut (0.25 mm) and (iv) Feed rate (0.05 mm/rev) shown in Table 7.

D Material Removal Rate

In machining operation, maximizing the material removal rate (MRR) is an important criterion. Experimental results of the MMR for turning 5 and 10 wt% TiB_2 reinforced with Al6063 composites with various cutting parameters are shown in Table 7, which also gives the S/N ratio for the material removal rate. The feed rate and the cutting speed are the most significant factor that influences the MRR with the value of 12.31 and 10.99 respectively for k_{10} tool and for K_{20}

tool the speed and feed are significant with 11.39 and 10.78 respectively.

Table 8. Shows the feed rate and the cutting speed values for K_{10} tool are significant with the factor (P) value for both are 0.000 in percentage. The contribution of the feed rate and the cutting speed are 37.62 and 34.00 respectively. The type of work material contribute about 7.54% and the interaction parameter between the cutting speed and feed rate (A * B) is significant with a factor of 0.016, and its contribution is 10.15%. The other interactions are insignificant [36]

Table 9. Shows the cutting speed and the feed rate values are significant with the factor (P) value for both are 0.001 and 0.002 respectively. The contribution of the cutting speed and the feed rate are 32.28 and 31.86. The depth of cut contributed about 16.64% and the interaction parameter between the cutting speed and feed rate (A * B) is significant with a factor of 0.158, and its contribution is 6.75%. The other interactions are insignificant [37].

Table 7. Average for S/N Ratio and Main Effect of MRR.

Process parameters	Design ation	For K_{10} Tool					For K_{20} Tool				
		Levels 1	2	3	Max-Min	Rank	Levels 1	2	3	Max-Min	Rank
Cutting	A	-8.42	-1.03	2.57	10.99	2	-8.66	-0.66	2.71	11.39	1
Feed rate	B	-9.42	-0.34	2.88	12.31	1	-8.42	-0.56	2.36	10.78	2
Depth of cut	C	-4.60	-1.55	-0.72	3.87	4	-6.43	0.53	-0.72	06.96	3
Wt% TiB2	D	-0.67	-0.82	-5.38	4.71	3	1.67	-1.09	-3.87	02.78	4

Table 8. ANOVA Analysis of S/N Ratio for MRR of K_{10} Tool

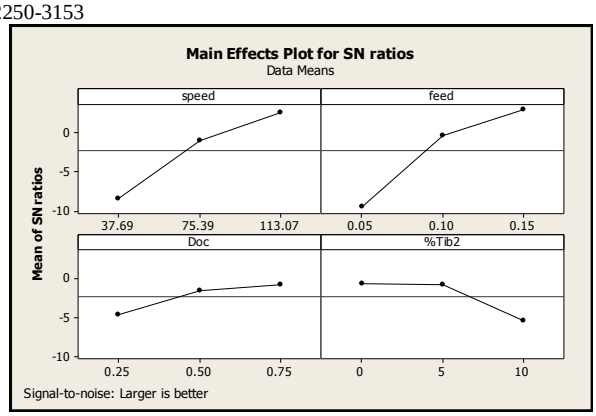
Process parameters	Design nation	DF	SS	MS	F	P	SS (%)
Cutting Speed (m/min)	A	2	6.87	3.43	40.13	0.000	34.00
Feed rate (mm/min)	B	2	7.59	3.79	44.39	0.000	37.62
Depth of cut (mm)	C	2	0.40	0.20	2.34	0.177	01.98
%TiB_2 material	D	2	1.52	0.76	8.90	0.016	07.54
Speed*Feed	A*B	4	2.05	0.51	5.98	0.027	10.15
Speed * DOC	A*C	4	0.31	0.08	0.91	0.513	01.55
Speed * %TiB_2	A*D	4	0.93	0.23	2.72	0.132	04.62
Error		6	0.51	0.09			02.54
Total		26	20.19				100.00

R-Sq = 97.46%

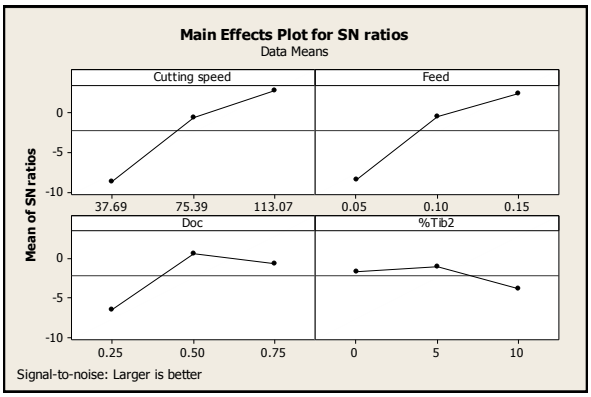
Table 9. ANOVA Analysis of S/N Ratio for MRR of K_{20} Tool

Process parameters	Design nation	DF	SS	MS	F	P	SS (%)
Cutting Speed (m/min)	A	2	6.17	3.09	23.37	0.001	32.28
Feed rate (mm/min)	B	2	6.09	3.05	23.08	0.002	31.85
Depth of cut (mm)	C	2	3.19	1.59	12.07	0.008	16.64
Wt %TiB_2 material	D	2	0.26	0.13	0.99	0.426	01.35
Speed * Feed	A*B	4	1.29	0.32	2.44	0.158	06.75
Speed * DOC	A*C	4	0.57	0.14	1.08	0.444	02.98
Speed * %TiB_2	A*D	4	0.75	0.19	1.43	0.332	03.92
Error		6	0.79	0.13			04.13
Total		26	19.12				100.00

R-Sq=95.86%



(a) K_{10} tool



(b) K_{20} tool

Figure 8(a) and (b) Main Effect Plots for S/N ratio of process parameters on MRR.

Figure 8 (a) shows that the S/N ratio of material removal rate for K_{10} tool, the A1 (low cutting speed) is the minimum value with -8.42 of S/N ratio, and increase dramatically to B2 (-1.03) and then to B3 (2.57). The main effect plot for the feed and depth of cut have same trend lines but the TiB_2 reinforced Al6063 composite work material shows inverse trend as wt% of TiB_2 increases the MRR decreases [36].

Figure 8(b) shows that the signal to noise ratio of material removal rate for K_{20} tool. The A1 (low cutting speed) is at minimum value with -8.66 of S/N ratio, and it increases dramatically to A2 (-0.66) and then to A3 (2.71). The main effect plot for the feed, and depth of cut have same trend lines but the TiB_2 reinforced Al6063 composite work material shows inverse trend as wt% of TiB_2 increases the MRR decreases[37].

From the above S/N ratio results for MRR, the cutting speed variation values, the K_{20} tool provides high values compared to K_{10} tool. From the analysis, the optimum parameters for the TiB_2/Al-6063 MMCs machining process are achieved for maximizing MRR. The optimal conditions arrived are:(i) Cutting speed (113.07 m/min),(ii) % weight fraction of TiB_2 (0),(iii) Depth of cut (0.75 mm) and (iv) Feed rate (0.15 mm/rev) shown in Table 7.

Table 7. Factors with optimum levels for Surface Roughness and MRR for both K_{10} and K_{20} type tools.

Process parameters	Optimum Level for Surface Roughness	Optimum values for Surface Roughness	Optimum Level for MRR	Optimum values for MRR
Cutting Speed (m/min)	A3	113.07	A3	113.07
Feed rate (mm/min)	B1	0.05	B3	0.15
Depth of cut (mm)	C1	0.25	C3	0.75
wt%TiB_2 + Al-6063 material	D1	0%	D1	0%

Table 8. Conformation Tests with optimum levels for Surface Roughness and MRR using K_{10} and K_{20} type tools.

Exp No	Cutting Speed (m/min)	Feed (mm/min)	Depth of cut (mm)	wt% TiB_2	For K_{10} Tool		For K_{20} Tool	
					Material Removal Rate (mm^3)/min)	Surface Roughness (R_a) µm	Material Removal Rate (mm^3)/min)	Surface Roughness (R_a) µm
1	113.07	0.05	0.25	0	--	0.376	--	0.353
2	113.07	0.15	0.75	0	3.028	--	3.037	--

E Conformation Test of Optimum Levels for Surface Roughness (R_a) and MRR

Table 8.Lists the results of the conformation test with optimal levels using two types of carbide Tools. For the surface roughness test optimum level are A3 (113.07 m/min), B1 (0.05 mm/min), C1 (0.25 mm) and D1 (Al-6063 matrix alloy) are adopted and the result show that the R_a values are 0.376 and 0.353 microns respectively for K_{10} and K_{20} carbide tool. For MRR the optimum level A3 (113.07 m/min), B3 (0.15 mm/min), C3 (0.75 mm) and D1 (Al-6063 matrix alloy) are adopted and the result show that the MRR values are 3.028 and 3.037 mm^3/min respectively for K_{10} and K_{20} carbide tool.

From the above result it shows that the K_{20} type carbide tool gives better machining conditions than the K_{10} type carbide tool. The low surface roughness value for the material Al-6063 matrix alloy (R_a) of about 0.353 microns and maximum MRR value of 3.037 mm^3/min.

IV Conclusions

1. Using Taguchi's experimental design, the parameters, which are having influence on surface Roughness and material removal rate on the machining of TiB_2/Al-6063 composites have been assessed.

2. Micro-structural examination showed that the TiB_2 distributions are more or less homogeneous and lower interface porosity could be observed.

3. Hardness of the aluminum alloy improved significantly by adding up of TiB_2 particles into it, while density of the composite also increased almost linearly with the weight fraction of particles..

4. The significant factors in turning 5 and 10 wt. % TiB_2/Al-6063 MMCs on surface roughness are feed rate and the cutting speed, with contributions 60.2% and 17.29 % respectively for K_{10} tool and for K_{20} tool the feed rate and the type of work material, with contribution about 38.52% and 13.22% respectively.

5. For material removal rate the feed rate and cutting speed are most significant factors, with contribution of 37.62% and 34.00 % respectively and the TiB_2 loading is contributed about 10.15% for K_{10} tool and for K_{20} tool the cutting speed and the feed rate are most significant factors, with contribution of 32.28% and 31.85% respectively. The depth of cut is contributed about 16.64%.

6. The optimal conditions of machining parameters for low surface finish for TiB2/Al-6063 composites are; speed of 113.07m/min; feed 0.05 mm/min, depth of cut of 0.25mm.

7. The optimal conditions of machining parameters for higher MRR for TiB2/Al-6063 composites are; speed of 113.07m/min; feed 0.15 mm/min, depth of cut of 0.75mm.

8. The optimal conditions for surface roughness and the MRR for TiB2/Al-6063 composites are shown in Table 7 and are same for two types of carbide tool.

9. The conformation test results with optimal machining conditions for surface roughness and the MRR are shown in Table 8 for both tools

10. The K_{20} type carbide tool gives better cutting conditions than the K_{10} type carbide tool with low surface roughness value of (R_a) 0.353 microns and maximum MRR value of 3.037 mm^3/min.

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Modeling and Performance Analysis of PID Controlled BLDC Motor and Different Schemes of PWM Controlled BLDC Motor

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Abstract- Brushless dc (BLDC) motor drives are continually gaining popularity in motion control applications. Therefore, it is necessary to have a low cost, but effective BLDC motor speed/torque regulator. They are used in Residential and commercial appliances such as refrigerators and air conditioning systems with conventional motor drive technology. A Brushless DC (BLDC) drives are known for higher efficiency and lower maintenance. This paper presents a "Modeling and performance analysis of PID controlled BLDC motor and different schemes of PWM controlled BLDC motor". This paper presents PID model of brushless dc (BLDC) motor with the use of MATLAB/ SIMULINK. The operational parameters of specific BLDC motor were modeled using the tuning methods which are used to develop subsequent simulations.

Index Terms- Brushless DC motor (BLDCM), Digital pulse width modulation (DPWM) PID controllers

I. INTRODUCTION

The use of the general type dc motor has its long history. It has been used in the industries for many years now. They provide simple means and precise way of control [1]. In addition, they have high efficiency and have a high starting torque versus falling speed characteristics which helps high starting torque and helps to prevent sudden load rise [2]. the brushless direct current (BLDC) motor are gaining grounds in the industries ,especially in the areas of appliances production ,aeronautics ,medicine, consumer and industrial automations and so on.

The BLDC are typically permanent synchronous motor, they are well driven by dc voltage. They have a commutation that is done mainly by electronics application.

Some of the man y advantages of a brushless dc motor over the conventional brushed dc motors are highlighted below [3]:

1. Better speed versus torque characteristics
2. High dynamic response
3. High efficiency
4. Long operating life
5. High speed ranges
6. Low maintenance (in terms of brushes cleaning; which is peculiar to the brushed dc motor).

The PID controller is applied in various fields of engineering, and it is also a very important tool in telecommunication system. If there is a system and stability is desired, then PID could be very useful. In practice, the design of the BLDCM drive involves a complex process such as modeling, control scheme selection, simulation and parameters tuning etc. An expert knowledge of the system is required for tuning the controller parameters of servo system to get the optimal performance. Recently, various modern control solutions are proposed for the speed control design of BLDC motor [4]. However, Conventional PID controller algorithm is simple, stable, easy adjustment and high reliability, Conventional speed control system used in conventional PID control [5]. But, in fact, most industrial processes with different degrees of nonlinear, parameter variability and uncertainty of mathematical model of the system. Tuning PID control parameters is very difficult, poor robustness, therefore, it's difficult to achieve the optimal state under field conditions in the actual production.. pwm control method is a better method of controlling, to the complex and unclear model systems, it can give simple and effective control. Proportional-integral (PI) control with hysteresis or pulse width modulation (PWM) switching is the most widely used speed control technique for BLDC motors with trapezoidal back EMF. It can be easily implemented on analog or digital components because it is well understood, simple, and in practice for a fairly long period of time. To enhance the performance or to reduce the cost has been focus of development work for a long time. Cost and implementation complexity are often the most important factors for design trade-offs between techniques, implementation, and strategy of motor control hardware. The aim of this paper is that it shows the dynamics response of speed with design the PID controller to control a speed of motor for keeping the motor speed to be constant when the load varies.

II. BRUSHLESS DC MOTOR AND MODEL CONCEPT

One of the major differences between the DC motor and the BLDC is implied from the name. the conventional dc motor has brushes that are attached to its stator while the "brushes" DC motor does not. Also, unlike the normal DC motor, the commutation of the BLDC could be done by electronic control [3]. Under the BLDC motor, the stator windings are energized in sequence for the motor to rotate. More also, there is no physical contact whatsoever between the stator and the rotor. Another vital part of the BLDC is the hall sensor(s); these hall sensors are systematically attached to the rotor and they are used as major sensing device by the Hall Effect sensor embedded into the stator [3]. This works based on the principal of Hall Effect.

The BLDC motor operates in many modes (phases), but the most common is the 3-phase. The 3- phase has better efficiency and gives quite low torque. Though, it has some cost implications, the 3-phase has a very good precision in control [6]. And this is needful in terms of the stator current.

Typically, the mathematical model of a Brushless DC motor is not totally different from the conventional DC motor. The major thing addition is the phase involved which affects the overall result of the BLDC model. The phase peculiarly affects the resistive and the inductive of the BLDC arrangement.

2.1 Mathematical model of a typical BLDC motor

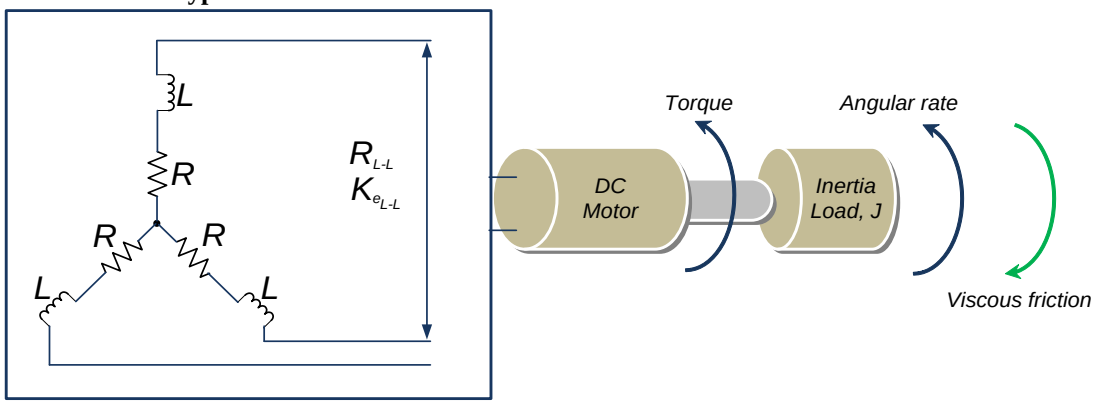

Fig.1. Brushless DC motor schematic diagram

For the mechanical time constant (with symmetrical arrangement), equation becomes:

$$\tau_m = \sum \frac{RJ}{K_e K_t} = \frac{J \sum R}{K_e K_t} \quad (2.1)$$

the electrical (time constant),

$$\tau_e = \sum \frac{L}{R} = \frac{L}{\sum R} \quad (2.2)$$

Therefore, since there is a symmetrical arrangement and a there phase, the mechanical (known) and electrical become:

Mechanical constant,

$$\tau_m = \frac{J.3R}{K_e K_t} \quad (2.3)$$

Electrical constant,

$$\tau_e = \frac{L}{3.R} \quad (2.4)$$

Where,

$K_e = \left[\frac{\text{v} - \text{secs}}{\text{rad}}\right]$: the electrical torque

$K_t = \left[\frac{\text{N} - \text{m}}{\text{A}}\right]$: the torque constant

Therefore, the equation for the BLDC is

$$G(s) = \frac{\frac{1}{K_e}}{\tau_m \cdot \tau_e \cdot s^2 + \tau_m \cdot s + 1} \quad (2.5)$$

III. MAXON BLDC MOTOR

3.1 Maxon EC 45 flat Ø 45 mm, brushless DC motor

The BLDC motor provided for this paper is the EC 45 flat Ø45 mm, brushless, 30 Watt from Maxon motors [7]. The order number of the motor is 200142. The parameters used in the modeling are extracted from the datasheet of this motor with corresponding relevant parameters used. Find below in Table 5.1 the major extracted parameters used for the modeling task.

	Maxon Motor Data	Unit	Value
	Values at nominal voltage		
1	Nominal Voltage	V	12.0
2	No load Speed	rpm	4370
3	No load Current	mA	151
4	Nominal Speed	rpm	2860
5	Nominal Torque (max. continuous torque)	mNm	59.0
6	Nominal Current (max. continuous current)	A	2.14
7	Stall Torque	mNm	255
8	Starting Current	A	10.0
9	Maximum Efficiency	%	77
	Characteristics		
10	***Terminal Resistance phase to phase***	Ω	1.1
11	***Terminal Inductance phase to phase***	mH	0.50

12	***Torque Constant***	mNm/A	24.5
13	***Speed Constant***	rpm/V	35.4
14	Speed/Torque Gradient	rpm/mNm	17.6
15	***Mechanical time constant***	ms	16.1
16	***Rotor Inertia***	gcm^2	82.5
17	*Number of phase*		3

Table.1. BLDC MOTOR PARAMETERS USED [7]

IV. BLDC MAXON MOTOR MATHEMATICAL MODEL

The mathematical model of the BLDC motor is modelled on the parameters from table given above.

$$G(s) = \frac{\frac{1}{K_e}}{\tau_m \cdot \tau_e \cdot s^2 + \tau_m \cdot s + 1} \quad \ldots\ldots\ldots\ldots \quad (4.1)$$

So the values for K_e, τ_m and τ_e need to calculated to obtain the motor model

$$\tau_e = \frac{L}{3.R}$$

$$\tau_e = \frac{0.50 \times 10^{-3}}{3 \times 1.10}$$

$$\tau_e = 151.51 \times 10^{-6} \quad \ldots\ldots\ldots\ldots\ldots\ldots \quad (4.2)$$

But τ_m is a function of R, J, K_e and Then

$$\tau_m = \frac{3.R_\emptyset.J}{K_e.K_t} = 0.0161$$

$$K_e = \frac{3.R_\emptyset.J}{\tau_m.K_t} = \frac{3 \times 1.1 \times 8.25 \times 10^{-6}}{0.0161 \times 24.5 \times 10^{-3}} = 0.06902 \frac{\text{v} - \text{secs}}{\text{rad}}$$

Therefore, the G(s) becomes:

$$G(s) = \frac{14.48}{151.51 \times 10^{-6} \times 0.0161 \cdot s^2 + 0.0171 \cdot s + 1}$$

$$G(s) = \frac{14.48}{2.44 \times 10^{-6} \cdot s^2 + 0.0161 \cdot s + 1} \quad \ldots\ldots\ldots\ldots\ldots\ldots \quad (4.3)$$

The G(s) derived above in the equation 4.3 is the open loop transfer function of the Brushless DC maxon motor using all necessarily sufficient parameters available.

4.1 Open Loop Analysis using MATLAB m- file

With the aid of the BLDC motor parameters provided, the open loop analysis is done by considering the stability factors and making the necessary plots for this analysis. Some of the plots include the step response, root locus, nyquist diagram, and bode plot diagram.

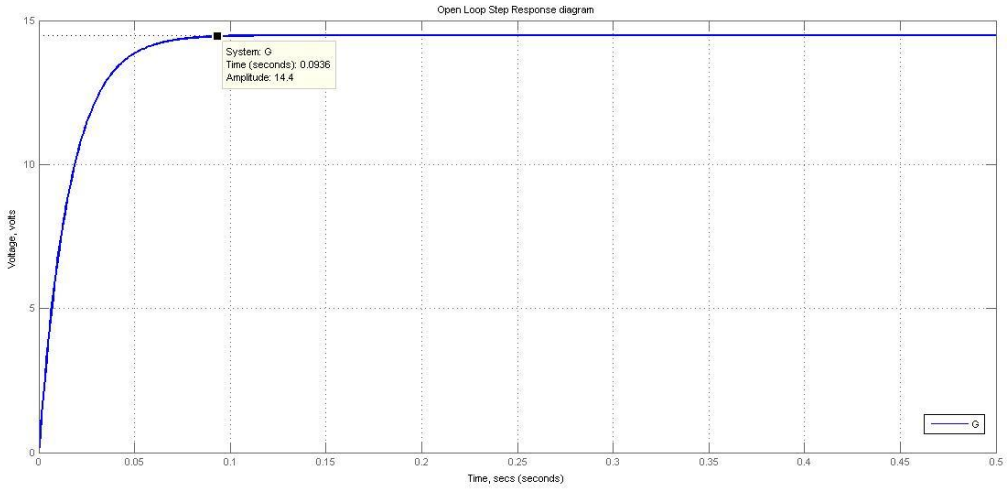

Figure.2. Open loop step response diagram

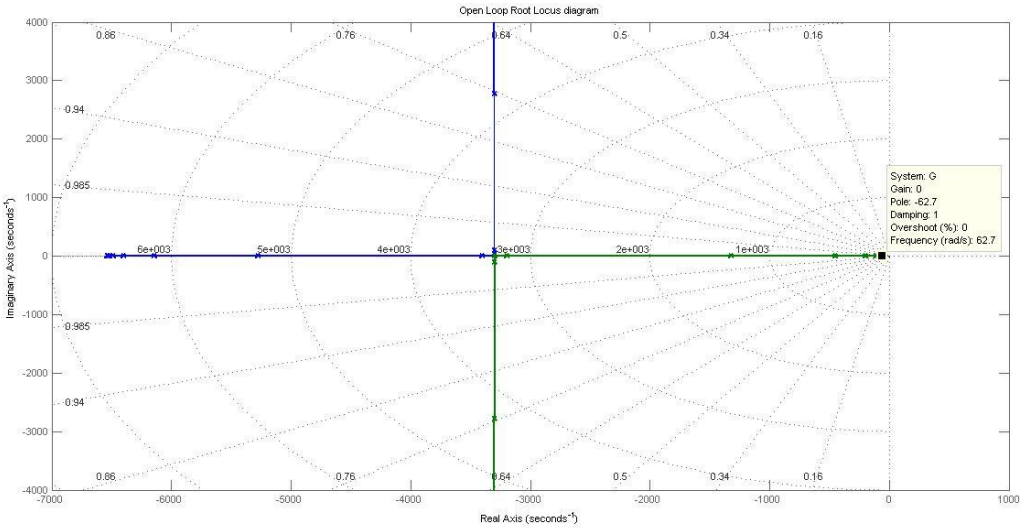

Figure.3. Open loop step root locus diagram

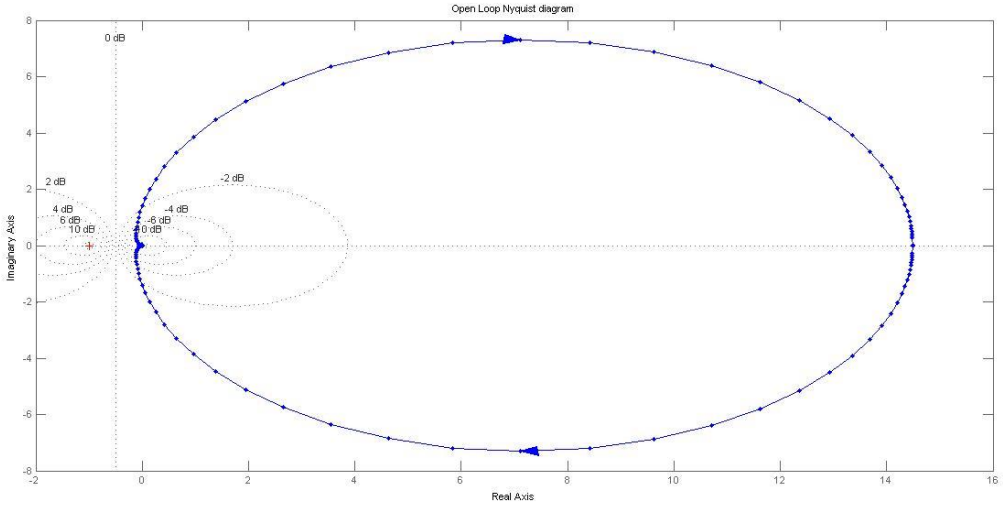

Figure. 4. Open Loop step Nyquist diagram

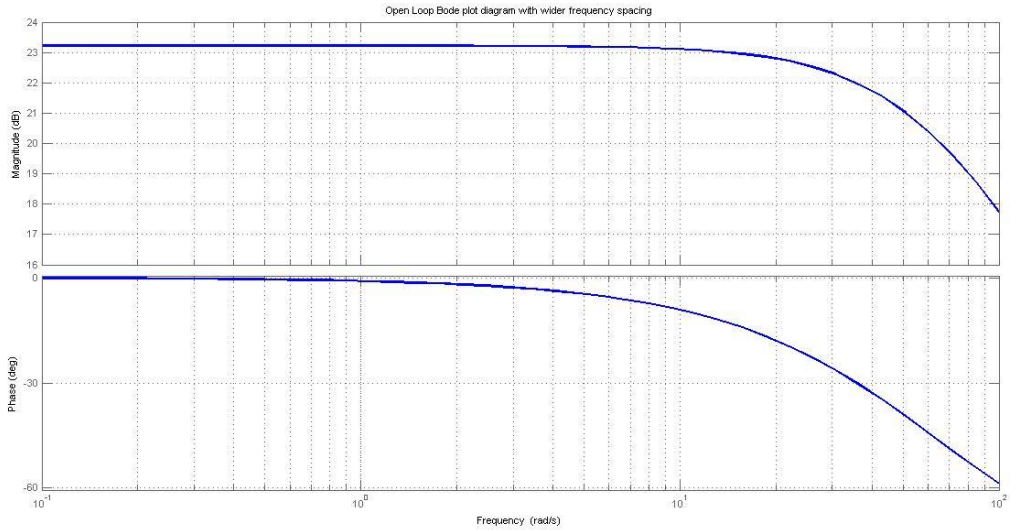

Figure. 5. Open Loop Step Bode Plot diagram

V. PID DESIGN CONCEPT

The Proportional-Integral-Derivative (PID) controller is about the most common and useful algorithm in control systems engineering [8]. In most cases, feedback loops are controlled using the PID algorithm. The main reason why feedback is very important in systems is to be able to attain a set-point irrespective of disturbances or any variation in characteristics of any form.

Consider the characteristics parameters – proportional (P), integral (I), and derivative (D) controls, as applied to the diagram below in figure 2, the system, S is to be controlled using the controller, C; where controller, C efficiency depends on the P, I and D parameters [7].

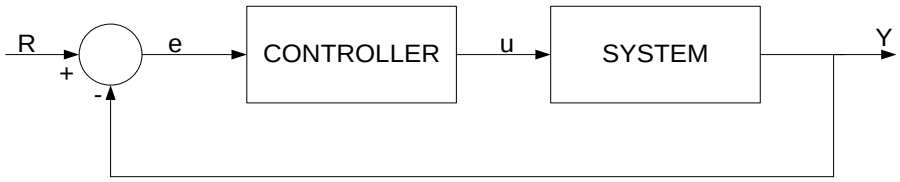

Figure. 6. A typical system with a controller

Typically, the function of the form shown in equation 5.1 is applicable in this kind of PID controller design.

$$K_p + \frac{K_I}{s} + K_D \cdot s = \frac{K_D s^2 + K_P s + K_I}{s} \qquad (5.1)$$

Where,

K_p = Proportional gain
K_I = Integral gain
K_D = Derivative gain

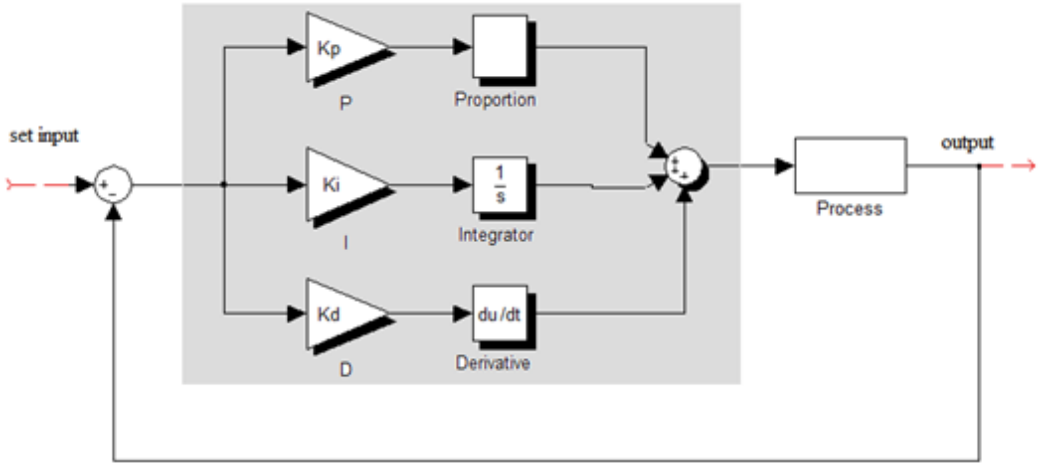

Figure. 7. PID parameters schematics

VI. PID CONTROLLER TUNING PARAMETERS

Under this section a critical analysis would be done on the PID tuning criteria and the parameters involved. Before a detail analysis is done, a quick look at the tuning methods is considered first and thereafter, specific tuning parameters are computed for the BLDC maxon motor. Some of the generally used tuning methods are the Trial and Error method, the Ziegler-Nichols method (1st), *Improved* Ziegler-Nichols method (2nd), Cohen-Coon method, Genetic Algorithms and so on. For this work, the Ziegler-Nichols tuning method would be given a priority.

6.1 Ziegler-Nichols tuning methods

The Ziegler-Nichols method used was done based on obtaining the open loop transfer function and thereafter obtaining the necessary parameter values needed for the various evaluation of the P, PI and PID parameters.

The open loop step response is characterized by two main parameters, the L (delay time parameter) and T (time constant). These two parameters are computed by drawing tangents to the open loop step response at its point of inflections (basically two points. The inflection points are particularly done so that there would be an intersection with the vertical (voltage axis, which correlates with the steady-state value) and horizontal (time axis) axes.

Based on the Ziegler-Nichols, the following were derived to obtain the control parameters based on the required model

	PID Type	K_P	$T_I = \frac{K_P}{K_I}$	$T_D = \frac{K_D}{K_P}$
1.	P	$\frac{T}{L}$	∞	0
2.	PI	$0.9 \times \frac{T}{L}$	$\frac{L}{0.3}$	0
3.	PID	$1.2 \times \frac{T}{L}$	$2 \times L$	$0.5 \times L$

Table.2. Ziegler-Nichols PID controller parameters model [9]

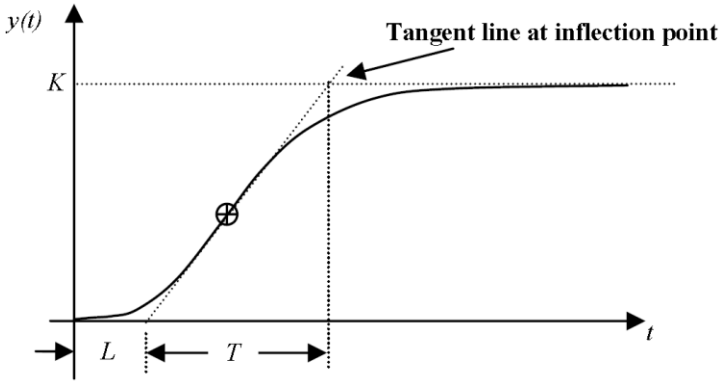

Figure. 8. Ziegler-Nichols step response tuning method [9]

From the Figur, the target is on how to evaluate the two parameters (L and T) needed. This is done as follows with the illustration.

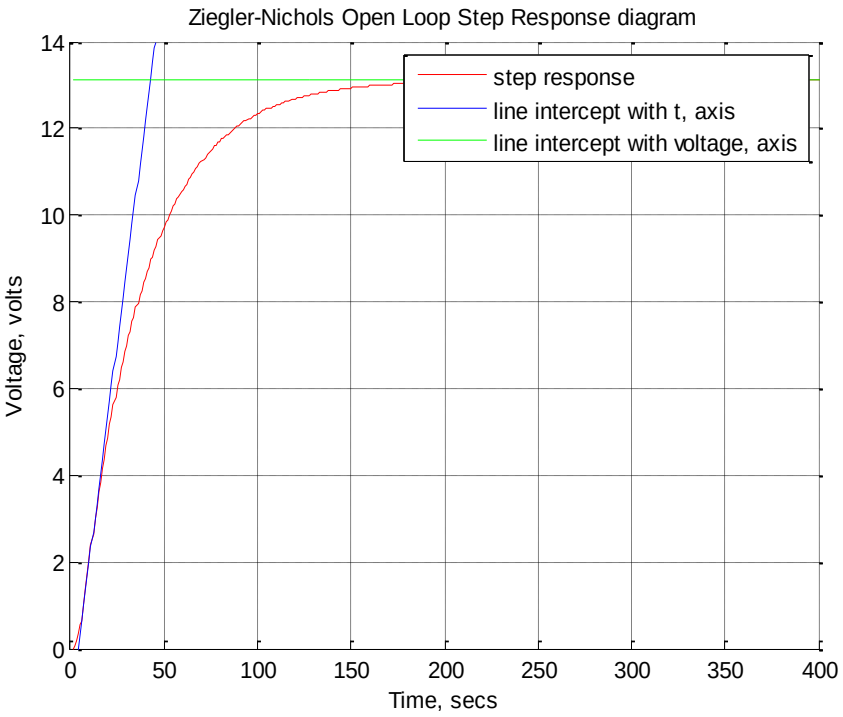

Figure .9 .Ziegler-Nichols open step response plot computation

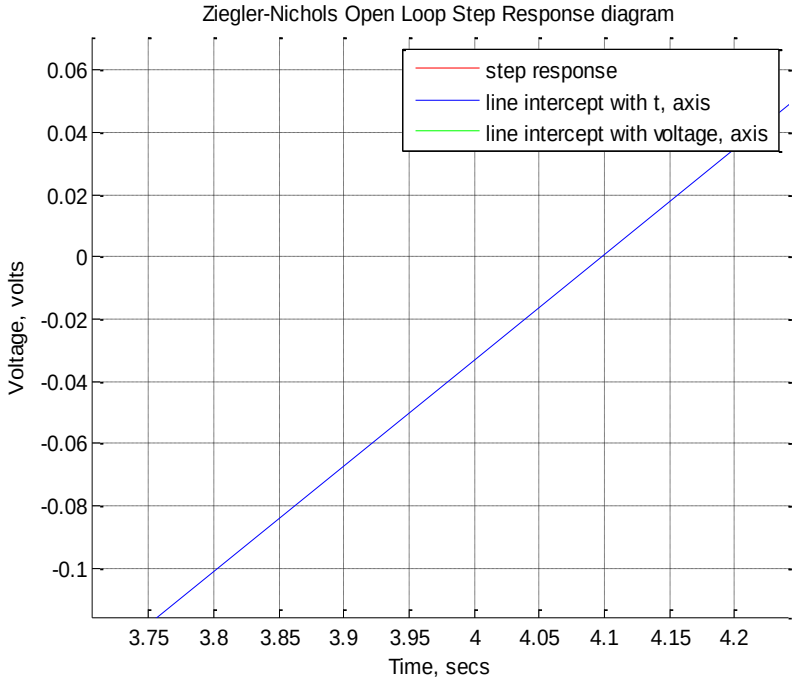

Figure.10. Ziegler-Nichols open step response horizontally zoomed

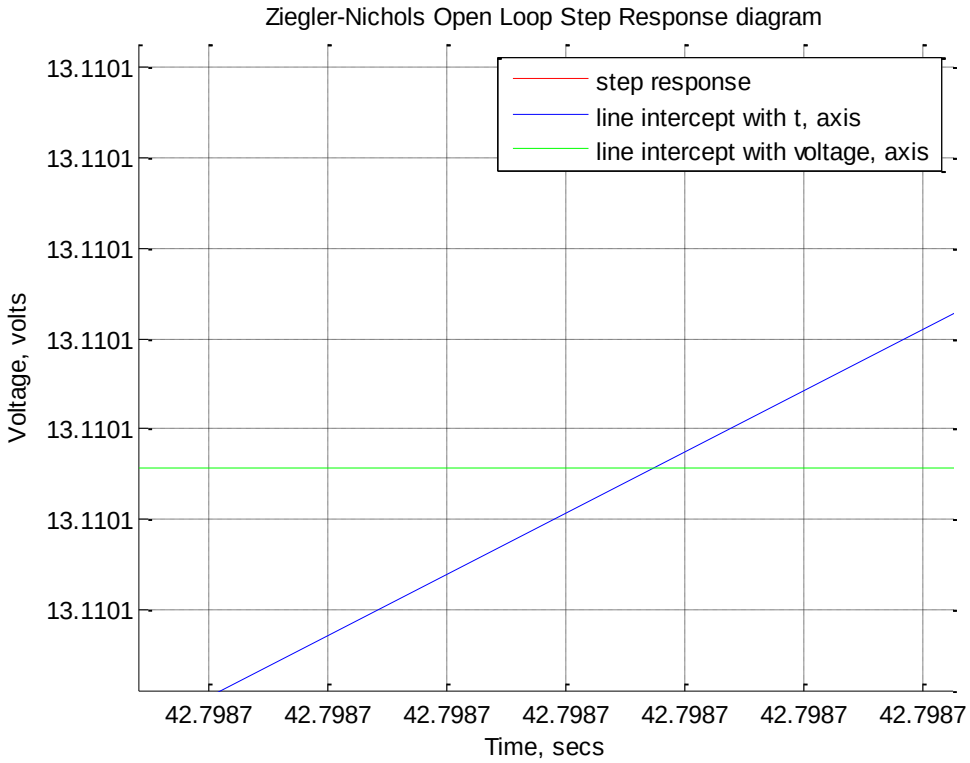

Figure.11. Ziegler-Nichols open step response vertically zoomed

Therefore, from the Figure , Figure and Figure6, the values of the L and T could be computed as follows:

Coordinate of the point of interception of the two lines $\approx$ (T^*, K) = (42.7987, 13.1101);

Where,

T^* is horizontal trace of the interception on the tangent lines drawn

L = 4.1;

K = 13.1101;

T = T^* – L = 4.1 = 42.7987 - 38.6987 $\approx$ 38.70

This implies that we have:

L = 0.0041;

K = 13.1101;

T = 0.0387

With the above computation, the P, PI and PID computation was done to get the best suited parameters combination desired.

So the updated table 9.1 would be table 9.2 shown below:

1.	P	9.439	∞	0
	PID Type	K_P	$T_I = \frac{K_P}{K_I}$	$T_D = \frac{K_D}{K_P}$
2.	PI	8.495	0.0137	0
3.	PID	11.327	0.0082	0.00205

Table.3. Results of the Ziegler-Nichols method for PID controller parameters

From table 9.2, the following parameters are obtained based on the equation format which is given below:

For PID only,

$$K_P + \frac{K_I}{s} + K_D \cdot s = 11.327 + \frac{1381.34}{s} + 0.0232 \cdot s \quad ($$

The outputs of the various PID combinations could be obtained as given below: using MATLAB Programming.

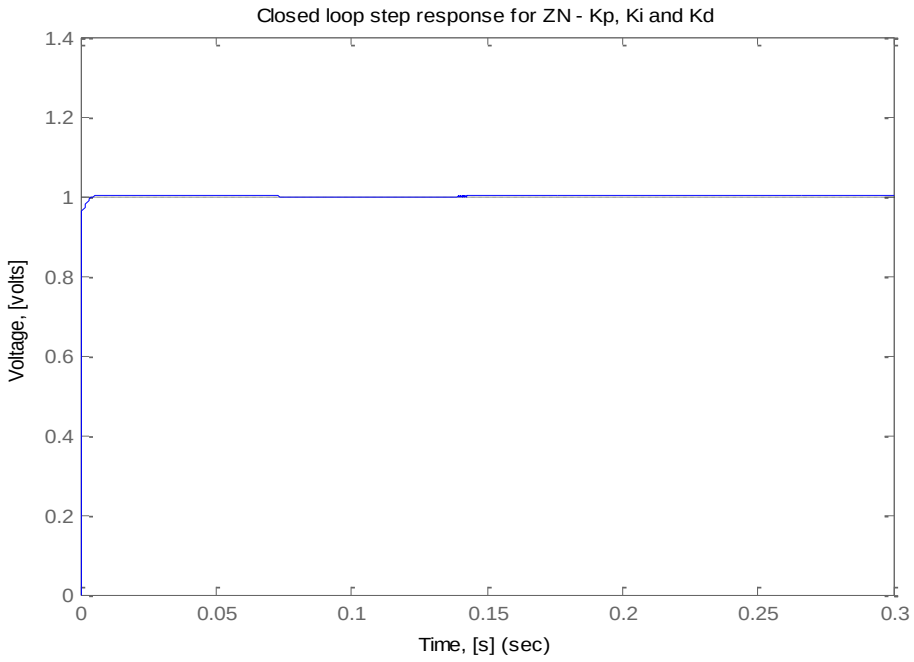

Figure.12. auto- scaled PID output for the Ziegler-Nichols tuning method

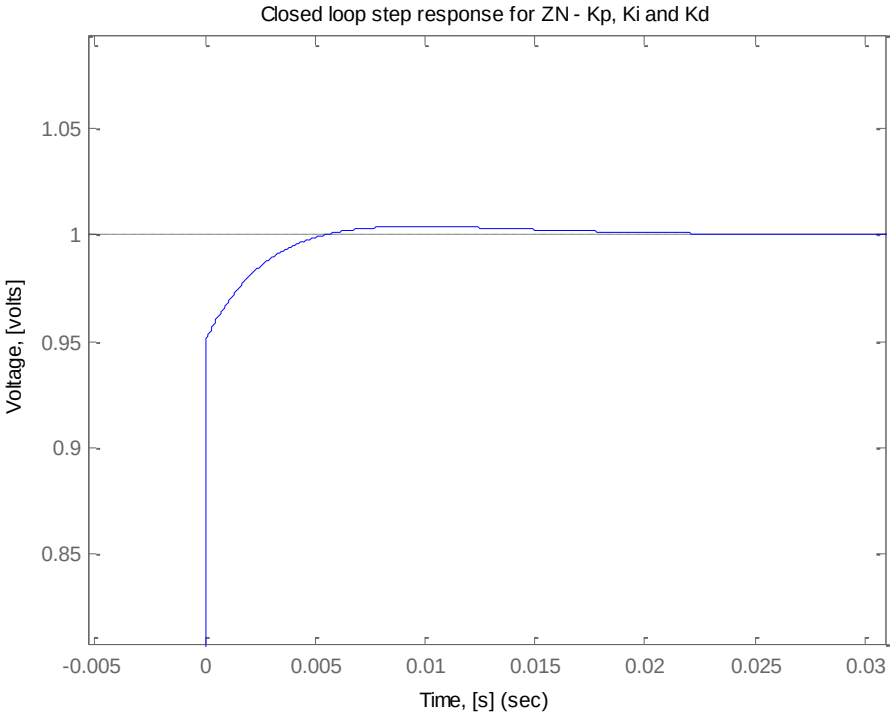

Figure.13. Auto-scaled PID output for the Ziegler-Nichols tuning method (zoomed overshoot point

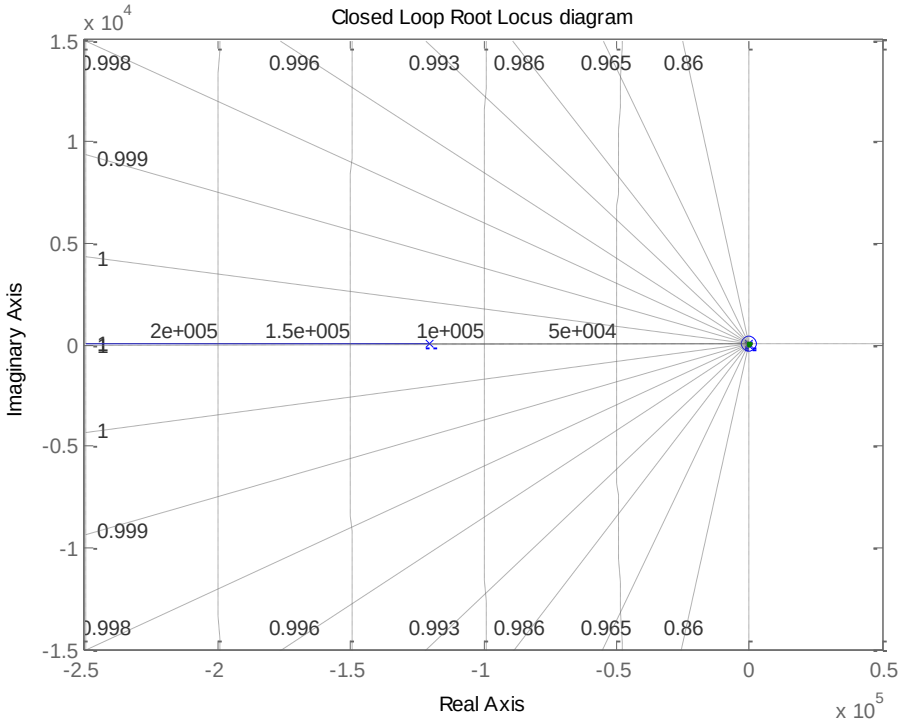

Figure.14. PID Ziegler-Nichols tuning method Root locus diagram

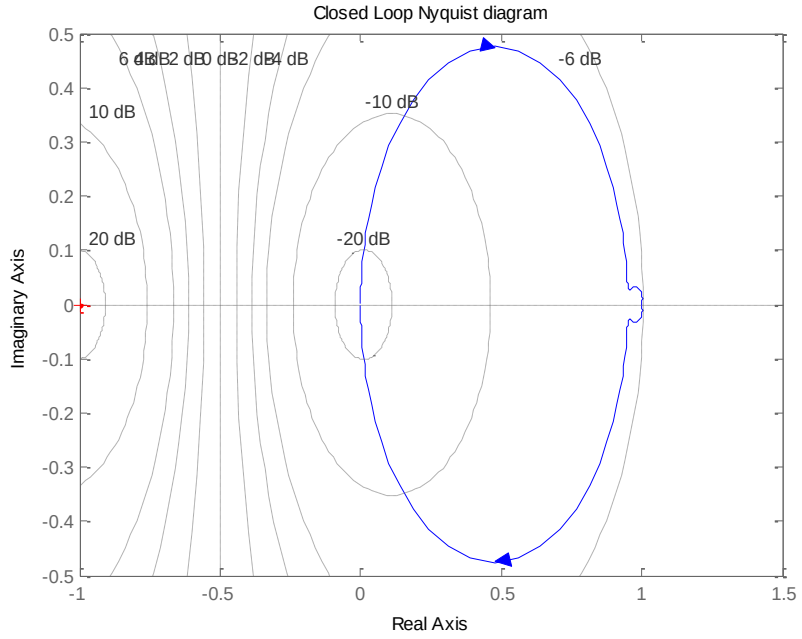

Figure.15. PID Ziegler-Nichols tuning method Nyquist diagram

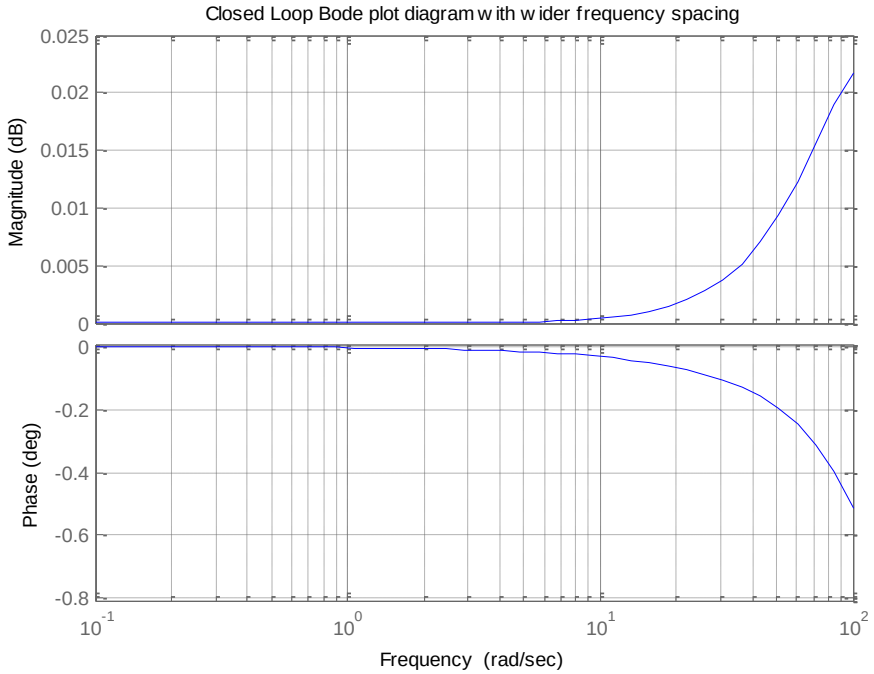

Figure.16. PID Ziegler-Nichols tuning Bode plot diagram

VII. CONVENTIONAL BLDC MOTOR DRIVE SYSTEM

For a three-phase BLDC application, a three-phase inverter bridge [10] is used. The typical inverter drive system for a BLDC motor is shown in Fig. 1. The three phase inverter operation can be divided into six modes (1-6) according to the current conduction states and conduction sequence. The switches in are operated such that each phase carries current only during the 120◦ period when the back EMF is constant.

Thus, there is a commutation event between phases for every 60◦ electrical, as shown in Fig. 2.Appropriate commutation therefore requires knowledge of rotor position, which can be directly detected using position sensors or estimated in sensorless manner by monitoring back EMF in the open phase . The three phase currents are controlled to take a form of quasi square waveform in order to synchronize with the trapezoidal back EMF to produce constant torque. This task is performed by speed/torque control loop in cooperation with rotor position sensor and hysteresis current controller.

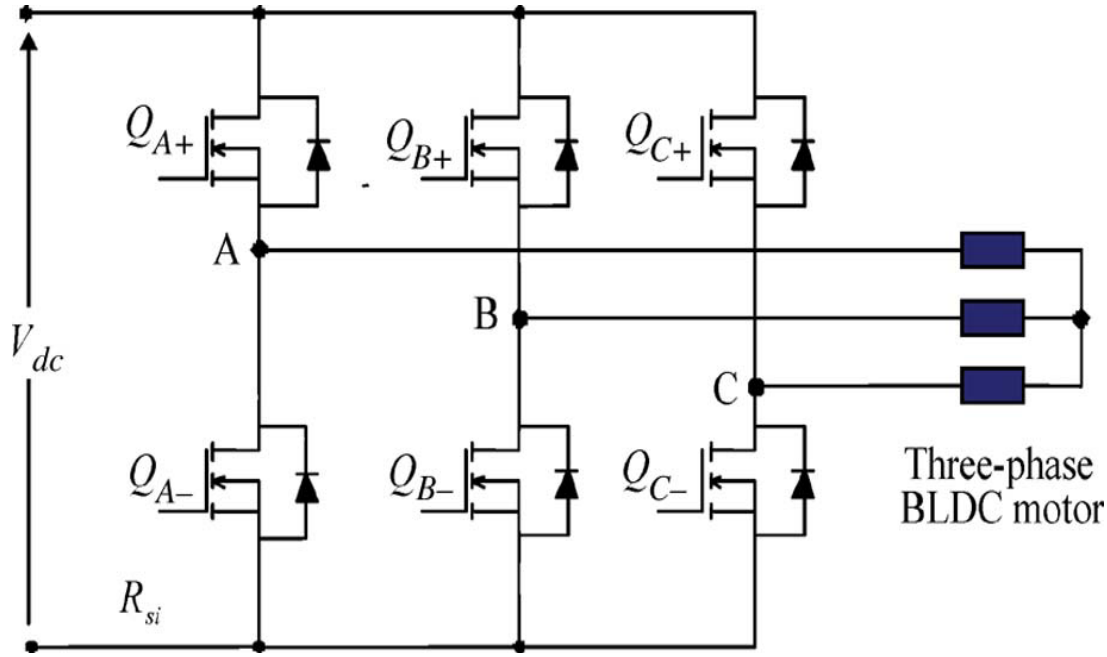

Fig.17. Typical inverter-fed BLDC motor drive system

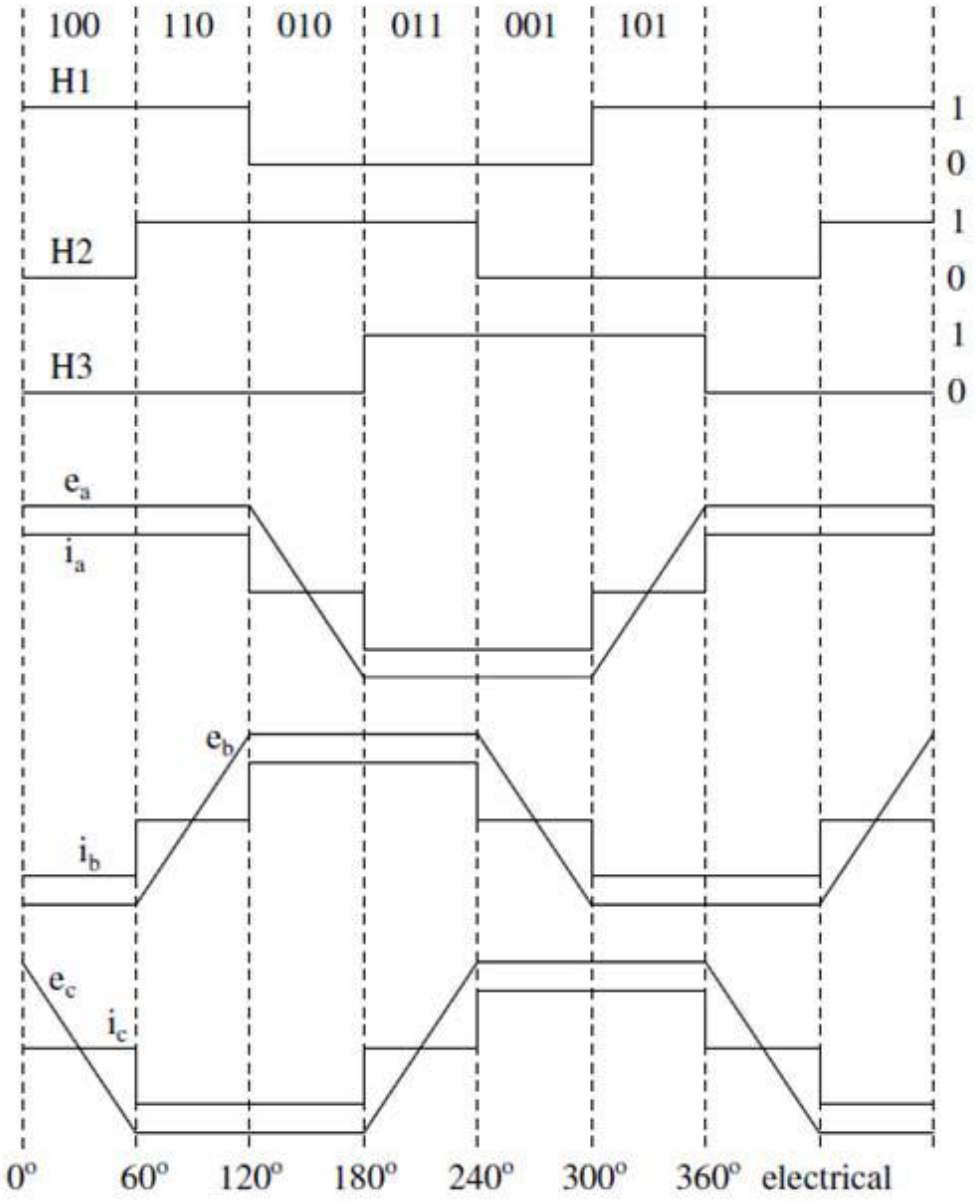

Fig.18. Hall sensor signal and Trapezoidal back EMF

Speed control in a BLDC involves changing the applied voltage across the motor phases [10]. This can be done using a sensored method based on the concept of pulse amplitude modulation, PWM, or hysteresis control. A common control algorithm for a permanent-magnet BLDC motor is PWM current control. It is based on the assumption of linear relationship between the phase current and the torque, similar to that in a brushed dc motor. Therefore, by adjusting phase current, the electromagnetic torque can be controlled to meet the requirement. Instantaneous current in the motor is regulated in each phase by a hysteresis regulator, which maintains the current within adjustable limits. The rotor position information is sensed to enable commutation logic, which has six outputs to control the upper and lower phase leg power switches. The current reference is determined by a PI regulator, which maintains the rotor average speed constant

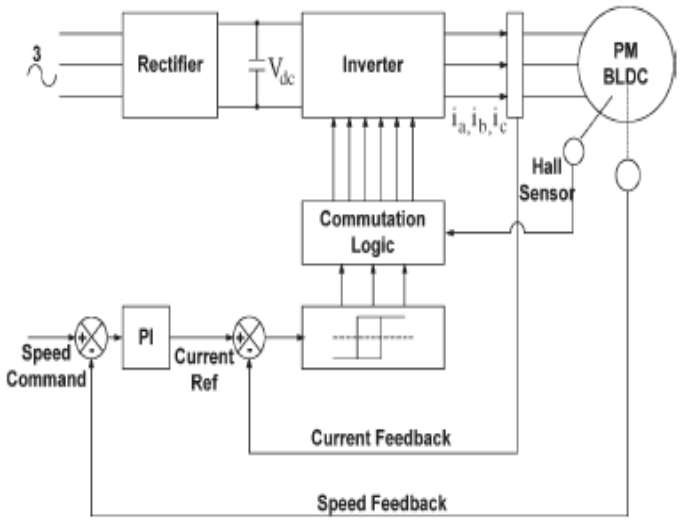

Fig.19. conventional PWM control scheme

VIII. DIGITAL PWM CONTROL SCHEME

The concept of this digital controller is very simple. Speed regulation is achieved by using two levels of duty cycles-a high duty (*DH*) and a low duty (*DL*)[11]. In essence, the controller treats the BLDC motor as a digital system, which may operate in two predefined states, namely state-1 and state-2[12]. These states are corresponding to two speeds *WL* (< than reference speed) and *WH* (> reference speed) and the speed regulation by appropriately altering the states.

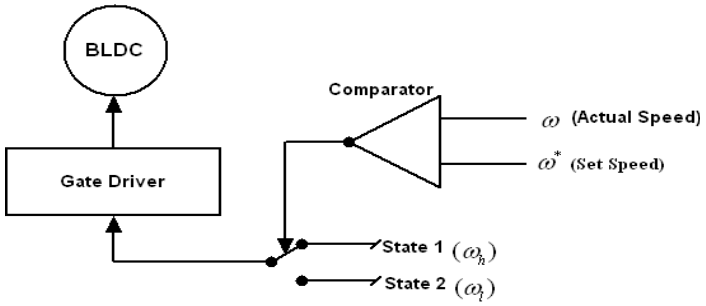

Fig.20. principal of operation of Digital controller

a) If the motor speed is less than commanded speed, then switch or stay at state-2 (*WH*)

b) If the motor speed is greater than commanded speed, then switch or stay at state-1 (*WL*)

Unlike a hysteresis current controller, a PWM control does not have an inherent current control capability. Hence, a current limiter has to be introduced [11]. A proportional controller provides the reference for the current limit. The current is always made to stay within a maximum and minimum limit. The maximum value of *I limit* is 1.5 times the rated motor current. This is because motors can handle 1.5 times the rated current for a short duration of time. The minimum value of *Ilimit* is defined as the ratio of a percentage (1%) of the rated torque to the torque constant.Fig.5. shows the block schematic of the digital PWM control of BLDC motor drives.

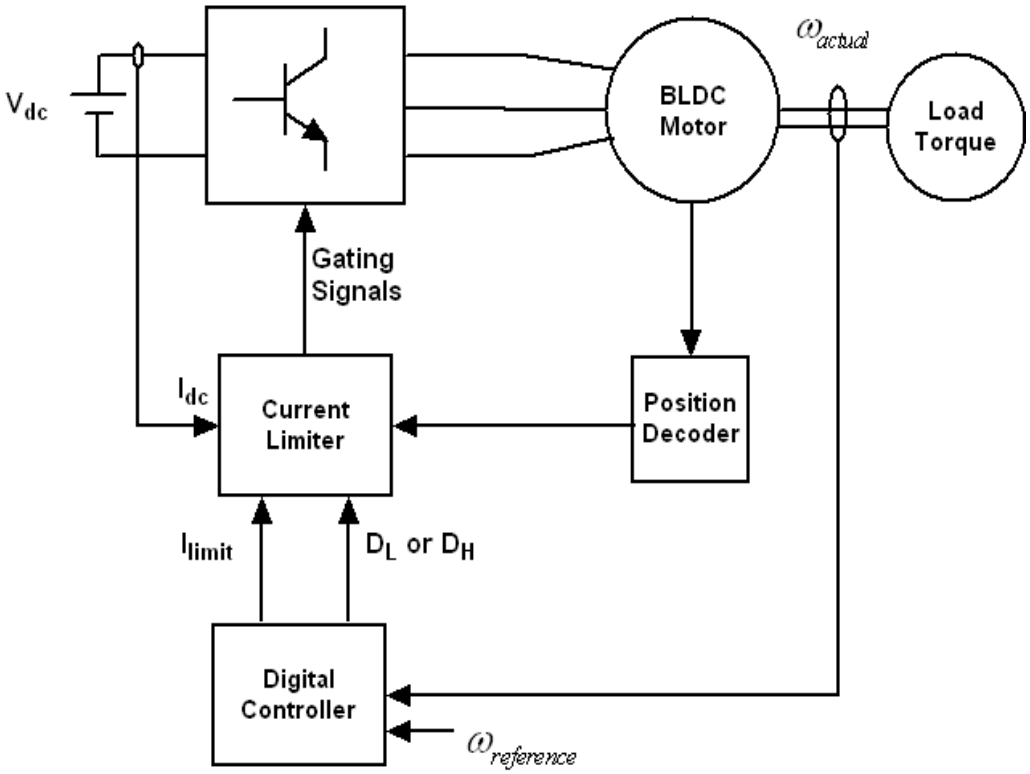

Fig.21. proposed Digital Control

IX. CONTROLLER DESIGN

The value of the duty ratio *D* can be obtained from the electrical and mechanical equations[11]. The value of *D* can be expressed as a function of the motor parameters. From the torque equation, we have

Tem= Jdw÷dt + Bw + Tl ……. (9.1)
Tem = Ktl …….. (9.2)

Where *Te*, w(t), *B* , *J* and *TL* denote developed electromagnetic torque, rotor angular velocity, viscous friction rotor moment of inertia and load torque respectively.

Equate (9.1) and (9.2), we get
Ktl = Jdw÷dt + Bw + Tl ……. (9.3)

where *Kt* = torque constant and *I* = average current. At steady state, (3) can be written in terms of steady-state angular velocity Wss as

I(wss) =(BWss+ Tl)÷Kt …….. (9.4)

At steady state angular velocity *Wss* , phase voltage *Van* can be expressed in terms of phase current I, winding resistance and velocity constant *ke*, ie given by

Van = IRa + KeWss …… (9.5)

The phase voltage in terms of dc-link voltage *V*dc and duty ratio *D* is
Van = DVdc …… (9.6)

Substituting the value of the steady-state current from (9.4) and phase voltage from (9.6) in (9.5),we get the value of duty ratio
D = [(Tl+BWss)÷Kt + KeWss]÷Vdc ……. (9.7)

By considering *WL* and *WH* we can get the *DL* and *DH* respectively.

The maximum deviation from the reference speed (*W**) due to the application of high duty *DH* is denoted by $\Delta\omega H$, and the maximum deviation from the reference speed due to the application of a low duty *DL* is denoted by $\Delta\omega L$. The speed response can be expressed as

W(t) = (Tem- Tl)÷B + [W-(Tem- Tl)÷B]e-(J\B)t …… (9.8)

X. CONCLUSION

The modeling and the simulation of PID control of BLDC motor speed and its torque results are tested.it started with the analysis and reasons why an absolute précised control is important in drives particularly the BLDC motor and then the mathematical modelling.in this paper the open loop analysis is done by considering the stability factor and making the necessary plot for this analysis. some of the plots include the step response, root locus, nyquist diagram, and bode plot diagram. Also study the different schemes of pwm controlled BLDC motor.

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Cysticercotic Encephalitis: A Case Report

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Abstract- Although encephalitic presentation of cysticercosis is rare, it should be included in differential diagnosis of acute encephalitic cases especially in endemic countries like India. Cysticercotic encephalitis is commonly seen in young females who typically present with features of raised intracranial pressure along with compromise of visual function. Early diagnosis and appropriate therapy can result in better outcome in this rare but commonly fatal neurological disorder.

Index Terms- neurocysticercosis encephalitis intracranial hypertension cortical blindness

I. Introduction

To our knowledge, this is the first report of a case of cysticercotic encephalitis in a child with few parenchymal lesions unlike previous rare reports of encephalitic presentation with miliary neurocysticercosis.

It re-emphasis that neurocysticercosis should be included in differential diagnosis of acute encephalitic cases especially in endemic countries like India.

It demonstrates the safety of cysticidal therapy even in presence of raised intracranial pressure caused by neurocysticercosis.

Cortical blindness is rarely reported as presenting feature of neurocysticercosis.

II. Case Summary

An 8 year old girl presented with complaints of bilateral frontal headache for 10 days, progressive loss of vision for last 4 days, one episode of seizure (right complex partial seizure) followed by altered sensorium 15 hours prior to admission. There was no history of seizures in the past, trauma, recent immunization, ear discharge or tuberculosis contact. The child on admission was afebrile, had GCS E2M3V2 along with frequent tonic posturing. Her pupils were normal in size and reacting normally to light. There were no signs of meningeal irritation. On motor examination, all four limbs had power of grade III/V, exaggerated deep tendon reflexes and right extensor plantar response. Rest of the systemic examination was essentially normal. Fundus examination revealed bilateral papilledema.

The child was intubated in view of low GCS and phenytoin loading dose (20mg/kg) followed by maintenance (5mg/kg/day) was administered along with mannitol and supportive therapy for intracranial hypertension. On investigation, her CBC, blood glucose, electrolytes, liver, renal function tests and blood ammonia levels were normal. Tuberculosis screening (Mantoux and CXR) was negative. A possibility of non convulsive status epilepticus was also considered. However, EEG showed low voltage recording suggestive of encephalopathy. CECT head done on day of admission showed hypodensity in left parietal region with no visible underlying lesion. MRI brain revealed a large hypointense lesion on T1W images in left occipito-parietal region which was hyperintense on T2W and FLAIR images. Small nodular lesions were seen in bilateral parietal regions as well which were hyperintense on T2W and FLAIR images suggestive of neurocysticercosis (NCC) with massive perilesional edema (Fig 1). Cerebrospinal fluid showed 8cells/mm^3 (100% lymphocytes) with protein, 45 mg/dl and sugar 56 mg/dl. CSF culture was sterile and CSF Adenosine Deaminase (ADA) levels were normal. Serum anticysticercus IgG antibody was positive. A final diagnosis of cysticercotic encephalitis was made.

Child was started on intravenous dexamethasone. Features of intracranial hypertension became passive 48 hours after admission. Her sensorium gradually improved and was extubated after 60 hours of admission. She was noticed to have right hemiparesis with decreased visual acuity (projection of rays, PR+ and perception of light, PL+ with no finger counting). Her pupillary reflex was normal. Dexamethasone was stopped and oral prednisolone (2mg/kg/day) was continued for 2 weeks. Albendazole therapy (15mg/kg/day) was started on 5^{th} day of steroid therapy and continued for 4 weeks. She had no seizures during her hospital stay.

At 3 months, minimal residual right hemiparesis was evident but cortical blindness persisted. Repeat neuroimaging revealed decrease in size of lesions with complete resolution of perilesional edema.

III. Discussion

Despite common occurrence of parenchymal neurocysticercosis, encephalitic presentation is rare. Typical manifestations include features of raised intracranial pressure along with compromised visual function. This encephalitic presentation of cysticercosis is more commonly seen in females, especially young females.[1]

To our knowledge, this is first case report of encephalitic presentation of cysticercosis in presence of few parenchymal lesions. Only a handful of clinical reports have been published of this relatively rare subacute cysticercotic infestation of the brain.[1-3] All these were seen in cases with multiple cysticercosis lesions unlike the present case with only three intraparenchymal cysts.[1-4] This case re-emphasis that neurocysticercosis should be included in differential diagnosis of acute encephalitic cases especially in endemic countries like India. In addition, encephalitic manifestation of neurocysticercosis doesn't seem to be related to presence of multiple lesions but, perhaps, significant perilesional edema. Neurocysticercosis resulting in cortical blindness has been rarely reported[2,5] though NCC related stroke

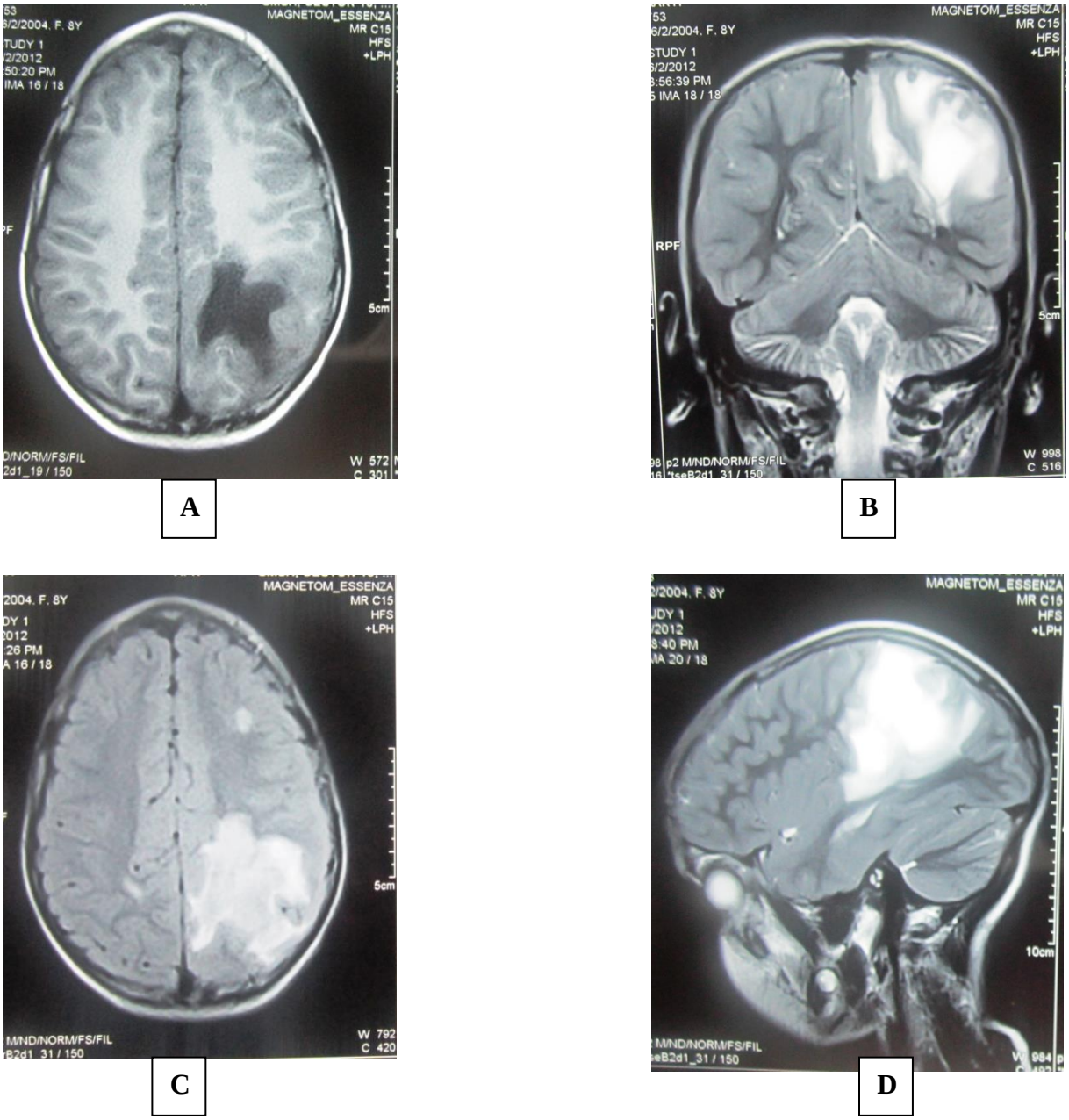

Fig 1 Legend: MRI Brain with axial T1W (A), axial FLAIR (B), coronal T2W (C) and sagittal T2W (D) cuts showing a large hypointense lesion on T1W images in left occipito-parietal region which is hyperintense on T2W and FLAIR images. Small nodular lesions are seen in bilateral parietal regions as well which are hyperintense on FLAIR images (B).

is a well established entity.[6] Cysticidal therapy has been often considered as contraindicated in presence of raised intracranial pressure.[7.] However, appropriate management of raised intracranial hypertension and judicious use of corticosteroids followed subsequently by cysticidal therapy proved effective as well as safe in this patient.

To summarize, cysticercotic encephalitis should be considered as differential for cases presenting as encephalitis in endemic countries like India. Prompt diagnosis and early as well as appropriate therapy can result in better outcome in this rare but commonly fatal neurological disorder.

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FPGA Implementation of Hybrid Test Data Compression Method Using Scan Chain Compaction and Dictionary-based Scheme

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***Abstract*-** The Goal in this paper is proposal of new test data compression method for reducing test data volume and test application time. The existing system in this paper is group the scan chain in the ascending order and it must be in odd number for group and the number of element in the groups. The proposed method consists of two steps: scan chain compaction (even or odd) and dictionary-based compression scheme. The scan chain compaction provides a minimum scan chain depth by using compaction of the compatible scan cells in the scan chain. The focus is to compacted scan chain is partitioned to the multiple internal scan chains it may be odd or even for using the fixed-length index dictionary-based compression scheme that provides the high compression ratio and the fast testing time. The proposed compression methods scan chain which is not arranged in ascending order and not in odd number. Hence it produces the output with four groups and it is implemented in FPGA using VHDL coding, where the existing system produces five groups, system delivers compressed patterns from the ATE to the chip and drives a large number of multiple internal scan chains using only a single ATE input and output. Experimental results for the test benches show that the test data volume and testing time for the proposed method are less than previous compression schemes.

***Index Terms*-** test data compression, test application time, full-scan circuit, scan chain compaction, dictionary-based compression scheme

I. INTRODUCTION

As the complexity of very large scale integration (VLSI) circuit increases, testing plays an important role in today's system design. One of the most important factors in driving up the test cost is increasing the amount of test data volume, which is a result of the large size of the designs and the new types of defects appearing in the advanced manufacturing process. A large amount of test data must be stored in the automatic test equipment (ATE) and transferred deep into the chip as fast as Possible. Since the channel capacity and the size of memory of ATE are limited, the test application time and the test power have been significantly increased. The test data compression technique offers a promising solution to the problem of increased the test data volume for SOC testing.

Several test data compression schemes have been proposed for reducing test data volumes. These previous schemes of compression techniques can be categorized into two groups: code-based schemes and structural schemes. In the code- based schemes, a number of techniques based on FDR coding, VIHC coding, RF-Huffman coding, and Selective-Huffman coding, have been proposed to reduce test data volume. Although these compression methods can achieve high compression ratio, the test application time is not sufficiently reduced, because of the long shifting time to drive into the full-scan chain. Additional synchronization and handshaking between the SOC and the ATE can be occurred in decompression decoder by using multiple clocks to increase testing speed. Several dictionary-based compression methods have been presented to reduce SOC test data volume. To get high compression results, these methods need the memory or large size hardware circuit to store dictionary entries.

In the structural compression method, the Illinois scan architecture use an external input port to drive multiple internal scan chain inputs without any logic gates; it provides the reduction of test data volume and testing time. However, additional fault simulation and test generation are necessary to achieve high fault coverage. Additionally, if the faults are hard to be detected by using the parallel mode of the Illinois scan architecture, the test vector needs to be driven serially into a single long scan chain. The scan chain reconfiguration method has been proposed in the form of an XOR-based. This method identifies the compatibility of scan cells in the scan chain, and then the scan chain network is constructed by XOR gates The scan chain has a shorter scan chain depth than a traditional scan chain. It can provide the reduction of test data volume and test application time simultaneously. However, if the test data cube does not contain enough compatible scan cells, the effect of reducing scan chain depth is restricted.

In this paper, we propose a new effective test data compression method using the scan chain compaction and the dictionary- based compression scheme. The first method, the scan chain compaction constructs a scan chain network which has minimum number of scan slices; the number of scan slices is scan depth, similarly in [11]. The small size of scan chain depth can reduce total data volume and test application time; because it needs less scan chain shifting operation to drive test data bits into the scan chain. However, the reduction of scan chain depth by the compaction is limited, if the test vector does not have enough compatible scan cells.

To achieve high compression ratio, we combine the dictionary-based compression scheme with the scan chain compaction. The scan chain network composed by the scan chain compaction is divided to the multiple internal scan chains to

reduce the scan shifting time using the dictionary-based compression scheme. The decoder of proposed dictionary-based scheme drives the test vector in to the whole multiple internal scan chain inputs at the same time, as it can help to reduce the test data volume and test application time. The hardware overhead can be kept reasonable size because we use the dictionary which has fixed-length indices and the small amount of the index entries which are already compressed by the scan chain compaction. The proposed compression scheme uses the test data vector generated by the traditional ATPG and delivers high compression ratio and fast application test while retaining the original faults coverage without any additional fault simulation and test generation. The proposed compression architecture is implemented with only one single scan-in and scan-out pin externally, although the multiple scan chains are consisted internally. Experimental results for large ISCAS-89 benchmark circuits show that proposed compression method could reduce test data volume and test application time significantly.

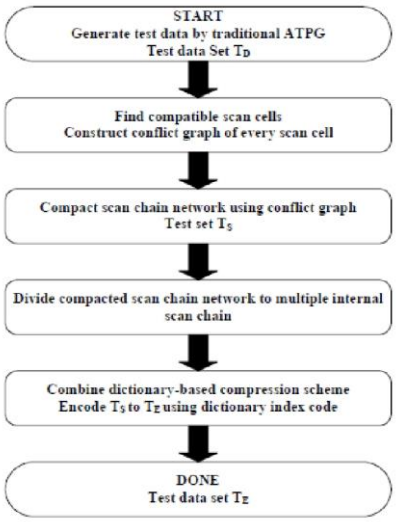


Fig. 1 Overall framework of the proposed scheme

II. OVERVIEW OF PROPOSED SCHEME

A. Definitions:

To present the proposed test data compression algorithm simply, several definitions are used

Definition 1. Scan Frame: A scan frame is a vector of inputs applied to the same scan cells in each test pattern,

Which contains the number of 0s, 1s and don't-cares (Xs).

Definition 2. Direct compatible: If scan frames of two scan cells in the scan chain never receive conflicting binary values in totality for the same test cube, these two scan chain are direct compatible, and can be driven the same test data.

Definition 3. Inverse compatible: If scan frames of two scan cells in the scan chain never receive identical binary values in totality for the same test cube, these two scan chain are inverse compatible, and can be driven the same test data by inserting inverter on the scan path.

B. Overall algorithm of the proposed scheme:

The overall framework of our compression scheme for scan based on circuit is shown in Fig. 1. The proposed scheme uses the test data cube of the full-scanned circuit, the test cube is generated by traditional ATPG without any additional fault simulation and test generation. Since the test vectors generated by the ATPG have large number of don't care bits, there are many compatible or inverse compatible scan cells in the full-scanned circuit. All compatible or inverse compatible scan cells are identified and composed to groups of scan cells which have same test values for all test vectors. To find the maximal number of scan cells in a group, the conflict graph for all scan cells is used. The scan chain can be constructed to the compacted scan chain network using the conflict graph of scan cells. The compacted scan chain network has a shorter scan chain depth than a conventional scan chain. The short scan chain depth can provide the reduction of the test data volume and test application time simultaneously. Although the scan chain is compacted to the scan chain network with short scan depth using the conflict graph, the reduction of scan chain depth is limited when the test vector does not have enough compatible scan cells. To alleviate this limit and achieve high

Compression ratio, the new dictionary-based compression scheme is combined with the scan chain compaction. To use the dictionary-based compression scheme, the compacted scan chain network is divided to the multiple internal scan chains. Each slice vector of the multiple internal scan chains is driven into the all scan chain inputs at the same time. The slice vector has a fixed length and which is a symbol of the dictionary. The symbol is encoded to the fixed length dictionary index that is shorter than the length of the symbol. Therefore the proposed dictionary-based compression is the vertical compression that presents not only the reduction of the test data volume and application time but also the use of only a single ATE channel. The original test vector TD can be compacted and encoded to the significantly compressed test vector TE by combining the scan chain compaction and the dictionary-based compression scheme.

III. SCAN CHAIN COMPACTION

The proposed scan chain compaction methodology aims at constructing a scan network with minimum scan depth. Such a solution perfectly exploits the structural relationship between scan cells to deliver the highest compression ratio while retaining the original faults coverage. There are many scan cells which are direct or inverse compatible with other scan cells in the full scan chain. These scan cells compose the compatible scan cell group,

all scan cells in the same group are always driven identical bits (direct compatible) or never identical bits (inverse compatible). All scan cells in the same group can be injected the same test data value at the same clock cycle. Therefore the compacted scan chain network can be constructed and each compatible scan cell group becomes one depth of the scan chain network.

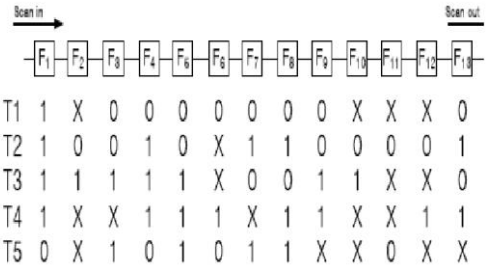


Fig. 2 Sample of single scan chain and test vector

The scan network expands and compacts depending on the compatible scan cell group sizes resulting of varying numbers of scan cells. To construct the scan chain network, the types of expansion and compaction mechanisms within the scan network are defined.

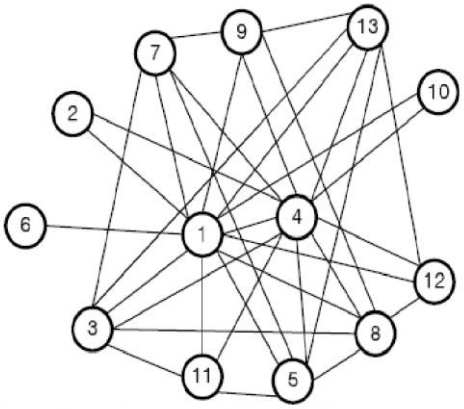
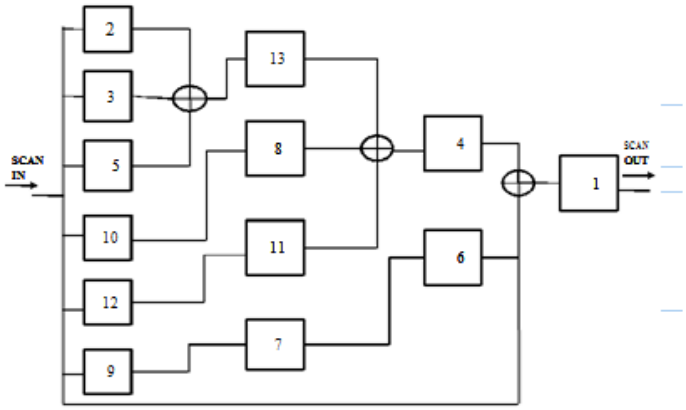


Fig. 3 Conflict graphs for scan cells in Fig 2.

The expansion mechanism in the form of a fan-out network allows in scan cells within the compatible scan cell group to always receive identical bits, although the group sizes may vary. The expansion expands the test data value from a single scan-in pin or small size groups to the large size groups with the fan-out network. The compaction mechanism uses XOR gates to compact the test data value from the larger size group to the next smaller size group. A series of such expansions and compactions construct the compacted scan chain network between a single scan-in and scan-out pin. The number of the compatible scan cell groups between scan-in and scan-out pins constitutes the scan depth, and as many shift cycles will be required for each test vector. We present two rules on group sizes to ensure the delivery of always identical bits to the scan cells within the same group, and avoid certain aliasing problem in the expansion and compaction mechanisms. First, XOR gates with only an odd number of inputs are used to compact a group into the subsequent group to avoid constant value of generation by the XOR gates. Secondly, the number of outputs of every expanding fan-out should be odd in order to avoid self-aliasing According to these two rules, each compatible scan cell group should consist of an odd number of scan cells. To achieve the highest compression ratio, the number of depth in the scan chain network is minimized, that depends on the number of compatible scan cell groups. The conflict graph is used to construct the most compact scan chain network. The conflict graph represents the relationship between scan cells for all scan cells in the scanned circuit. The vertices of the graph represent each scan cell, an edge between two vertices exists if and only if two scan cells corresponding to the vertices are incompatible that means these two scan cells cannot belong to the same group. Fig. 3 shows the conflict graph for all scan cells of the single scan chain in Fig. 2.As mentioned in Section II, scan-cell compatibility can be the form of direct or inverse compatibility. Although the proposed method so far has concentrated on the direct compatible case, inverse

Compatibility is also equally beneficial. If two inverters are inserted into the scan path, the inverse compatible can still be treated as compatible.



T1	1	0	0	0
T2	1	0	1	1
T3	1	1	1	0
T4	1	1	1	1
T5	0	1	0	1

Fig.5.sample of constructed scan chain network

Then we construct the conflict graph of these scan cells showed in Fig. 3 to find the compatible scan cell groups by the use of matching lines, {F2, F3 F5, F9, F10, F12}, {F7, F8, F11, F13}, {F4, F6}, {F1}. These ordered scan cell groups are stitched via expanding fan-outs and compacting XOR gates, the sample compacted scan chain network produced by the proposed scheme are given in Fig. 5. The compacted scan network in Fig. 5 reduces the depth of scan chain from 13 to 4, then the test data volume is reduced to 20 (=4* 5) bits. Similarly, the test application time for each test vector is reduced from 13 clock cycles to 4 clock cycles. The area overhead of the compacted scan chain network is 3-input XOR gates that are used for the compaction mechanism.

IV. DICTIONARY-BASED COMPRESSION SCHEME

The compacted scan chain network, which is mentioned in the previous section, reduces the scan depth of the scan chain, and reduces the test cost. However, if the test data cube does not contain enough compatible scan cells, the effect of the Short scan depth is restricted. Then we combine the dictionary- based compression scheme with the compacted scan chain network to achieve the high compression ratio and the fast application time. To address the dictionary-based compression scheme, the compacted scan chain network which has a single scan-in and single scan-out, is partitioned to the multiple scan chains. We divide the scan network to proper number of scan chains, and balance the depth of each scan chain. If the scan chains are unbalanced, additional dummy scan cells are inserted to make that all the scan chains have the same scan depth. The multiple scan chain composed from the compacted scan chain network presents the reduction of the test application time, however the test data volume is unchanged and additional scan inputs and outputs are required. The hardware overhead for the dictionary logic is a weak point of the proposed scheme like other dictionary-based schemes. However, the hardware overhead can be kept reasonable size because we use a small amount of the test data vector which is already compressed by the scan chain compaction and the dictionary has fixed-length indices.

The proposed dictionary-based compression scheme compresses the test data cube of the multiple scan chains in vertically. The decompress decoder of the scheme receives the compressed test data from the ATE, decompresses the compressed test data, and drives the test vector into the all multiple scan chain input at the same time, it can help to reduce the test data volume and test application time. Fig. 6 illustrates the block diagram of the proposed compression scheme dictionary-based test data compression scheme. In the scheme, an ATE scan-in is driven into the m multiple scan chains of the compacted scan network through the dictionary-based decoder.

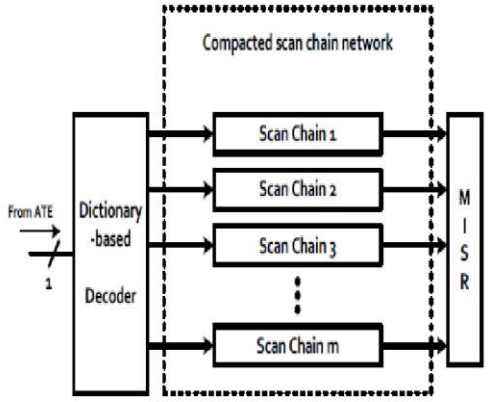


Fig. 6 Illustration of the proposed compression scheme

Each scan slice vector becomes an m-bits symbol, which is encoded to codeword. During the test application, the codeword is shifted into the decoder, after that m-bits symbol is immediately generated by the decoder and driven into the multiple scan chain inputs. The test responses shifted from each scan chain outputs can be compacted by using MISR or other techniques. The proposed decompression architecture assume a l-to-m scan configuration, in which the number of internal scan chains is m times the number of external scan I/O pins, so significant reductions in the test data volume and test application time can be achieved. The dictionary logic is implemented by using the combination logic circuit. The size of dictionary depends on the length of dictionary index.

Moreover, the proposed dictionary-based compression scheme does not require multiple clock cycles and additional synchronization and handshaking between the CUT and the ATE.

V. CONCLUSION

In this paper, we proposed an effective hybrid test data compression method using the scan chain compaction and the dictionary- based compression scheme. We have shown how proposed test data compression method can be used to reduce the test data volume and test application time for SOC. The Proposed compression method delivers compression patterns from the ATE to the chip and drives a large number of multiple internal scan chains using only a single ATE input and output. The test data volume is compressed twice with the scan chain compaction and the dictionary-based compression, then they can be achieved the highest test data compression ratio and the fastest application time. The proposed compression scheme does not require multiple clock cycles, additional synchronization, and handshaking between the circuit and the ATE. Experimental results for the ISCAS -89 test benches show that the test data volume and testing time for the additional synchronization, and handshaking between the circuit and the ATE. Experimental results for the ISCAS- 89 test benches show that the test data volume and testing time for the proposed

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Evaluation of Bioactive Phytoconstituents in *Linum Usitatissimum l.* by GC-MS

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Abstract- In this study, flaxseed was subjected to identification of bioactive compounds by using Gas Chromatography-Mass Spectrum technique. Flaxseed was extracted with 99%ethanol. Extracted sample was injected, according to, the bioactive compounds are screened. The results revealed the presence of seventeen compounds of which squalene (45.27 %) and 9, 12, 15-octadecatrienoic acid (z,z,z) (24.67) were the phytocomponents with high peak areas.

Index Terms- Bioactive Compounds, GC-MS, Flaxseed, Phytocomponents.

I. INTRODUCTION

Flaxseed is the seed from the flax plant (Linum usitatissimum L.), which is a member of the Linaceae family. The plant is not a new crop and native to West Asia and the Mediterranean[2]. Common flax was one of the first crops domesticated by man. Flax originated in India, and from its hardiness and usefulness, is generally diffused over the globe. Ancient centres of flax-growing are mountainous areas of India and China; In India flax was cultivated as a fiber crop earlier than cotton. As early as in the 4th or 5th millennium B.C. flax was cultivated for its fiber in Mesopotamia, Assyria and Egypt. Wild narrow-leaved flax and semi-cultured procumbent flaxes grow in Transcaucasia. Many monuments of Ancient Egypt reflect cultivation of flax and spinning and weaving of its fibres. In Russia flax has been cultivated since the birth of the Russian nation[3].

Annual: stem-cylindric, erect, simple below, 0.6-1.2 m, often solitary corymbosely branched above: leaves narrow sub-3-nerved, linear or lanceolate without stipular glands: flowers 2.5cm diam, in broad cymes: sepalsovate, acuminate, 3-nerved, glandular, margins ciliate or not, petals blue, styles quite free: stigmas linear-clavate: fruit---is a rounded capsule about one – fourth inch or more in diam, surrounded by the persistant sepals. It possesses 10 locules, each with one seed. The seed is oval, lenticular, and pointed at one end. Its length varies between 4 and 6 mm. and breadth between 2 and 3 mm. Each capsule contains ten seeds[5].

Flaxseed (Linum usitatissimum L.)Diet has been suggested to be the most important environmental factor. Flaxseed (FS), a whole grain used as a nutritional supplement, has gained popularity because it is a rich source of natural antioxidants. FS has high concentrations of omega-3 fatty acids and lignans. Omega-3 fatty acids reduce inflammation and may be helpful in treating a variety of autoimmune diseases[1]. By virtue of the presence of physiologically active food components that may provide health benefits beyond basic nutrition, flaxseed is often grouped into one of several categories: ‘‘functional food’’, ‘‘bioactive food’’ and an ‘‘endocrine active food’’[4]. Hence, the main objectivity of this study to identify the phytocomponents to do the further research parts in cancer

II. RESEARCH ELABORATIONS

Extraction procedure: seeds of flaxseeds were bought from Nilgris market, Coimbatore. 10gm powdered flaxseed was soaked in 20ml of Absolute alcohol overnight and then filtered through a Whatman® No. 41 filter paper (pore size 20 - 25_m) along with 2gm Sodium sulfate to remove the sediments and traces of water in the filtrate. Before filtering, the filter paper along with sodium sulphate was wetted with absolute alcohol. The filtrate is then concentrated by bubbling nitrogen gas into the solution and was concentrated to 1ml. The extract contains both polar and non-polar phytocomponents.

Gas Chromatography–Mass Spectrometry (Gc/Ms) Analysis

GC/MS analysis of this extract was performed using a Perkin Elmer GC Claurus 500 system and Gas Chromatograph interfaced to a Mass Spectrometer (GC/MS) equipped with a Elite-1 fused silica capillary column (30 m × 0.25 mm ID. ×1 µMdf, composed of 100% Dimethyl poly siloxane). For GC/MS detection, an electron ionization system with ionization energy of 70 eV was used. Helium gas (99.999%) was used as the carrier gas at a constant flow rate of 1 ml/min. and an injection volume of 2 µl was employed (split ratio of 10:1). Injector temperature 250°C; Ion-source temperature 280°C. The oven temperature was programmed from 110°C (isothermal for 2 min.), with an increase of 10°C/min, to 200°C, then 5°C/min to 280°C, ending with a 9 min. isothermal at 280°C. Mass spectra were taken at 70 eV; a scan interval of 0.5 seconds and fragments from 45 to 450 Da. Total GC running time was 36 min. The relative percentages were calculated.

Characterisation of Compounds

Interpretation on mass-spectrum GC-MS was conducted using the database of National institute Standard and Technology (NIST) having more 62,000 patterns. The spectrum of the unknown components was compared with the spectrum of known components stored in the NIST library. The name, molecular weight and structure of the components of the test materials were ascertained.

III. RESULT AND DISCUSSION

Table I: PHYTOCOMPONENTS IN ALCOHOL EXTRACT

No	RT	Name of the compound	Molecular formula	MW	Peak Area %
1.	3.59	Butane ,1,1-diethoxy -2-methyl-	$C_9H_{20}O_2$	160	0.45
2.	4.02	Hexanoic acid,ethyl ester	$C_8H_{16}O_2$	144	0.25
3.	5.00	Propane , 1,1,3-triethoxy-	$C_9H_{20}O_3$	176	0.17
4.	7.45	2-Furancarboxaldehyde,5-(hydroxymethyl)-	$C_6H_6O_3$	126	2.72
5	8.52	2-methoxy-4-venylphenol	$C_9H_{10}O_3$	150	0.32
6	9.61	Cyclohexane,1,2,4-trimethoxy-,stereoismer	$C_9H_{18}O_3$	174	0.49
7	10.29	Sucrose	$C_{12}H_{22}O_{11}$	342	9.80
8.	11.78	8-Azabicyclol [3,2,1]ocatane,8-acetyl-	$C_9H_{15}NO$	153	2.18
9.	17.38	n-Hexadecanoic acid	$C_{16}H_{32}O_2$	256	7.44
10	17.76	Hexadecanoic acid, ethyl ester	$C_{18}H_{36}O_2$	284	0.85
11	20.12	Palmidrol	$C_{18}H_{37}NO$	299	8.94
12	20.22	9,12,15-octadecatrienoic acid,(z,z,z)-	$C_{18}H_{30}O_2$	278	24.67
13	20.52	Oleic acid	$C_{18}H_{34}O_2$	282	10.16
14	22.33	Hexadecanal,2-Methyl-	$C_{17}H_{34}O$	254	1.69
15	22.93	Pyrolidine,1-(1-oxo-7,10-hexadecadienyl)-	$C_{20}H_{35}NO$	305	17.60
16	25.06	1-Monolinoleoglycerol trimethylsithyl ether	$C_{27}H_{54}O_4Si_2$	498	4.56
17	31.24	Squalene	$C_{30}H_{50}$	410	45.27

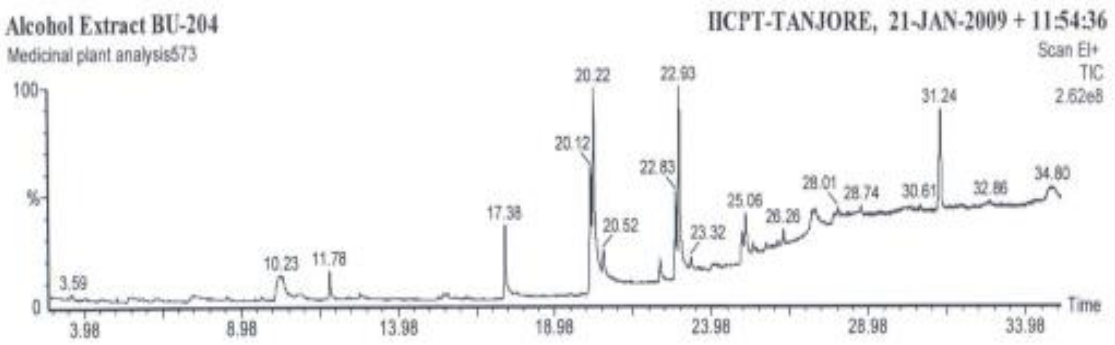

Figure 1: Phytocomponents of alcoholic extract of flaxseed

The studies on the active principles in the flaxseed ethanolic extract by GC-MS analysis clearly showed the presence of seventeen compounds (Tab-1).The active principles with their retention time (RT).molecular formula, molecular weight (MW), and concentration (peak area%)are presented in Table-1.The GC-MS chromatogram of the seventeen peak of the compounds detected was shown in Figure-1.Chromatogram GC-MS analysis of the ethanol extract of flaxseed have been showed the presence of 5 major peaks and the components corresponding to the peaks were determined.

IV. DISCUSSION

The spectral analysis of the ethanol extract of flaxseed revealed the presence of compounds by forming 5 major peaks were squalene (45.27%), 9, 12, 15, octadecatrienoic acid,(z,z,z)- (24.6%), pyrrolidine, 1-(1-oxo-7,10-hexadecadienyl)- (17.60), oleic acid (10.16%) and sucrose (9.80 %) respectively. The triterpene has also been found to have protective activity against several carcinogens. Substances related to squalene, including β-carotene, coenzyme Q10 (ubiquinone) and vitamins A, E, and K[10].The primary therapeutic use of squalene currently is as an adjunctive therapy in a variety of cancers. Although epidemiological, experimental and animal evidence suggests anti-cancer properties, to date no human trials have been conducted to verify the role this nutrient might have in cancer therapy regimens[9,7]. oleic acid (omega-9) is found in animal and vegetable oils and is the major component that is responsible for health benefits of the Mediterranean diet, rich in vegetables and fruits. Although researchers were aware that a Mediterranean diet can reduce the risk of breast cancer and other illnesses such as

heart disease, until now they did not know how[6]. In this study, flaxseed was extracted with absolute alcohol and subjected to screening of bioactive compounds by Gas chromatography- Mass spectrum technique. The presence of 5 major peak have shows the anticancer properties.

V. CONCLUSION

This study denotes the various phytocomponents in flaxseed and it has some anticancer properties. Further studies should be needed to study anticancer properties in breast cancer.

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Formulation and Sensory Evaluation of Yoghurt-based Weaning Foods Manufactured from Mung bean, Soybean and Brown Rice

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Abstract- Three yoghurt-based weaning foods (WF1, WF2 and WF3) designed for 1-3 years old toddlers which prepared by blending of mung beans (*Vigna radiata*), soybean (*Glycine max*) and brown rice (*Oryza sativa*), were evaluated for their chemical, physical, microbial and organoleptic properties. Proximate compositions of the yoghurt-based weaning foods were varied significantly ($P<0.05$) and fulfilled the nutritional requirements given by the Protein Advisory Group Recommendations. Water binding capacity and water solubility index were varied significantly ($P<0.05$). Total plate counts obtained separately for yoghurt, extruded grain flour mixtures and freshly prepared yoghurt-based weaning foods were within the recommendations. According to the sensory evaluation data, WF1 scored higher on average by the panelists than WF2 and WF3 in terms of appearance, aroma, mouth feel and color. However, WF2 was scored highly on average by the panelists in terms of taste and overall acceptability. Therefore, results suggests that yoghurt-based weaning foods are a potential tool for eliminating the protein-energy malnutrition among the children in developing countries which provide sufficient energy with good quality proteins in adequate levels through an cost effective way than conventional liquid-based staple cereal flour mixtures.
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Index Terms- formulated weaning foods, mung bean, soybean, toddler, yoghurt

I. INTRODUCTION

Malnutrition has become one of the major health problems facing by the developing countries which contributes to infant mortality, poor physical and intellectual development of children which lowers the resistance to diseases [1]. Throughout the developing world, malnutrition affects about 800 million people which approximately accounts for 20 percent of the world population [2,3]. For instance, Sri Lanka Demographic and Health Survey (2006/07) highlighted that 18% of Sri Lankan children are stunt, 15% are wasted, 22% are underweighted and 4% are severely underweighted [4]. High price of commercially available weaning foods, vegetables, animal proteins and the non-availability of low priced nutritious foods, combined with bad feeding practices and late introduction of supplementary foods, are mostly responsible for the observed malnourishment among children in Asia [5]. Weaning is the process of complete transition from breast feeding to a semi solid diet. Weaning foods are generally introduced between the ages of six months to three years where the breast feeding itself no longer meets the increasing nutritional requirements of the child [6]. Therefore, there is a high possibility of occurring Protein Energy Malnutrition (PEM) during this transitional phase when children are weaned from liquid to semi-solid or fully adult foods where the growing body of children needs a nutritionally balanced and calorie dense supplementary foods such as weaning foods in addition to mother`s milk [1,3]. This can be more severe if abrupt weaning is practiced where the family menu is directly introduced to the infant that leads to malnutrition, growth retardation and higher rates of mortality [5].

In developing countries, most of the complementary foods are based on local staple foods mainly produced from cereals and given in liquid gruel form for infants [6, 7, 8]. To be suitable for the feeding of young children, these cereal-based weaning foods are prepared in liquid form by dilution with a large quantity of water, thereby resulting in more volume but with a low energy and low nutrient dense food [7]. These cereal-based gruel forms are poor in nutritional value as they are lack in essential amino acids such as threonine, lysine and tryptophan [6]. Moreover, the poor quality of protein and high viscosity of such gruel makes it difficult for the child to consume enough to meet both energy and protein requirements which leads to occurrence of protein-energy malnutrition [6, 8]. Food based approaches used in combination with nutritional education programs can be used as a strategy to overcome the nutrient deficiencies. One such strategy is blending of legumes with cereals or in other words, fortification of legumes into solely cereal-based diets. Therefore, locally available legumes: mung bean (*Vigna radiata*) and soybean (*Glycine max*) can be used due to their high protein and iron content. As these grain legumes are relatively low cost source of iron and protein, they can be used to prepare supplementary foods for children in low income families [6, 9]. Legumes are high in lysine but lack in sulfur-containing amino acids. On the other hand, cereals are lack in lysine but high in sulfur-containing amino acids [10]. Soybean is often used to improve the

protein quality of cereal blends, due to its high levels of protein (40%) and fat (20%) content [8]. Moreover, soybean is rich in lysine which deficient in most cereals however, lack in sulfur containing amino acids as common in legumes. Therefore, fortification of legumes with cereals may overcome the nutritional deficiencies while improving the nutritional quality of the cereal-based weaning foods. Processing methods such as malting and extrusion has been reported to improve nutritional value of the food blends. Extrusion cooking is a food processing technology that rapidly mixes and cooks the material at temperatures of over 100 ºC and dry the product in a relatively short time [11]. This thermal process improves the nutritional quality by the inactivation of antinutritonal factors associated in legumes such as protease inhibitors, hemagglutinnins and polyphenolic compounds while eliminating vegetative microorganisms [12, 13]. On the other hand, malting converts cereal starch into fermentable sugars and improves availability of amino acids [8]. More nutritional benefits are expectable when extruded-legume fortified-cereal blends are incorporated with yoghurt. As it is a source of animal proteins rich in essential amino acids and proteins with high biological value, it may eliminates any nutrient deficiency associated with extruded grain mixtures in addition to its health benefits. For example, as the milk proteins, fat, and lactose components undergo partial hydrolysis during fermentation, yoghurt is an easily digested product of milk for infants which overcome the problems associated with lactose intolerance [14]. In addition, yoghurt based weaning food enables more nutrients in less amount comparatively to the general liquid gruel form. Therefore, the objective of the current study was to formulate an energy dense yoghurt-based weaning food rich in nutrition for 1-3 years old toddlers by blending of mung beans, soybeans with brown rice using low cost extrusion cooking.

II. MATERIALS AND METHODS

Raw Materials

Brown rice (*Oryza sativa*), mung beans (*Vigna radiata*), soybean (*Glycine max*), cocoa powder and full cream milk powder were purchased from the local market. Bovine milk for yoghurt preparation was procured from Mawela Farm, Department of Animal Science, Faculty of Agriculture, University of Peradeniya, Sri Lanka.

Preparation of Raw Materials

Mung beans were washed, soaked (8 h), dried (at 50 ºC for 24 h) and roasted under an open flame (160 ºC) until become golden brown, and partially milled using a Ferrell-Ross® roller mill (Ferrell-Ross Corp, Oklanoma city, Okla, USA). Brown rice were washed, dried in a dehydrator (55 ºC for 1 h) and partially milled using a Ferrell-Ross® roller mill (Ferrell-Ross Corp, Oklanoma city, Okla, USA). Soybean grains were soaked (8 h), drained by placing on a nylon sieve (1 h), dried at 50 ºC for 24 h and roasted under an open flame (160 ºC) until become golden brown coloured. Then de-hulled using a Bauer 148-2 de-huller (Bauer bros.co, USA) and partially milled using a Ferrell-Ross® roller mill (Ferrell-Ross Corp, Oklanoma city, Okla, USA).

Formulation of Weaning Foods

Three composite weaning foods (WF1, WF2 and WF3) were formulated using varying amounts of raw materials and yoghurt (Table 1) by considering nutrient and caloric values of each ingredient in order to meet the nutrient requirement of toddlers according to the recommendations given by the World Health Organization in which 100 g portion of each weaning food enables to provide ⅓ of the daily energy and carbohydrate requirement, ⅔ of the daily protein requirement and ¼ of the daily fat requirement of a growing toddler. Each grain mixture corresponding to each weaning food was extruded separately and mixed with corresponding amount of yoghurt at serving (Table 1).

Table 1: Composition of different composite weaning foods

Ingredient	WF1	WF2	WF3
Brown rice	2	2	2
Mung beans	11	17	22
Soybeans	4	3	3
Dried milk powder	3	3	3
Yoghurt	80	75	70
Total	**100**	**100**	**100**

Extrusion Process

Prepared mung bean, soybean and brown rice were blended according to the relevant proportions as stated in the Table 1. Each blend was extruded separately using a co- rotating twin screw extruder (die size 0.25 inches) with a smooth barrel. The resulted extruded products were milled into flour using a Fitz® 832D Fitz mill (The Fitzpatrick Company, Russia) and sieved through a 0.3 mm sieve. Thereafter, 3% spray dried milk powder and 5% coco powder were mixed with each blend.

Preparation of Yoghurt

Milk was standardized with cream in order to obtain a final fat content of 4%. Homogenized and pasteurized (95 ºC for 5 min) standardized milk was inoculated with a commercial starter culture containing *Streptococcus thermophilus* and *Lactobacillus delbrueckii subsp. bulgaricus* at a rate of 2% (w/v) after addition of sugar and gelatin. The mix was incubated in an ACP incubator (ACP Co Ltd, Japan) at 42 ± 2 ºC and was terminated at pH 4.6 and stored under refrigerated conditions (4 ± 1 ºC).

Chemical Analysis

Each grain mixture was mixed with corresponding amount of yoghurt as presented in the Table 1 and proximate composition of each weaning food in gruel form was determined according to the AOAC protocols (1995) [15]. Carbohydrate content was determined by the difference between total dry matter and the sum of other proximate components (crude protein, crude fat, crude fiber, ash) while gross energy content was measured by the summation of caloric value of each nutrient using energy conversion factors given by Bangoura and Zhou (2007) [13].

Physical Properties Determination

The water binding capacity and water solubility index were determined according to the methods previously described by Griffith-L, Castell-Perez and Griffith-M. (1998) [16].

Analysis of Shelf-life

pH of the yoghurt stored under refrigerated storage (4 ± 1 ºC) was determined at 4 days intervals for 28 d using Hannah 211electric pH meter (HANNA, TOA Electronics, Ltd, Tokyo, Japan). Titratable acidity of the yoghurt was measured by titrating 9 mL of sample with 0.1 N NaOH solution using phenolphthalein as indicator, at 4 d intervals during 28 d of storage.

Microbial Analysis

The microbial analysis (Total plate count and Coliform count) was conducted separately for yoghurt and composite grain flour mixtures on the day of manufacturing, and for freshly prepared gruel form after blending of each grain mix with corresponding amount of yoghurt. Total Plate Count (TPC) was detected in each of the composite grain flour mixtures and freshly prepared yoghurt based weaning foods immediately after preparation. A series of dilutions (10^{-1} to 10^{-6}) were prepared by dissolving 1 g of each sample with 9 mL of distilled water. Then 0.1 mL of each diluent was plated out on plate count agar and incubated at 37 ºC for 48 h in a Fisher 322 incubator (Scientific Company, USA). The colonies were counted manually and the results were expressed as the number of Colony Forming Units (CFU) per gram. Coliform counts in yoghurt, composite grain flour mixtures and freshly prepared yoghurt based weaning foods immediately after preparation were detected using pour plate technique. Point one milliliters (0.1 mL) of each 10^{-1} dilution series of the corresponding sample was plated out in Violet Red Bile (VRB) agar and incubated at 30 ºC for 24 h. Colonies were identified as enumerate blue and red colonies associated with entrapped gas regardless of size or intensity of color according to the method described by Ahima (2009) and expressed as CFU/g[17].

Sensory Evaluation

Each extruded grain mix was blended with the corresponding amounts of yoghurt and the sensory evaluation of yoghurt-based weaning foods in gruel form was conducted by 30 untrained panelists drew randomly from the Faculty of Agriculture, University of Peradeniya, Sri Lanka. Each panelist received 3 samples of freshly prepared weaning foods to taste, evaluate and comment on sensory characteristics. They were asked to evaluate the appearance, aroma, taste, mouth feel, colour and overall acceptability based on a 7 point hedonic scale; like extremely =7, like very much = 6, like slightly =5, neither like nor dislike = 4, dislike slightly = 3, dislike very much = 2, and dislike extremely = 1.

Statistical Analysis

The experiment was conducted as Complete Randomized Design (CRD). One-way ANOVA (Analysis of variance) was performed to analyze data on physical and chemical analysis and means were separated by Least Significant Difference (LSD) procedure using SAS statistical software version 9.1 (SAS institute. Cary, NC, USA). Sensory data were analyzed by Friedman`s Test using SPSS statistical software version 17 (SPSS Inc., Chicago, IL, USA). All values were reported as mean ± standard error mean (SEM) and significances were determined at $P < 0.05$. All determinations were conducted in triplicates (n=3).

III. RESULTS AND DISCUSSION

Nutritional composition of the weaning foods

The nutritional composition of each weaning food resulted from the proximate analysis was presented in the Table 2. Dry matter (DM) content of the yoghurt based formulated weaning foods were varied significantly (P<0.05) where DM content of WF3 was greater than that of the WF1 and WF2. Results revealed that percentage of DM increased as the proportion of grain flour mixture increased. Moisture content of the three yoghurt-based weaning foods were varied significantly (P<0.05). As expected, the WF1 which contained the highest proportion of yoghurt in its formulation (80%) exhibited the highest moisture content. Moreover, the moisture content tended to decrease as the proportion of yoghurt decreased in the formulation.

Table 2: Nutritional Composition of different weaning foods in dry weight basis (WF1, WF2, and WF3)

Nutrient (%)	WF1	WF2	WF3
Dry matter	42.53 ± 0.04^{b}	42.53 ± 0.04^{b}	50.18 ± 0.05^{a}
Moisture	56.42 ± 1.00^{a}	53.53 ± 1.16^{b}	50.75 ± 0.35^{c}
Crude protein	15.22 ± 0.32^{b}	16.28 ± 0.37^{a}	16.19 ± 0.28^{a}
Crude fiber	0.71 ± 0.19^{c}	0.93 ± 0.10^{a}	0.81 ± 0.12^{b}
Crude fat	12.43 ± 0.37^{a}	12.38 ± 0.31^{a}	12.38 ± 0.31^{a}
Carbohydrates	69.07 ± 0.20^{a}	68.55 ± 0.30^{b}	68.10 ± 0.35^{b}
Ash	1.70 ± 0.05^{a}	1.69 ± 0.12^{a}	1.66 ± 0.30^{a}
Gross Energy (kcal)	1854.36 ± 1.20^{a}	1845.08 ± 0.41^{ab}	1835.99 ± 0.61^{b}

Note: Values in the same raw with different superscripts are significantly different at (P<0.05)

The percentage of crude protein (CP) among the formulated weaning foods was varied significantly where the CP content of WF2 and WF3 was greater (P<0.05) than that of the WF1. This may be due to the higher portion of mung bean in WF2 and WF3. Wikramanayake (1996) has stated that the processing techniques used to prepare the weaning foods such as roasting assist in breakdown of lipocytes to release fat and protein [10]. Moreover, according to Griffit et al. (1998) roasting improved sensory qualities and aided in inactivation of destructive enzymes, which improves the storage and nutritional quality of the product and reduce trypsin inhibitor activity when seed temperatures reached 90–100 °C where the lipoxygenase activity loss at temperatures of 75–80 °C [16]. Thus, it can be reasonably argued that the proteins originated from the roasted mung beans and soybeans which used to prepare weaning foods are in better quality. Moreover, according to the Codex Alimentarius Commission Recommendations (1982), a minimum protein content of 15% in food is required for maximum complementation of amino acids for a satisfactory growth of a toddler and therefore all of the yoghurt-based weaning foods tested were agreed to so-called recommendation [18]. On the other hand, a combination of plant proteins and animal proteins may fulfill the requirements of amino acids especially, essential amino acids comparatively to a solely vegetable source. In the current study, all of the evaluated yogurt-based weaning foods were consisted of both plant protein sources (soya, brown rice and mung beans) and an animal protein source (yoghurt). Therefore, it can be suggested that both quality and biological value of protein in the evaluated weaning foods are much higher than a solely cereal based weaning food. Crude fat content of the yoghurt based weaning foods was not varied significantly (P>0.05). Fat contents of the yoghurt based weaning foods evaluated were above the value of 10% recommended by the Protein Advisory Group (1972) Recommendations. Fat goes oxidative deterioration during storage and therefore, fat content of a particular food affects greatly to its shelf stability. Hence, a food sample with high fat content is more liable to spoilage than one with a lower fat content [1]. However, the high fat content in the experimental weaning foods is not a matter since the yoghurt and extruded grain mixture will be packed separately with intension to blend together when fed to a toddler in freshly prepared form. Therefore, the shelf life of these products is more likely to be depended on the shelf life of the yoghurt rather than on the fat content of the blended form. Yoghurt and soybean are the main sources of fat in the weaning foods tested. Soybean oil agrees with the recommendations given by the FAO/WHO (1998) in which vegetable oils are allowed to be incorporated in foods for infants and children, which will not only increase the energy density, but also acts as a carrier for fat soluble vitamins (K, A, D, E) while providing essential fatty acids. In addition, soybean and cereals contain unsaturated fatty acids which do not increase the cholesterol content in the blood; hence it can be recommended for children without causing any harmful effect on blood cholesterol level [10].

Ash content resulted from the proximate analysis reflects the mineral content of each weaning food. There was no significant difference (P>0.05) found among the yoghurt based weaning foods with relevant to ash content. However, it was within the acceptable range of <5% recommended by the Protein Advisory Group (1972). Crude fiber (CF) content among the formulated weaning foods

were varied significantly ($P<0.05$) and was less than the maximum fiber content of 5% of the Protein Advisory Group Recommendations (1972) [19]. It was expected to increase the fiber content of the weaning foods as the proportion of extruded grain flour content increased. However, CF content of WF3 which contained 30% of extruded grain flour mixture was significantly lower than that of the WF2 which contained only 25% of extruded grain flour mixture. According to Ghasemzadeh and Ghavide (2011), incorporation of de-hulled legumes instead of whole legume grains, reported low fiber contents [20]. Accordingly, significantly lower CF content associated with the WF3 is most probably due to the addition of de-hulled mung beans although WF3 contained highest percentage (22%) of mung beans in composition. Carbohydrate content of the WF1 was higher ($P<0.05$) than that of the WF2 and WF3. In this study, brown rice was used as the main carbohydrate supplement which produce adequate amount of carbohydrates in order to fulfill the minimum carbohydrate requirement of 65% as recommended by the Protein Advisory Group`s recommendations (1972) for 1-3 year old toddler [19]. Gross energy/ caloric content of the evaluated yoghurt-based weaning foods were varied significantly. Gross energy/ caloric content of the WF1 found to be the highest ($P<0.05$) among the three yoghurt based weaning foods tested. However, no significant differences were found between the gross energy contents of the WF1 and WF2, and WF2 and WF3. Calories in a diet are provided by protein, fat and carbohydrates [10]. WF1 contained comparatively low CP content; therefore, the high caloric content associated with the WF1 more likely to be due to the comparatively high crude fat and carbohydrate content associated with it (Table 2). The caloric contents of protein, fat and carbohydrates were ranged 65.65 - 68.37, 113.12 – 118.51, and 257.33 – 259.64 kcal, respectively among the three yoghurt based weaning foods (data not shown). In other words, contribution of protein, fat and carbohydrates to the total caloric content of the evaluated yoghurt-based weaning foods were ranged from approximately 14-16%, 25-27% and 58-59%, respectively. Furthermore, according to the Protein Advisory Group recommendations, the contribution of fat and carbohydrates to the total caloric value should be <30% and within 50-60%, respectively. Thus, contribution of protein, fat and carbohydrates to the total caloric content of the evaluated yoghurt-based weaning foods were fulfilled the requirement of the Protein Advisory Group recommendations (1972).

Physical Property Measurements

Water binding capacity (WBC) and Water solubility index (WSI) of the formulated weaning foods are presented in the Table 3. The WBC of the three yoghurt based weaning foods were varied significantly ($P<0.05$). However, the values were lower than the values reported by Ahmadzadeh-Ghavidel and Prakash (2010) [21]. On the other hand, Griffith et al. (1998) observed low WBC associated with high fat, high protein with low carbohydrate levels of the weaning foods. In addition, Ghasemzadeh and Ghavide (2011) reported higher values for WBC as the starch content increased [20]. Therefore, the low WBC observed in the weaning foods is most probably due to the high protein and fat content with low starch content in the formulated weaning foods. WSI of the formulated weaning foods were varied significantly ($P<0.05$) and displayed a negative relationship ($r = –0.916$) with the WBC. In other words, WSI tended to increase as WBC decreases. High water solubility results a fine paste and improved mouth feel. The extruded products usually increase the water solubility as cooking increase the susceptibility of grain starch to glucoamylase hydrolysis which leads to gelatinization during processing [16].

Table 3: Water binding capacity (WBC) and Water Solubility Index (WSI) of formulated weaning foods in gruel form (WF1, Wf2, and WF 3)

Weaning Food	WBC (mL/g)	WSI (%)
WF1	3.06 ± 0.04^{a}	17.24 ± 0.23^{c}
WF2	1.67 ± 0.06^{c}	30.28 ± 0.06^{a}
WF3	2.49 ± 0.05^{b}	27.55 ± 0.05^{b}

Note: Values in the same column with different superscripts are significantly different at $P<0.05$

Microbial Analyses

The TPC and coliform counts of yoghurt, composite grain flour mixtures and yoghurt based formulated weaning foods are summarized in the Table 4.

Table 4: Total Plate Count (TPC) Data and Coliform Counts of yoghurt, grain mixtures and formulated weaning

Microbial count	Yoghurt	Grain Mixtures			Formulated weaning foods in gruel form		
		WF1	WF2	WF3	WF1	WF2	WF3
TPC (CFU/g)	ND	2.80×10^{6}	2.75×10^{6}	3.20×10^{6}	5.94×10^{4}	5.62×10^{4}	6.34×10^{4}
Coliform	0	0	0	0	0	0	0

Note: ND = Not determined; WF = weaning food

According to the USDA specifications and Sri Lanka Standard (SLS) Institute recommendations, the coliform count in yoghurt should be less than 10/g and 1 CFU/g, respectively [22, 23]. Coliform counts were not detected in yoghurt and composite flour mixtures and therefore, had fulfilled both recommendations. On the other hand, according to the recommendations given by the UK Food Protection Agency and Food Standards of Australia and New Zealand (FSANZ), the TPC and Coliform count of cereal flour mixtures should be $< 10^7$ CFU/g and < 3 CFU/g, respectively [24]. In addition, the acceptable Total Plate Count (TPC) and Coliform count for ready to eat food items should be $< 10^5$ CFU/g and < 3 CFU/g, respectively. Therefore, formulated weaning foods tested in this study were within the limits of acceptable range. Absence of coliform counts most probably due to the hygienic practices and good management practices employed throughout the manufacturing process including preparation of yoghurt, composite flour mixtures and yoghurt-based gruel forms.

Shelf life examination of yoghurt

The variation in pH of the prepared yoghurt has been illustrated in the Figure 1 which was measured at 4day intervals over 28 days of refrigerated storage (4 ± 1 ºC)**.**pH value of the yoghurt was continuously decreased throughout the storage period which was varied from 4.62 to 4.31, 4.58 to 4.29 and 4.64 to 4.28 in replicate 1, 2 and 3, respectively.

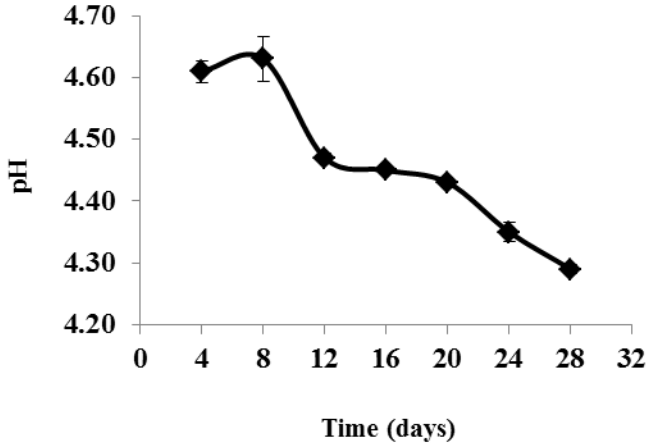

Figure 1 Variation in pH value of yoghurt during 28 days of storage under refrigerated conditions (n=3)

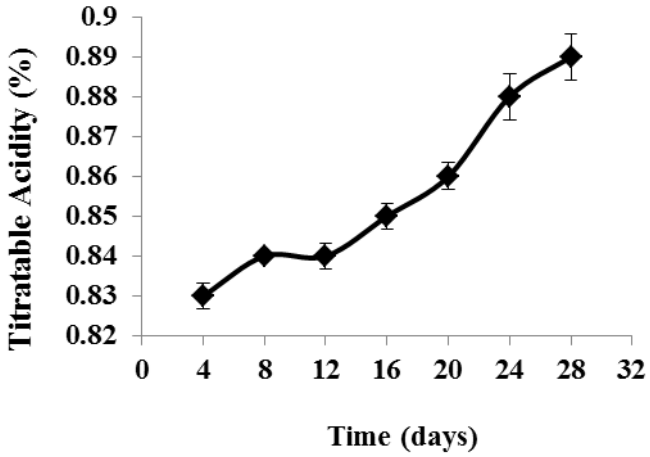

Figure 2 Variation in titratable acidity (%) of yoghurt during the 28 days of refrigerated storage (n=3)

On the 20th day, sour taste and unpleasant odor were observed in every replicate. Therefore, it was concluded that the yoghurt was unacceptable for consumption after 20 days of refrigerated storage. In addition, the average variation in titratable acidity (%) of the

yoghurt which was measured at 4day intervals over 28 days under refrigerated storage has been illustrated in the Figure 2.Titratable acidity was continuously increased throughout the refrigerated storage period of the yoghurt which was maximized after 28 days (0.89±0.01). These values were within the recommended range of0.8- 1.25%according to the Sri Lankan Standards (1989) given for yoghurt [23]. However,most of the yoghurts available in the market have a shelf life of two weeks (14 d).Therefore**,** based on the pH and titratable acidity values, it can be suggested that the specially prepared high fat yoghurt for current study can be stored safely under refrigerated conditions for two weeks (14 d).

Sensory evaluation

The scores obtained for sensory attributes: appearance, aroma, taste, mouth feel, color and overall acceptability demonstrated no significant differences (P>0.05) among the 3 formulated weaning foods tested (Table 5). WF1 scored higher on average by the panelists than WF2 and WF3 in terms of appearance, aroma, mouth feel and color. WF1, WF2 and WF3 contained 80%, 75% and 70% of yoghurt, respectively (Table 1). Therefore, it can be concluded that the higher yoghurt content may have a positive influence on appearance, aroma, mouth feel and color.

Among the three formulated weaning foods, WF2 was scored highly on average by the panelists in terms of taste and overall acceptability. The average scores of WF1, WF2 and WF3 for taste was WF1 < WF3 <WF2. It is obvious that higher yoghurt content does not have positive influence on taste and something else responsible for the taste. Both WF2 and WF3 had equal amounts of brown rice and soybeans. Therefore, both brown rice and soybeans could not be responsible for the taste difference observed between WF2 and WF3. Therefore, the taste difference is more likely to be due to the difference in mung bean content in which WF2 and WF3 contained 17% and 22% of mung bean, respectively (Table 1) which was higher than that of the WF1 (11%). On the other hand, WF3 which contained highest mung bean content was ranked second on taste attribute which reflect the fact that yoghurt and mung bean content solely does not govern the taste of the weaning foods. Taste is more likely to be resulted due to a combined effect of the levels of yoghurt and mung beans where moderate levels of both yoghurt and mung bean had given the highest scores on taste attribute (WF2). Moreover, WF3 had received the lowest scores on average for all sensory attributes tested, except aroma and taste. Therefore, it is obvious that higher grain content may have a negative impact on the sensory properties, in particular appearance, mouth feel, color and overall acceptability. In addition, it can be concluded that a combined effect of the level of yoghurt and grain mixtures plays a significant role to determine the overall acceptability of the weaning food.

Table 5: Mean scores of testing panelists (n=30) for sensory properties of formulated weaning foods (WF1, WF2, and WF3) in freshly prepared gruel form

Characteristic	WF1	WF2	WF3
Appearance	5.10 ± 0.24	4.97 ± 0.25	4.80 ± 0.33
Aroma	5.03 ± 0.21	4.90 ± 0.22	4.93 ± 0.28
Taste	4.63 ± 0.27	5.03 ± 0.23	4.97 ± 0.23
Mouth feel	5.33 ± 0.19	5.20 ± 0.21	4.73 ± 0.23
Color	5.13 ± 0.22	4.93 ± 0.22	4.57 ± 0.33
Overall acceptability	4.87 ± 0.22	5.00 ± 0.20	4.63 ± 0.27

Data expressed as Mean ± SEM

Comments from the panelist were also evaluated during the sensory evaluation. Some panelists criticized the sour taste observed in some products, mainly WF1, and the clumpiness resulted due to the absent of homogenized mixing of grain mixtures and yoghurt. Complaints regarding the sourness were not observed in WF2 and WF3. Therefore, it can be concluded that the sourness associated with the WF1 could be due to the high yoghurt content (80%) which leads to the panelists to score WF1 as the lowest on taste attribute. Addition of flavor such as vanilla, strawberry and chocolate, and addition of color were among the suggestions given by the panelists in order to improve the acceptability of the weaning foods and to mask the sour taste resulted from yoghurt.

Although the overall sensory scores of formulated weaning foods remained low, it can be concluded that the WF2 is the best weaning food according to the organoleptic properties, especially due to its taste and high overall acceptability.

IV. CONCLUSION

The proximate compositions of the three extruded yoghurt-based weaning foods were varied significantly and fulfilled the requirements given by the Protein Advisory Group Recommendations (1972). Water Binding Capacity was highest (3.06±0.04) in WF1 whereas Water Solubility Index was highest (30.28±0.06) in WF2 which were significantly greater (P<0.05) than that of the other weaning foods. The specially prepared yoghurt was found to be safe for consumption for 2 weeks based on the titratable acidity and pH values determined during the 28 day storage period under refrigerated conditions (4±1 ºC). Total Plate Counts of the yoghurt,

extruded grain flour mixtures and freshly prepared yoghurt-based weaning foods obtained separately, were within the acceptable range recommended by the Food Standards of Australia and New Zealand (FSANZ). According to the sensory evaluation data, WF1 scored higher on average by the panelists than WF2 and WF3 in terms of appearance, aroma, mouth feel and color. However, WF2 was scored highly on average by the panelists in terms of taste and overall acceptability. Based on the results, it can be suggested that fortification of legumes into the conventional cereal based weaning foods and diluting in yoghurt instead of liquid, results a product with extreme nutritional quality, cost effective and could be a possible and effective tool in order to overcome the malnutrition among children in the developing countries.

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Revisiting Defences against Large Scale Online Password Guessing Attacks

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Abstract- Brute force and dictionary attacks on password protected remote login services are increasing rapidly. Letting legitimate user's login conveniently while preventing such attacks is difficult. Automated Turing Tests (ATTs) are effective and easy to implement but cause reasonable amount of inconvenience to the user. We discuss the existing and proposed login protocols designed to prevent large scale online dictionary attacks. We propose Password Guessing Resistant Protocol (PGRP), which is derived upon revisiting prior proposals designed to restrict such attacks. PGRP reduces the total number of login attempts from unknown remote host while trusted or legitimate users can make several failed login attempts before being challenged by ATT.

I. Introduction

With increasing number of online users in the real world, maintaining privacy details and protecting them with a password also becomes difficult. Here we involve developing a secure application to prevent our privacy information by using Password Guessing Resistant Protocol (PGRP). Password guessing attacks can be classified into two:

1. Brute Force Attack: A Brute Force attack is a type of password guessing attack which consists of trying every possible code, combination, or password until the correct one is found. A brute force attack is a very slow type of attack because of the many possible combinations of characters in the password. However, brute force is effective; given enough time and processing power, all passwords can eventually be identified.

2. Dictionary Attack: A dictionary attack is another type of password guessing attack which uses a dictionary of common words to identify the user's password.A dictionary attack is a method of breaking into a password protected server by systematically entering every word in a dictionary as a password.

Existing System: The use of passwords is a necessity in computer security but passwords are often easy to guess by automated programs or tools running dictionary attacks. In the existing system, an automated test is implemented that humans can pass, but current computer programs can't pass. Any program that has high success over these tests can be used to guess passwords cause security risks. An example of such a test is a 'captcha'. A captcha is a test used in computing which ensures that the response is generated by a person and not by a tool. The process usually involves a computer asking a user to complete a simple test which can ensure a successful login. These tests are designed to be easy for a computer to generate, but difficult for a computer to solve, so that if a correct solution is received, it can be presumed to have been entered by a human. Following figure (Fig.1) is an example of the captcha.

Proposed System: Password Guessing Resistant Protocol (PGRP), which is our proposed system, enforces ATTs after a few failed login attempts are made from unknown systems. We define trusted or known systems as those from which a successful login has occurred within a fixed period of time. These are identified by their IP addresses saved on the login server as a white list. PGRP accommodates both graphical user interfaces and character interfaces, while the previous protocols deal exclusively with the former. PGRP uses IP addresses for tracking legitimate users. The proposed system is more sensitive against brute force and dictionary attacks while also allows a number of failed login attempts for legitimate users.

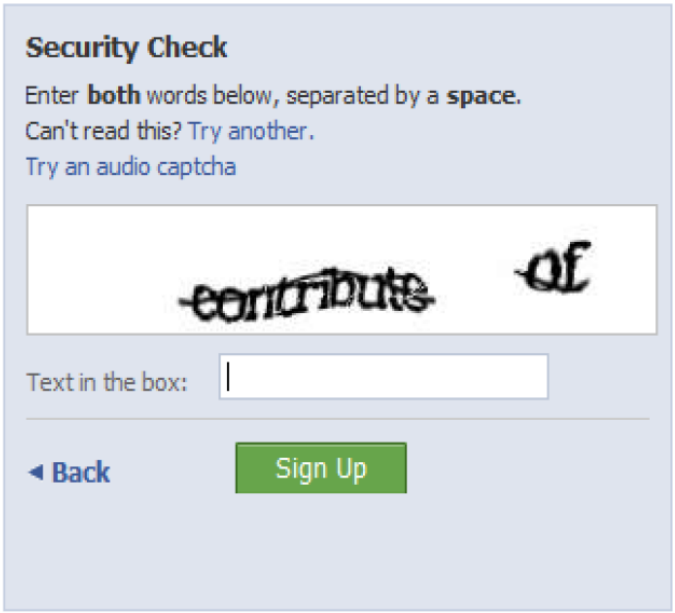

Fig.1: An example of the captcha

The proposed system is much more convenient than the existing system and consists of minimal steps for legitimate user to login.

Two processes involved in this:

1) If a trusted system fails the first login attempt then it is given two more chances (totally three chances). If the user fails in the third attempt to login then the intimation will be given.

2) If an unknown system fails in the first login attempt then it will not be given any more chances and intimation A flowchart

of the algorithm of the discussed protocol is shown below. character-based er. It t will be given.

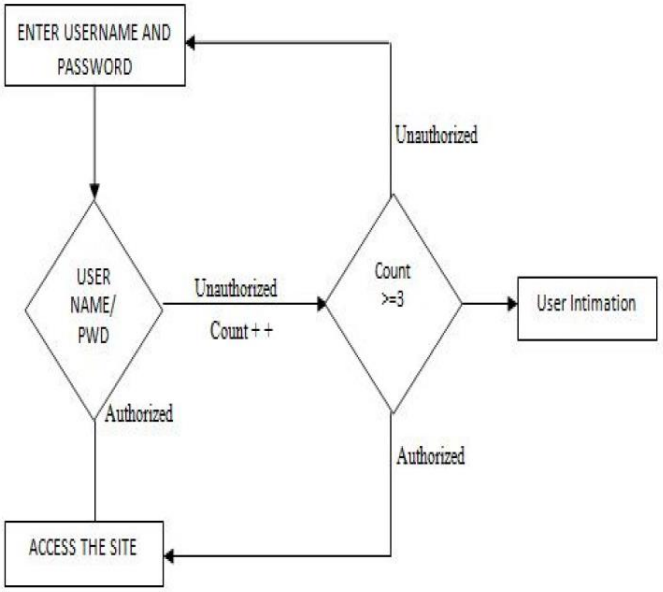

Fig.2: flowchart of proposed system

II. OVERVIEW

The general idea behind PGRP is that user does not have to face an ATT challenge for the following two conditions

1) When the number of failed login attempts for a given username is very

2) When the remote host has successfully logged before reaching the threshold limit of failed login attempts. In contrast to previous protocols, PGRP uses either IP addresses, cooki identify systems from which users have been successfully authenticated. number of failed login attempts for a specific username is below a threshold, the user is not required to answer an ATT challenge even if the login attempt is from a new machine for the first time.

III. CONCLUSION

Password guessing attacks have been increasing rapidly. To put an end to this we use PGRP. PGRP will restrict the number of attempt made by a system or a machine and allow the legitimate user to have a full secured access over their account. PGRP appears suitable for organizations of both small and large number of user accounts. PGRP can restrict brute force attack and dictionary attack, so it enhances the security of user's account.

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Return Stroke VLF Electromagnetic Wave of Oblique Lightning Channel

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***Abstract*-**In this paper, the theories of vertical return stroke, in which Bruce and Golde [1] current is supposed to flow, have been extended to oblique orientations of the channel. The VLF electric field and power density have been calculated for a distance 100 km from the source as function of the channel parameters and of the channel orientation. It is shown that VLF electromagnetic energy radiated from the oblique channel is of Gaussian shape with the maximum appearing at certain frequency depends on orientation of channel.

***Index Terms*-**Current model, current moment, Electromagnetic waves, Frequency spectrum, Lightning discharge, return stroke, VLF electric field.

I. INTRODUCTION

The exact physical processes of electromagnetic wave, generation from lightning discharges, extending from a few hertz in the ELF range to beyond the visible region are not yet fully understood. Using various models for lightning discharge current, many workers [1 – 12] have studied certain characteristics of generating sources and radiated electromagnetic spectrum from them. The effect of orientation of cloud to ground lightning discharges on the radiated energy spectrum has not been thoroughly studied except the some researchers [13 – 15]. In these studies, changes in the field strength and polarization of radiated signals in the extended frequency range have been shown. Visual observations and photographs [16 – 18] have made it obvious that the return strokes are not always vertical. In reality lightning channels are oblique and tortuous. This is a better approximation to the nature. The verticality of return stroke depends on the location of the centre of the charged cloud and the conducting region of the ground. Few researchers [19 - 21] have reported that there are errors found in the magnetic direction caused by non verticality of the return strokes.

The polar diagram of the radiated electromagnetic energy has the usual figure of the eight shapes symmetric along the direction of the return stroke channel, accordingly changes with the changes in the orientation of the return stroke and gives rise to a variability of maximum radiated energy from one return stroke to another return stroke [22]. Thus it is found that the orientation of return stroke controls the magnitude of the field and frequency of the VLF waves, is one of the most important parameters to study the conditions of formation of whistlers. To give a better approximation near to nature, the theory of the vertical electrically conducting channel [22] which simulates return stroke has been extended to oblique orientation of lightning channel. The effect of perfectly conducting ground is taken into account for the calculation of the radiated electric field and power density.

II. RADIATED ELECTRIC FIELD FROM OBLIQUE RETURN STROKE

In order to study the characteristics of radiated electric field, we take the well accepted Bruce and Golde [1] return stroke current model which is given as

$$I_t = I_0[exp(-\alpha t) - exp(-\beta t)] \quad (1)$$

Where, I_0, α and β are constants which vary from one stroke to another.

Srivastava and Tantry [23] chose I_0 = 22kA, $\alpha = 1.4 \times 10^4 s^{-1}$ and $\beta = 50 \times 10^4 s^{-1}$ for the study of VLF characteristics of electromagnetic radiation.

Return stroke velocity model is given by Srivastava [24] as

$$V_t = V_0[\exp(-at) - \exp(-bt)] \quad (2)$$

Where, $V_0 = 3 \times 10^8 \text{ ms}^{-1}$, $a = 6 \times 10^4 s^{-1}$ and $b = 7 \times 10^5 s^{-1}$.

An expression very similar to Eq. (2) has been developed by Rai[25].

Few researchers [26 – 28] have reported that the Eq. (2) explains the spectral details of the radiated electromagnetic energy, magnetic field and the increase of electric field with increasing order of the strokes in a multi stroke flash of lightning.

The current moment associated with return stroke channel determines the strength and spectral features of the radiated electromagnetic wave. The current moment associated with the oblique return stroke is defined as

$$M_c = I_t \int_o^t V_t \, dt = \mathrm{I_t l_t} \cos \Phi \qquad (3)$$

Where, $\mathrm{l_t}$ is the length of the return stroke channel from the ground to charge centre of the thunder cloud (Fig. 1), and Φ is the angle made by the return stroke from the vertical.

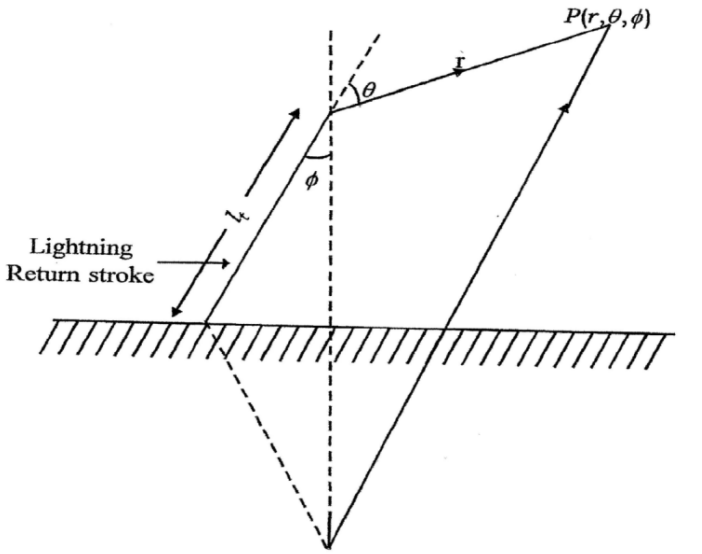

Fig. (1) Geometrical parameter of an oblique return stroke used in Eq. (5)

From simple antenna theory, the component of the radiated electric field at a distance, r from the lightning discharge channel is written as

$$E_d(\mathrm{r, t}) = \frac{1}{4\pi\varepsilon_0}\frac{1}{c^2 r}\frac{dM_c(t)}{dt} \qquad (4)$$

Where, ε_0 (= 8.85 x$10^{-12}\mathrm{Fm}^{-1}$) is the permittivity of free space, and c is the velocity of light.

The value of $\mathrm{M_c}\,(t)$ is substituted from Eqs. (1 – 3) into Eq. (4), the direct component of the radiated electric field is obtained as

$$E_{\mathrm{d}}\,(\mathrm{r, t}, \theta, \Phi) = \frac{30\mathrm{I_0 V_0}\cos\Phi\sin(\theta+\Phi)}{\mathrm{cabr}}\, f(\mathrm{t}) \qquad (5)$$

Where, θ is the observation angle made by the line joining the observation point to the mean direction of the discharge channel and the factor f(t) is similar to that given by Srivastava and Tantry [23]

$$f(t) = (b-a)\{\alpha \exp(-\alpha t) - \beta \exp(-\beta t)\} -$$

$$b\{(a+\alpha)exp[-(a+\alpha)t]-(a+\beta)exp[-(a+\beta)t]\}+$$

$$a\{(b+\alpha)exp[-(b+\alpha)t]-(b+\beta)exp[-(b+\beta)]t\} \quad (6)$$

The image contribution to the electric field of the return stroke due to perfectly conducting ground is given as

$$E_I(r,t,\theta,\Phi)=FE_d(r,t,\theta,\Phi) \quad (7)$$

Where, F is the ground attenuation factor which depends on frequency and distance. As a first approximation, the ground attenuation factor for low frequency waves at large distances of observation remains almost constant and close to unity [29]. Under this approximation, the radiated electric field at the point of observation is expressed as [30]

$$E(r,t,\theta,\Phi)=E_1(r,t,\theta,\Phi)+E_d(r,t,\theta,\Phi)=2E_d(r,t,\theta,\Phi) \quad (8)$$

Using Eq. (5) into Eq. (8), the radiated electric field is rewritten as

$$E(r,t,\theta,\Phi)=\frac{60I_0V_0\cos\Phi\sin(\theta+\Phi)}{cabr}f(t) \quad (9)$$

III. FREQUENCY SPECTRUM

Magnitude of the frequency spectrum of direct radiated electric field is written as

$$|E_d(\omega)|=\left|\int_0^\infty E_d(r,t,\theta,\Phi)\,exp(-i\omega t)\,dt\right| \quad (10)$$

In terms of direct radiated electric field, the electromagnetic power density is written as

$$P_d(\omega)=|E_d(\omega)|^2/\eta \quad (11)$$

Where, ω is the frequency, in rad s^{-1}, at which power is radiated.

By including the effect of perfectly conducting ground the radiated power density is written as

$$P_d(\omega)=|E_d(\omega)|^2/\eta=4|E_d(\omega)|^2/\eta=4P_d(\omega) \quad (12)$$

Substituting for E_d (r, t, θ, Φ) from Eq. (5) into Eq. (9), we obtain

$$|E_d(\omega)|=\frac{30I_0V_0\cos\Phi\sin(\theta+\Phi)}{cabr}\left[\left\{\left(\frac{\beta^2(b-\alpha)}{\beta^2+\omega^2}+\frac{b(a+\alpha)^2}{(a+\alpha)^2+\omega^2}+\frac{a(b+\beta)^2}{(b+\beta)^2+\omega^2}\right)-\left(\frac{\alpha^2(b-\alpha)}{\alpha^2+\omega^2}+\frac{b(a+\beta)^2}{(a+\beta)^2+\omega^2}+\frac{a(b+\alpha)^2}{(b+\alpha)^2+\omega^2}\right)\right\}^2+\omega^2\left\{\left(\frac{\beta(b-\alpha)}{\alpha^2+\omega^2}+\frac{b(a+\beta)}{(a+\beta)^2+\omega^2}+\frac{a(b+\alpha)}{(b+\beta)^2+\omega^2}\right)-\left(\frac{\beta(b-\alpha)}{\beta^2+\omega^2}+\frac{b(a+\alpha)}{(a+\alpha)^2+\omega^2}+\frac{a(b+\beta)}{(b+\beta)^2+\omega^2}\right)\right\}^2\right]^{1/2} \quad (13)$$

Taking the effect of perfectly conducting ground, the expression for frequency spectrum of the radiated electric field is written as

$$|E(\omega)|=2|E_d(\omega)| \quad (14)$$

IV. RESULTS AND DISCUSSIONS

So far the theory developed for this study shows that the effect of orientation of lightning channel are introduced into the expressions of VLF electric field and into the radiated power density through the factors $\cos\Phi\sin(\theta+\Phi)$ and $\cos^2\Phi\sin^2(\theta+\Phi)$, respectively.

Calculation of the radiated electric field from the current source has been done by using Eq. (9). The variation of computed VLF electric field arriving at a distance of 100 km and making different observation angles with the channel for different orientation

of return stroke is shown in Fig.(2). It is observed that maximum radiated electric field is at observation angle $\theta = 90^0$ and at orientation of the

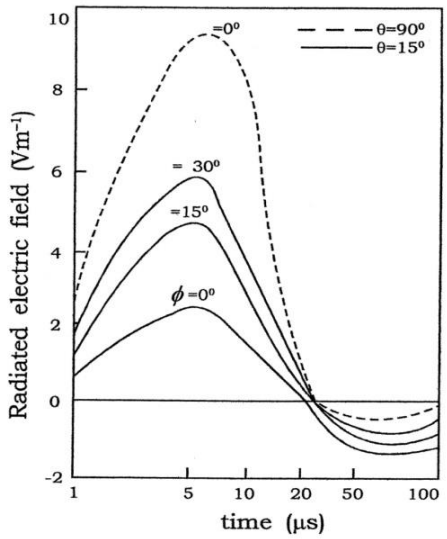

Fig.(2) Variation of the radiated electric field from oblique return stroke with time at distance, 100 km for different orientation of channel.

channel $\Phi = 0^0$ which is the consistent with the dipolar characteristics of the source. The electric field is seen to have maximum value at 5μs for every orientation of oblique channel and for every observation angle. After 5μs, VLF electric field decreases with increasing time and finally exhibits an excursion to negative values. It is clear from the figure that at same angle of observation, $\theta = 15^0$ and at observation distance, r = 100 km the peak value of the VLF electric field decreases with decreasing orientation of channel. Zero crossing of the radiated electric field depends on orientation of return stroke.

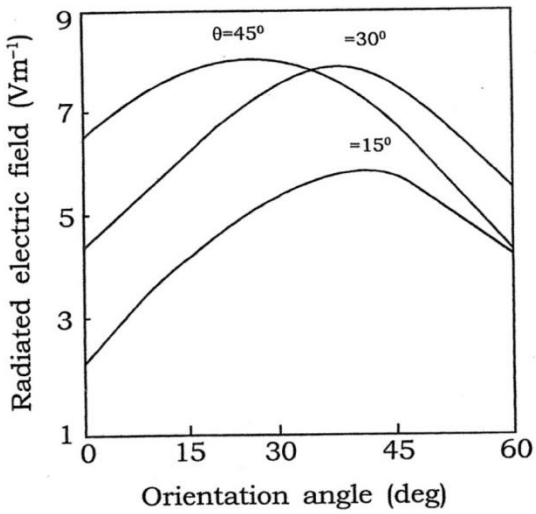

Fig.(3) Variation of radiated electric field from return stroke with orientation at observation distance, 100 km and at t = 5 μs.

The variation of the radiated electric field with orientation of oblique lightning channel for observation angles, θ = $15^0, 30^0$ and 45^0 at time 5μs and at observation distance, r = 100 km is depicted in Fig. (3). The radiated electric field increases with increasing orientation and is seen to attain a maximum value and there after decreases. At orientation, $\Phi = 31.5^0$ value of the radiated electric field is the same for observation angles, $\theta = 30^0$ and 45^0; after orientation 31.5^0 the electric field decreases more rapidly.

Orientation of the return stroke is considered from the vertical line joining the base of the channel. Therefore, the front of the draining process continuously moves away from the observation point. Thus, the polar diagram of radiated electromagnetic energy accordingly changes and gives rise to variability of the level of the radiated electric field [22].

In order to study the variation of the radiated electric field with frequency, computation is done by using expression (14) for different orientations of oblique return stroke. In this calculation the observation distance is kept at 100 km and angle of observation is assumed to be $\theta = 90^0$. The variations of the radiated electric field with frequency are shown in Fig. (4).The radiated VLF electric field variation with frequency has a Gaussian shape with peak at certain frequency. The orientation of return stroke governs the shape of the Gaussian curves.

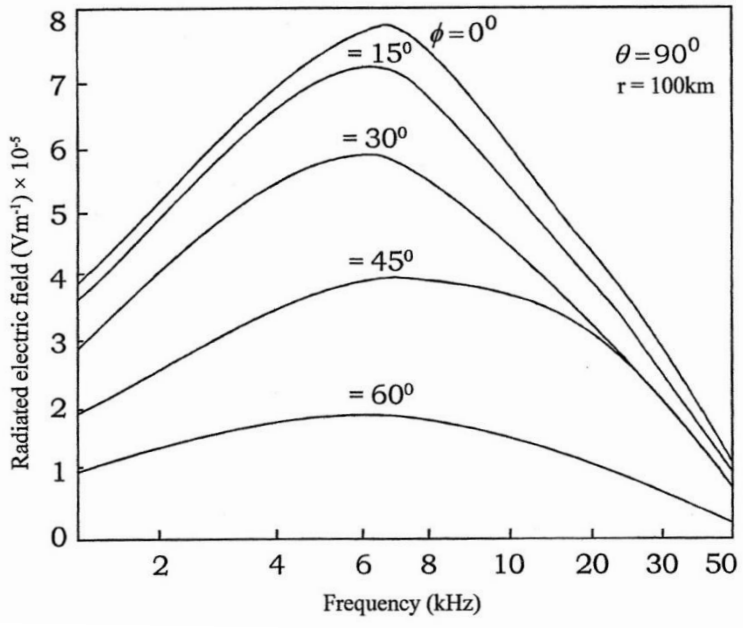

Fig. (4) Variation of radiated electric field of oblique return stroke with frequency at r = 100 km.

The radiated electric field increases with frequency and has its maximum value at 6 kHz for every orientation and there after decreases with frequency. The peak value of the field decreases with increasing orientation.

Fig. (5) shows the variations of the peak value of the radiated electric field from lightning return stroke with orientation at observation distance, r = 100 km and at frequency 6 kHz. The radiated electric field at frequency 6 kHz decreases from its maximum value with increasing orientation of lightning channel. This variation is might be due to the dependency of length of return stroke on the orientation of channel and change in the observation angle from θ to $\theta + \Phi$.

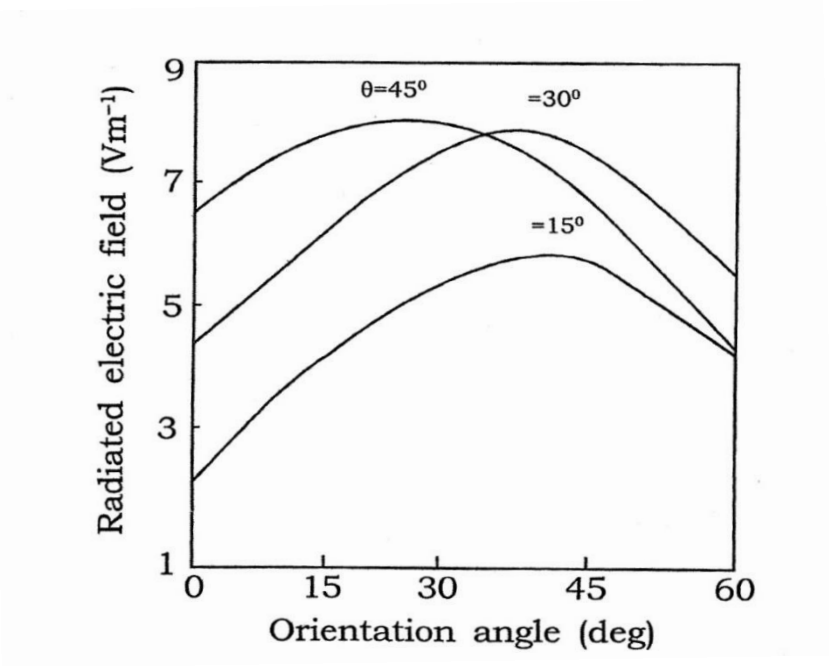

Fig. (5) Variation of radiated electric field from oblique return stroke with orientation at observation distance, r = 100 km and at frequency, f = 6 kHz.

Variations of the radiated power density with frequency for different orientations are shown in Fig. (6). In this computation the observation distance is kept at 100 km and observation angle is taken as $\theta = 90^0$. The power density variation with frequency for different orientations of lightning channel has Gaussian shape.

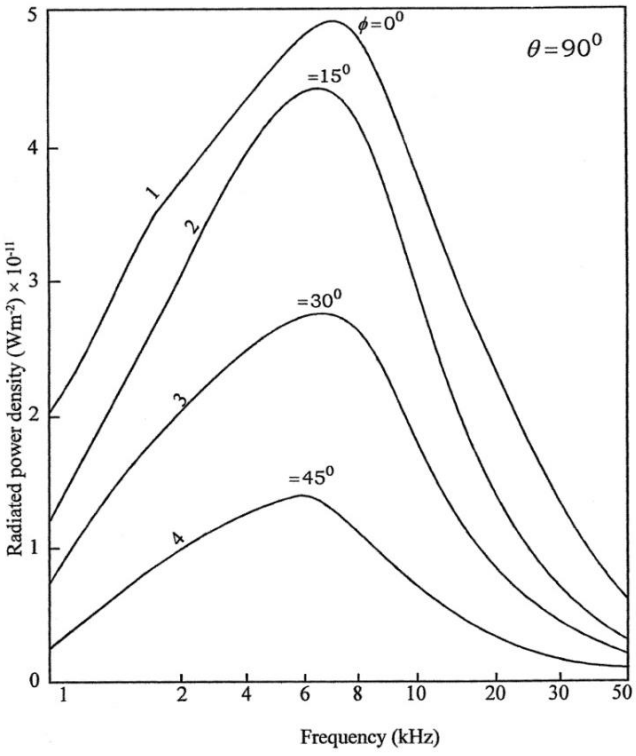

Fig. (6) Frequency spectrum of radiated power density of oblique return stroke at observation distance, r = 100 km.

Considering the half power density points, the band width of the radiated spectrum has been calculated. It is found that a band of frequencies containing different ranges is emitted out from lightning strokes depends on the orientation of the channel. The band widths of the radiated VLF waves as obtained from the Fig. (6) are given as

1. (1.7 - 15.0) kHz for $\Phi = 0^0$
2. (1.8 - 12.5) kHz for $\Phi = 15^0$
3. (2.0 - 09.6) kHz for $\Phi = 30^0$, and
4. (2.1 - 08.8) kHz for $\Phi = 45^0$.

The lower cut off frequency increases with increasing orientation of lightning channel while upper cut off frequencydecreases with increasing of orientation. Thus, it can be argued that the lightning return stroke, where input energy is supplied by thunder cloud, shows the behavior of amplification. Prasad and Singh [22] have argued that the lower and upper frequency cut offs observed in whistler sonograms can be used to derive approximate values of constants, I_0, α and β.

The maximum value of the power density is found at 6 kHz for each value of orientation, and it is found to decrease with increasing orientation.

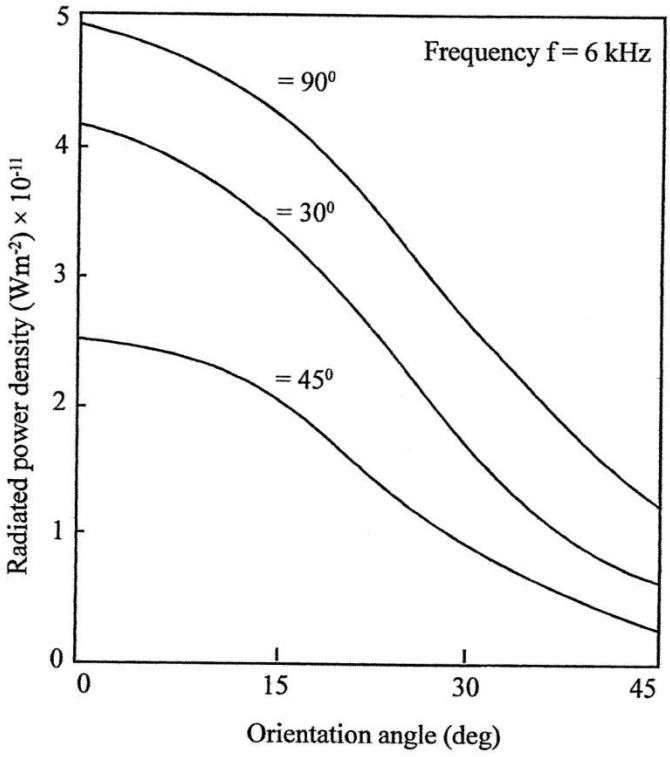

Fig. (7) Variation of radiated power density associated with oblique return stroke with orientation at observation distance, r =100 km.

The variation of the radiated power density at frequency, 6 kHz with orientation of the oblique return stroke is shown in Fig. (7). In this computation, observation distance is kept at 100 km and observation angles are taken as $\theta = 90^0$, 45^0 and 30^0. It is clear from this figure that the changes in the radiated power density associated with return stroke depend on the orientation of lightning channel.

V. CONCLUSIONS

In this study an attempt is made to calculate the radiated VLF electric field and power density from an oblique lightning return stroke at distance 100 km from the source. The aim of this study is to show that the orientation of oblique lightning channel which simulates return stroke plays an important role in controlling the shape and waveforms of the radiated VLF electric field and power density.

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A Comparative Study of Anthropometric Measurements, Physique and Body Composition of Interversity level Jumper Girls

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***Abstract*-** In the present cross sectional study of anthropometric measurements, physique and body composition of all Interversity level Jumper girls, specialising in high jump, long jump and triple jump have been reported. The data were collected from Dec.1997to Jan2000.The mean, standard deviation, analysis of variance and post-hoc test were applied for evaluation.

The result of the study shows that the high jumpers were found to be higher in mean of height and lower in mean value of weight as compare to long and triple jumpers but the test ANOVA shows non -significant differences in all anthropometric measurements except calf skin fold. In case of derived measurements the long jumper, high jumper and triple jumper are significantly different in height/weight ratio, Pondreal index and ectomorphy component.

***Index Terms*-** Anthropometry, % body fat, physique, lean body mass, jumpers.

I. INTRODUCTION

High Jump is an event where the competitors must jump over a bar at measured heights. Women's high jump event was a part of the track and field athletics programme at the 1928 Summer Olympics. It was the first appearance of the event marked the debut of women's Olympic athletics. The competition was held on Sunday, August 5, 1928.Twenty high jumpers from nine nations competed. Ethel Cather Wood of Canada won Gold medal and her winning record was 1.59 mts and at present high jump world record is 2.09 mts set by Stefka Kostadinova in 1987, and also the longest held record in the event. The long jump is a field event in which athletes combine speed, strength, and agility in an attempt to leap as far as possible from a takeoff point. The first world record in the women's long jump was recognized by the Federation Sportive Feminine International (FSFI) in 1922. The FSFI was absorbed by the International Association of Athletics Federations in 1936.The long jump was entered in Olympics in 1948 in London summer Olympics. Olga Gyarmati was a retired Hungarian athlete who competed at three Olympic Games in four different events. Her greatest success was winning the inaugural Olympic Women's Long Jump competition in London in 1948. Triple Jump incorporates a 'hop', 'step' and 'jump' where the distance is measured from the board where the athlete takes off from in the hop. The women's triple jump was introduced into the Atlanta Olympics in 1996.Inessa Mykolajivna Kravets Shulyak, was won the first gold Medal. She jumped the world record at the 1995World Championshipsin Gothenburg with 15.50m after studying a picture of Jonathan Edwards.

Anthropometry is a scientific specialized closely allied to physical education, sports science, sports Medicine, human biology, physical anthropology and several medicine disciplines. The knowledge of anthropometry is increasingly being appreciated by the sports administrators. Assessment of human physical performance through anthropometry helps to evaluate the physical structure and the performance of individual. So anthropometry is a science which deals with human body measurements and sports person are selected on the basis of bodily characteristics for a particular sport or event. Every game requires a particular type of a body and unspecific body types in relation to the sports events may be hindrance in the improvement and achievement of an athlete's performance. According to Sodhi and Sidhu (1984), "physique refers to the shape, the size and the form of an individual. All the three factors are intimately linked with each other and are manifestation of the internal structure and the tissue components which in turn, are influenced by environmental and genetic factors. The athletes in a particular sport must possess such typical characteristics which are of advantage to them during the game. The attainment of these characteristics will help an athlete to perform better during the competition". Therefore, it has been observed that apart from other factors the performance of sportsman, and sports women in any sport and game is influenced by various specific characteristics of physique, body composition, psychological traits and physiological functions which help him to attain better performance.(Cureton,1951;Tanner,1964;Hirata,1966and1979;de Garay e et al.,1974,Wolanski,1979,Carter,1982;Singh et al.,1987;Mokha and Sidhu,1988; Sandhu,1993) The present study is likely to impart new knowledge with regard to the anthropometric measurements, physique and body composition of all India interuniversity female jumpers.

II. MATERIAL AND METHODS

In the present investigation the All India Inter University jumpers have been studied for anthropometric measurements, physique and body composition .70 jumpers were studied which included 25 high jumpers, 20 long *jumpers* and 25 triple jumpers. Only those subjects were studied who had participated or held some positions in All India Interuniversity competition. The anthropometric measurements i.e. Height, Weight, Sitting Height, Upper arm circumference, fore arm circumference, Thigh circumference, calf circumference, waist circumference, Biceps skin fold, Triceps skin fold, Fore arm skin fold, Subscapular skin

fold, Superiallic skin fold and Calf skin fold , Humerus bicondylar diameter, Fermur diameter, Wrist and Ankle diameter were taken by using the standard techniques. Harpanden skinfold caliper was used for measuring skinfold. From the anthropometric measurements so taken, body mass index, Pondreal index, percent body fat, height/weight ratio, humerus dia/femur dia, wrist dia/ankle dia. and sitting height/height ratio and Somatotype components (endomorphy-mesomorphy-ectomorphy) has been calculated.

III. RESULT AND DISCUSSION

In the present study the anthropometric measurements have been taken on the categories of jumper's viz. high, long and triple jumpers. Out of the three categories of jumpers; on an overage, the high jumpers are the tallest with longest trunks and largest arm spans and lightest with less muscle at thigh, calf and the fore arm. They also posse's medium bony diameters of femur, wrist and ankle .The body mass index (18.39) and pond real index (22.38) is the least above said jumpers of this series. The long jumpers on the other side are heaviest with most massive musculature at fore arm. They also poses maximum body mass index (19.96) and pond real index (23.15). The long jumpers among three categories have the shortest stature, hand span and bony diameters at hummers, wrist, femur and ankle. Body fat at upper and lower extremities is the least. Among all the jumpers the triple jumpers posses highest hand span with massive bony upper and lower extremities and muscular thigh and calf .Their trunk and arm span and shortest with less muscular upper arms and waists of the whole series of jumpers.(Table 1$2)

When all the jumpers are compared for anthropometric measurements with each other, the values of ANOVA indicate significant F ratio in calf skin fold and pondreal index. Post-hoc test shows significant differences between long and triple jumpers.

Body Proportions:-

Table 2 indicates that height/weight ratio of the high jumpers is large (44.73), relations to weights. However, long jumpers are on the lowest (43.31) side and triple jumpers (43.87) fall in between in this regard. The value of ANOVAs indicates significant 'F' ratio at 0.05 levels (Table 4) .The Post-hoc test shows significant differences between long jumpers and high jumpers.

A resume of the study of height /weight ratio clearly indicates that high jumpers have less weight in relation to height. With this, they definitely have an advantage over the others because they have to lift their body against the gravitational pull. The lighter the weight the better are the chance of performance .Moreover, if the body height is large they will try to lift the body of higher levels.

In case of long jumpers the proportion of height/weight ratio indicates, inverse trend. They have more body weight in relation to height. Perhaps the reason is that when athlete is participating in a long jump he is allowed to run some distance to gain momentum .The larger the body weight the larger is the momentum which helps the individual to take his body to large distance.

The height /weight of triple jumpers fall in between the two, because the triple jump is a special event in which both momentum and lifting of the body up ward is required .Their other body proportions do not indicate and any difference wrist/ humerus diameters ankle /femur diameter and sitting height /height.

Body Composition

When the body composition of the jumpers of the present series is studied, it is found that high jumpers are the boniest with mean Ossa being 8.10 kg. The triple jumpers and long jumpers follow them with 8.03 and 7.64 values. In derma component in the jumpers, the triple jumpers are leading with the value 7.79 kg and the high jumpers and long jumpers follow them with 7.30 kg and 7.04 kg values. The body density, on an average in the jumpers is almost equally distributed. In case of % body fat again triple jumpers are most fatty value being 15.32% followed by high jumpers and long jumpers with 14.55% and 13.69 %.(Table 2)

This shows that the high jumpers are most bony with medium derma and percentage of body fat but almost equal body density among the jumper's series. This type of body composition of high jumpers with massive bony and medium musculature and fat might be helpful for getting the peak performance. The strong bones and medium fat component is useful in giving good performance in high jump. On the other hand, the triple jumpers have largest derma and % body fat with medium bony weight, because such types of bodies may be able to generate more momentum which is useful in performing better triple jumps.

Long jumpers have almost equal units of Ossa and derma components but less percentage of body fat among the jumper series. In other words, the main body weight of jumpers is constituted by bones with moderate percentage of body fat. This type of body composition might be helpful for increasing the momentum and maintaining a co-ordination in long jump execution for getting peak performance in long jump

Somatotypes

While studying the somatotypes of the present series, it is found that High Jumpers are the most ectomorphic, with mean ectomorphic component being 4.17. Triple jumpers and long jumpers follow them with 3.53 and 3.20 values (Table 2). The endomorphic and mesomorphic components of the jumpers are almost equality distributed. On an average the high jumpers are endomorphic ectomorph. This shows that the high jumpers have thin and lean body types which help them to jump to a higher level of the bar which require lifting of relatively less body weight, whereas the long jumper have almost same endomorphic and mesomorphic components. Both the endomorphic and mesomorphic components seem to help increase the momentum which results in the better performance in long jump. On the other hand the triple jumpers have endomorphic-ectomorph somatotypes. The ectomorphic component is helpful for lifting the body during the process of triple jumpers.

IV. COMPARISON WITH OLYMPIANS AND OTHER JUMPERS

We have compared the stature and body weight between and jumpers of presents series with Olympians of Mexico city 1968(de Garay et al., 1974) [4], Olympians of Montreal 1976 (Carter J.E.L.,1982) [5] , Olympians of Tokyo 1964 (Hirata K.I.,1979) [6] Olympians of Munich 1972 (Novak et al., 1972)[7], Brazilian jumpers (Guimaraes et al., 1978 , 1980) [8], state jumpers of San Deigo (Westlake , 1967)[9] . Similarly, the pond real index and height/weight ratio of present series of jumper has been compared with Olympians of Montreal, Tokyo and Munich studied by Hirata in 1979.While comparing our data with Olympians, it has been found that the Olympians posses larger height, weight and pondreal Index. Table 7 $ 8 depicts the picture more vividly.

MEAN, STANDARD DEVIATION AND STANDARD ERROR OF MEAN OF ANTHROPOMETRIC MEASUREMENTS (Table-1)

	Long Jump			**High Jump**			**Triple Jump**		
No. of Jumpers	20			25			25		
Measurements	Mean	SD	SE (M)	Mean	SD	SE (M)	Mean	SD	SE (M)
Height	160.09	5.92	1.32	163.68	6.99	1.40	161.79	5.89	1.18
Sitting Height(cm)	79.65	3.01	0.67	80.92	4.03	0.81	79.43	3.54	0.71
Weight (kg)	51.08	7.78	1.74	49.43	7.08	1.42	50.40	5.22	1.04
Arm Span	163.63	7.62	1.70	166.72	6.91	1.38	163.02	6.90	1.38
Hand Span (cm)	19.30	1.09	0.24	19.82	1.34	0.27	20.12	1.55	0.31
Humerus Diameter (cm)	5.79	0.37	0.08	5.90	0.35	0.07	5.87	0.35	0.07
Wrist Diameter (cm)	4.68	0.24	0.05	4.82	029	0.06	4.87	0.39	0.08
Femur Diameter (cm)	8.52	0.69	0.15	8.67	0.51	0.10	8.68	0.38	0.08
Ankle Diameter (cm)	6.18	0.37	0.08	6.23	0.33	0.07	6.28	0.30	0.06
Upper Arm Cir. (cm)	22.77	2.19	0.49	22.90	3.55	0.71	22.54	1.67	0.33
Fore Arm Cir. (cm)	21.25	1.31	0.29	21.18	1.26	0.25	21.23	1.16	0.23
Thigh Cir. (cm)	49.44	4.14	0.93	48.92	3.84	0.77	51.22	3.17	0.63
Calf Circumference (cm)	31.23	1.79	0.40	31.02	2.08	0.42	31.95	1.97	0.39
Waist Cir. (cm)	68.84	6.04	1.35	67.55	6.42	1.28	67.34	4.68	0.94
Biceps Skinfold (mm)	5.32	1.53	0.34	5.54	1.77	0.35	6.12	1.99	0.40
Triceps Skinfold (mm)	10.55	3.77	0.84	11.47	3.42	0.68	11.14	3.75	0.75
Fore Arm Skinfold(mm)	6.92	3.32	0.74	6.95	2.19	0.44	7.48	1.88	0.38
Superilliac Skinfold(mm)	7.08	1.89	0.42	7.86	2.80	0.56	8.80	2.90	0.58
Subscapular Skinf. (mm)	8.85	2.18	0.49	9.84	3.05	0.61	10.92	3.52	0.70
Calf Skinfold (mm)	11.06	5.14	1.15	14.41	6.18	1.24	17.69	3.78	0.76

MEAN, STANDARD DEVIATION AND STANDARD ERROR OF MEAN OF DERIVED MEASUREMENTS (Table-2)

	Long Jump			**High Jump**			**Triple Jump**		
No. of Jumpers	20			25			25		
Derived Measurements	Mean	SD	SE (M)	Mean	SD	SE (M)	Mean	SD	SE (M)
Endomorphy	2.65	0.77	017	2.94	0.87	0.17	3.12	0.81	0.16
Mesomorphy	2.55	0.97	0.22	2.19	1.02	0.20	2.44	0.80	0.16
Ectomorphy	3.20	1.39	0.31	4.17	1.11	0.22	3.53	1.04	0.21
Ossa	7.64	0.88	0.20	8.10	1.02	0.20	8.03	0.75	0.15
Derma	10.30	2.41	0.41	10.46	2.81	0.45	10.96	2.37	0.39
Body density	1.05	0.01	0.00	1.05	0.01	0.00	1.05	0.01	0.00
Percent body fat	20.05	3.02	0.57	20.93	3.52	0.62	21.59	3.13	0.55
Body Mass Index	19.96	3.10	0.69	18.39	1.92	0.38	19.25	1.67	0.33
Pondereal Index	23.15	1.23	0.27	22.38	0.78	0.15	22.82	0.75	0.15
Height/weight Ratio	43.31	2.16	0.48	44.73	1.52	0.30	43.87	1.42	0.28
Wrist/Humerus Dia.	81.18	5.84	1.31	81.79	5.04	1.01	83.19	7.75	1.55
Ankle/Femur Diameter	7308	7.71	1.73	71.94	3.85	0.77	72.42	4.15	0.83
Sitting Height/Height	49.77	1.37	0.31	49.44	1.39	0.28	49.10	1.61	0.32

ANALYSIS OF VARIANCE OF ANTHROPOMETRIC MEASUREMENTS (Table 3)

Measurements	Source of Variation	Sum of Squares	d. f.	Mean sum of Squares	F-Ratio
Height	Between Samples	144.95	2	72.48	1.818
	Within Samples	2671.00	67	39.87	
Sitting Height	Between Samples	31.639	2	15.81	1.226
	Within Samples	864.22	67	12.90	
Weight	Between Samples	31.06	2	15.53	.346
	Within Samples	3005.66	67	44.86	
Humerus Bic. Dia.	Between Samples	.148	2	7.393E-02	.580
	Within Samples	8.542	67	.127	
Femur Bic. Diameter	Between Samples	.358	2	.179	.637
	Within Samples	18.83	67	.281	
Wrist Diameter	Between samples	.390	2	.195	1.933
	Within Samples	6.760	67	.101	
Ankle Diameter	Between Samples	.102	2	5.091E-02	.465
	Within Samples	7.336	67	.109	
Hand Span	Between Samples	7.557	2	3.779	2.044
	Within Samples	123.837	67	1.848	
Arm Span	Between Samples	192.792	2	96.396	1.905
	Within Samples	3391.068	67	50.613	
Upper Arm Cir. (cms)	Between Samples	1.662	2	.831	.121
	Within Samples	459.689	67	6.861	
Fore Arm Cir. (cms)	Between Samples	4.593E-02	2	2.296E-02	.015
	Within Samples	102.834	67	1.535	
Thigh Cir. (cms)	Between Samples	71.989	2	35.995	2.621
	Within Samples	920.182	67	13.734	
Calf Cir.	Between Samples	11.572	2	5.786	1.503
	Within Samples	257.854	67	3.849	
Waist Cir.	Between Samples	28.381	2	14.191	.430
	Within Samples	2209.390	67	32.876	
Bicpes Skinfold	Between Samples	7.966	2	3.983	1.242
	Within Samples	214.907	67	3.208	
Triceps Skinfold	Between Samples	9.466	2	4.733	.357
	Within Samples	888.464	67	13.261	
Fore Arm Skin fold	Between Samples	4.703	2	2.351	.682
	Within Samples	408.560	67	6.098	
Superilliac Skinfold	Between Samples	33.711	2	16.856	2.467
	Within Samples	457.747	67	6.832	
Subscapular Skinfold	Between Samples	48.416	2	24.208	2.656
	Within Samples	610.657	67	9.114	
Calf Siknfold	Between Samples	489.911	2	244.955	9.320**
	Within Samples	1760.933	67	26.283	

ANALYSIS OF VARIANCE OF DERIVED MEASUREMENTS (Table 4)

Derived Measurements	Source of Variation	Sum of Squares	d.f.	Mean sum of Squares	F-Ratio
Endomorphy	Between Simple	2.530	2	1.265	1.878
	With Samples	45.125	67	.674	
Mesomorphy	Between Samples	1.617	2	.809	.928
	With Samples	58.377	67	.871	
Ectomorphy	Between Samples	11.010	2	5.505	3.995*
	With Samples	92.331	67	1.378	
OSSA	Between Samples	2.640	2	1.320	1.673
	With Samples	52.881	67	.789	
DERMA	Between Samples	5.445	2	3.322	.420
	With Samples	434.707	67	4.996	
Body Density	Between Samples	1.401E-04	2	9.31E-05	1.165
	With Samples	4.028E-03	67	4.967E-05	
Percent Body Fat	Between Samples	26.286	2	14.707	1.245
	With Samples	707.319	67	7.898	
Body Mass Index	Between Samples	27.757	2	13.878	2.755
	With Samples	337.518	67	5.038	
Pondreal Index	Between Samples	6.688	2	3.344	3.973*
	With Samples	56.393	67	.842	
Height / Weight Ration	Between Samples	23.344	2	11.672	4.076*
	With Samples	191.865	67	2.864	
Wrist/Humerus Diameter	Between Samples	48.872	2	24.436	.606
	With Samples	2701.703	67	40.324	
Ankle/Femur Diameter	Between Samples	14.310	2	7.155	.252
	With Samples	1899.927	67	28.357	
Sitting HeightHeight	Between Samples	4.951	2	2.476	1.152
	With Samples	143.994	67	2.149	

- Significant at 0.05% level

Results of Post-Hoc Test in Anthropometric Measurements between different Events (Table 5)

Measurements	Long/ High	Long/ Triple	High/ Triple	Measurements	Long/ High	Long/ Triple	High/ Triple
Height	3.595	1.707	1.888	Calf Circumference	211	.713	.208
Sitting Height	1.270	.218	1.488	Waist Cir.	1.292	1.500	.208
Weight	1.648	.680	.968	Biceps Skinfold	.229	.805	.576
Humerus Bic. Dia.	.114	7.800	3.600	Triceps Skinfold	.918	.590	.2387
Wrist Dia.	.135	.183	4.800	Fore Arms Skinfold	2.800	.556	.528
Femur Dia.	.152	.164	1.200	Superilliac Skinfold	.785	1.729	.944
Ankle Diameter	4.300	9.500	5.200	Subscapular Skinfold	.999	2.079	1.080
Upper Arm Cir.	.126	.234	.360	Calf Skinfold	3.348	6.628*	3.280
Fore arm Cir.	6.100	1.700	4.400	Arm Span	3.090	.606	3.696
Thigh Cir.	.524	1.780	2.304	Hand Span	.521	.821	.300

Results of Post-Hoc Test In Derived Measurements (Table 6)

Variable	Long/High Jumpers	Long/Triple Jumpers	High/Triple Jumpers
Endomorphy	**.290**	**.476**	**.185**
Mesomorphy	**.363**	**.105**	**.258**
Ectomorphy	**.946***	**.331**	**.632**
OSSA	**.458**	**.393**	**6.504**
DERMA	**.171**	**.659**	**.487**
Body Density	**2.030**	**3.550**	**1.520**
Percent Body Fat	**.878**	**1.537**	**.650**
Body Mass Index	1.569	.713	.855
Pondreal Index	.766*	.328	.438
Height/Weight Ratio	1.422*	.557	.864
Wrist/Humerus Diameter	.608	2.006	1.398
Ankle/Femur Diameter	1.134	.658	.475
Sitting Height/ Height	.360	.665	.335

Comparison of Stature, Body Weight, Pondreal Index and Height/Weight Ratio of Present Study (High Jumpers) Table 7

Counteries	No.	Height	Weight	Height/Weight Ratio	Pondreal Index
Present Study	25	163.7	50.4	44.7	22.1
Brazil (Guimaraes et al. 1978,80)	11	174.4	56.0	45.7	
Maxico City 1968(de aray et al., 1974)	4	175.5	58.9	45.11*	
Montreal 1976(Carter,J.E.L.,1982)	40	177.0	62.5	44.60*	22.20 (Gold Medalist) 22.39(Avg.)
Tokyo1964(Hirata, K.I., 1979)	24	172.0	62.2	43.41*	23.02
Munich 1972 (Novak et al.,1972)	45	175.3	62.8	44.10*	22.70

Comparison of Stature, Body Weight, Pondreal Index and Height/Weight Ratio of Present Study (Long Jumpers) Table 8

Countries	No.	Height	Weight	Height/Weight Ratio	Pondreal Index
Present Study	20	160.9	51.1	43.31	23.15
Brazil (Guimaraes et al. 1978,80)	16	164.5	51.9	44.10	
Maxico City 1968(de Garay et al., 1974)	8	166.4	55.2	43.70*	
Montreal 1976(Carter,J.E.L.,1982)	41	170.0	59.5	43.55*	26.10 (Gold Medalist) 25.13(Avg.)
Tokyo1964(Hirata, K.I., 1979)	32	167.8	58.2	43.30*	23.03
Munich 1972 (Novak et al.,1972)	34	169.5	59.5	43.42*	25.14

- Calculated from the Avg. Values (Table 7 & 8)

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A Review on Ball Burnishing Process

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Abstract- Burnishing is a very simple and effective method for improvement in surface finish and can be carried out using existing machines, such as lathe. On account of its high productivity, it also saves more on production costs than other conventional processes such as super finishing, honing and grinding. Moreover, the burnished surface has a high wear resistance and better fatigue life.

A literature survey being specifically focused on Ball burnishing process is done .It gives a thorough idea about various workpiece materials, various cutting tools and machine tools, process parameters ,lubricants, variable measured and methodology used as well as the prominent levels of each, being observed in the researches till today.

Index Terms- Superfinishing, Burnishing, Materials, Process Parameters

I. INTRODUCTION

To ensure reliable performance and prolonged service life of modern machinery, its components require to be manufactured not only with high dimensional and geometrical accuracy but also with high surface finish. The surface finish has a vital role in influencing functional characteristics like wear resistance, fatigue strength, corrosion resistance and power loss due to friction. Unfortunately, normal machining methods like turning, milling or even classical grinding can't meet this stringent requirement.

Table 1.1 illustrates gradual improvement of surface roughness produced by various processes ranging from precision turning to super finishing including lapping and honing.

A. Burnishing

The burnishing process consists of pressing hardened steel rolls or balls into the surface of the workpiece and imparting a feed motion to the same. Ball burnishing of a cylindrical surface is illustrated in Fig.1.1

During burnishing considerable residual compressive stress is induced in the surface of the workpiece and thereby fatigue strength and wear resistance of the surface layer increase. This process will smooth and harden the surface, creating a finish which will last longer than one that hasn't been burnished.

B. Classification of Burnishing Processes

Burnishing process can be typically classified into two categories as:

1. Based on deformation element
   a. Ball burnishing
      i. Flexible
      ii. Rigid
   b. Roller burnishing
2. Based on the motion of the tool, on the surface
   a. Normal or ordinary
   b. Impact
   c. Vibratory

C. Ball Burnishing

In the process the deformation element is hard ball. Alumina carbide ceramic, cemented carbide, silicon nitride ceramic, silicon carbide ceramic, bearing steel is the material used for ball. As ball acts as tool in deformation the surfaces layer, for the given normal force it gives high specific pressure, more fatigue strength, micro hardness & depth of work hardening layer as compared to roller burnishing. As there is a point & rolling friction between the ball & the work piece, the deformation zone is located adjacent to the ball on the work piece.Fig.1.1 represents scheme of ball burnishing process. Burnishing tools are also now widely applied in nonautomotive applications for a variety of benefits; to produce better and longer lasting seal surfaces; to improve wear life; to reduce friction and noise levels in running parts; and to enhance cosmetic appearance. Examples include valves, pistons of hydraulic or pneumatic cylinders, lawn and garden equipment components, shafts for pumps, shafts running in bushings, bearing bores, and plumbing fixtures [27].

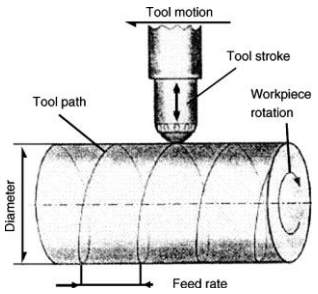


Fig.1.1. Scheme of ball burnishing [9]

II. WORKPIECE MATERIALS THEIR PROPERTIES AND APPLICATION

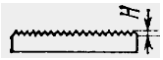
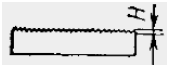
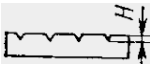
A. Aluminium and its alloys :

Properties –

U M Shirsat and B B Ahuja (2004) **[1]** used aluminium because Pure aluminium can't be heat treated properly however some of the aluminium alloys can be heat treated to improve their mechanical properties up to certain extent. Aluminium material can't be machined properly on conventional and even CNC machines due to poor machinability.

M.H. El-Axir,and A.M. Abodiena (2008) **[6]** used Aluminum alloy 2014 as an experimental work material. This material was selected because of its importance in industry.In the study of M.M. El-Khabeery & M.H. El-Axir(2001) **[16]**,

Table 1.1. Comparison of Finishing Processes

Process	Diagram of resulting surface	Height of micro irregularity (µm)
Precision Turning		1.25-12.50
Grinding		0.90-5.00
Honing		0.13-1.25
Lapping		0.08-0.25
Super finishing		0.01-0.25

workpieces of 6061 aluminum alloy were used. This material was chosen because of its importance in industry and its susceptibility to degradation when burnished, through surface and subsurface damage. Aluminum alloys are particularly well suited for parts and structures requiring high strength-to-weight ratio.

Applications-

Aluminum alloys are probably the best known materials used extensively in aircraft and truck wheels, M.M. El-Khabeery & M.H. El-Axir(2001) **[16]**.

N.S.M. El-Tayeb (2007)**[20]** selected Aluminium 6061 because of its wide range of applications in the industry such as aircraft fittings, truck wheels, brake disc, hinge pins, couplings, brake pistons and hydraulic pistons. The strength and good mechanical properties make the AA 7075 aluminum alloy appropriate for the use in aerospace industry, Ugur Esme (2010) **[15].**

B. Steel and its alloys:

Properties-

The hardness of PDS5 tool steel is about HRC33 (HS46). One specific advantage of this material is that after machining, the mold can be directly used without further heat treatment due to its special pre-treatment, Fang-Jung Shiou and Chien-Hua Chen (2003) **[4]** .L.N. Lo´pez de Lacalle et.al (2005) **[7]** used Tempered steel :This is a versatile through- hardening 5% Cr mould and die steel, with good wear resistance, polishability, and toughness. Mean hardness is 52 HRC.

A.A. Ibrahim et.al.(2009) **[9]** In his work, mild steel was used as a workpiece material. This material was selected because of its importance in industry and its susceptibility to degradation when burnished, through surface and subsurface damage.

Applications-

Fang-Jung Shiou and Chien-Hua Chen (2003) **[4]** used PDS5 tool steel in their study (equivalent to AISI P20), adequate for the mold of large plastic injection products in the field of automobile and domestic appliances. L.N. Lo´pez de Lacalle et.al (2005) **[7]** used Heat treated steel :The first application was on a pre-hardened steel (AISI P20) at 290–335 HB (31– 37 HRC,980–1130 MPa). This steel is used for the construction of injection moulds.Feng Lei Li et.al(2012) **[11]** used AA 7075 and AISI 5140 for validation . They are always used to produce some important components under the working conditions of alternate loading, and burnishing can increase the fatigue strength of them significantly.

M.H. El-Axir (2000) **[17]** used Steel-37 (0.20% C; 0.30% Si; 0.80% Mg; 0.05% P; 0.05% S) of hardness 220 *H*v. It was selected because of its importance in industry and its susceptibility to degradation when burnished, through surface and subsurface damage.“En-8” material, which is commonly used in shafts, press-rods and actuators. Consequently, it can be visualized that work-hardening phenomenon ob-served present in En-8, whereas for soft materials like Aluminium, material flow is reported due to plastic deformation, which is due to the soft and ductile nature of the material, C.S. Jawalar &R.S.Walia (2009) **[18].** AISI 4140 was chosen as a workpiece material by Y. Tian & Y. C. Shin (2007) **[19]**due to its wide applications.Burnishing experiments are conducted on turned mild steel work piece, which is ductile and available commercially in the form of round bars, M.Rao J. N. ,C. K. Reddy & P. V. R. Rao (2011) **[27].**

C. Polymers:

K.O. Low and K.J.Wong(2011) **[12]**used polymers in their study as polyoxymethylene (POM) and polyurethane (PUR).The two polymers were selected as each come from the two different groups of polymers, thermoplastic and thermosets. The POM is a semi- crystalline thermoplastic supplied by TNG Sdn. Bhd, Malaysia, while the PUR is a rigid thermosetting polymer supplied by RS Components Sdn. Bhd, Malaysia. The polymers were received as solid rods.

D. Titanium:

S. Thamizhmnaii et.al(2008) **[23]** used Titanium alloy Ti-6Al-4V . It was cut from square block and machined to 45 x 45 x 120 mm in size. The initial surface hardness was 272 HV.

E. Brass :

Commercially pure aluminium (99.85% Al) and brass (high speed machining brass, AS 1573- 385 and JIS H3250 C3604 standards), were used throughout the experiments. The aluminium and brass workpieces were received as bars with a diameter of 32 mm. These bars were cut to appropriate lengths then annealed in an electric furnace at 400°C for alumininum, and 580°C for brass for 1 h. After annealing, the workpieces were turned to proper diameters according to the experimental tests required. The surface hardness of the metals after annealing and turning was as follows: *HK*it=35.48 kg/mm2 for aluminium;

Table 2.1. Materials, Their Different Alloys and Studies Appeared in Literature

Materials	Alloys	References
Aluminium	Cast Al-Cu alloy Al 7075 T6 AA2014 AA 7178 AA 7075 Al Alloy 6061-T6 Al 6061	U M Shirsat and B B Ahuja (2004) **[1]** ,A.M. Hassan and A.M.S. Momani (2000) **[2]** A.M. Hassan and A.M.Maqableh(2000)**[3]** ,H. Basak and H. H. Goktas (2009) **[5]** ,. El-Axir,and A.M. Abodiena (2008) **[6]** ,Aysun Sagbas (2011) **[10]** ,Feng Lei Li et.al(2012) **[11]** , Ugur Esme (2010) **[15]** , M.M. El-Khabeery & M.H. El- Axir(2001) **[16]** ,N.S.M. El-Tayeb (2007)**[20]** S. Thamizhmanii, B. Saparudin &S.Hasan(2007)**[21]** J.N. M. Rao, A.C.K.Reddy and P.V. R.Rao (2010) **[25]**
Steel	Plastic Injection Molding Steel Heat treated and Tempered steel Hardened Steel Mild Steel AISI 5140 Steel-37 Steel En24 X5CrNiMo17-12-2 O1 Alloy Steel En-8	Fang-Jung Shiou and Chien-Hua Chen (2003) **[4]** , Lo´pez de Lacalle et.al (2005) **[7]** ,L. Luca, S. N.-Ventzel and I. Marinescu(2005) **[8]** , A.A. Ibrahim et.al.(2009) **[9]** ,Feng Lei Li et.al(2012) **[11]** ,Wit Grzesik and Krzysztof Zak(2012) **[13],**Dabeer P.S. and Purohit G.K.(2010) **[14]** , M.H. El-Axir(2000) **[17]** , Y. Tian & Y. C. Shin (2007) **[19],**Binu C. Yeldose, B. Ramamoorthy(2008) **[22]** ,Tomasz Dyl(2011) **[24]** ,K.S. Rababa and M.M. Al-mahasne(2011) **[26],**M.Rao J. N. ,C. K. Reddy & P. V. R. Rao (2011) **[27],**C.S. Jawalar &R.S.Walia (2009) **[18]**
Polymers	Polyoxymethylene (POM) and Polyurethene(PUR)	K.O. Low and K.J.Wong(2011) **[12]**
Brass	Heat-treated brass AS 1573-385	A.M. Hassan and A.M.Maqableh(2000) **[3],** A.M. Hassan and A.M.S. Momani (2000) **[2]** ,S. Thamizhmanii, B. Saparudin &S.Hasan(2007) **[21]**
Titanium	Ti6Al4V	S. Thamizhmnaii et.al(2008) **[23]**
Nickel	MP35 N	Yinggang Tian & Yung C. Shin (2007) **[19]**
Copper	--	S. Thamizhmanii, B. Saparudin &S.Hasan(2007) **[21]**

*HK*it=106.27 kg/mm2 for brass, A.M. Hassan and A.M.S. Momani (2000) **[2].**

F. Nickel :

Y. Tian & Y. C. Shin (2007) **[19],** for his study MP35N was selected as one of the Workpiece materials because of its unique hardening mechanism and excellent properties, such as high strength, excellent ductility and good corrosion resistance.

G. Copper :

S. Thamizhmanii, B. Saparudin &S.Hasan(2007) **[21]** used three types of non ferrous work pieces namely Aluminum, Brass and Copper were used. These three materials are commercially available as square bars. The initial size of all the materials was 45 mm square and 100 mm long.

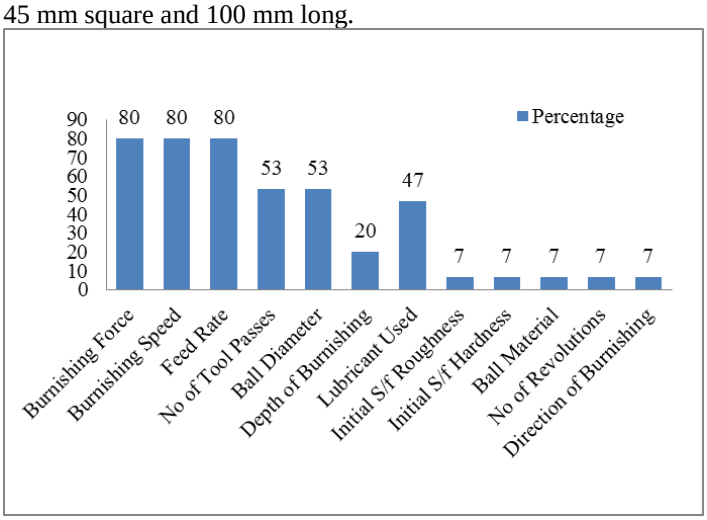


Figure 3.1: Parameter Percentage Importance According to the Authors Who are Studied (Ball Burnishing)

III. ANALYSIS OF PROCESS PARAMETERS

Work on the burnishing process has been carried out by various authors on Conventional Lathe [1-3, 17-18, 20, 27], CNC Lathe [6, 10, 11, 15, 19] and CNC Milling [4, 7, 21, 23]. Table summarizes the Parameters that different authors considered in the Ball Burnishing Process. Fig.3.1 represents the percentage importance of the ball burnishing parameters according to the authors which are studied in the literature.

Parameters like Burnishing force, Speed and Feed were considered the most i.e.80%, secondly the no of tool passes and ball diameter (53%),then Lubricant used (47%),the lesser was Depth of Burnishing and Initial s/f roughness, initial s/f hardness ,Ball material, No of revolutions and Direction of Burnishing were the least considered (7%), by the authors.

IV. LUBRICANTS AND THEIR PERFORMANCE

This section focuses on different lubricants appeared in the literature along with their results being seen on surface finish. At different values of force, speed and feed, kerosene gave best surface finish as compared to SAE-30 oil,5% and 10% graphite in SAE-30 oil by weight[1].In order to study the effect of lubricants on the ball burnishing process, different types of commercial lubricants were used by A.M. Hassan and A.M.Maqableh(2000) [3].

Lubrication, in this case using the machining coolant. In the burnishing process, coolant forms an elasto-hydrodynamic film between the ball and the workpiece's surface, avoiding the adhesive contact of materials on the ceramic ball and removing frictional heat [7].Since dry burnishing conditions produced poor surface finish, it was decided to apply a suitable lubricant during all tests which was emulation-type Soluble oil mixed with water [9].Continuous lubrication (kinematic viscosity of 1.4m²/s at 40 ºC) was supplied from external reservoir to the balls during all burnishing operations [12]. Super finishing passes were done using the ceramic honing stone 99A320N10V and cooling medium containing 85% kerosene and 15% machine oil. Burnishing was performed under static ball–workpiece interaction using special burnishing tool equipped with bearingizing ball and controlled spring-based pressure system to generate force [13].

Amongst the lubricants, use of diesel effects in maximum hardness followed by kerosene and soluble oil.

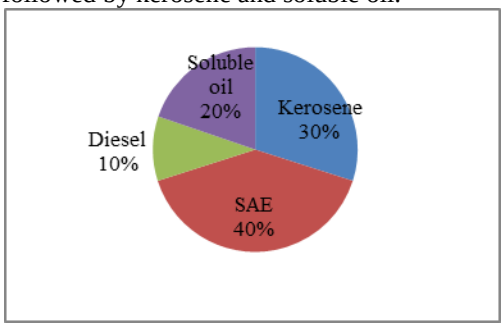


Fig 4.1. Lubricants and Their Performance

Hence in micro-hardness of En-8 Burnishing, number of passes contributes maximum (56.1%) followed by lubricant 22.9%, speed (18.37%) and feed has the least contribution [18]. For lubricated tribo-tests, a light viscosity lubricant (SAE 20W/40) was used. SEM examination of the worn surface reveals that interposing lubricant during tribo-test acts as a cooler and polishing agent, resulting in smoother surface compared to the burnished surface [20].

It is required that the lubricant has considerable effect on roller burnishing process. There will be higher reduction in surface roughness with increased burnishing force, in the presence of lubricant than in dry condition. A light oil such as kerosene oil produces better surface finish values than heavy oils such as SAE 40 engine oil on aluminium work piece [25]. Fig 4.1 shows the percentage usage of different lubricants by the researchers.

V. MEASURED VARIABLES AND RESULTS OBTAINED

The following section explores about various measured variables and summarizes few results from literature which gives an idea about the nature of measured variable.

A. Surface Roughness

Roughness is due to the inherent kinematic differences of the cutting process. Various parameters of surface roughness i. e. Ra, Rz, Rmax are measured by using Surface Roughness Tester, Centre line average (C. L. A.) or Ra value is the arithmetic average roughness height. Maximum reduction in surface roughness is observed in first five passes on mild steel by Roller Burnishing operation [27].

B. Surface Microhardness

The microhardness of the surface considerably decreases with increasing burnishing speed within the range used in this work. It is believed that the deforming action of the roller decreases with an increase in burnishing speed. Microhardness increases with an increase in the number of passes and/or depth of penetration. The highest surface microhardness was obtained with the combination of a high number of passes and high depth of penetration used in this work [16].The surface hardness of mild steel specimens increases with increase in the burnishing force up to 42 kgf. Further increase of burnishing force results in the decrease of surface hardness on mild steel specimens. The maximum surface hardness obtained is 70 HRB [27].

C. Surface out-of-roundness

Surface roundness plays an important role on the required tolerance and fit especially during part assembly. A roundness error is considered as one of the important geometrical errors for cylindrical components because it has negative effects on accuracy and other important factors such as wears in rotating elements. Also, it is well known that only plastic deformation takes place in the surface during the burnishing process, which, in turn, causes variation in the produced roundness [9].

- Therefore, surface roundness before and after burnishing was measured using ROUNDTEST RA-112-122. For better results the arithmetic average of three readings has been calculated [9].
- Surface roundness (maximum deviation) before and after internal burnishing was measured using coordinate measuring machine CMM. Similarly for better results the arithmetic average of three readings has been calculated [6].

D. Friction and Wear characteristics

The friction characteristics are assessed in terms of coefficient of friction of the polymers sliding against stainless steel. The coefficient of friction initially decreased with increasing burnishing force, attaining up to 32.9% and 28.8% reduction for POM and PUR respectively at burnishing force of 124 N, and then it increased [12].

Tribo-tests were performed using disc-on-ring and cross cylinder techniques to investigate the effect of burnishing parameters on friction and wear resistance under dry and lubricated contact conditions[20].

E. Tensile test

Tension test is probably the most fundamental type of mechanical test you can perform on material. Tensile tests are simple, relatively inexpensive, and fully standardized. By pulling on the material, As the material is being pulled, you will find its strength along with how much it will elongate. A curve will result showing how it reacted to the forces being applied. The point of failure is of much interest and is typically called UTS on the chart. The test was carried out on work piece at strain rate 1×10-3, the load-deflection curve was obtained from which the true stress-strain diagram is graphical for each alloy steel[26].

F. Residual Stress

An increase in burnishing speed within the range used in this work produces a significant decrease in the maximum compressive residual stress and may be changed gradually to tensile. This is because of the decrease in the deforming action of the roller burnishing tool with the increase in burnishing speed. It is well known that the generation of residual stress depends upon the interplay of many factors — such as in homogeneous plastic deformation induced by mechanical and thermal events associated with the process. The increase in the deformation action of the roller burnishing tool at low speed produces high plastic deformation at the burnished surface which leads to compressive residual stress [16].

Table 6.1.Summary of Ball Burnishing Process

Reference	Work Material	Machine Tool	Process Parameter	Measured Variable	Analysis Method	Remark
U M Shirsat and B B Ahuja (2004) **[1]**	Aluminium	Kirloskar Turn master T-40 Lathe	Burnishing Force, Speed, Feed Rate, Lubricant	s/f roughness and micro hardness	Mathematical model, ANOVA	-About 60-70% improved s/f finish obtained -At different values of force ,speed and feed ,Kerosene gave best s/f finish
A.M. Hassan and A.M.S. Momani (2000) **[2]**	Pure Al and Brass	Colchester master 2500 lathe	Burnishing Force, speed, Feed Rate , No of Tool Passes, Ball Diameter	s/f roughness and micro hardness, fatigue life and weight loss due to corrosion	Design of Experiment (DOE)	Shot peening and burnishing used on metallic components to improve certain properties
A.M. Hassan and A.M.Maqableh(2000) **[3]**	Brass and Cast Al-Cu alloy	Colchester master 2500 lathe	Ball Dia., Lubricant, Initial S/f Roughness , Initial S/f Hardness	s/f roughness and micro hardness and weight loss	DOE	An increase in ball diameter will cause a decrease in: (i) the final surface roughness; (ii) the total amount of the increase in hardness; (iii) the wear resistance
Fang-Jung Shiou and Chien-Hua Chen (2003) **[4]**	Plastic Injection Molding Steel	Three axis CNC Milling	Burnishing Force , Speed, Feed Rate, Ball Material	Optimal burnishing parameters	Taguchi's L18 orthogonal table, ANOVA ,full factorial	The optimal burnishing parameters for the plastic injection mold steel PDS5 were the combination of the tungsten carbide ball, the burnishing speed of 200 mm/min, the burnishing force of 300 N, and the feed of 40μm
Hudayim Basak and H. Haldun Goktas (2009) **[5]**	Al alloy (Al 7075 T6)	---	Feed Rate , No of Tool Passes, Ball Diameter, Burnishing Force	s/f roughness and micro hardness	Fuzzy logic	Best s/f roughness obtained at 300 N ,0.3 feed and 3 passes
M.H. El-Axir,and A.M. Abodiena (2008)	Al alloy 2014	A CNC lathe machine (Model, Biglia B56/1)	Speed, Feed Rate , No of Tool Passes, Depth of	Out of roundness(μm) and micro hardness (HV)	RSM with Box and Hunter method	s/f micro hardness and out of roundness better with depth of penetration range (0.025-0.045mm) and 3-4 passes

[6]			Burnishing			
L.N. Lo´pez de Lacalle et.al (2005) **[7]**	Heat treated and Tempered steel	5-axis High Speed Milling Centre	Burnishing Force , Speed, No of tool passes, Depth of Burnishing, Lubricant	Max height (Rt),Max Roughness (Rz),Mean roughness(Ra), micro hardness, microstructure photographs	DOE	Max pressure 30 MPa leads to highest quality improvement for the materials of 35-55 HRC
L. Luca, S. N.-Ventzel and I. Marinescu(2005) **[8]**	Hardened Steel	---	Burnishing Force , Speed, Feed Rate, Ball Diameter	Final roughness	Factorial Experiment	Result showed significant influence of pressure as well as original roughness after hard turning
A.A. Ibrahim et.al.(2009) **[9]**	Mild Steel	---	Burnishing Force , Speed, Feed Rate, Ball Diameter, Lubricant	s/f finish and s/f out of roundness	Self organising fuzzy logic controller, DOE	Experimental and fuzzy results showed that an increase in burnishing speed up to 1.5 m/s leads to a decrease in the burnished out-of-roundness whereas the increase in burnishing speed more than 1.5m/sec results in an increase in out-of-roundness
Aysun Sagbas (2011) **[10]**	Al alloy 7178	Industrial type of CNC lathe	Burnishing Force , Speed, Feed Rate , No of Tool Passes	s/f roughness	Desirability function approach (DFA) with RSM, Quadratic model, ANOVA	-Significant factors on s/f roughness were Burnishing force and no of passes -RSM helps determining appropriate burnishing conditions
Feng Lei Li et.al(2012) **[11]**	AA 7075 and AISI 5140	C6132A Lathe machine	Burnishing Force , Speed, Depth of Burnishing, Lubricant	s/f roughness, microscopic topography	Analytical modelling, DOE	s/f roughness proportional to burnishing force to 2/3 and ½ power in roller and ball burnishing respectively
K.O. Low and K.J.Wong(2011) **[12]**	Polyoxymethylene (POM) and Polyurethene (PUR)	---	Burnishing Force , Speed, Feed Rate , No of Tool Passes, Ball Diameter, Lubricant	s/f roughness, wear and friction characteristics	SEM images, DOE	Lowest s/f roughness 0.44 µm and 0.46 µm achieved for POM and PUR respectively
Wit Grzesik and Krzysztof Zak(2012) **[13]**	AISI 5140	---	Burnishing Force, Feed, Ball Diameter, Lubricant	s/f roughness	Optical images ,DOE	Improvement of s/f roughness is about 40%

Dabeer P.S. and Purohit G.K.(2010) **[14]**	Aluminium	---	Burnishing Force, Speed, No of Tool Passes, Ball Diameter	s/f roughness	Mathematical model, RSM	Optimum s/f finish obtained at 425rpm speed,7 mm ball diameter,70 N force and 2 tool passes
Ugur Esme (2010) **[15]**	AA 7075	FANUC GT-250B CNC lathe	Burnishing Force , Speed, Feed Rate , No of Tool Passes	s/f roughness and micro hardness	Grey based Taguchi method, ANOVA	s/f roughness and micro hardness are greatly improved

VI. SUMMARY AND DISCUSSIONS

Table 6.1 depicts the summary of literature reviewed for the present study which focuses on following points;

1. A review on burnishing thus gives us an overview of Materials, Burnishing tools and Machine tools along with process parameter and lubricant comparison.
2. It can be seen that Aluminum and steel materials are most preferred by the authors whereas a much scope is there to work on Polymers.
3. Amongst the ball burnishing process parameters Burnishing force, speed and feed were considered the most compared to Ball material, No of revolutions and Direction of burnishing.
4. SAE-oils are mostly used as Lubricants but use of diesel effects maximum hardness followed by Kerosene and then soluble oil.
5. Surface roughness and Microhardness are the popular measured variables.
6. Mathematical modeling and design of experiment are the commonly used methodologies. In the recent years, researchers are using Taguchi Method, Response surface Methodology and Fuzzy Logic for optimizing Burnishing Process parameters.

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Investigation of Genetic Relatedness of Ten Syngonium Cultivars using RAPD markers

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Abstract- For the first time, genetic variability and differentiation among Ten Synconium white butterfly populations originating from originating from tissue cultured plants maintained in pro-trays under the poly house & green house conditions in Indo American hybrid Seed Company, Bangalore were examined. Random Amplified Polymorphic DNA (RAPD) marker data were obtained and analysed with respect to genetic diversity. The ornamental crops Syngonium generate substantial revenue worldwide in the trade of ornamentals. The understanding in the pattern of genetic diversity of these plants has important implications in breeding programs and in the improvement of plant species. Initially RAPD analysis was used to assess the level of polymorphism, similarities and relationships among Syngonium populations collected from IAHS. RAPD analysis was done by screening 10 varieties of Syngonium using four primers. Totally 10,12, 23 and 25 polymorphic bands were observed in following four primers such as E20, C02, C05 and D18. Totally maximum and minimum bands were amplified in E20 (80 bands) CO2 (40) primers. While, the number of faded bands also increased in both primers. Based upon the E20 and D18 primer contains similar as well as maximum polymorphic bans also been observed the above said similar primer. Cluster analysis was done based on the UPGMA method using NTSYS software. Dendrogram analysis clearly noticed the Similarity Co-efficient for cultivars in the cluster I, II and II varied from 0.10-0.33, 0.78-1.0, 0.20-0.80 and 0.20-0.90 for E20, CO2, CO5 and E18 respectively.

Index Terms- RAPD markers, Genetic diversity, Protein markers, Cluster analysis, Dendrogram

I. Introduction

Syngonium a genus of ornamental plants in the family Araceae, are woody vines growing to heights of 10–20 m or more in trees. *Syngonium* species also called as arrowhead are often grown as house plants. There are several variegated cultivars, the main differences being in the position and extent of the cream or white markings. Some leaves are almost entirely white, pink or yellow. Propagation is by cuttings or air layering (Bijender Singh *et al*., 2011). The *Syngonium* variety ranking about seventh in the number of foliage plants shipped to market plays a major place in the trade of ornamental plants being provided with molecular and somatic cell technologies which provides an essential underpinning to conventional breeding approaches. Molecular procedures facilitate the identification of novel germplasms for incorporation into breeding programs involving sexual hybridization. Genetic diversity analysis of *Syngonium* varieties helps in the genetic improvement by genetic manipulation of targets including modification of stature and floral characteristics and improved tolerance to pests, diseases, and environmental stress (Simi *et al*., 2012).

Knowledge of genetic relationships in crops is important for genetic resource conservation, plant breeding, variety protection and genetic evaluation (Weising *et al*., 2005). These differences include insertions, deletions, translocations, duplications and point mutations. They do not, however, encompass the activity of specific genes. Information on genetic diversity is essential in optimizing both conservation and utilization strategies (Devos and Gale, 1992). Genetic diversity analysis is an important area of applied research in ornamentals owing to the relatively high value of individual plants compared with the agricultural crops. Knowledge of genetic relationships in crops is important for genetic resource conservation, plant breeding, variety protection and genetic evaluation **(**Michele *et al.*, 2003**).** Co- migration of non-homologous fragments can be a problem at higher taxonomic levels, but is probably negligible at the intraspecific level or when the studied species are closely related (Ashok, 2010). RAPD markers have, despite the above-mentioned limitations, been widely used in plant genetics and breeding, e.g., to estimate genetic diversity, detect phylogenetic relationships, and create genetic maps (Weising *et al.*, 1995, 2005). RAPD markers can analyze genetic diversity and genetic relationships within rather large plant sets due to their technical simplicity and cost-effectiveness. RAPDs have been successfully used in conventional DNA fingerprinting in plants (Caetano-Anolles *et al.*, 1991**).** However a commonly experienced problem with RAPD analysis is its poor reproducibility (Devos and Gale, 1992). However no studies have been conducted to determine the genetic relatedness of Synconium cultivars, hence this work has been undertaken for following objectives; Extraction of high quality DNA from ornamentals – Syngonium and Dieffenbachia using CTAB method, DNA profiling of ornamentals by RAPD markers and analysis of genetic diversity among ornamentals by constructing dendrogram

II. Materials and Methods

2mg of leaf sample was taken and crushed by adding 600µl Extraction buffer containing 100mM Tris-Cl, 1.4M NaCl, 20mM EDTA, 2%CTAB, 2%Mercaptoethanol and 1%PVP using pestle & mortar. It was then taken in eppendorf tubes. The grinded mixture was incubated at 50^0C for 10 mins and 500µl of chloroform isoamyl alcohol (24:1) was added. The mixture was

kept in shaker for 15 mins and centrifuged at 12,000rpm for 12 mins at RT. About 70µl of supernatant was transferred to new tube, and then equal volume of isopropanol and 0.5V of 5M NaCl was added. The samples were incubated for 2-3hrs at -20^{0}C to precipitate the DNA. Tubes were centrifuged at 12,000rpm for 12mins at 4^{0}C to make the pellet. The pellet was washed with 200µl 70% Ethanol (spinned at 12,000rpm for 4mins at RT) and air dried. The DNA pellet was dissolved in 50µl 1x TE buffer and stored at -40^{0}C

The present study was undertaken with the aim of analyzing the genetic variation found among and between Syngonium white butterfly varieties using DNA and protein markers.

All the experiments for the genetic diversity analysis were carried out at Biotech Research Development Lab, Indo American hybrid Seed Company, Bangalore. All the materials and methods adopted during the study were described below,

Plant Material

Total ten varieties of *Syngonium sps.,* namely *white butterfly, novaka, pixie, bronze, neon red, dark green with pink, pink vein, silver moon*, S. *podophyllum* and *marianne* were used in this study. All these plant leaves were collected from the tissue cultured plants maintained in pro-trays under the poly house & green house conditions in Indo American hybrid Seed Company, Bangalore, India. Samples of young immature leaves were collected randomly from each of the plant variety using blade & butter.

Genomic DNA extraction

The plant tissues of ornamentals are notoriously difficult material for DNA isolation due to the presence of various Secondary plant products. A number of DNA isolation methods have been developed for different plant groups. A method developed by Doyle & Doyle, 1996 called CTAB method is followed for the genomic DNA extraction. In this method, CTAB extraction buffer is used.

CTAB is a cationic detergent which aids in the lyses of cell membranes and will form complexes with nucleic acids. Sodium chloride aids in the formation of nucleic acid – CTAB complexes, by masking the negative charge of DNA. EDTA will chelate the magnesium ion which is an essential cofactor for DNase, thereby inhibits its activity and β-Mercaptoethanol is a reducing agent which protects DNA from degradation by various oxidants. RNA, protein, polysaccharides, pigments and tannins in plant cells will be removed by treating the extract with RNase, chloroform and phenol, respectively.

After protein is removed DNA is purified by precipitation with ethanol. Isopropanol and absolute alcohol removes water molecules from DNA.

Protocol

2mg of leaf sample was taken and crushed by adding 600µl Extraction buffer containing 100mM Tris-Cl, 1.4M NaCl, 20mM EDTA, 2%CTAB, 2%Mercaptoethanol and 1%PVP using pestle & mortar. It was then taken in eppendorf tubes. The grinded mixture was incubated at 50^{0}C for 10 mins and 500µl of chloroform isoamyl alcohol (24:1) was added. The mixture was kept in shaker for 15mins and centrifuged at 12,000rpm for 12 mins at RT. About 70µl of supernatant was transferred to new tube, and then equal volume of isopropanol and 0.5V of 5M NaCl was added. The samples were incubated for 2-3hrs at -20^{0}C to precipitate the DNA. Tubes were centrifuged at 12,000rpm for 12mins at 4^{0}C to make the pellet. The pellet was washed with 200µl 70% Ethanol (spinned at 12,000rpm for 4mins at RT) and air dried. The DNA pellet was dissolved in 50µl 1x TE buffer and stored at -20^{0}C

Quality and quantity check of DNA

DNA was checked for its purity and intactness and then quantified. The genomic DNA was run on 0.8% agarose gel stained with ethidium bromide following the protocol of Sambrook *et al*., (1989) and was visualized in a gel documentation system (GelDoc-It 310 Imaging System., UVP, CA, USA). For quality check 3µl of genomic DNA mixed with 2µl gel loading dye was used.

Intact and pure genomic DNA as assessed by agarose gel electrophoresis was quantified with spectrophotometer (UltraSpec III). The absorbance for all accession was measured at 260 nm and 280nm and then the ratio of OD_{260}/OD_{280} was calculated.

PCR Amplification

PCR amplification for RAPD analysis was done using 10 primers supplied by Genei, to get polymorphic bands. The RAPD primers used for diversity analysis are C02, C05, D18 and E20. PCR were performed in 15 µl reaction mixture with 2µl of template DNA using different primers to check for better amplification.

III. RESULTS

In this study, fingerprinted a set of ten Synconium white butterfly populations originating from poly house and green house conditions in Indo American hybrid Seed Company, by means of RAPD markers in order to obtain the molecular data on their genetic background. Thus, we have demonstrated the reliability of RAPD analysis to detect DNA polymorphisms and relationships within ten varieties of *Syngonium* populations. The selected primers were highly discriminating since they were characterized by relatively high collective Rp rate (30.00), a high number of polymorphic markers and electrophoretic banding patterns. Totally 10,12, 23 and 25 polymorphic bands were observed in following four primers such as E20, C02, C05 and D18 respectively (Table-2). Out of seven, three primers didn't given any fair amplified product so these three primers were useful for successfully provided deep insights on the genetic background of the above mentioned populations. In RAPD literature, the presence of primers does not allow amplification to occur (Caetano-Annoles 1994), whereas others primers yielding faint banding profiles (Moreno *et al*. 1995, Ortiz *et al*. 1997) were reported. In addition, according to the results forwarded by Devos and Gale (1992), Penner *et al*. (1993) and This *et al*. (1997), some primers are more efficient than others in producing stable and reproducible profiles. The genetic divergence of the populations under investigation was confirmed at the DNA level. In fact, the UPGMA cluster analysis permitted the discrimination of all the genotypes and their sorting into 3 main groups in E20

primer followed by two in CO_2 and cumulative nature in D18 primer (2, 4, 8, 10). Thus, population clustering was made dependently from the original region. In the same way, the occurrence of RAPD fingerprints unique to each of the studied populations (Table 2) suggests that RAPD markers may constitute rapid molecular tools for assigning one genotype to its origin. On the other hand, RAPD markers are suitable to perform genetic variation analyses at both intra and inter population levels (Syankumar and Sasikuma 2007; Hadian *et al*. 2008, Zahid *et al*., 2009). According to Shannon's index, high levels of genetic diversity were detected between the populations rather than within populations (Table 2). The existence of low genetic diversity within the species has been mostly attributed to self pollination, unless other environmental pressures are influencing genetic diversity (Archak *et al*. 2002).

In E20 RAPD primer showed fragments ranging from 500bp to 2000bp through all the ten plants. The size of the DNA fragment 1000bp uniformly present in all the nine plants except wb8species Fig. 1. Although, higher 3000bp amplified bands have visibly noticed in wb8 plant with faded manner. Amoung the 10 plants of butterfly varieties analysed clearly (wb4 and wb8) showed indifference among the other plants, while the others resembled the same. Meanwhile intensity of bands such as 2 and 3 were different from other bands. Inorder to that this E20 primer illustrated cluster analysis showed poor similarity coefficient observed on both wb9 and wb10 plant species of butterfly Fig.2. The second primer of CO2 amplification result was showed in Fig.2. It was clearly proved the specified average smallest range to largest amplifying efficiency such as 100 to 1500bp. Despite, dominantly two adjacent thickened bands such as 600 and 1500bp were seen on wb1 to wb10. Moreover, dendrogram analysis clearly noticed highest similarity Co-efficient shared between wb5, 6 and wb9 plants. Additionally poor similarity also been noted on wb4 plants in CO2 primer Fig. 4.

The intra diversity analysis of Syngonium variety white generated 8 allelic banding patterns. The bands 3 to 7 are more intense while the band 1, 2 and 8 are faded nature. The bands 5 and 7 are present in all plants except wb9 plant and less faded in wb6. In cluster analysis identically 0.20% of similarity coefficient was noted on both wb7 and wb8 plant variety. Finally the fourth primer D18 expressed RAPD profile showed that 300 and 600bp were observed in all the ten experimental plants. Although higher molecular weighed fragment of 2000bp have seen only in wb8 plant. Furthermore very smallest range of band 250bp also specifically appeared in wb5 plant Fig.7. While the consistent similarity coefficient noticed between wb7 and wb8 plant (Fig.8).

Fig. 1: RAPD profile of Syngonium variety White butterfly with E20 primer

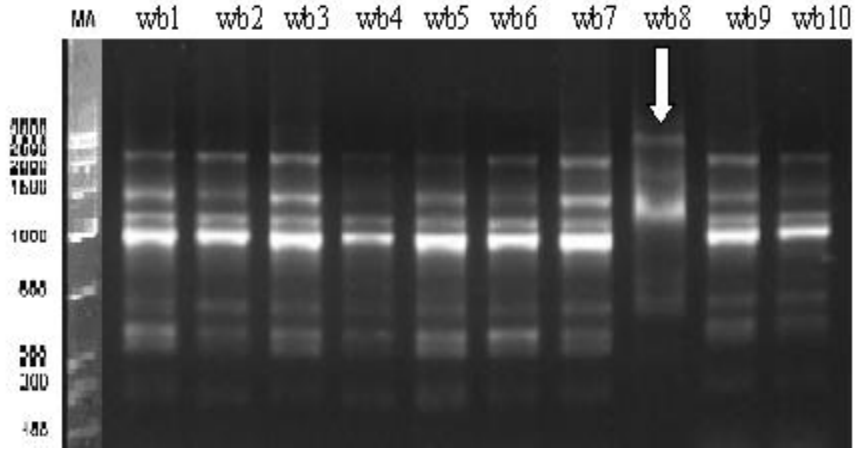

Fig. 1: RAPD profile for Syngonium variety White butterfly with E_{20} primer

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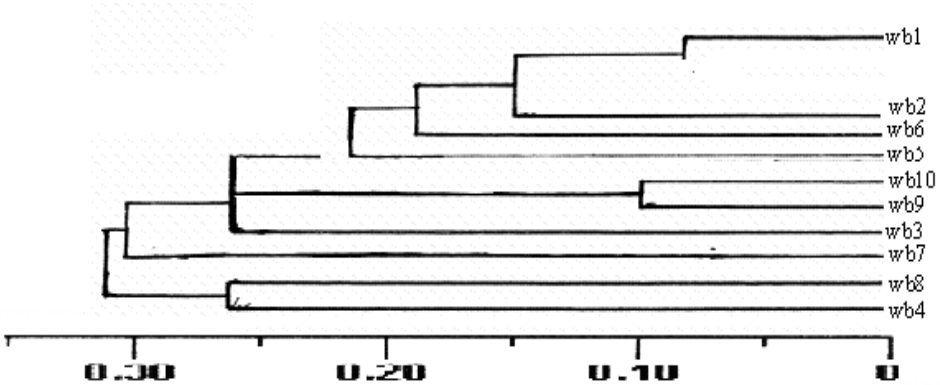
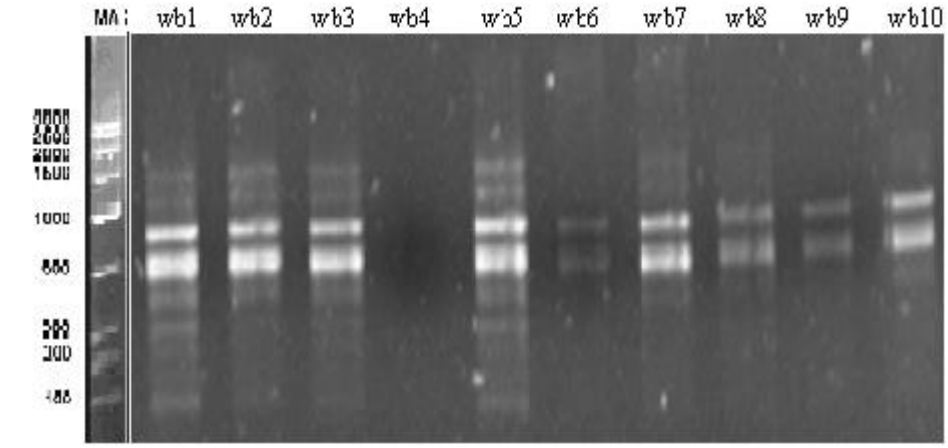


Fig. 2: UPGMA Cluster Analysis using simple matching similarity coefficients with E_{20} primer

Fig. 3: RAPD banding profile generated using C02 primer for Syngonium variety



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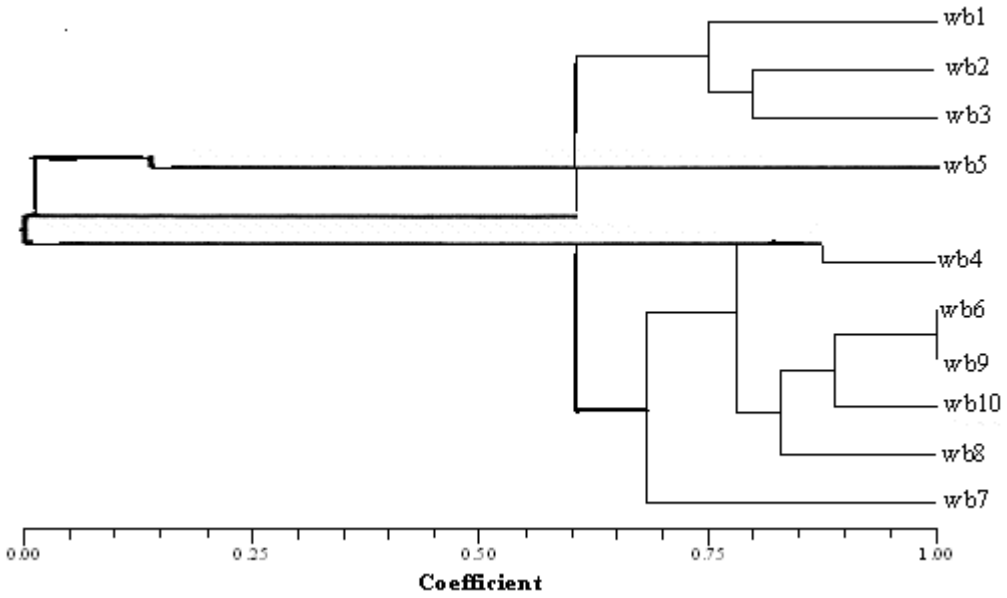

Fig. 4: UPGMA Cluster Analysis using simple matching similarity coefficients with CO_2

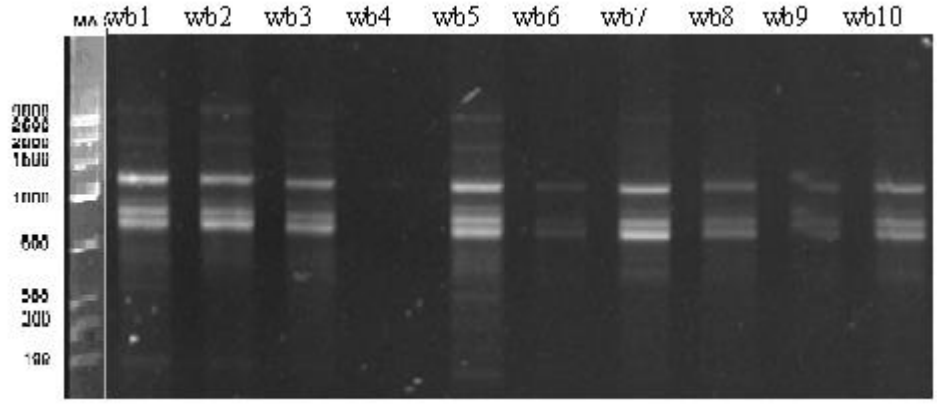

Fig. 5: RAPD banding profile generated using C05 primer for Syngonium variety

. **(B)**

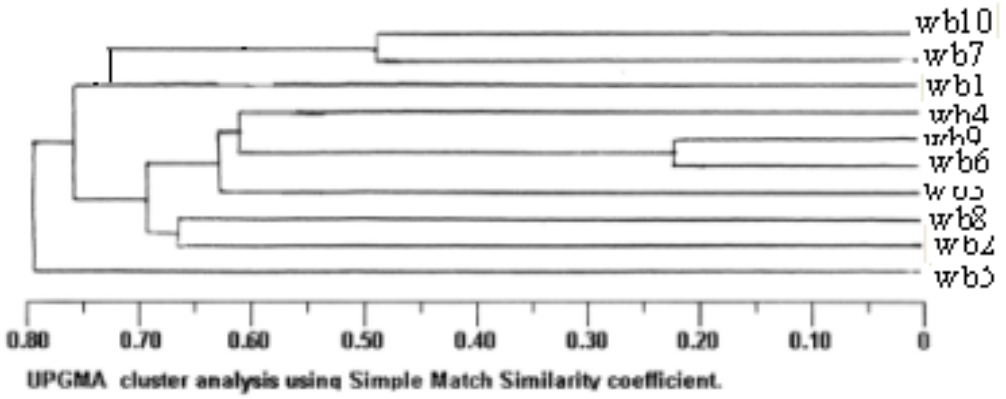

Fig. 6: UPGMA Cluster Analysis using simple matching similarity coefficient C05 primer

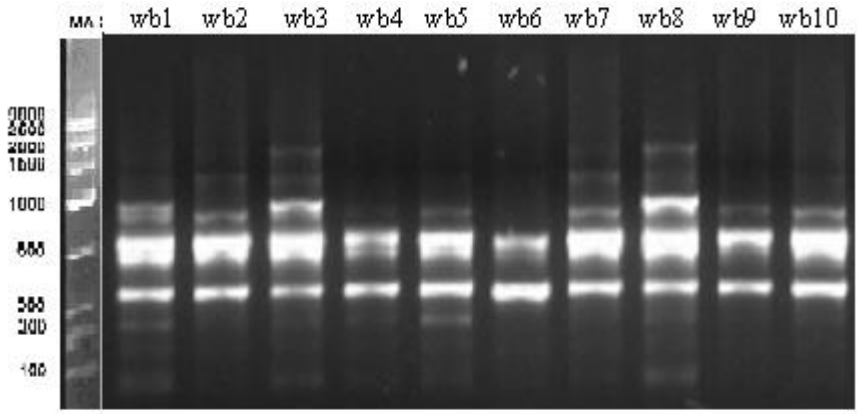

Fig. 7: RAPD banding profile generated using D18 primer for Syngonium variety

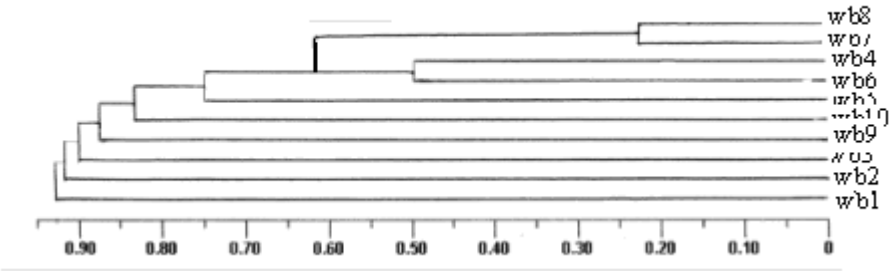

Fig. 8: UPGMA cluster analysis within the ten *Synchonium* white butterfly populations established by means D18 RAPD primer.

Table-1: Genetic diversity analysis through RAPD markers and Shannon's index among ten *Synchonium* white butterfly populations

No	Primer	Sequence (5-3´)	Total Bands	No. of faded bands	Polymorphic bands
1.	E20	5'AACGGTGACC3'	80	35	10
2.	C02	5'GTGAGGCGTC3'	40	27	12
3.	C05	5'GATGACCGCC3'	46	18	23
4.	E18	5'GGACTGCAGA3'	48	25	23

IV. CONCLUSION

To conclude, this is the first report on the assessment of genetic diversity and population differentiation analysis in ten Syngonium white butterfly populations originating from tissue cultured plants maintained in pro-trays under the poly house & green house conditions in Indo American hybrid Seed Company, Bangalore, using RAPD markers. This study successfully provided deep insights on the genetic background of the above mentioned populations. Considerable genetic diversity has been detected within the species. However, using more informative RAPD markers and including more populations would allow detecting the genetic background resources in local species of this similar variety too. Therefore, to prevent further substantial loss of genetic diversity, we should conserve as many populations as possible. In this context, that an appropriate strategy for sampling may be formulated when ex situ conservation is required. Besides, the best in situ conservation strategy for Syngoium population is to preserve its natural habitats. We further aim to include wild germplasm in order to select cultivars of economically importance and to transfer important genetic traits from wild to cultivated species by marker-assisted-selection. The dendrogram generated showed the Dieffenbachia and Syngonium varieties show a genetic similarity of 60%. The overall percentage of genetic diversity that exits between all the varieties is 40%.

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Effect of Oxygenated Fuel Dibutyl Ether on Diesel Engine

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Abstract- Stringent emission norms and environment degradation due to pollutants from the automotive vehicles lead us to find the suitable alternative for the petro-diesel. The oxygenated fuel used is dibutyl ether(DBE). While injection pressure is 200 bar and injection timing is 20.5^0BTDC we get higher brake thermal efficiency and lower brake specific fuel consumption for DBE10 compare to diesel.

Index Terms- injection pressure, oxygenated fuel, brake thermal efficiency

I. INTRODUCTION

Diesel engines have been widely used in recent decades as an alternative power source for light-duty vehicles because of the economic and environmental reasons. Therefore, the global diesel fuel consumption has increased with the pollutions sourced from diesel engines [1,2]. The unburned or partially burned (total) hydrocarbon (THC) emissions, smoke (soot) or particulate matter (PM), nitrogen oxides (NOx), sulfur oxides (SOx) emitted from compression ignition (CI) engines and particularly carbon dioxide (CO2) create severe environmental problems [2-4], which have been tried to be reduced by the stringent emission legislations. The different alternatives such as the investigation of viable alternative fuels and the reformulation of conventional fuels have been evaluated for meeting the emission standards and future energy demand [5,6]. The reformulation of diesel fuel contains the reduction of the sulfur and aromatic contents or the oxygen addition to diesel fuel [7]. A lot of works have been performed to show the effects of using alternative diesel fuels and additives including synthetic diesel fuels, biodiesels, alcohols and ethers.

The properties of DBE are almost similar to diesel, expect DBE as higher cetane number and lower density compare to diesel. DBE as higher oxygen content.

II. RESEARCH PROCESS

An effort is made in this study to evaluate the effect of varying the injection timing on the performance and emissions of a 5.2 kW engine fuelled with dibutyl ether (DBE10) for establishing the appropriate injection pressure and also to study the effect of oxygenated fuel.

III. STUDIES AND FINDINGS

Properties of diesel and simarouba biodiesel

Property comparison of Diesel and Simarouba bio diesel are shown in table 1

SL.NO.	Characteristics	Diesel	DBE 100%	DBE10
1	Calorific value (KJ/Kg)	43000	42800	42998
2	Viscosity at 40^0C	2.6-5	0.23	2.7
3	Cetane number	50	91	51
4	Flash point (^{0}C)	55	25	55
5	Specific gravity	0.84	0.767	0.839

IV. EXPERIMENTAL SET UP, PROCEDURE AND OBSERVATION

The experiment aims at determining appropriate proportions of biodiesel & diesel for which higher efficiency is obtainable. Hence .experiments are carried out at constant speed, comparing the performance of compression ignition engine operated on blends of diesel. The S20 blend is checked under loads 20%,40%,60% and 80% with injection timing 15.1^0 by varying injection pressure 200

bar,250 bar,300 bar and compression ratio 17.5. The samples are prepared by using the 1000 ml measuring jar and a 10 ml graduated test tube.

Fig.1 shows the schematic diagram of the complete experimental setup for determining the effects of waste cooking oil as bio diesel fuel additives on the performance and emission characteristics of compression ignition engine. It consists of a single cylinder four stroke water cooled compression ignition engine connected to an eddy current dynamometer. It is provided with temperature sensors for the measurement of jacket water, calorimeter water, and calorimeter exhaust gas inlet and outlet temperature. It is also provided with pressure sensors for the measurement of combustion gas pressure and fuel injection pressure. An encoder is fixed for crank angle record. The signals from these sensors are interfaced with a computer to an engine indicator to display P-Θ, P-V and fuel injection pressure versus crank angle plots. The provision is also made for the measurement of volumetric fuel flow. The built in program in the system calculates indicated power, brake power, thermal efficiency, volumetric efficiency and heat balance. The software package is fully configurable and averaged P-Θ diagram, P-V plot and liquid fuel injection pressure diagram can be obtained for various operating conditions.

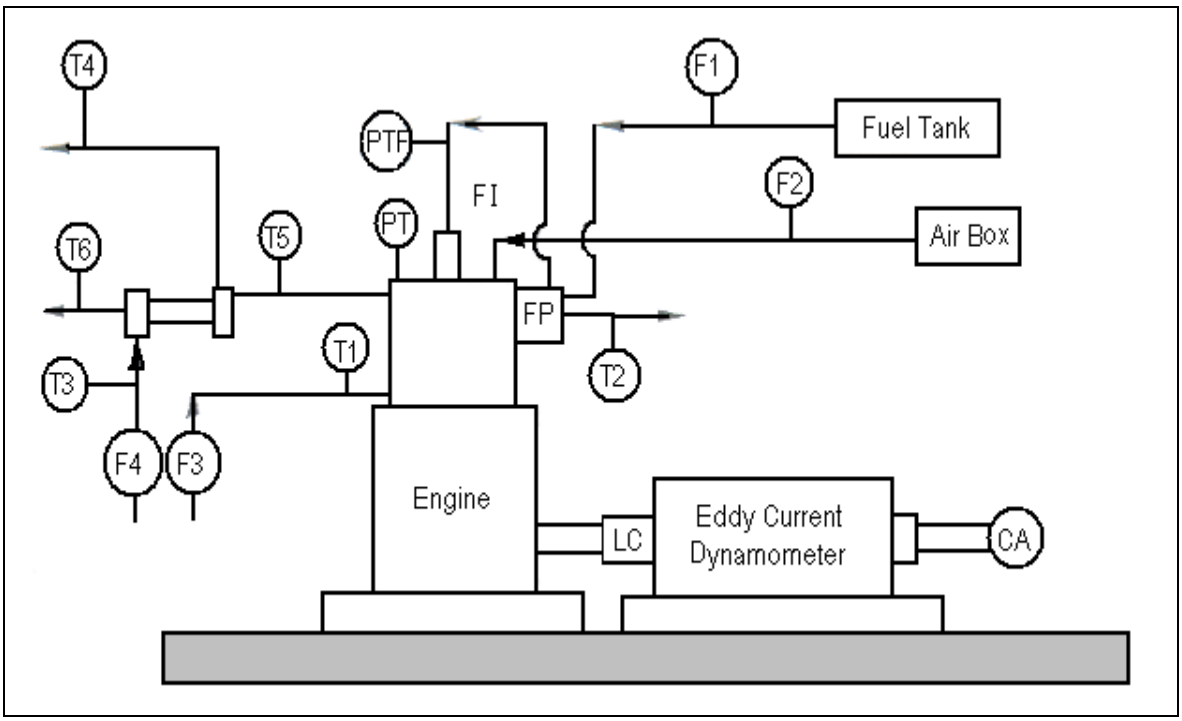


Fig. 3.1 Schematic Diagram of the Experimental Set-up

PT Combustion Chamber Pressure Sensor
PTF Fuel Injection Pressure Sensor
FI Fuel Injector
FP Fuel Pump
T1 Jacket Water Inlet Temperature
T2 Jacket Water Outlet Temperature
T3 Inlet Water Temperature at Calorimeter
T4 Outlet Water Temperature at Calorimeter
T5 Exhaust Gas Temperature before Calorimeter
F1 Liquid fuel flow rate
F2 Air Flow Rate
F3 Jacket water flow rate
F4 Calorimeter water flow rate
LC Load Cell
CA Crank Angle Encoder
EGC Exhaust Gas Calorimeter

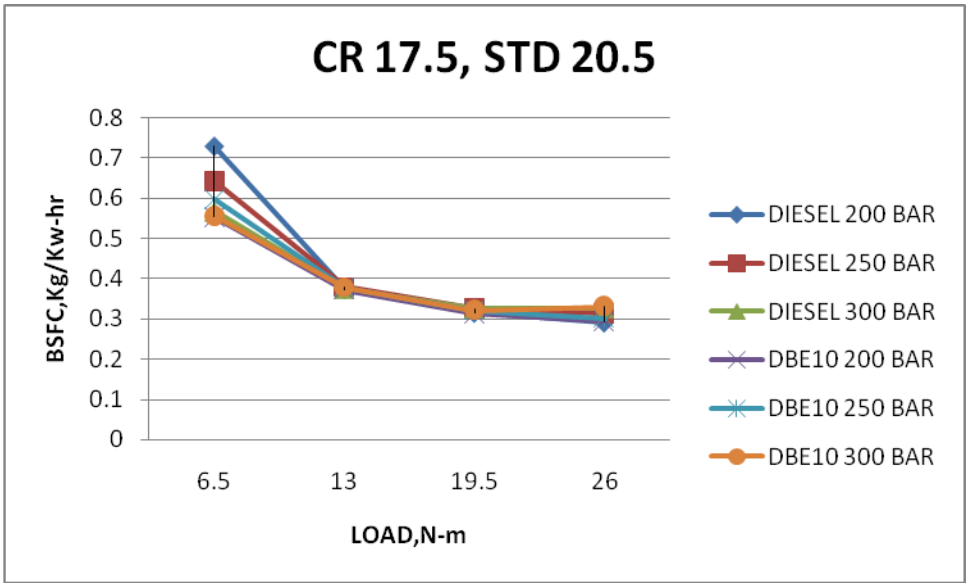
T6 Exhaust Gas Temperature after Calorimeter

ENGINE SPECIFICATIONS

SL.NO	Engine parameters	specification
1	Engine type	TV1(Kirloskar ,four stroke)
2	Rated power	5.2 KW at 1500 rpm
3	Bore	87.5 mm
4	Stroke	110 mm
5	Cubic capacity	661 cc
6	Compression ratio	17.5:1
7	Injection pressure	200 bar
8	Injection timing	20.5^0 BTDC

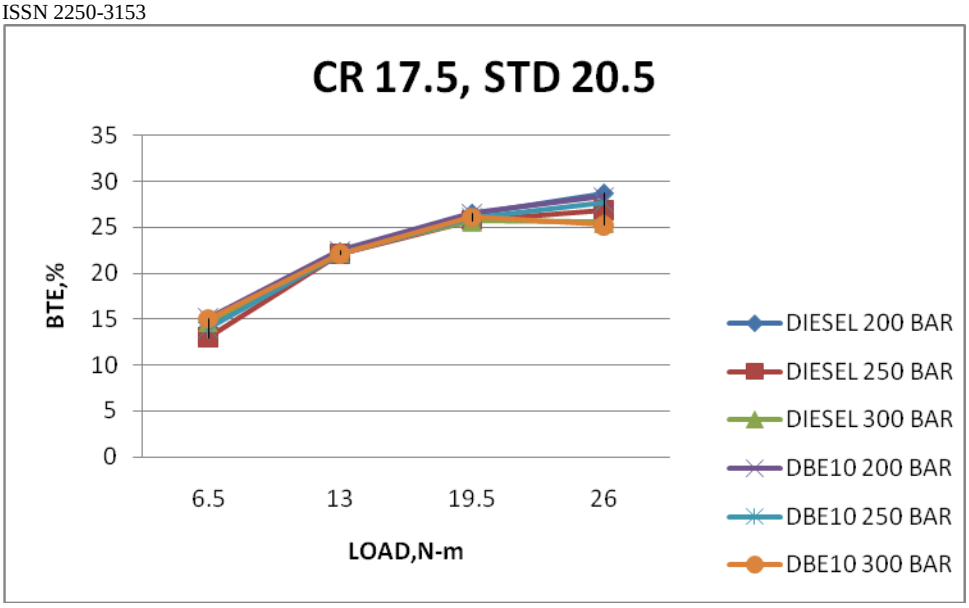
V. RESULTS AND DISCUSSIONS

Brake specific fuel consumption



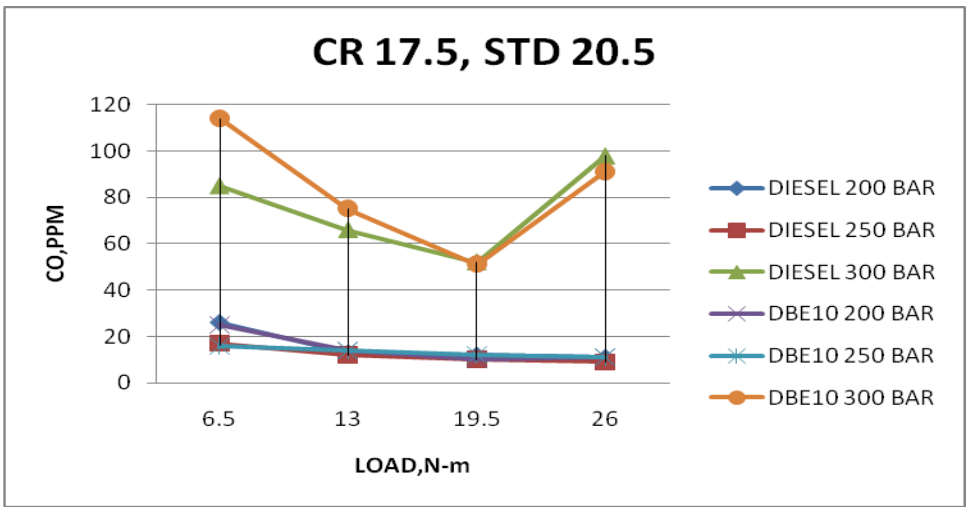
At 200 bar brake fuel consumption of DBE10 is lower than diesel.

Brake thermal efficiency



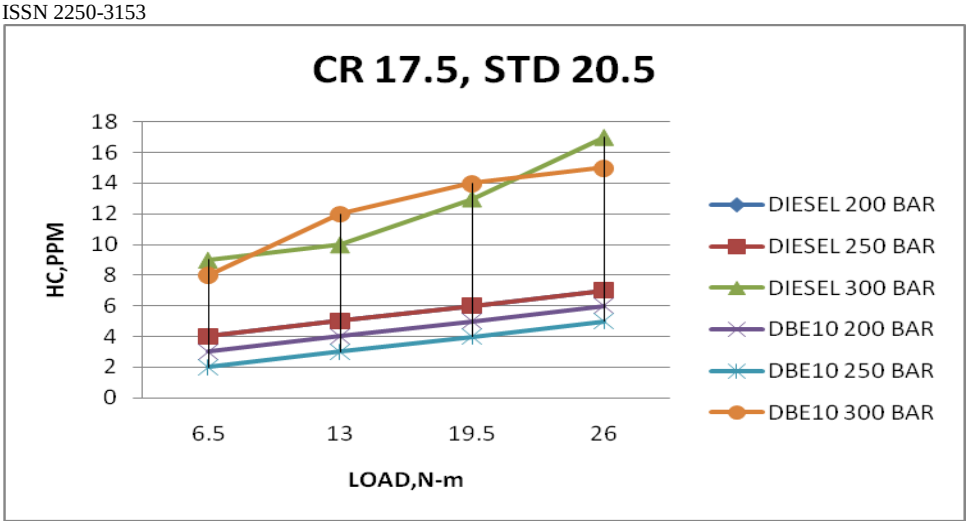
DBE10 gives better brake thermal efficiency compare to diesel in most of the cases.

Carbon monoxide



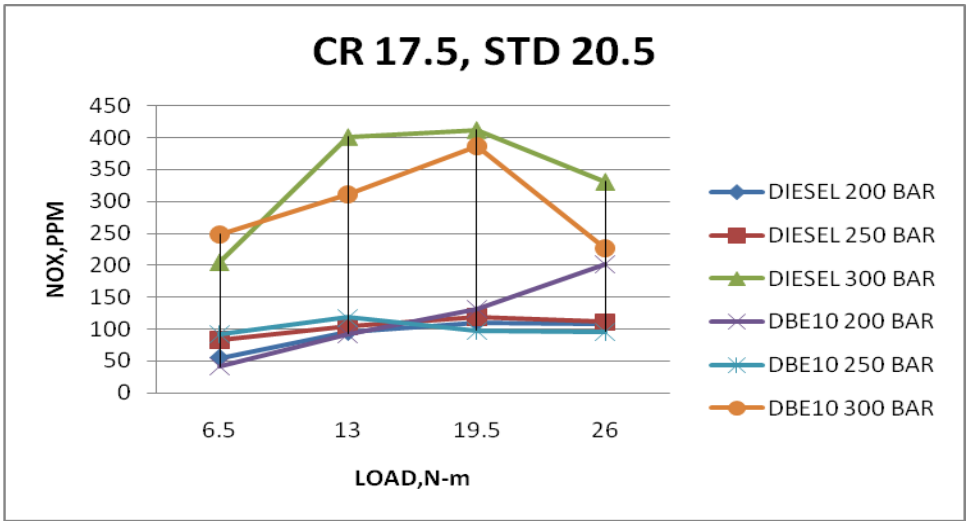
At 250 bar both DBE10 and diesel gives lower carbon monoxide emission.

Hydrocarbons



Lowest hydrocarbons emisson is observed at 250 bar for DBE10 compare to diesel.

Oxides of nitrogen



Lowest NOX emission is observed at 250 bar for DBE10 compare to diesel.

APPENDIX

BSFC-BRAKE SPECIFIC FUEL CONSUMPTION

BTE -BRAKE THERMAL EFFICIENCY

BTDC-BEFORE TOP DEAD CENTRE

CO- CARBON MONOXIDE

HC- HYDROCARBONS

NOX- OXIDES OF NOTROGEN

PPM- PARTS PER MILLENNIUM

DBE10-DIESEL+10 DROPS OF DIBUTYL ETHER

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UV Irradiation Assisted Photocatalytic Decolorization of Direct Red 23 in Aqueous Solution

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Abstract- Study on the photocatalytic decolorization of aqueous solution of Direct Red 23, a commercial azo-reactive textile dye, in the presence of UV light radiation. The effect of different reaction parameters on the photo catalyst of DR 23 was assessed best operational parameters were observed as photons: catalyst loading 1gm-3.0gm/50ml, pH 3.5-11.5, temperature 303k,dye concentration10mg/50ml-70mg/50ml and 2.30hrs light irradiation on photo decolorization. FTIR is employed successfully to study the behavior of dowex-11 Resin and after the immobilized by Methylene blue dye. FTIR spectrum is recorded in the range 400-4000cm-1.Activity of photo catalyst MBIR-11 remains impervious on continuous use.

Index Terms- Photocatalytic decolorization process, Methylene blue immobilized resin dowex-11, operational parameters, Direct Red 23, UV radiation.

I. INTRODUCTION

Now days, industrial waste water, detergents and a wide variety of textile dyes and other industrial dyestuffs constitutes one of the largest group of organic compounds that represent an continuously increasing the environmental dangers. This waste water causes damages to the ecological system of the receiving surface water capacity and certain a lot of disturbance to the ground water resources. Many dyes are used in the textile industries are stable to light and are not biodegradable. In order to reduce the risk of environmental pollution from such waste water[1]. Among these biodegradation, chlorination, ozonation and adsorption are the most commonly used conventional process. Effluents dyes are usually resistant to aerobic degradation and carcinogenic compounds may be generated during the anaerobic treatment, for example: - aromatic amines from azo dyes; in these respects, bio-treatment alone has been found to be ineffective for the treatment of dye effluents[2].

Decolorization of azo dye effluents has therefore received increasing attention for the removal of dye pollutants, traditional physical technique (adsorption on activated carbon, reverse osmosis, ion exchange on synthetic adsorbent resins, ultra filtration, coagulation by chemical agents, etc.) can generally be used efficiently[3-6]. Decolorization is a result of two mechanisms: adsorption and ion exchange (Slokar and Le Marechal(1998), and is influenced by many physio-chemical factors, such as, dye/sorbent interaction, adsorbent surface area, particle size, contact time, pH, and temperature (Kumar et al., 1998). Adsorption also does not end result in the formation of harmful substance[7,8].

Photocatalytic oxidation of azo dyes and other dyes have been investigated by a number of researchers. Photocatalytic oxidation processes can oxidize a wide variety of toxic and persistent organic compounds to harmless inorganics such as mineral acids, carbon dioxide and water (Dominguez et al., 1998)[9]. Also, in this process form several byproducts such as halides, inorganic acids, metals and organic aldehydes depending on the initial materials and the extent of decolorization (Robinson et al., 2001)[10].

In order to remove hazardous materials like dyes, adsorption is a method which has gain considerable attention in the recent few years adsorption is such and simple technique[11].Recently developments of advanced oxidation processes (AOPs) have led to new improvements of the oxidative degradation of the organic compounds. UV radiation in the presence of H_2O_2has yielded encouraging results of color removal from azo reactive dye containing water[12]. Advanced oxidation processes such as catalytic oxidation using photon Fenton, Fenton and visible/solar light system[13,14] H_2O_2/UV processes and titanium oxide photocatalysis[15,16].

Rate of degradation of acid green 16 was studied by Sakthivelet al[17]using ZnO irradiated with sun light. Yingma and Jian Nian Yao prepared a thin film on TiO2 and used it to study photodegradation of Rh-B. Ciping et al. observed that the formation of free radical intermediates in photoreactions with ZnO dispersion. Richard *et al.* reported that the oxidizing species involved in photocatalytic transformation on ZnO are either hydroxyl radicals or holes[18-20].

Lizamaet al[21] reported the photocatalytic decolorization of reactive blue 19(RB 19) in aqueous solution in the presence of TiO_2 or ZnO as catalyst and concluded that ZnO is a more efficient photocatalyst than TiO_2 in the removal the color of RB-19. Rambabu Pachwarya and R.C. Meena[22] have been investigated the degradation of textile azo dye Amido black- 10 B by newly developed photocatalyst (Methylene blue immobilized resin Dowex 11).

Direct Red 23 is the commercial textile azo dye which is caused some harmful effects. Therefore the removals of such type of dyes from effluents become environmentally important. So in this paper we studied, photocatalyst (MBIR-11) was observed for the decolorization efficiency of aqueous solution of a textile dye, Direct Red 23 in different conditions and observe the effect of many parameters on the rate removal efficiency.

II. MATERIALS AND METHODS

Molecular Formula	= C35H25N7Na2O10S2
Molecular Weight	= 813.72
Solubility	= Soluble in water
Appearance	= Red
Class	= Azo

λ_{max} = 508 nm

And double distilled water was used for preparation of various solutions.

Structure of dye

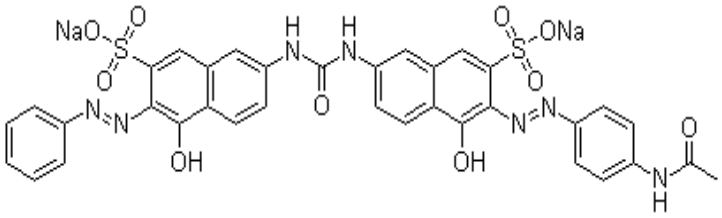


Fig.1 Structure of Direct Red 23

Apparatus

For the decolorization, the photocatalytic degradation was carried out in specially designed glass reactor containing an azo dye solution (volume 50 ml) and a defined amount of a photo catalyst and the reactor solution was illuminated with a 200W mercury lamp having wavelength 510nm. Constant stirring of solution in the reactor was insured by using magnetic stirrers. The temperature was maintained constant throughout the reaction time and we used borosilicate glass reactors. Solution was illuminated by 200W mercury lamp placed above reactor. At 10 min interval, 10 ml solution was taken out from reaction mixture. Filtered the catalyst particles and investigated the transparency of colored water with the help of 160UV/Visible spectrophotometer and lamp was positioned above the reactor.

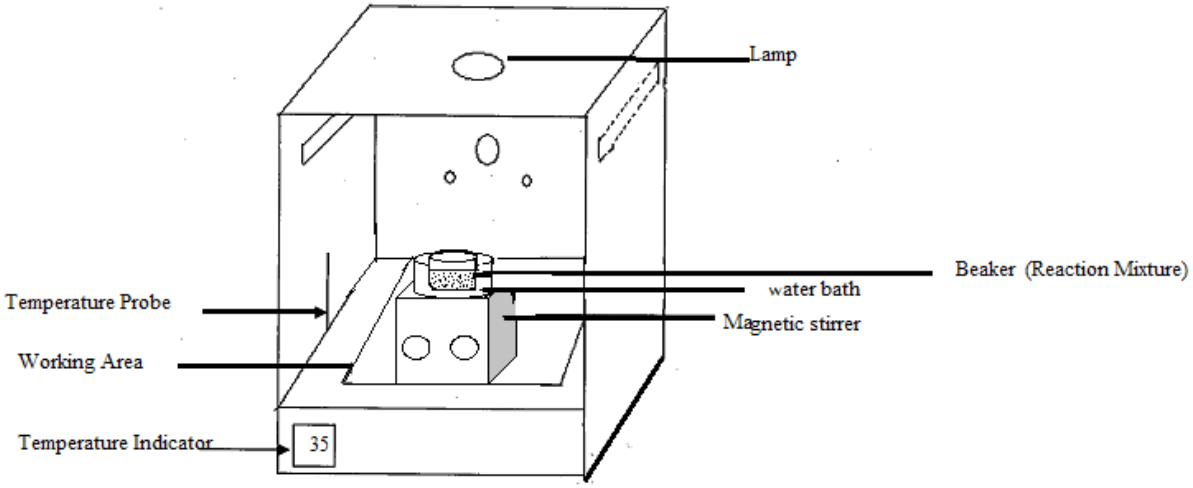


Fig.2 Experimental Setup of Photocatalytic process

Photocatalytic Experiments

To 50ml of dye solution, photocatalyst such as MBIR 11 was added to irradiation. Observation were finding out under solar light as well as UV light. The aqueous solution of dye was magnetically stirred throughout the experiment. After 10 min. intervals aliquot was taken out with the help of a pipette and then filtered the solution. Then rate of decolorization efficiency was investigated. The decolorization efficiency (%) has been calculated as:

$$\text{Efficiency (\%)} = \frac{Co - C}{Co} \times 100$$

Where Co is the initial concentration of dye and C is the concentration of dye after photo irradiation.

Similar experiments were carried out by varying the concentration of dye (10 mg- 70 mg/l), radical quencher, pH of the solution (pH 3-11), catalyst loading (1- 3gm) and light intensity.

III. RESULTS AND DISCUSSION

Mechanism of degradation of azo dye

We used Methylene blue immobilized resin Dowex-11 is newly developed photocatalyst. Due to its photosensitive nature, when light radiation is irradiated on its then electronic transition occurs from valence band (VB) to conduction band (CB) and through (ISC) electrons reach into triplet state of methylene blue. After it intermolecular electronic transition create between resin, methylene blue dye molecules, Direct red 23 dye molecules, water molecules and dissolved oxygen, consequential through chain process, hydroxyl radicals ($^{\bullet}OH$) and super oxide ions (O_2^-) are formed and these are highly oxidizing in nature, by the action of hydroxyl radicals and super oxide ions (O_2^-) on azo dyes, are transformed in simple organic compounds like CO_2, H_2O etc. The analytical analysis is confirmed the CO_2 and H_2O are present in products. On the basis of the experimental data, the following tentative mechanism may be proposed

$$\text{Dye} \xrightarrow{\text{Light Radiations}} \text{Dye}^{*1}$$

$Dye^{*1} \xrightarrow{ISC} Dye^{*3}$

$Dye^{*3} + I\text{-}R \longrightarrow Dye^{+} + I\text{-}R\ (e^{-})$

$I\text{-}R\ (e^{-}) + O_2 \longrightarrow I\text{-}R + O_2^{-}$

$Dye^{+} + {}^{-}OH \longrightarrow Dye + {}^{\bullet}OH$

$Dye / Dye^{+} + {}^{\bullet}O \longrightarrow Products$

I-R = Immobilized Resin

Dye = Direct Red 23

The generation of hydroxyl radicals and Super oxide ions (O_2^-) can explain better with the assist of proposed diagram. This proposed diagram shows the action of photocatalyst and process of generation of oxidative intermediates. Photocatalytic action shows in fig.3

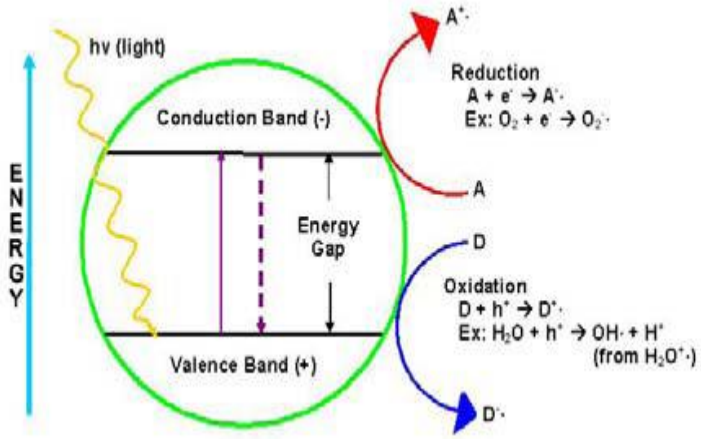
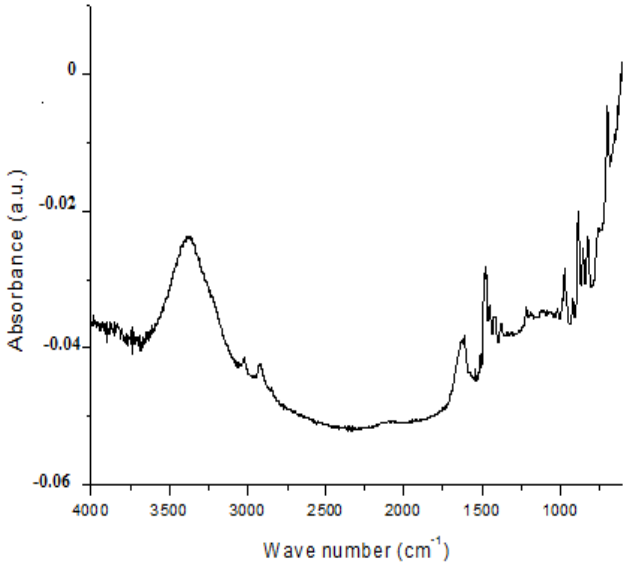


Fig.3 Photocatalytic Action

FTIR spectra of photocatalyst(MBIR-11)

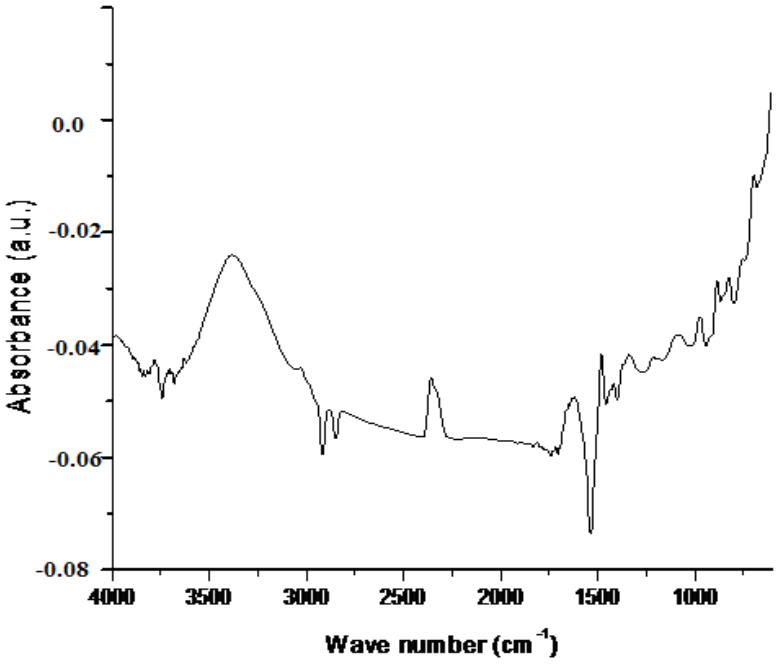
The IR spectrum of Dowex-11 resin (Pure) show peak in the region 3100-3000 cm^{-1} is attribute to =C-H stretching vibration. Another strong and sharp peak show in the region 1650-1600 cm^{-1} is denote –C=C- stretching vibration. And two peak in the region 1600-1450 cm^{-1} show aromatic –C=C- stretching vibration. Following peak is shown in fig. 4(a).

In order to examine the differences between immobilized Resin Dowex-11and pure resin Dowex 11, the FT-IR spectra were applied to the study. The FTIR spectrum of immobilized Resin Dowex-11exhibited many alterations from that of pure Resin Dowex-11. The major differences where 3650-3400cm−1 to >N-H stretching vibration and 3100-2900 cm^{-1} stretching vibration of =C-H (Methylene) group. Due to immobilization and another strong and sharp peak of N-H bending in the region 1550-1510 cm^{-1}, these peaks are shown in fig. 4(b). These same spectra of after complete experiment show that, the dye Acid red 73 is not present in pores of resin.



(a)IR spectra of Dowex-11 resin without grafting

Fig. 4 (a) FTIR spectra of Dowex-11 resin without immobilization



(b) IR spectra of Dowex-11 resin after immobilization by methylene blue dye

Fig. 4 (b) FTIR spectra of immobilized Dowex-11 resin

Factors influencing the photocatalytic decolorization

Effect of initial dye concentration

It is important to study of the photocatalytic reaction rate on the substrate concentration. It is generally noted that the decolorization efficiency increases with the increases in dye concentration to a certain level and a further increases in the dye concentration leads to degradation rate of the dye. Due to the probability of •OH radical's formation on the catalyst surface and to the probability of •OH radicals reacting with dye molecules. The presumed reason is that at high dye concentrations the generation of •OH radicals on the surface of catalyst is reduce since the active sites are covered of dye ions. The major portion of degradation occurs in the region near to the irradiated side (termed as reaction zone) where the irradiation intensity is much higher than in the other side[23]. Thus at higher dye concentration, decolorization decreases at sufficiently long distance from the light source or the reaction zone due to the retardation in the penetration of light. Hence, it is concluded that as initial concentration of the dye increases the requirement of catalyst surface needed for the decolorization also increases (Fig. 5).

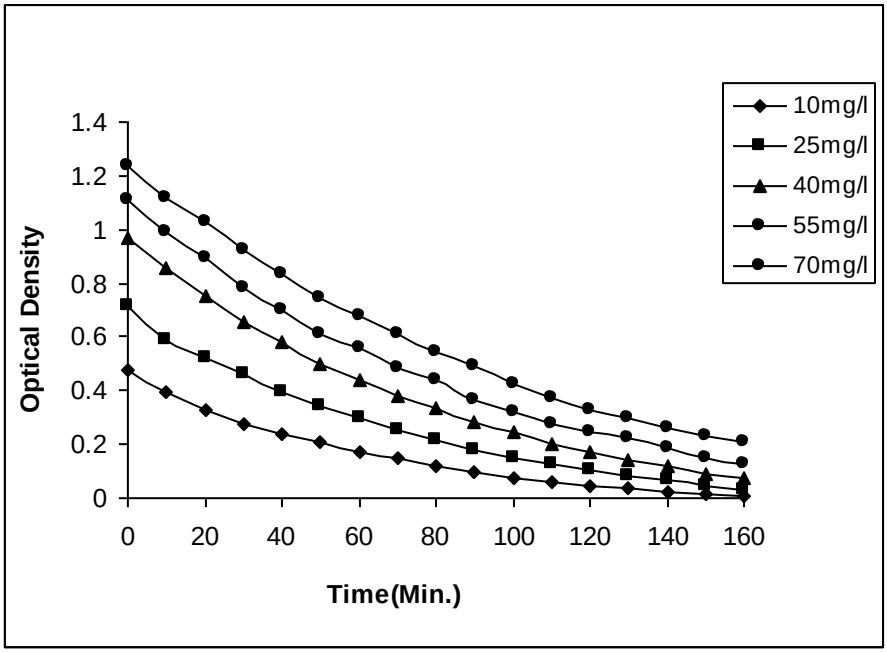


Fig.5 Effect of variation in dye concentration

Effect of pH

The results shows that rate of degradation is very low in high acidic pH range lower then pH 3.5. As well as pH increases rate of degradation also increases when pH reaches to basic range the rate of degradation increases fast, in pH range 7.5 to 9 very good rate of degradation. The increase in rate of Photocatalytic degradation may be due to more availability of $^{-}$OH ions (Fig.6).

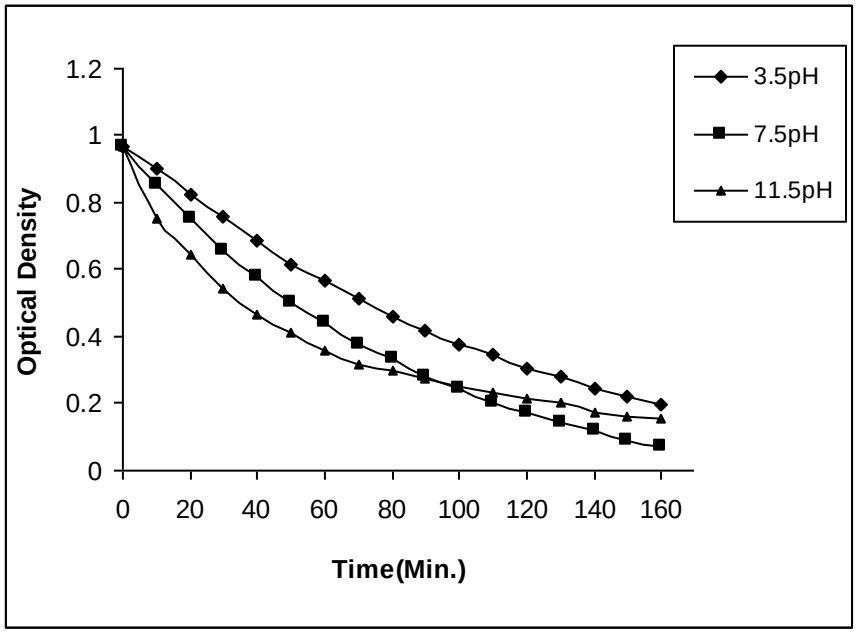


Fig.6 Effect of variation in pH on O.D.

Effect of variation in light intensity

Light intensity is most important factor which affects the rate of degradation. We find out that as light intensity increases the rate of degradation of dye molecules also increases up some extant after it no change observe in rate of degradation. This change in rate of degradation of dye molecules by variation in light intensity as light intensity increases number of photons increases to reach the catalyst surface so number exited catalyst molecules increases and resultant increase the number of holes, hydroxyl radicals and super oxide ions (O_2^-) and rate of degradation of dye molecules increase. We observe that after some extant of increase in light intensity there is no effect on rate of degradation on further increases in light intensity. This may cause that maximum number of photons which required for excitation are available in fix range irradiating light intensity after it if we further increase light intensity no any considerable change observed in rate of degradation because there is no requirement of more photons for excitation. Because all catalyst molecules become active in fix light intensity range after it we increase light intensity to any range, the rate of degradation remains unchanged (fig.7).

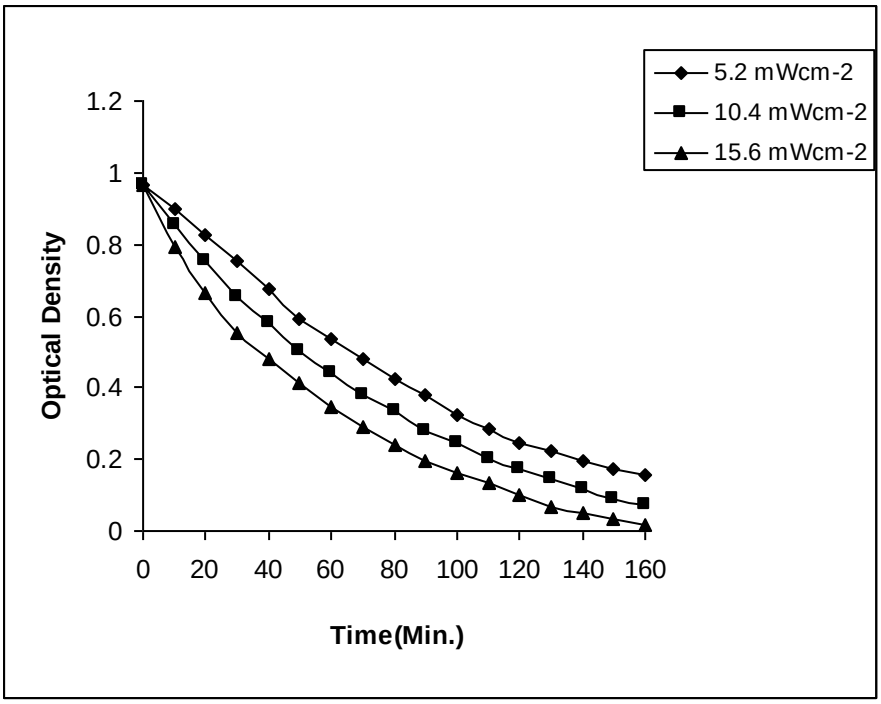


Fig.7 Effect of Variation in Light Intensity on O.D.

Effect of variation in catalyst loading

Increase in the rate of degradation with increase in amount of catalyst is due to availability of more catalyst surface area for absorption of quanta and interaction of molecules of reaction mixture with catalyst then hydroxyl radicals and super oxide ions (O_2^-) are increase. These are principle oxidizing intermediate in advance oxidation process resultant increases degradation of dye solution (Fig.8).

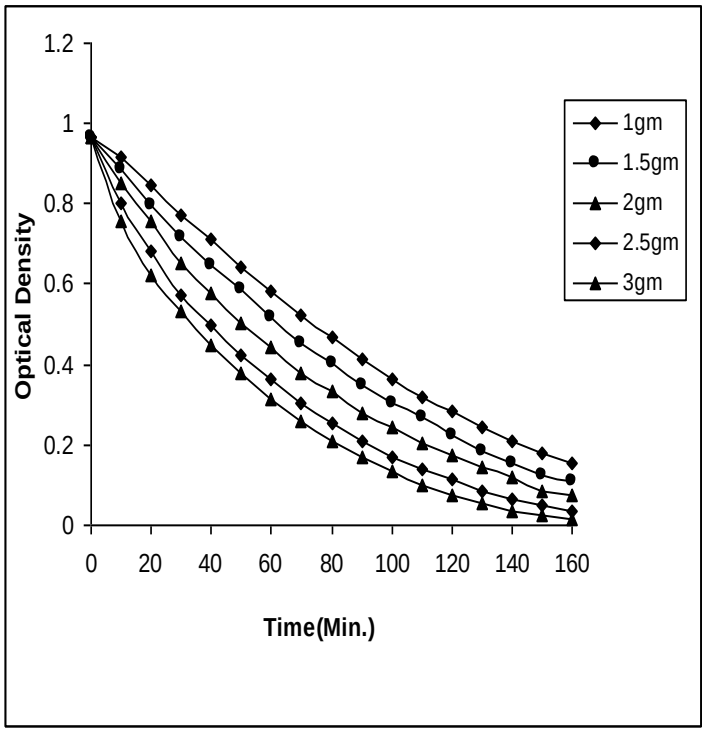


Fig.8 Effect of variation in catalyst loading on O.D.

IV. CONCLUSION

After long examination we conclude that the photo catalyst (MBIR Dowex 11) has very good potential of degradation of azo dyes into simple mineralize products. Photo degradation of the dye was very less when photolysis was carried out in absence of the catalyst and negligible in absence of light.

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Pollution and Conservation of Ganga River in Modern India

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Abstract- According to a World Bank Sponsored Study (State of Environment Report- U.P.) (In: Mallikarjun, 2003), pollution levels in the Ganga are contributing 9-12% of total disease burden in Uttar Pradesh (U.P.). The coliform bacteria levels are in excess of 2 lakh MPN as against the national water quality standard of 5000 (Mallikarjun, 2003). The report estimated total health damage on account of water pollution in up to is around 6.4 million daily (Disability Adjusted Life Year).

According to the CPCB survey report, the total municipal sewage generated in the identified 25 towns in 1985 was of the order of1340 million litres per day (mld). Apart from this sewage, 260 mld of industrial wastewater, runoff from 6 million tons of fertilizers and 9,000 tonnes of pesticides used in agriculture within the basin, large quantities of solid waste, including thousands of animal carcasses and human corpses were being released into the river every day. Out of this, works corresponding to 873 mld only (65%) were taken up under the first phase of GAP. The remaining sewage was to be taken up under the 2nd phase of GAP which is already in progress. The Action Plan primarily addressed itself to the interception and diversion for treatment of the targeted municipal sewage of 873 mld.

According to report of Water Resources Planning Commission (May, 2009), the programme GAP and NRCP has been positive. Water quality monitoring done by reputed independent institutions indicates some improvement in the water quality over pre-GAP period. The water quality analysis of samples collected at 16 stations on River Ganga during 1986 and 2008 shows improvement in Dissolved Oxygen (DO) levels at 4 locations namely up and down streams of Allahabad and Varanasi. All the 16 stations except Patna downstream and Rajmahal show reduction in Biological Oxygen Demand (BOD) values. The BOD level show marked reduction in Allahabad and Varanasi indicating improvement in the water quality over pre-GAP period. However, at 7 of these 16 sites, BOD level does not meet standard for bathing water. The situation is much better for DO for which at only one site the bathing standard is not met. On the other hand in terms of total coliform count only at one place the bathing standard is met. The count exceeds by many times the bathing standard.

I. INTRODUCTION

The Ganga is a holy and historical river of India and Bangladesh. The Ganges calls at the join of Devprayag the Bhagirathi and Alaknanda rivers. The Bhagirathi flows at the foot of Gangotri Glacier, at Gaumukh, at an elevation of 3,892 m (12,769 ft.). The Bhagirathi is considered to be the true source in Hindu culture and the Alaknanda is a longer. It has total length 2,525 km river rises in the western Himalayas in the Indian state of Uttarakhand, and flows south and east through the Gangetic Plain of North India and go to Bay of Bengal through Bangladesh, It is the longest river of India and is the second greatest river in the world by water discharge. The Ganges basin is the most heavily populated river basin in the world, with over 400 million people and a population density of about 1,000 inhabitants per square mile (390 /km^2). The Ganga was ranked among the five most polluted rivers of the world in 2007. The Ganga Action Plan, an environmental initiative to clean up the river, has been a major failure thus far, lack of good environmental planning, Indian traditions and beliefs, and lack of support from religious authorities. Kumbh Mela is a one of the most reason of pollution of Ganga river in India.

II. CAUSES OF POLLUTION IN GANGA

It provides water to drinking purpose and irrigation in agriculture about 40% of India's population in 11 states. After 27 years and Rs. 1000 crore expenditure on Ganga river, it has a critical situation. In modern times, it is known for being much polluted, 30 polluted nalas flows in Ganga river from Varanasi city within seven kilometers.

A. Human waste

The river flows through 29 cities in which cities population living above ten lakh. A large proportion damp the solid and liquid wastes in Ganga river like domestic usage (bathing, laundry and public defecation), Sewage wastes, unburnt dead bodies through in Ganga river. Patna and Varanasi cities are more responsible to water pollution in Ganga and 80 % sewage wastes are responsible to water pollution of Ganga.

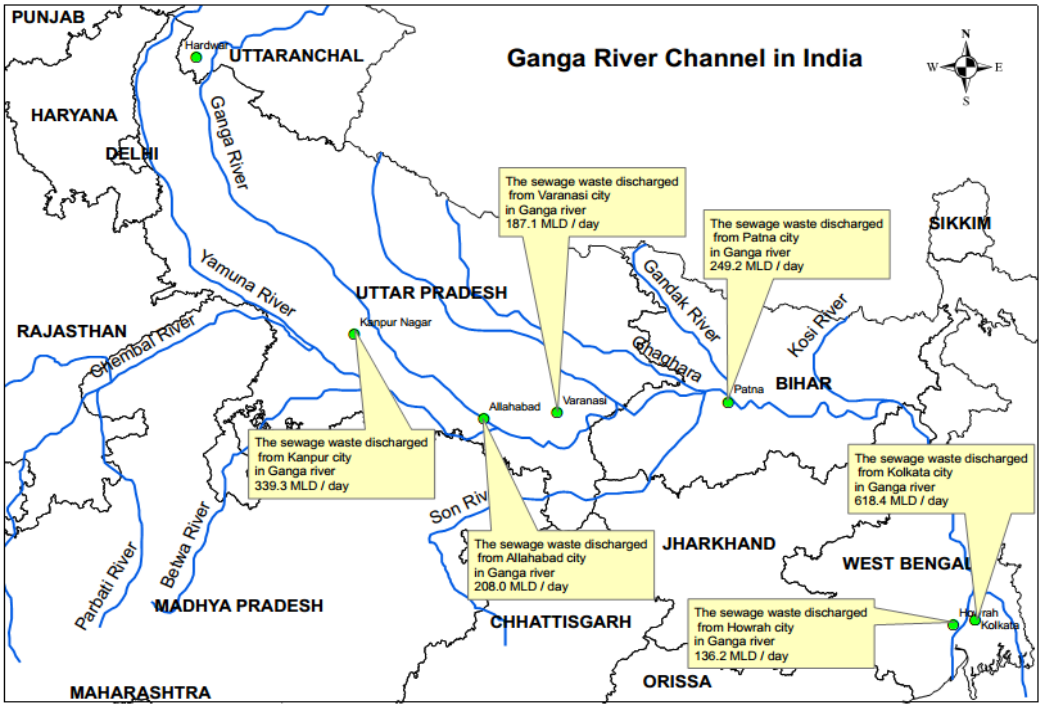

Figure 1: Showing the Ganga river channel and sewage waste discharge volume per day in 2009 (CPCB) from major cities of India

B. Industrial waste

Countless industries lies on the bank of the Ganga river from Uttrakhand to West Bengal like chemical plants, textile mills, paper mills, fertilizer plants and hospitals waste. These industries are 20 % responsible to water pollution and run off solid waste and liquid waste in the Ganga river. It is very dangers to water quality, their chemical properties and riverine life.

C. Religious factor

Festivals are very important and heartiest to every person of India. During festival seasons a lot of peoples come to Ganga Snans to cleanse themselves. After death of the people dump their asthia in Ganga river it is a tradition of India because they think that Ganga gives mukti from the human world. Khumbha Mela is a very big festival of the world and billion peoples come to Ganga Snans at Allahabad, Hardwar in India. They through some materials like food, waste or leaves in the Ganges for spiritualistic reasons.

III. IMPACT OF POLLUTION

A. Riverine life

The Ganga river pollution increased day by day and from this pollution marine life have been going to lost in near future and this polluted water disturb the ecosystem of the river. And irrigation and Hydroelectric dams give struggle to life in their life cycle.

B. Bio life

Some dams are constructed along the Ganges basin. Dams are collected a huge volume of water and this is hazard for wild life which are moving in Ganga river. The Kotli Bhel dam at Devprayag will submerge about 1200 hectors of forest. In India wildlife has been warning that the wild animals will find it difficult to cope with the changed situation.

C. Human beings

An analysis of the Ganges water in 2006 showed significant associations between water-borne/enteric disease occurrence and the use of the river for bathing, laundry, washing, eating, cleaning utensils, and brushing teeth. Exposure factors such as washing clothes, bathing and lack of sewerage, toilets at residence, children defecating outdoors, poor sanitation, low income and low education levels also showed significant associations with enteric disease outcome. Water in the Ganges has been correlated to contracting dysentery, cholera, hepatitis, as well as severe diarrhea which continue to be one of the leading causes of death of children in India.

IV. GANGA ACTION PLAN

The Ganga Action Plan or GAP was a program launched by Rajiv Gandhi in April 1986 in order to reduce the pollution load on the river. Under GAP I, pollution abatement schemes were taken up in 25 Class-I towns in three States of U.P., Bihar and West Bengal. GAP I was declared complete on 31.03.2000 with an expenditure of Rs. 452 crore.

As GAP I addressed only a part of the pollution load of Ganga, GAP II was launched in stages between 1993 and 1996, 59 towns along the main stem of river Ganga in five States of Uttarakhand, U.P., Jharkhand, Bihar and West Bengal are covered under the Plan and included the following tributaries of the Ganges, Yamuna, Gomti, Damodar and Mahananda.

V. NATIONAL RIVER GANGA BASIN AUTHORITY (NRGBA)

National River Ganga Basin Authority (NRGBA) was established by the Central Government of India, on 20 February 2009 under Section 3 (3) of the Environment Protection Act, 1986. It also declared Ganges as the "National River" of India. The chair includes the Prime Minister of India and Chief Ministers of states through which the Ganges flows.

VI. STRUGGLE TO GANGA RIVER

Noted environmentalist Swami Gyan Swarup Sanand, who has been fasting to press for his demand of conservation of Ganga River, has once again been forcefully admitted to hospital here, sources said on Saturday night. Posted on: 06 May 2012.

VII. KUMBHA MELA ISSUE OF WATER AND POLLUTION IN RIVER GANGA

According to Hindustan Newspaper, January 11, 2013, the Prime Minister has been monitoring the availability of adequate water from Tehri Dam in river Ganga at Allahabad during the Kumbh Mela. Directions have been given to control the pollution load flowing in river Yamuna during the Kumbh Mela period.

- Tehri Hydro Development Corporation (THDCIL) has agreed to release 250 cumecs water from 21st December 2012 to 20th February 2013 and 220 cumecs water from 21st February 2013 to 28th February 2013 in view of demand of water for Allahabad 'Kumbh Snans'. Instructions have also been given by PMO that Delhi Jal Board should ensure that the performance of the 72 MGD STP (Sewage Treatment Plant) at Keshavpur renovated /commissioned recently is stabilized so that it functions optimally and the effluent meets the norms. The Delhi Government has been asked to ensure that the performance of the STPs and CETPs (Common Effluent Treatment Plants) is optimized to meet the effluent quality norms.
- At Sangam, Allahabad, the Biochemical Oxygen Demand (BOD) of Yamuna and Ganga is generally less than 6 mg/ltr but the main issue is of the color of effluents discharged by the pulp and paper industries into the river Ram Ganga and Kali (both tributaries of Ganga). Monitoring of water quality in river Ram Ganga and river Kali and their tributaries is being initiated on a daily basis by the State Boards of Uttrakhand and Uttar Pradesh with the coordination of CPCB. Action will be taken against the industries for violating the norms.
- Spiritual dip in holy Ganga at Kumbh is not clean. The pollution level in the sacred river has risen since Kumbh started at the historical city of Allahabad on January 14, 2013 and the water is no fit for bathing purposes, latest evaluation by country's pollution watchdog the Central Pollution. The level of the Biochemical Oxygen Demand (BOD) levels - used to measure of the level of organic pollution in the water - had increased to 7.4 milligram per litre at the main bathing place, known as Sangam, since the Kumbh started.
- A day before the Kumbh, the pollution level was 4.4 milligram per litre slightly more than the national standard for bathing quality of water of 3 milligram per litre. "Higher the BOD level worse it is for one's skin," said a CPCB expert. High exposure to dirty water can result in skin rashness and allergies.The official reason for the sudden rise of contaminants in the river was sudden increase in flow of human waste because of increased bathing during Kumbh. Around 10 million people have already visited the Kumbh and the UP government has employed around 10,000 sweepers to keep the city clean. Off the record officials admit that their drive to check sewage from industries in Ganga upstream of Allahabad has not worked as dirty sewage was still flowing into the river.
- The Board has been asked by the environment ministry to monitor the pollution level in Ganga under its National Ganga Basin River Authority and conduct periodic check on polluting industries along the river bank. But, the dirt in the river is not a deterrent for people to take a dip at Allahabad. Hindus believe that the Ganga water has ability to clean and purify itself, a claim not scientifically proven. And, this belief has driven millions to the world biggest Hindu congregation and another 15 million are expected to visit in the 55-day long festival to end on March 10.

VIII. CPCB ACTION FOR PREVENTION OF POLLUTION OF RIVER GANGA

- Training cum Awareness programme on Saltless Preservation of Hides/ skins was organized by CPCB at Lucknow and Kanpur, which was attended by representatives from slaughter houses, tannery & allied units and officers of UPPCB. The programme was oriented towards the ongoing efforts pursuing basin-wise approach for reduction of dissolved solids in wastewater from leather processing industries in particular by invoking salt less preservation of hides/ skins.
- CPCB has initiated a Techno-Economic Feasibility for setting up of Common Recovery Plant & Common Effluent Treatment Plant for Pulp & Paper Industries identified clusters at Muzaffar nagar, Moradabad and Meerut. CPCB also made a reconnaissance survey from Gomukh to Uluberia (West Bengal) for identified the point source and its impact on River. This reconnaissance survey is conducted in association with Shri Rajinder Singh, Member, NGRBA.
- CPCB issued direction to UPPCB and Uttarakhand PCB in the matter of Prevention and Control of Pollution from agro based Pulp & Paper Sector Mills. As a result 31 industries have been issued directions in U.P., 25 digester sealed at Uttarakhand, 8 industries were directed and 4 were stop chemical pulping. CPCB conducted monitoring of 26 industrial units in the stretch of river Ganga between Kannauj to Varanasi in the month of September 2010. Of these 7 were found closed during inspection, 2 were

complying to the prescribed discharge norms, 9 were requiring minor improvements, 4 have been issued directions (under section 5 of Environment Protection Act 1986) for closure, 3 have been issued directions for corrective measures (under section 5 of Environment Protection Act 1986) and 1 have been issued Show Cause notice for closure (under section 5 of Environment Protection Act 1986).

IX. CONCLUSION

This study is dedicated to Ganga river which is flow in northern India and it is a proud to Indian population. This study is successful with analysis with previous study on the Ganga river and here positive results to Ganga clean.

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ABBREVIATION

CPCB- Central Pollution Control Board
MLD- Million Litres per day
Gap- Ganga Action Plan
DO- Dissolved Oxygen
BOD- Biological Oxygen Demand
CETP- Common Effluent Treatment Plants
STP- Sewage Treatment Plant
PMO- Prime Minister Office
UPPCB- Utter Pradesh Pollution Control Board

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A Study of Data Allocation Problem for Guilt Model Assessment in Data Leakage Detection Using Clouding Computing

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***Abstract*—** In the virtual and widely distributed network, the process of handover sensitive data from the distributor to the trusted third parties always occurs regularly in this modern world. It needs to safeguard the security and durability of service based on the demand of usersA data distributor has given sensitive data to a set of supposedly trusted agents (third parties). Some of the data are leaked and found in an unauthorized place (e.g., on the web or somebody's laptop). The distributor must assess the likelihood that the leaked data came from one or more agents, as opposed to having been independently gathered by other means. We propose data allocation strategies (across the agents) that improve the probability of identifying leakages. These methods do not rely on alterations of the released data (e.g., watermarks). In some cases, we can also inject "realistic but fake" data records to further improve our chances of detecting leakage and identifying the guilty party. This paper presents a data leakage detection system using various allocation strategies and which assess the likelihood that the leaked data came from one or more agents. For secure transactions, allowing only authorized users to access sensitive data through access control policies shall prevent data leakage by sharing information only with trusted parties and also the data should be detected from leaking by means of adding fake record`s in the data set and which improves probability of identifying leakages in the system. Then, finally it is decided to implement this mechanism on a cloud server.

Index Terms-- data leakage, data security, fake records, cloud environment.

I. INTRODUCTION

In this paper, we develop a model for finding the guilty agents. We also present algorithms for distributing objects to agents, in a way that improves our chances of identifying a leaker. Finally, we also consider the option of adding ―fake objects to the distributed set. Such objects do not correspond to real entities but appear realistic to the agents. If it turns out that an agent was given one or more fake objects that were leaked, then the distributor can be more confident that agent was guilty. We also consider optimization in which leaked data is compared with original data and accordingly the third party who leaked the data is guessed. We will also be using approximation technique to encounter guilty agents. We proposed one model that can handle all the requests from customers and there is no limit on number of customers. The model gives the data allocation strategies to improve the probability of identifying leakages. Also there is application where there is a distributor, distributing and managing the files that contain sensitive information to users when they send request.

Data leakage happens every day when confidential business information such as customer or patient data, source code or design specifications, price lists, intellectual property and trade secrets, and forecasts and budgets in spreadsheets are leaked out. When these are leaked out it leaves the company unprotected and goes outside the jurisdiction of the corporation. This uncontrolled data leakage puts business in a vulnerable position. Once this data is no longer within the domain, then the company is at serious risk.

At this point the distributor can assess the likelihood that the leaked data came from one or more agents, as opposed to having been independently gathered by other means. If the distributor sees enough evidence that an agent leaked data then they may stop doing business with him, or may initiate legal proceedings. Mainly it has one constraints and one objective. The Distributor's constraint satisfies the agent, by providing number of object they request that satisfy their conditions.

II. LTERATURE SURVEY

The guilt detection approach we present is related to the data provenance problem [3]: tracing the lineage of S objects implies essentially the detection of the guilty agents. and assume some prior knowledge on the way a data view is created out of data sources. objects and sets is more general .As far as the data allocation strategies are concerned; our work is mostly relevant to watermarking that is used as a means of establishing original ownership of distributed objects. [3] Finally, there are also lots of other works on mechanisms that allow only authorized users to access sensitive data through access control policies [9], [2]. Such approaches prevent in some sense data leakage by sharing information only with trusted parties. However, these policies are restrictive and may make it impossible to satisfy agent's requests. Maintaining the Integrity of the Specifications

III. NEED FOR DATA ALLOCATION STRATEGIES

Information systems are generally defined by the company's data and the material and software resources that allow a company to store the data and circulate this data. Information systems are essential to companies and must be protected as highest priority. Organization securities generally consists in ensuring that an organization's material and software resources are used only for their intended purposes and also it needs to provide Information privacy, or data privacy and that is the relationship between collection and dissemination of data, technology, the public expectation of privacy, and the legal and political issues surrounding them.

Privacy concerns exist wherever personally identifiable information is collected and stored in digital form or otherwise. Improper or non-existent disclosure control can be the root cause for privacy issues.

Using the data allocation strategies, the distributor intelligently give data to agents in order to improve the chances of detecting guilty agent. Fake objects are added to identify the guilty party. If it turns out an agent was given one or more fake objects that were leaked, then the distributor can be more confident that agent was guilty and when the distributor sees enough evidence that an agent leaked data then they may stop doing business with him, or may initiate legal proceedings.

In this section we describe allocation strategies that solve exactly or approximately the scalar versions of approximation equation. We resort to approximate solutions in cases where it is inefficient to solve accurately the optimization problem.

1. Explicit Data Requests

In case of explicit data request with fake not allowed, the distributor is not allowed to add fake objects to the distributed data. So Data allocation is fully defined by the agent's data request. In case of explicit data request with fake allowed, the distributor cannot remove or alter the requests R from the agent. However distributor can add the fake object. In algorithm for data allocation for explicit request, the input to this is a set of request ,……, from n agents and different conditions for requests.

Step 1: Calculate total fake records as sum of fake Records allowed.

Step 2: While total fake objects > 0

Step3: Select agent that will yield the greatest improvement in the sum objective

i.e. $i = \text{argma } x((1/|Ri|)-(1/(|Ri+1|))) \Sigma Ri \cap Rj$

Step 4: Create fake record

Step 5: Add this fake record to the agent and also to fake record set.

SteP6: Decrement fake record from total fake record set. Algorithm makes a greedy choice by selecting the agent that will yield the greatest improvement in the sum-objective.

2. Sample Data Requests

With sample data requests, each agent Ui may receive any T subset out of different object allocations. In every allocation, the distributor can permute T objects and keep the same chances of guilty agent detection. The reason is that the guilt probability depends only on which agents have received the leaked objects and not on the identity of the leaked objects. The distributor gives the data to agents such that he can easily detect the guilty agent in case of leakage of data. To improve the chances of detecting guilty agent, he injects fake objects into the distributed dataset. These fake objects are created in such a manner that, agent cannot distinguish it from original objects. One can maintain the separate dataset of fake objects or can create it on demand. In this paper we have used the dataset of fake tuples. For example, distributor sends the tuples to agents A1 and A2 as R1= {t1, t2} and R2= {t1}.

If the leaked dataset is L= {t1}, then agent A2 appears more guilty than A1. So to minimize the overlap, we insert the fake objects in to one of the agent's dataset. Practically server (Distributor) has given sensitive data to agent. In that distributor can send data with fake information. And that fake information does not affect to Original Data. Fake formation cannot identify by client. it also finds the data leakage from which agent (client)

IV. METHODOLOGY

a. Problem Definition

The distributor owns the sensitive data set T= {t1, t2… tn}. The agent Ai request the data objects from distributor. The objects in T could be of any type and size, e.g. they could be tuples in a relation, or relations in a database. The distributor gives the subset of data to each agent. After giving objects to agents, the distributor discovers that a set L of T has leaked. This means some third party has been caught in possession of L. The agent Ai receives a subset Ri of objects T determined either by implicit request or an explicit request.

Implicit Request Ri = Implicit (T, mi) : Any subset of mi records from T can be given to agent Ai

Explicit Request Ri = Explicit (T, Condi) : Agent Ai receives all T objects that satisfy Condition.

b. Data Allocation Problem

1. Fake Object

The distributor may be able to add fake objects to the distributed data in order to improve his effectiveness in detecting guilty agents. However, fake objects may impact the correctness of what agents do, so they may not always be allowable. Our use of fake objects is inspired by the use of ―trace‖ records in mailing lists. In this case, company A sells to company B a mailing list to be used once (e.g., to send advertisements). Company A adds trace records that contain addresses owned by company A. Thus, each time company B uses the purchased mailing list, A receives copies of the mailing. These records are a type of fake objects that help identify improper use of data. The distributor creates and adds fake objects to the data that he distributes to agents. Depending upon the addition of fake tuples into the agent's request, data allocation problem is divided into four cases as:

i. Explicit request with fake tuples (EF)

ii. Explicit request without fake tuples (E~F)

iii. Implicit request with fake tuples (IF)

iv. Implicit request without fake tuples (I~F).

Implicit Request Ri = Implicit (T, mi) : Any subset of mi records from T can be given to agent Ai

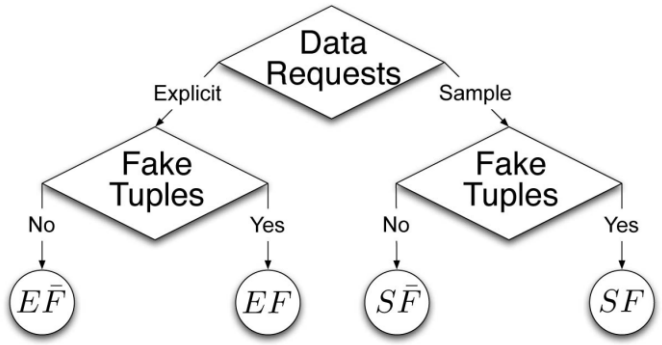

Fig.1: Leakage problem instances.

2. Optimization Problem:

The distributor's data allocation to agents has one constraint and one objective. The distributor's constraint is to satisfy agents' requests, by providing them with the number of objects they request or with all available objects that satisfy their conditions. His objective is to be able to detect an agent who leaks any portion of his data.

c. Guilt Model Assessment

Let L denote the leaked data set that may be leaked intentionally or guessed by the target user. Since agent having some of the leaked data of L, may be susceptible for leaking the data. But he may argue that he is innocent and that the L data were obtained by target through some other means. Our goal is to assess the likelihood that the leaked data came from the agents as opposed to other resources. E.g. if one of the object of L was given to only agent A1, we may suspect A1 more. So probability that agent A1 is guilty for leaking data set L is denoted as Pr{Gi| L}.

Algorithm1:

Allocation of Data Explicitly:

Input: -

i. T= {t1, t2, t3, .tn}-Distributor's Dataset

ii. R- Request of the agent

iii. Cond- Condition given by the agent

iv. m= number of tuples given to an agent m<n, selected randomly

Output: - D- Data sent to agent

1. D=Φ, T'=Φ
2. For i=1 to n do
3. If(t .fields==cond) then
4. T'=T'U{ t i}
5. For i=0 to i<m do
6. D=DU{ti}
7. T'=T'-{ ti}
8. If T'=Φ then
9. Goto step 2
10. Allocate dataset D to particular agent
11. Repeat the steps for every agent

To improve the chances of finding guilty agent we can also add the fake tuples to their data sets.

Algorithm2:
Addition of fake tuples:
Input:
i. D- Dataset of agent
ii. F- Set of fake tuples
iii. Cond- Condition given by agent
iv. b- number of fake objects to be sent.

Output:- D- Dataset with fake tuples
1. While b>0 do
2. f= select Fake Object at random from set F
3. D= DU {f}
4. F= F-{f}
5. b=b-1

Similarly, we can distribute the dataset for implicit request of agent. For implicit request the subset of distributor's dataset is selected randomly. Thus with the implicit data request we get different subsets. Hence there are different data allocations. An object allocation that satisfies requests and ignores the distributor's objective to give each agent unique subset of T of size m.

V. BASICS OF CLOUD COMPUTING

Key to the definition of cloud computing is the ―cloud itself. For our purposes,
The cloud is a large group of interconnected computers. These computers can be personal computers or network servers; they can be public or private. For example, Google hosts a cloud that consists of both smallish PCs and larger servers. Google's cloud is a private on(that is, Google owns it) that is publicly accessible (by Google's users).
This cloud of computers extends beyond a single company or enterprise. The applications and data served by the cloud are available to broad group of users, cross-enterprise and cross-platform. Access is via the Internet. Any authorized user can access these docs and apps from any computer over any Internet connection. And, to the user, the technology and infrastructure behind the cloud is invisible. It isn't apparent (and, in most cases doesn't matter)whether cloud services are based on HTTP, HTML, XML, Java script, or other specific technologies.
From Google's perspective, there are six key properties of cloud computing:

- ***Cloud Computing is user-centric***. Once you as a user are connected to the cloud, whatever is stored there -- documents, messages, images, applications, whatever – becomes yours. In addition, not only is the data yours, but you can also share it with others. In effect, any device that accesses your data in the cloud also becomes yours.
- ***Cloud computing is task-centric***. Instead of focusing on the application and what it can do, the focus is on what you need done and how the application can do it for you., Traditional applications—word processing, spreadsheets, email, and so on – are becoming less important than the documents they create.
- ***Cloud computing is powerful***. Connecting hundreds or thousands of computers together in a cloud creates a wealth of computing power impossible with a single desktop PC.
- ***Cloud computing is accessible.*** Because data is stored in the cloud, users can instantly retrieve more information from multiple repositories. You're not limited to a single source of data, as you are with a desktop PC.
- ***Cloud computing is intelligent***. With all the various data stored on the computers in the cloud, data mining and analysis are necessary to access that information in an intelligent manner.
- ***Cloud computing is programmable***. Many of the tasks necessary with cloud computing must be automated. For example, to protect the integrity of the data, information stored on a single computer in the cloud must be replicated on other computers in the cloud. If that one computer goes offline, the cloud's programming automatically redistributes that computer's data to a new computers in the cloud.

Computing in the cloud may provide additional infrastructure and flexibility.

1. Databases in cloud computing environment

In the past, a large database had to be housed onsite, typically on a large server. That limited database access to users either located in the same physical location or connected to the company's internal database and excluded, in most instances, traveling workers and users in remote offices.
Today, thanks to cloud computing technology, the underlying data of a database can be stored in the cloud, on collections of web server instead of housed in a single physical location.
This enables users both inside and outside the company to access the same data, day or night, which increases the usefulness of the data. It's a way to make data universal.

2. Lineage Tracing General Data warehouse Transformations [2]

Yingwei Cui and Jennifer Widom focus on transformation or modification of data happening automatically due to mining of data or while storing the data in the warehouse.

In a warehousing environment, the *data lineage* problem is that of tracing warehouse data items back to the original source items from which they were derived. It formally defines the lineage tracing problem in the presence of general data warehouse transformations, and they present algorithms for lineage tracing in this environment.

3. Databases in the Cloud: a Work in Progress [5]

Edward P. Holden, Jai W. Kang, Dianne P. Bills, MukhtarIlyassov focus on trial of using cloud computing in the delivery of the Database Architecture and Implementation in the cloud.

It describes a curricular initiative in cloud computing intended to keep our information technology curriculum at the forefront of technology. Currently, IT degrees offer extensive database concentrations at both the undergraduate and graduate levels. Supporting this curriculum requires extensive lab facilities where students can experiment with different aspects of database architecture, implementation, and administration. A *disruptive technology* is defined as a new, and often an initially less capable technological solution, that displaces an existing technology because it is lower in cost. Cloud computing fits this definition in that it is poised to replace the traditional model of purchased-software on locally maintained hardware platforms.

Cloud computing is the use of *virtual computing technology* that is scalable to a given application's specific requirements, without local investment in extensive infrastructure, because the computing resources are provided by various vendors as a service over the Internet.

VI. RESULT

In our scenarios we have taken a set of 500 objects and requests from every agent are accepted. There is no limit on number of agents, as we are considering here their trust values. The flow of our system is given as below:

1. Agent's Request: Either Explicit or Implicit.
2. Leaked dataset given as an input to the system.
3. The list of all agents having common tuples as that of leaked tuples is found and the corresponding guilt probabilities are calculated.
4. It shows that as the overlap with the leaked dataset minimizes the chances of finding guilty agent increases.

The basic approaches for leakage identification system in various areas and there by proposing a multi-angle approach in handling the situational issues were all studied in detailed.

When the occurrence of handover sensitive data takes place it should always watermarks each object so that it could able to trace its origins with absolute certainty, however certain data cannot admit watermarks then it is possible to assess the likelihood that an agent is responsible for a leak, based on the overlap of the data with the leaked data and also based on the probability that objects can be guessed by any other methodologies Data leakage is a silent type of threat. Your employee as an insider can intentionally or accidentally leak sensitive information. This sensitive information can be electronically distributed via e-mail, Web sites, FTP, instant messaging, spread sheets, databases, and any other electronic means available – all without your knowledge.

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Enhancement of Merkle-Hellman Knapsack Cryptosystem by use of Discrete Logarithmics

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Abstract- The Merkle-Hellman invented in 1978 is based on the superincreasing subset problem. Ralph Merkle and Martin Hellman used the subset problem to create a cryptosystem to encrypt data. A super-increasing knapsack vector s is created and the super-increasing property is hidden by creating a second vector M by modular multiplication and permutation. The vector M is the public key of the cryptosystem and s is used to decrypt the message.

This paper demonstrates how to strengthen the encrypted message being sent by use of discrete logarithmics so that only the intended recipient of the message is able to decipher the message.

Index Terms- Security; cryptography; cryptosystem; discrete logarithm; euler's totient function; knapsack problem; superincreasing vector.

I. INTRODUCTION

The knapsack problem is an NP complete problem in combinatorial optimization. The knapsack problem selects the most useful items from a number of items given that the knapsack or the rucksack has a certain capacity.

Knapsack problems are widely used to model solutions industrial problems such as public-key cryptography.

The 0-1 knapsack problem states that if there is a knapsack with a given capacity and a certain number of items that need to be put in the knapsack. Each item has a value and a weight associated with it. The knapsack problem selects the items that can be put in the knapsack so that the value of all the items is maximized and the weight does not increase the total capacity of the knapsack.

This can be denoted as –

Maximize $\sum^{n}_{i=0} p_i x_i$ (1)

Subject to $\sum^{n}_{i=0} w_i x_i \leq W$ (2)

$$x_i = \begin{cases} 1, \text{ if the item is included in the knapsack} \\ 0, \text{ if the item is not included in the knapsack} \end{cases} \quad (3)$$

where,

'p' is the value associated with each item i
'w' is the weight associated with each item i
'W' is the maximum capacity of the knapsack
'n' is the number of items

The subset sum problem is a special case of the knapsack problem [5]. This problem finds a group of integers from a list vector V, where V = (v1, v2, v3, … vn), the subset of elements in the vector V which have a given sum S. It also determines if a vector X = (x1, x2, x3… xn) exists where x_i element of {0,1} so that V*X = S [5].

Ralph Merkle and Martin Hellman used the subset problem to create a cryptosystem to encrypt data. A super-increasing knapsack vector s is created and the super-increasing property is hidden by creating a second vector M by modular multiplication and permutation. The vector M is the public key of the cryptosystem and s is used to decrypt the message [2].

II. EXISTING SYSTEM

The existing system uses only Merkle-Hellman Knapsack Cryptosystem to encrypt the messages. Though the excrypted message is difficult to break, it can be found out if we get to know the pattern of the message being sent.

III. PROPOSED SYSTEM

Here we propose a cryptosystem which makes use of the Merkle-Hellman structure and also makes use of discrete logarithmics to encyrpt the message. Now the hacker has to know the super-increasing sequence, the private key and public key in order to decrypt the message which strengthens the message to be sent.

IV. ENCRYPTING MESSAGES

The proposed cryptosystem performs encryption in two steps.

First the plain text is broken down into each character and the characters are converted to their binary equivalent. These characters are then encrypted through the Merkle-Hellman encryption scheme whose main idea is to create a subset problem which can be solved easily and then to hide the super-increasing nature by modular multiplication and permutation.

Secondly, these encrypted characters are further encrypted though the use of discrete logarithmics based on RSA concepts.
We choose two prime numbers and calculate 'n' as the product of these two prime numbers, euler's totient function $\emptyset$ (n) as the

result of (p-1) * (q-1). We choose another number 'e' which is relatively prime to the other two prime numbers which we had earlier chosen and gcd (e, $\emptyset$(n)) =1.

Using these details we find out the value of 'd' such that d $\equiv e^{-1}$ (mod $\emptyset$ (n))

This (e, n) acts as the private key and the pair (d, n) acts as the public key.

Thus the formula to encrypt the message using discrete logarithmics is $C = M^e \bmod n$ (4).

A. Mathematical Explanation

The first step is to choose an initial vector IV and a key of length 7 bits. These are used to perform the first encryption process.

The second step is to convert all the characters of the message into binary. The binary sequence is represented by the variable b.

The third step is to perform CBC to get the temporary cipher text.

The fourth step is to choose a superincreasing sequence of positive integers. A superincreasing sequence is one where every number is greater than the sum of all preceding numbers.

s = (s1, s2, s3, …. sn) (4)

The fourth step is to choose two numbers – an integer (a), which is greater than the sum of all numbers in the sequence's' and its co-prime (r).

The sequence's' and the numbers 'a' and 'r' collectively form the private key of the cryptosystem.

All the elements – s1, s2, s3, …. sn of the sequence 's' are multiplied with the number 'r' and the modulus of the multiple is taken by dividing with the number a.

Therefore,

$p_i = r * s_i \bmod (a)$. (5)

All elements p1, p2, p3, …. pn of the sequence p are multiplied with the corresponding elements of the binary sequence b. The numbers are then added to create the encrypted message Mi.

The sequence M = (M1, M2, M3… Mn) forms the cipher text of the cryptosystem.

B. Example

– Encrypting the string "get"

Step 1. Finding character equivalent

The first step is to convert all the characters in the string into their binary equivalent –

g = 1 1 0 0 1 1 1
e = 1 1 0 0 1 0 1
t = 1 1 1 0 1 0 0

Step 2: For 2nd character

The next step is to choose a super-increasing sequence. In this case the sequence is –

s = (3, 5, 15, 25, 54, 110, 225)

The binary sequence is c = (c1, c2, c3 ….. cn)
The two numbers chosen are – 439 and 10.
a = 439, r = 10

The sequence p = p1, p2,… pn
Where $p_i = r * s_i \bmod a$

The message is encrypted by multiplying all the elements of sequence p with the corresponding elements of sequence **c** and adding the resulting sum.

Therefore, the encrypted message
$M = \sum^{n}_{i=0} p_i * b_i$ (6)

p1 = 3 * 10 mod 439 = 30
p2 = 5 * 10 mod 439 = 50
p3 = 15 * 10 mod 439 = 150
p4 = 25 * 10 mod 439 = 250
p5 = 54 * 10 mod 439 = 101
p6 = 110 * 10 mod 439 = 222
p7 = 225 * 10 mod 439 = 55

Encrypting the character g –
p = (30, 50, 150, 250, 101, 222, 55)
The binary equivalent of 'g' is 1 1 0 0 1 1 1
M_g = 30 + 50 + 101 + 222 + 55 = 458

This is encrypted using logarithmic functions. The logarithmic functions are based on the RSA algorithmic concepts.

Here the prime numbers chosen are 53 and 31. The value of 'e' is chosen as 7. Therefore'd' becomes 223. The value of n is the product of the two relatively prime integers. Thus here n = 1643.

Thus the message is encrypted according to the formula, $C = M^e \bmod n$ (7)
Where, (e, n) also act as private keys

Thus the encrypted code for the first character becomes $C_g = 458^7 \bmod 1643 = 344$

Step 3: For 2nd character
The second character is 'e'.
Its binary equivalent is 1 1 0 0 1 0 1

Encrypting the character e –
p = (30, 50, 150, 250, 101, 222, 55)

The binary equivalent of 'e' is 1 1 0 0 1 0 1
M_e = 30 + 50 + 101 + 55 = 236

This is encrypted using logarithmic functions.
Thus the encrypted code for the second character becomes $C_e = 236^7 \bmod 1643 = 937$

Step 4: For 3rd character
The second character is't'.

Its binary equivalent is 1 1 1 0 1 0 0

Encrypting the character e –
p = (30, 50, 150, 250, 101, 222, 55)

The binary equivalent of 'e' is 1 1 1 0 1 0 0
M_t = 30 + 50 + 150 + 101 = 331

This is encrypted using logarithmic functions.
Thus the encrypted code for the second character becomes $C_t = 331^7 \bmod 1643 = 1499$

These codes are combined together and sent to the receiver. Thus the message to be tranmitted to the receiver is 034409371499.

V. DECRYPTING MESSAGES

During the decryption process, the blocks of encrypted code are separated. On these blocks first logarithmic decryption is performed using the RSA concepts.

The formula used is $M = C^d \bmod n$ (8)

The values of the prime numbers, e, d and n are same as that used during encryption.

The output of it is decrypted using Merkle-Hellman Knapsack cryptosystem decryption process.

A. Mathematical Explanation

To decrypt the message M, the recipient of the message would have to find the bitstream which satisfies the Equation [1]–

$M = \sum_{i=0}^{n} p_i * b_i$ (9)

To solve the equation (8), the user would need the private key (s, a, r).

The first step is to calculate the modular multiplicative inverse of 'r' in r mod a [4].

This is calculated using the Extended Euclidean algorithm. This is denoted by r-1.

The second step is to multiply each element of the encrypted message (M) with r-1 mod a.

The largest number in the set which is smaller than the resulting number is subtracted from the number.

This continues until the number is reduced to zero[1].

This temporary code is then fed into the CBC cryptosystem where each temporary block is passed through the decryption algorithm. The result is XORed with the preceeding ciphertext block to produce the plaintext block. [7]

The decryption technique is just the processes taking place in reverse order.

B. *Example*

Decrypting the message: C=034409371499

Step 1.

Separate the cipher text into groups of 4 digits from the starting position.
Therefore, C1 = 0344
C2 = 0937
C3 = 1499

Step 2: Decrypting 1st encrypted code.

Perform discrete logatithmic decryption schemes on C1=0344 using the concepts of RSA.

Thus, the prime numbers chosen were 53 and 31. The value of 'e' was chosen as 7. Therefore'd' becomes 223. The value of n is the product of the two relatively prime integers.Thus here n = 1643.

Thus the message is encrypted according to the formula, $M = C^d \bmod n$ (10)

Where, (d, n) also act as public keys

Thus, $M1 = 344^{223} \bmod 1643 = 458$

The modular inverse of 10 in 10 mod 439 is calculated using the extended Euclidean algorithms and was found out to be 44.

The encrypted message M1 is 458 and
s = 3, 5, 15, 25, 54,110, 225.

Again, 458 * 44 mod 439 = 397

The largest number in the sequence s, which is smaller than 397 is 225.

397 – 225 = 172
172 – 110 = 62
62 – 54 = 8
8 – 5 = 3
3 – 3 = 0

Thus, the binary sequence becomes 1 1 0 0 1 1 1.
The character equivalent to this binary sequence is 'g'.

Step 3: Decrypting 2nd encrypted code.

Perform discrete logatithmic decryption schemes on C2=0937 using the concepts of RSA.

Thus, $M2 = 937^{223} \bmod 1643 = 236$

The encrypted message M2 is 236 and
s = 3, 5, 15, 25, 54,110, 225.
Again, 236 * 44 mod 439 = 287

The largest number in the sequence s, which is smaller than 287 is 225.
287 – 225 = 62
62 – 54 = 8
8 – 5 = 3
3 – 3 = 0

Thus, the binary sequence becomes 1 1 0 0 1 0 1.
The character equivalent to this binary sequence is ‘e’.

Step 4: Decrypting 3rd encrypted code.
Perform discrete logatithmic decryption schemes on C3=1499 using the concepts of RSA.

Thus, M3 = 1499^{223} mod 1643 = 331

The encrypted message M3 is 331and
s = 3, 5, 15, 25, 54,110, 225.
Again, 331 * 44 mod 439 = 77

The largest number in the sequence s, which is smaller than 77 is 54.

77 – 54 = 23
23 – 15 = 8
8 – 5 = 3
3 – 3 = 0

Thus, the binary sequence becomes 1 1 1 0 1 0 0.
The character equivalent to this binary sequence is‘t’.
These decrypted characters are combined together and the resultant output is ‘get’.

Thus the original message “get” is got back.

VI. EXPERIMENTAL RESULTS

To demonstrate the proposed system we use Java platform and BlueJ version 1.3.5 as the software. The number of lines used in coding for developing the cryptosystem is 354.

We take any string as inputs and we get the cipher text as output.

This encryption process is demonstrated in the figure 3 given below.Here we take the plain text”get” as input. We get the cipher text as “034409371499”. Here the first four digits represent the encrypted first character; the next four digits represent the encrypted second character and henceforth.

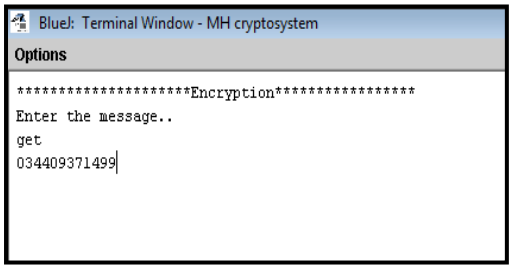

Figure 3: Output of Encryption Process

The decryption process is demonstrated in the figure 4 given below. The cipher text received is “034409371499”. The digits are extracted four at a time and the numbers are decrypted to get the original message back.

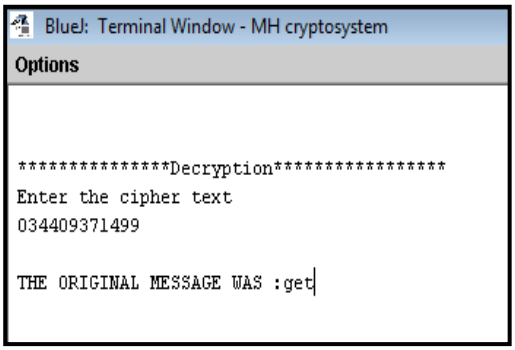

Figure 4: Output of Decryption Process

The original message and the decrypted message matched.Thus the cryptosystem works successfully.

VII. CONCLUSION

This paper explained how to encrypt and decrypt data by enhacning the working of Merkle-Hellman Knapsack cryptosystem through the use of discrete logarithmic concepts. The whole cryptosystem was demonstrated by encrypted a string “get” and then decrypting it. The decrypted string matched the original string.

VIII. FUTURE SCOPE

The future scopes include use of shifting and hashing algorithms. The encryption algorithm can be strengthened by combining it with other encryption schemes also.

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Short Term Optimal Generation Scheduling of Multi-Chain Hydrothermal System Using Constriction Factor Based Particle Swarm Optimization Technique (CFPSO)

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Abstract- In this paper, the particle swarm optimization technique with constriction factor is proposed to solve short term multi chain hydrothermal scheduling problem with non smooth fuel cost objective functions. The performance of the proposed algorithm is demonstrated on hydrothermal test system comprising of three thermal units and four hydro power plants. A wide range of thermal and hydraulic constraints such as power balance constraint, minimum and maximum limits of hydro and thermal units, water discharge rate limits, reservoir volume limits, initial and end reservoir storage volume constraint and water dynamic balance constraint are taken into consideration. The simulation results of the proposed technique are compared with those obtained from other methods such as, simulated annealing (SA) and evolutionary programming (EP) to reveal the validity and verify the feasibility of the proposed method. The test results show that the proposed algorithm achieves qualitative solution with less computational time when compared to the other methods.

Index Terms- Hydrothermal Generation Scheduling, Valve Point Loading Effect, Particle Swarm Optimization (PSO), Constriction Factor.

I. INTRODUCTION

The hydrothermal generation scheduling plays an important role in the operation and planning of a power system. Since the operating cost of thermal power plant is very high compared to the operating cost of hydro power plant, the integrated operation of the hydro and thermal plants in the same grid has become the more economical [1]. The main objective of the short term hydro thermal scheduling problem is to determine the optimal generation schedule of the thermal and hydro units to minimize the total production cost over the scheduling time horizon (typically one day or one week) subjected to a variety of thermal and hydraulic constraints. The hydrothermal generation scheduling is mainly concerned with both hydro unit scheduling and thermal unit dispatching. The hydrothermal generation scheduling problem is more difficult than the scheduling of thermal power systems. Since there is no fuel cost associated with the hydro power generation, the problem of minimizing the total production cost of hydrothermal scheduling problem is achieved by minimizing the fuel cost of thermal power plants under the constraints of water available for the hydro power generation in a given period of time [2]. In short term hydrothermal scheduling problem, the reservoir levels at the start and the end of the optimization period and the hydraulic inflows are assumed known. In addition, the generating unit limits and the load demand over the scheduling interval are known. Several mathematical optimization techniques have been used to solve short term hydrothermal scheduling problems [3]. In the past, hydrothermal scheduling problem is solved using classical mathematical optimization methods such as dynamic programming method [4-5], lagrangian relaxation method [6-7], mixed integer programming [8], interior point method [9], gradient search method and Newton raphson method [2]. In these conventional methods simplifying assumptions are made in order to make the optimization problem more tractable. Thus, most of conventional optimization techniques are unable to produce optimal or near optimal solution of this kind of problems. The computational time of these methods increases with the increase of the dimensionality of the problem. The most common optimization techniques based upon artificial intelligence concepts such as evolutionary programming [10-11], simulated annealing [12-14], differential evolution [15], artificial neural network [16-18], genetic algorithm [19 -22] and particle

swarm optimization [23-27] have been given attention by many researchers due to their ability to find an almost global or near global optimal solution for short term hydrothermal scheduling problems with operating constraints. Major problem associated with these techniques is that appropriate control parameters are required. Sometimes these techniques take large computational time due to improper selection of the control parameters. The PSO is a population based optimization technique first proposed by Kennedy and Eberhart in 1995. In PSO, each particle is a candidate solution to the problem. Each particle in PSO makes its decision based on its own experience together with other particles experiences. Particles approach to the optimum solution through its present velocity, previous experience and the best experience of its neighbours [28]. Compared to other evolutionary computation techniques, PSO can solve the problems quickly with high quality solution and stable convergence characteristic, whereas it is easily implemented.

II. PROBLEM FORMULATION

The main objective of short term hydro thermal scheduling problem is to minimize the total fuel cost of thermal power plants over the optimization period while satisfying all thermal and hydraulic constraints. The objective function to be minimized can be represented as follows:

$$FT = \sum_{t=1}^{T}\sum_{i=1}^{N} ntFit(Pgit) \quad (1)$$

In general, the fuel cost function of thermal generating unit i at time interval t can be expressed as a quadratic function of real power generation as follows:

$$Fit(Pgit)=aiP^{2}git+biPgit+ci \quad (2)$$

Where P_{git} is the real output power of thermal generating unit i at time interval t in (MW), Fit (Pgit) is the operating fuel cost of thermal unit i in ($/hr), FT is the total fuel cost of the system in ($), T is the total number of time intervals for the scheduling horizon, nt is the numbers of hours in scheduling time interval t, N is the total number of thermal generating units, a_i, b_i and ci are the fuel cost coefficients of thermal generating unit i.

The generating units with multi-valve steam turbines exhibit a greater variation in the fuel cost function. The valve opening process of multi-valve steam turbines result in ripples in fuel cost curve [29]. Due to the valve point effects, the real input-output characteristic contains higher order non linearity and discontinuity which result in non smooth and non convex fuel cost functions. The valve point effects are taken into consideration by adding rectified sinusoidal cost function to the original fuel cost function described in (2). The fuel cost function of thermal power plant with valve point loading effect can be expressed as:

$$Fit^{v}(Pgit)=aiP^{2}git+biPgit+ci+\left|ei\times\sin(fi\times(Pgit^{min}-Pgit))\right| \quad (3)$$

Where Fit^{v} (Pgit) is the fuel cost function of thermal unit i including the valve point loading effect and fi, ei are the fuel cost coefficients of generating unit i with valve point loading effect.

The minimization of the objective function of short term hydrothermal scheduling problem is subject to a number of thermal and hydraulic constraints. These constraints include the following:

1) Real Power Balance Constraint:

For power balance, an equality constraint should be satisfied. The total active power generation from the hydro and thermal plants must equal to the total load demand plus transmission line losses at each time interval over the scheduling period.

$$\sum_{i=1}^{N} Pgit+\sum_{j=1}^{M} Phjt=PDt+PLt \quad (4)$$

Where, PDt is the total load demand during the time interval t in (MW), Phjt is the power generation of hydro unit j at time interval t in (MW), Pgit is the power generation of thermal generating unit i at time interval t in (MW), M is the number of hydro units and PLt represents the total transmission line losses during the time interval t in (MW). For simplicity, the transmission power loss is neglected in this paper.

2) Thermal Generator Limit Constraint:

The output power generation of thermal power plant must lie in between its minimum and maximum limits. The inequality constraint for each thermal generator can be expressed as:

$$Pgi^{min} \le Pgit \le Pgi^{max} \quad (5)$$

Where Pgi^{min} and Pgi^{max} are the minimum and maximum power outputs of thermal generating unit i in (MW), respectively. The maximum output power of thermal

generator i is limited by thermal consideration and minimum power generation is limited by the flame instability of a boiler.

3) Hydro Generator Limit Constraint:

The output power generation hydro power plant must lie in between its minimum and maximum bounds. The inequality constraint for each hydro generator can be defined as:

$$Phj^{min} \le Phjt \le Phj^{max} \quad (6)$$

Where Phj^{min} is the minimum power generation of hydro generating unit j in (MW) and Phj^{max} is the maximum power generation of hydro generating unit j in (MW).

4) Reservoir Storage Volume Constraint:

The operating volume of reservoir storage limit must lie in between the minimum and maximum capacity limits.

$$Vhj^{min} \le Vhjt \le Vhj^{max} \quad (7)$$

Where Vhj^{min} is the minimum storage volume of reservoir j and Vhj^{max} is the maximum storage volumes of reservoir j.

5) Water Discharge Rate Limit Constraint:

The water Discharge rate of hydro turbine must lie in between its minimum and maximum operating limits.

$$qhj^{min} \le qhjt \le qhj^{max} \quad (8)$$

Where qhj^{min} and qhj^{max} are the minimum and maximum water discharge rate of reservoir j, respectively

6) Initial and End Reservoir Storage Volume Constraint:

This constraint implies that the desired volume of water to be discharged by each reservoir over the scheduling period should be in limit.

$$Vhjt^{0} = Vhj^{begin} = Vhj^{max} \quad (9)$$

$$Vhjt^{T} = Vhj^{end} \quad (10)$$

Where Vhj^{begin} and Vhj^{end} are the initial and final storage volumes of reservoir j, respectively

7) Water Dynamic Balance Constraint:

The water continuity equation relates the previous interval water storage in reservoirs with the current storage including delay in water transportation between successive reservoirs. The water continuity equation can be represented as:

$$Vhjt = Vhj,t-1 + Ihjt - qhjt - shjt + \sum_{u=1}^{R_{uj}} (qu,t-\tau_{uj} + Su,t-\tau_{uj}) \quad (11)$$

Where I_{hjt} is water inflow rate of reservoir j at time interval t, S_{hjt} is the spillage from reservoir j at time interval t, τuj is the water transport delay from reservoir u to reservoir j and Ruj is the number of upstream hydro reservoirs directly above the reservoir j.

8) Hydro Plant Power Generation Characteristic:

The hydro power generation is a function of the net hydraulic head, water discharge rate and the reservoir volume. This can be expressed as follows:

$$Phjt = f(qhjt, vhjt) \text{ and } vhjk = f(hjk) \quad (12)$$

The hydro power generation can be expressed in terms of reservoir volume instead of using the reservoir effective head, and the frequently used functional is:

$$Phjt = c1jV^{2}hjt + c2jq^{2}hjt + c3jVhjtqhjt + c4jVhjt + c5jqhjt + c6j \quad (13)$$

Where c_{1j}, c_{2j}, c_{3j}, c_{4j}, c_{5j} and c_{6j} are the Power generation coefficients of hydro generating unit j

III. PARTICLE SWARM OPTIMIZATION WITH CONSTRICTION FACTOR

A. Overview of Particle Swarm Optimization

Particle swarm optimization (PSO) is a population based stochastic optimization technique, inspired by social behavior of bird flocking or fish schooling. It is one of the most modern heuristic algorithms, which can be used to solve non linear and non continuous optimization problems. PSO shares many similarities with evolutionary computation techniques such as genetic algorithm (GA). The system is initialized with a population of random solutions and searches for optima by updating generations. However, unlike GA, PSO has no evolution operators such as mutation and crossover. The PSO algorithm searches in parallel using a group of random particles. Each particle in a swarm corresponds to a candidate solution to the problem. Particles in a swarm approach to the optimum solution through its present velocity, its previous experience and the experience of its neighbors. In every generation, each particle in a swarm is updated by two best values. The first one is the best solution (best fitness) it has

achieved so far. This value is called Pbest. Another best value that is tracked by the particle swarm optimizer is the best value, obtained so far by any particle in the population. This best value is a global best and called gbest. Each particle moves its position in the search space and updates its velocity according to its own flying experience and neighbor's flying experience. After finding the two best values, the particle update its velocity according to equation (14).

$$V_i^{k+1}=\omega\times V_i^k+c_1\times r_1\times(Pbest_i^k-X_i^k)+c_2\times r_2\times(gbest^k-X_i^k) \quad (14)$$

Where V_i^k is the velocity of particle i at iteration k, X_i^k is the position of particle i at iteration k, ω is the inertia weight factor, $c_1$and c_2 are the acceleration coefficients, $r_1$and r_2 are positive random numbers between 0 and 1, $Pbest_i^k$ is the best position of particle i at iteration k and $gbest^k$ is the best position of the group at iteration k.

In the velocity updating process, the acceleration constants c_1, c_2 and the inertia weight factor are predefined and the random numbers $r_1$and r_2 are uniformly distributed in the range of [0,1]. Suitable selection of inertia weight in equation (14) provides a balance between local and global searches, thus requiring less iteration on average to find a sufficiently optimal solution. A low value of inertia weight implies a local search, while a high value leads to global search. As originally developed, the inertia weight factor often is decreased linearly from about 0.9 to 0.4 during a run. It was proposed in [30]. In general, the inertia weight ω is set according to the following equation:

$$\omega=\omega_{max}-\frac{\omega_{max}-\omega_{min}}{Iter_{max}}\times Iter \quad (15)$$

Where ω_{min} and ω_{max} are the minimum and maximum value of inertia weight factor, $Iter_{max}$ corresponds to the maximum iteration number and Iter is the current iteration number.

The current position (searching point in the solution space) can be modified by using the following equation:

$$X_i^{k+1}=X_i^k+V_i^{k+1} \quad (16)$$

The velocity of particle i at iteration k must lie in the range:

$$V_{i\,min}\le V_i^k\le V_{i\,max} \quad (17)$$

The parameter V_{max} determines the resolution or fitness, with which regions are to be searched between the present position and the target position. If V_{max} is too high, the PSO facilitates a global search and particles may fly past good solutions. Conversely, if V_{max} is too small, the PSO facilitates a local search and particles may not explore sufficiently beyond locally good solutions. In many experiences with PSO, V_{max} was often set at 10-20% of the dynamic range on each dimension.

The constants c_1 and c_2 in equation (14) pull each particle towards Pbest and gbest positions. Thus, adjustment of these constants changes the amount of tension in the system. Low values allow particles to roam far from target regions, while high values result in abrupt movement toward target regions. Figure 1 shows the search mechanism of particle swarm optimization technique using the modified velocity, best position of particle i and best position of the group.

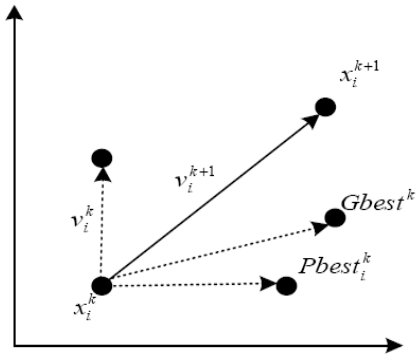

Fig.1. Updating the position mechanism of PSO technique

B. Constriction Factor Approach

After the original particle swarm proposed by Kennedy and Eberhart, a lot of improved particle swarms were introduced. The particle swarm with constriction factor is very typical. Recent work done by Clerc [31] indicates that the use of a constriction factor may be necessary to insure convergence of the particle swarm optimization algorithm. In order to insure convergence of the particle swarm optimization algorithm, the velocity of the constriction factor approach can be represented as follows:

$$V_i^{k+1}=K\times[\omega\times V_i^k+c_1\times r_1\times(Pbest_i^k-X_i^k)+c_2\times r_2\times(gbest^k-X_i^k)] \quad (18)$$

Where K is the constriction factor and given by:

$$K=\frac{2}{\left|2-\varphi-\sqrt{\varphi^2-4\varphi}\right|} \quad (19)$$

Where: $\varphi=c_1+c_2$, $\varphi>4$

The convergence characteristic of the particle swarm optimization technique can be controlled by φ. In the

constriction factor approach, φ must be greater than 4.0 to guarantee the stability of the PSO algorithm. However, as φ increases the constriction factor decreases and diversification is reduced, yielding slower response. Typically, when the constriction factor is used, φ is set to 4.1 (i.e. $c_1 = c_2 = 2.05$) and the constant multiplier k is 0.729. The constriction factor approach can generate higher quality solutions than the basic PSO technique.

IV. ALGORITHM FOR SHORT TERM HYDROTHERMAL SCHEDULING PROBLEM USING CFPSO TECHNIQUE

The sequential steps of solving short term hydro thermal scheduling problem by using constriction factor based PSO algorithm are explained as follows:

Step 1: Read the system input data, namely fuel cost curve coefficients, power generation limits of hydro and thermal units, number of thermal units, number of hydro units, power demands, power generation coefficients of hydro power plants, upper and lower limits of reservoir volumes, discharge rate limits and water inflow rate through the hydro turbines.

Step 2: Initialize a population of particles with random positions according to the minimum and maximum operating limits of each unit (upper and lower bounds of power output of thermal generating units and upper and lower bounds of water discharge rate of hydro units). These initial particles must be feasible candidate solutions that satisfy the practical operation constraints of all thermal and hydro units.

Step 3: Initialize the velocity of particles in the range between $[-V_i^{max}, +V_i^{max}]$.

Step 4: Calculate the reservoir storage of j^{th} hydro power plant in the dependent interval by using the water balance continuity equation defined in (11).

Step5: Check the inequality constraint of reservoir storage volume by the following equation:

$$V_{hjt} = \begin{cases} V_{hjt} & \text{if } V_{hj}^{\min} \le V_{hjt} \le V_{hj}^{\max} \\ V_{hj}^{\min} & \text{if } V_{hjt} \le V_{hj}^{\min} \\ V_{hj}^{\max} & \text{if } V_{hjt} \ge V_{hj}^{\max} \end{cases} \quad (20)$$

Step 6: Calculate the hydro power generation from the equation given in (13).

Step 7: Calculate the thermal demand by subtracting the generation of hydro units from the total load demand. The thermal demand (total load – hydro generation) must be covered by the thermal units. The thermal generations are calculated from the power balance equation given in (4).

Step 8: Check the inequality constraint of thermal power generated according to the following equation:

$$P_{git} = \begin{cases} P_{git} & \text{if } P_{gi}^{\min} \le P_{git} \le P_{gi}^{\max} \\ P_{gi}^{\min} & \text{if } P_{git} \le P_{gi}^{\min} \\ P_{gi}^{\max} & \text{if } P_{git} \ge P_{gi}^{\max} \end{cases} \quad (21)$$

Step 9: Evaluate the fitness value of each particle in the population using the objective function given in equation (1).

Step 10: If the evaluation value of each particle is better than the previous Pbest, then set Pbest equal to the current value.

Step 11: Select the particle with the best fitness value of all the particles in the population as the gbest.

Step 12: Update the velocity of each particle according to equation (18).

Step 13: Check the velocity of each particle according to the following equation:

$$V_i^{k+1} = \begin{cases} V_i^{k+1} & \text{if } V_i^{\min} \le V_i^{k+1} \le V_i^{\max} \\ V_i^{\min} & \text{if } V_i^{k+1} \le V_i^{\min} \\ V_i^{\max} & \text{if } V_i^{k+1} \ge V_i^{\max} \end{cases} \quad (22)$$

Step 14: The position of each particle is modified according to equation (16).

Step 15: Check the inequality constraints of the modified position.

Step 16: If the stopping criterion is reached (i.e. usually maximum number of iterations) go to step 17, otherwise go to step 4.

Step 17: The particle that generates the latest gbest is the optimal generation power of each unit with minimum total fuel cost of the thermal power plants.

Step 18: Print the outputs of hydrothermal scheduling and stop.

V. CASE STUDY AND SIMULATION RESULTS

To verify the feasibility and effectiveness of the proposed algorithm, a hydrothermal power system consists of a multi chain cascade of four hydro units and three thermal units were tested. The effect of valve point loading has been taken into account in this case study to illustrate the robustness of the proposed method. The transport time delay between cascaded reservoirs is also considered in this case study. The scheduling time period is one day with 24 intervals of one hour each. The data of test system are taken from [17] and [18]. The multi chain hydro sub system configuration is shown in figure 2. The water time transport delays between connected reservoirs are given in table I. In this case study, the output power of hydro power plants is represented as a function of the reservoir storage and the water discharge rates. The hydro power generation coefficients are given in table II. The reservoir storage limits, discharge rate limits, initial and end reservoir storage volume conditions and the generation limits of hydro power plants are shown in table III while table IV shows the reservoir inflows of multi chain hydro power plants. The fuel cost coefficients and the minimum and maximum limits of three thermal generating units are given in table V. The load demand over the 24 hours is given in table VI. The proposed algorithm has been implemented in MATLAB language and executed on an Intel Core i3, 2.27 GHz personal computer with a 3.0 GB of RAM. The control parameters of CFPSO algorithm used to solve short term hydro thermal scheduling problem are given in table VII. The optimal solution obtained from the proposed algorithm is achieved in 50 trial runs. The resultant optimal schedule of thermal and hydro power plants obtained from the CFPSO technique for each time interval is shown in table VIII. Table IX presents the fuel cost of each thermal unit and the total fuel cost of thermal power plants obtained from the proposed algorithm for each time interval while table X shows the optimal hourly water discharge of hydro power plants obtained from the CFPSO method. The optimal hourly storage volumes of hydro reservoirs obtained from the proposed algorithm are given in table XI.

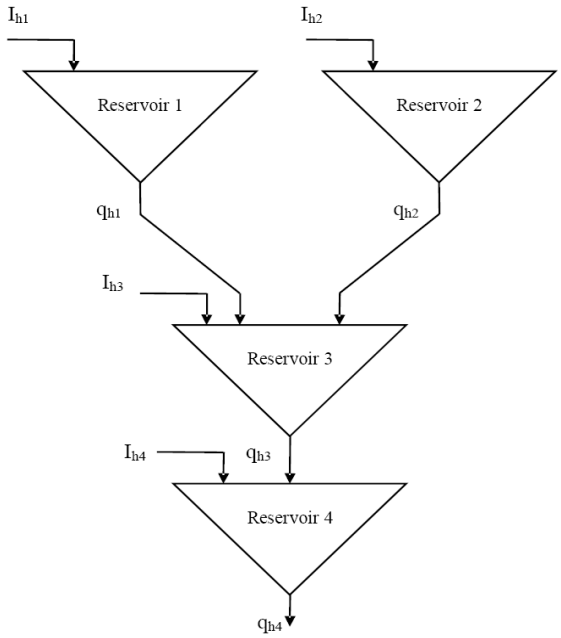

Fig.2. Multi chain hydro sub system networks

Table I: Water time transport delays between connected reservoirs

Plant	**1**	**2**	**3**	**4**
R_u	0	0	2	1
τ_u	2	3	4	0

Ru : Number of upstream hydro power plants
τu : Time delay to immediate downstream hydro power plant

Table II: Hydro power generation coefficients

Plant	**C_1**	**C_2**	**C_3**	**C_4**	**C_5**	**C_6**
1	-0.0042	-0.4200	0.0300	0.9000	10.000	-50.000
2	-0.0040	-0.3000	0.0150	1.1400	9.5000	-70.000
3	-0.0016	-0.3000	0.0140	0.5500	5.5000	-40.000
4	-0.0030	-0.3100	0.0270	1.4400	14.000	-90.000

Table III: Reservoir storage capacity limits, plant discharge limits, plant generation limits and reservoir end conditions ($\times 10^4 m^3$)

Plant	**V_h^{min}**	**V_h^{max}**	**V_h^{ini}**	**V_h^{end}**	**q_h^{min}**	**q_h^{max}**	**P_h^{min}**	**P_h^{max}**
1	80	150	100	120	5	15	0	500
2	60	120	80	70	6	15	0	500
3	100	240	170	170	10	30	0	500

4	70	160	120	140	13	25	0	500

Table IV: Reservoir inflows of multi chain hydro plants ($\times 10^4 m^3$)

Hour	Reservoir				Hour	Reservoir			
	1	2	3	4		1	2	3	4
1	10	8	8.1	2.8	13	11	8	4	0
2	9	8	8.2	2.4	14	12	9	3	0
3	8	9	4	1.6	15	11	9	3	0
4	7	9	2	0	16	10	8	2	0
5	6	8	3	0	17	9	7	2	0
6	7	7	4	0	18	8	6	2	0
7	8	6	3	0	19	7	7	1	0
8	9	7	2	0	20	6	8	1	0
9	10	8	1	0	21	7	9	2	0
10	11	9	1	0	22	8	9	2	0
11	12	9	1	0	23	9	8	1	0
12	10	8	2	0	24	10	8	0	0

Table V: Fuel cost coefficients and operating limits of thermal units

Unit	ai	bi	ci	e_i	f_i	P_{gi}^{min}	$P_{gi,}^{max}$
1	0.0012	2.45	100	160	0.038	20	175
2	0.0010	2.32	120	180	0.037	40	300
3	0.0015	2.10	150	200	0.035	50	500

Table VI: Load demand for 24 hour

Hour	P_D (MW)	Hour	P_D (MW)	Hour	P_D (MW)	Hour	P_D (MW)
1	750	7	950	13	1110	19	1070
2	780	8	1010	14	1030	20	1050
3	700	9	1090	15	1010	21	910
4	650	10	1080	16	1060	22	860
5	670	11	1100	17	1050	23	850
6	800	12	1150	18	1120	24	800

Table VII: Control parameters of particle swarm optimization

CFPSO parameters	Value
Population size	50
Maximum number of generations	300
Acceleration coefficients(c_1/c_2)	2.05
Minimum inertia weight (ω_{min})	0.4
Minimum inertia weight (ω_{max})	0.9
Constriction factor (k)	0.729

Table VIII: Hourly optimal hydrothermal schedule using constriction factor based particle swarm optimization (CFPSO)

Hour	Thermal generation (MW)			Hydro generation (MW)			
	P_{g1}	P_{g2}	P_{g3}	P_{h1}	P_{h2}	P_{h3}	P_{h4}
1	102.3522	209.8194	57.6422	60.1722	80.3207	38.6494	201.0440
2	20.0000	126.8176	230.7566	73.0700	79.3509	55.3298	194.6751
3	105.4454	130.2316	139.7551	54.0153	55.8002	42.4402	172.3121
4	25.1898	128.3247	141.6169	86.1289	65.3077	48.1490	155.2830
5	123.6643	116.0352	140.8527	54.2512	43.3706	23.7179	168.1081
6	20.2832	300.0000	144.4642	54.0606	73.2636	41.5883	166.3402
7	32.7205	300.0000	230.9010	88.9708	71.1724	55.5877	170.6477
8	101.6320	296.3523	234.4262	77.8782	70.3955	54.2548	175.0610
9	104.6402	295.1020	365.9320	56.0490	37.4051	44.0579	186.8139
10	110.1216	300.0000	319.4361	64.1774	44.9308	40.0597	201.2744
11	102.9433	299.8210	324.6830	96.2948	46.6031	38.5205	191.1343
12	29.9546	300.0000	410.6102	102.7084	56.2583	57.3524	193.1162
13	20.0000	294.0590	408.0650	87.5439	45.7874	54.3512	200.1934
14	20.1798	294.8191	319.1150	81.9074	51.3624	52.8587	209.7574
15	65.0533	297.0703	229.3150	94.6490	50.6550	49.3154	223.9421

16	116.1536	139.0801	406.3149	84.1369	53.9792	42.2257	218.1095
17	103.0538	209.8115	317.8150	99.4313	47.9614	52.1143	219.8126
18	35.3345	298.2462	320.2436	102.2590	69.0529	60.3747	234.4891
19	102.0183	211.1061	321.2727	84.0163	40.2404	52.7194	258.6312
20	100.0383	212.6210	313.3650	58.2941	42.5457	50.6354	272.5005
21	29.9704	295.1772	140.3611	79.0149	64.9985	37.0795	263.3983
22	109.9750	134.5710	232.0451	57.9149	42.6570	42.0930	240.7441
23	103.0293	125.5876	230.0580	65.3415	42.4109	45.5238	238.0490
24	22.6076	209.6222	140.0572	67.0476	49.5320	42.4138	268.7197

Table IX: Hourly fuel cost of each thermal unit and total fuel cost of the system using CFPSO technique

Hour	F1 ($/hr)	F2 ($/hr)	F3 ($/hr)	FT ($/hr)
1	365.2875	650.8290	328.8924	1345.009
2	149.4800	443.0075	723.1182	1315.606
3	388.5058	474.3245	472.8158	1335.646
4	193.8263	456.8758	490.4691	1141.171
5	535.8437	460.7023	483.1984	1479.744
6	151.9094	940.9172	517.4616	1610.288
7	255.8136	940.9172	724.5312	1921.262
8	367.7239	906.2017	758.9062	2032.832
9	381.4544	894.2413	1318.932	2594.627
10	429.0324	940.9172	974.9719	2344.922
11	366.5683	939.2243	1027.5630	2333.356
12	233.5563	940.9172	1276.1740	2450.648
13	149.4800	893.1142	1263.5300	2306.124
14	151.0224	891.5338	974.0437	2016.600
15	422.8659	913.0624	711.8716	2047.800
16	479.1878	552.1295	1269.9400	2301.257
17	367.5381	650.8123	979.1672	1997.518
18	276.1141	924.2753	983.0938	2183.483
19	366.4173	662.9219	993.4389	2022.778
20	373.0982	677.1374	996.4685	2046.704
21	233.6855	894.9607	478.5194	1607.165
22	427.7779	513.3079	735.7177	1676.804
23	367.3231	431.6617	716.2805	1515.265
24	171.8302	651.5545	475.6256	1299.011

Table X: Hourly hydro plant discharge using CFPSO technique

Hour	Hydro plant discharges ($\times 10^4 m^3/hr$)			
	q_{h1}	q_{h2}	q_{h3}	q_{h4}
1	5.7990	12.9505	20.5398	13.1229
2	7.4559	14.9805	12.8725	13.9983
3	5.0000	8.2127	17.9687	13.0000
4	9.6117	10.3248	16.4797	13.0000
5	5.0528	6.1585	22.5614	13.4225
6	5.0000	14.3987	18.7684	13.2641
7	10.1422	14.1917	11.5845	13.0000
8	8.1422	13.8035	15.0767	13.0000
9	5.1849	6.0528	19.1534	13.0000
10	6.0564	6.9771	20.4840	14.0564
11	11.1414	7.0141	20.9517	13.0528
12	13.5567	8.7984	13.7150	13.0000
13	9.4322	6.8451	16.3067	13.0000
14	8.3885	7.5845	17.3808	13.1585
15	10.6153	7.2676	19.0765	13.8979
16	8.6736	7.7958	20.6765	13.1849
17	11.8519	6.7976	18.0101	13.0000
18	13.4021	12.1095	13.0102	14.2676
19	8.9637	6.1907	18.3101	17.0670
20	5.4306	6.3603	19.0075	18.9034
21	8.1710	10.8396	22.3957	17.3716
22	5.3689	6.2854	21.1805	14.6610
23	6.1983	6.0528	19.9457	13.8475
24	6.3625	7.0287	20.5701	17.4578

Table XI: Hourly storage volume of hydro reservoirs using CFPSO technique

Hour	Reservoir storage volume ($\times 10^4 m^3$)			
	V_{h1}	V_{h2}	V_{h3}	V_{h4}
0	100.0000	80.0000	170.0000	120.0000
1	104.2010	75.0495	157.5602	109.6771
2	105.7451	68.0690	152.8877	98.0788
3	108.7449	68.8563	138.9190	86.6788
4	106.1334	67.5315	143.1888	73.6789
5	107.0806	69.3730	146.0638	80.7961
6	109.0805	61.9743	144.5081	80.4045
7	106.9384	60.0000	155.8601	85.3732
8	106.7962	60.0000	154.9947	88.8529
9	111.6113	61.9472	156.2400	98.4143
10	116.5549	63.9701	161.0899	103.1263
11	117.4135	65.9560	163.0839	101.6580
12	113.8568	65.1576	162.6066	103.7347
13	115.4264	66.3125	163.3334	109.8881
14	119.0379	67.7280	167.1081	117.2136
15	120.4226	69.4604	173.3867	124.2674
16	121.7490	69.6646	170.9875	124.7975
17	118.8971	69.8670	170.9504	128.1042
18	113.4950	63.7575	177.8232	131.2175
19	111.5313	64.5668	176.9825	133.2270
20	112.1007	66.2065	177.6245	135.0001
21	110.9297	64.3669	182.7404	135.6386
22	113.5608	67.0815	178.7143	133.9877
23	116.3625	69.0287	171.5595	138.4503
24	120.0000	70.0000	170.0000	140.0000

In order to verify and validate the effectiveness of the proposed technique, its simulation results will be compared with the results obtained from the simulated annealing and evolutionary programming techniques. Table XII shows the comparison of total fuel cost and computation time of the proposed methods among other methods. From table XII, it is observed that the constriction factor based PSO algorithm give high quality solution with less computation time compared to other methods. Figure 3 shows the hourly hydro plant power generation including total hydro generation by using proposed method, the hourly thermal plant power generation including

total thermal generation is given in figure 4, the hourly hydro plant discharges using proposed technique are shown in figure 5 while figure 6 presents the hourly reservoir storage volumes using proposed algorithm.

Table XII: comparison of total fuel cost and computation time of the proposed technique among GA, SA and EP techniques

Method	Total fuel cost ($)	CPU Time (Sec)
CFPSO	44925.62	183.64
SA [27]	45466.000	246.19
EP [27]	47306.000	9879.45

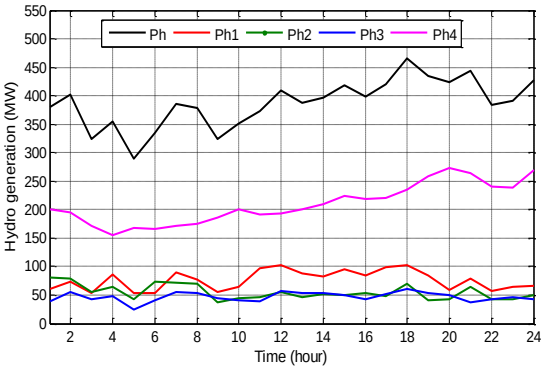

Fig.3. Hourly hydro plant power generation

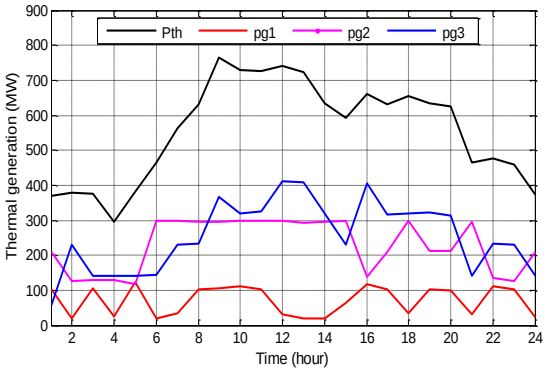

Fig.4. Hourly thermal plant power generation

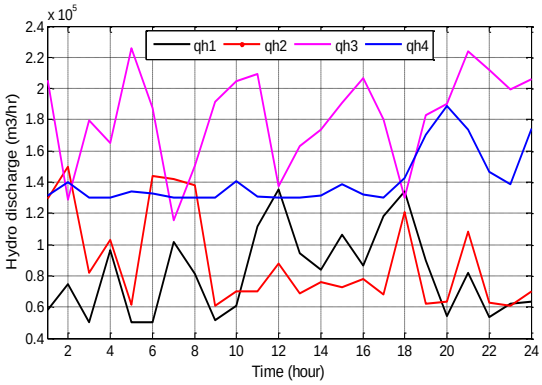

Fig.5. Hourly hydro plant discharge trajectories

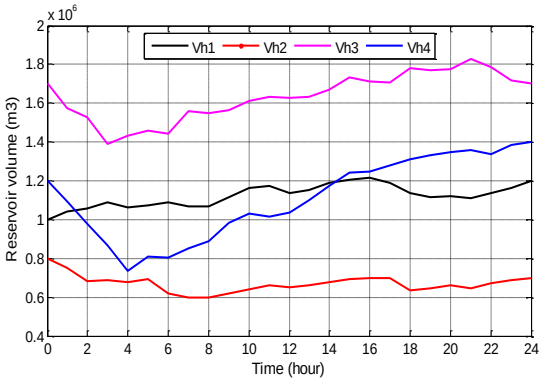

Fig.6. Hourly hydro reservoir storage volume trajectories

VI. CONCLUSIONS

In this paper, particle swarm optimization technique with constriction factor has been proposed for solving short term multi chain hydrothermal scheduling problem. To demonstrate the performance efficiency of the proposed algorithm, it has been applied on test system consists of a multi chain cascade of four hydro units and three thermal units. The effect of valve point loading is considered in this paper to demonstrate the robustness of the proposed technique. The results obtained by the proposed technique have been compared with other evolutionary computation techniques such as simulated annealing (SA) and evolutionary programming (EP) to verify the feasibility of the proposed method. The numerical results show that the proposed algorithm give a cheaper total fuel cost than those obtained from the other techniques. From the tabulated results, it is clear that the computational time of the proposed algorithm is much less than the other methods. Thus, the proposed approach can converge to the minimum fuel cost faster than the other approaches. Finally, the Simulation results demonstrate that the proposed method is a powerful optimization tool for solving

hydrothermal scheduling problems with non smooth objective functions.

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Estimation of Survival Distribution Using R Software

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Abstract- Survival analysis is widely applied in many fields such as biology, medicine, and public health. A typical analysis of survival data involves the modeling of time-to-event data, such as the time until death. To find Survival probabilities of the given observations censored or not, use Kaplan-Meier methods of estimation. But this method gives survival probabilities at any specific time. So this method does not compare total survival experience of two survival group. The non-parametric test, log rank test takes the complete follow up period and testing the significance difference between two survival distributions. In this paper, R software is used for finding survival (remission) probabilities and testing survival (remission) distributions using log rank test for 30 Resected Melanoma Patients.

Index Terms- Kaplan-Meier estimation, log rank test, R Software, Resected Melanoma Patients

I. Introduction

R Package is used to do Survival analysis. A lot of functions for survival analysis are in the package **Survival**. First install R and load the package **survival**. A step function with jumps at the observed event times will be obtained by using Kaplan-Meier method to estimate the survival function. This estimation takes censored and uncensored observations information to find out survival probabilities. Survival up to any point of time is calculated as the product of the conditional probabilities of surviving each time interval. This method of survival distribution is also obtained using R package.

The problem of testing survival distributions arises often in medical research. Survival curve gives rough idea about the distributions. But researcher expects significant difference in two or more treatments to prolong life of maintain health. So statistical test like log rank is necessary to find out significance difference exists or not between two survival distributions.

II. Non-Parametric Estimation and Non-Parametric Test

2.1. Non Parametric estimation – Kaplan-Meier method of estimation.

Let n be the total number of individuals whose survival times, censored or not, are available. Relabeling the survival times in order of increasing magnitude such that $t_1 \leq t_2 \leq \cdots .. t_n$ and the values of r are consecutive integers 1,2,….,n if there are no censored observation . If there are censored observations, they are not. Then the survival probabilities are calculated using S (t) = $\prod_{t_r \leq t} \frac{(n-r)}{(n-r+1)}$, where r runs through those positive integers for which $t_r \leq t$ and t_r is uncensored.

2.2 Non-Parametric Test - The Log-rank Test

Let d_t be the number of deaths at time t and n_{1t} and n_{2t} be the numbers of patients still exposed to risk of dying at time up to t in the two treatment groups. The expected deaths for groups 1 and 2 at time t are

$$e_{1t} = \frac{n_{1t}}{n_{1t} + n_{2t}} * d_t, \quad e_{2t} = \frac{n_{2t}}{n_{1t} + n_{2t}} * d_t$$

Then the total numbers of expected deaths in the two groups $E_1 = \sum e_{1t}, E_2 = \sum e_{2t}$. Let O_1 and O_2 be the observed numbers and E_1 and E_2 the expected numbers of death in two treatment groups.

$$The\ Test\ statistic\ \chi^2 = \frac{(O_1 - E_1)^2}{E_1} + \frac{(O_2 - E_2)^2}{E_2}$$ has approximately the chi-square distribution with one degree of freedom. A large χ^2 value (e.g., $\geq \chi^2_{1,05}$) would lead to the rejection of the null hypothesis in favor of the alternative that the two treatments are not equally effective at α= 0.05.

2.3 *R Software*

R software is used to find survival probabilities, survival curves and testing significant difference between two survival distributions using log rank test. First install package survival using

>install.packages('survival')

To load libraries, use

>library(survival)

III. Example

Thirty melanoma patients were studied to compare the immunotherapics BCG (Bacillus Calmette –Guerin) and Coryne bacterium parvum for their abilities to prolong remission time.

First create data set with remission time event (Censor = 0, uncensored = 1), sex (Male = 1, female = 2),treat(BCG = 1, C.Parvum = 2)

To find survival probabilities using Kaplan-Meier method of estimation use

> library(survival)
Loading required package: splines

>rtime=c(33.7,3.8,6.3,2.3,6.4,23.8,1.8,5.5,16.6,33.7,17.1,4.3,26.9,21.4,18.1,5.8,3.0,11.0,22.1,23.0,6.8,10.8,2.8,9.2,15.9,4.5,9.2,8.2,8.2,7.8)

>

revent=c(0,1,1,1,1,0,1,1,0,0,0,1,0,0,0,1,1,0,1,0,1,0,1,1,1,1,1,0,0,0)

> sex=c(2,2,1,2,1,2,2,1,1,2,2,1,1,1,1,2,1,2,2,1,1,2,2,1,1,2,2,1,1,1,1,2,2,2)

>

treat=c(1,1,1,1,1,1,1,1,1,1,2,2,2,2,2,2,2,2,2,2,2,2,2,2,2,2,2,2,2,2)

> data=data.frame (rtime, revent,sex,treat)

> fit1=survfit(Surv(rtime,revent)~1,data=data)

> summary(fit1)

Call: survfit(formula = Surv(rtime, revent) ~ 1, data = data)

Table 1: Remission Probabilities of melanoma patients obtained through K-M estimate

time	n.risk	n.event	survival	std.err	lower 95% CI	upper 95% CI
1.8	30	1	0.967	0.0328	0.905	1.000
2.3	29	1	0.933	0.0455	0.848	1.000
2.8	28	1	0.900	0.0548	0.799	1.000
3.0	27	1	0.867	0.0621	0.753	0.997
3.8	26	1	0.833	0.0680	0.710	0.978
4.3	25	1	0.800	0.0730	0.669	0.957
4.5	24	1	0.767	0.0772	0.629	0.934
5.5	23	1	0.733	0.0807	0.591	0.910
5.8	22	1	0.700	0.0837	0.554	0.885
6.3	21	1	0.667	0.0861	0.518	0.859
6.4	20	1	0.633	0.0880	0.482	0.832
6.8	19	1	0.600	0.0894	0.448	0.804
9.2	15	2	0.520	0.0937	0.365	0.740
15.9	11	1	0.473	0.0964	0.317	0.705
22.1	6	1	0.394	0.1078	0.230	0.674

This table shows corresponding survival (remission) probability, Standard error, lower and upper limits for 95% condifidence interval For Survival curves use

> plot(fit1,xlab="time in months",ylab="Survival Probability")

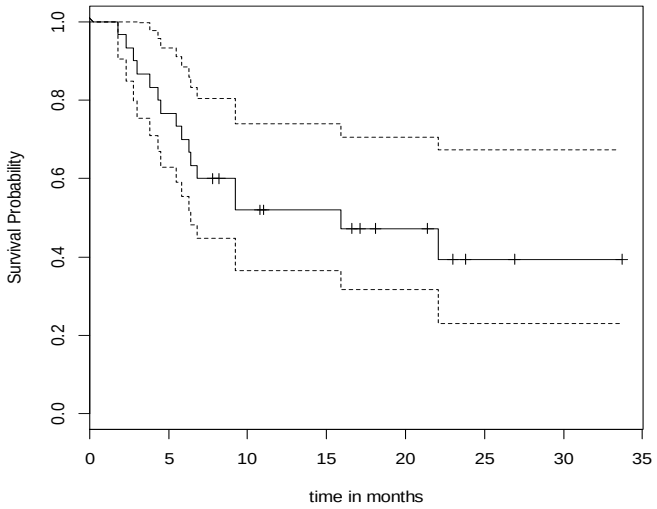

Figure 1. KM survival curves with 95% Confidence interval for melanoma patients

> fit3=survfit(Surv(rtime,revent==1)~sex,data=data)
summary(fit3)
Call: survfit(formula = Surv(rtime, revent == 1) ~ sex, data = data)

Table 2: Remission Probabilities of melanoma patients obtained through K-M estimate related with sex

sex=1

time	*n.risk*	*n.event*	*survival*	*std.err*	*lower 95% CI*	*upper 95% CI*
3.0	*15*	*1*	*0.933*	*0.0644*	*0.815*	*1.000*
4.3	*14*	*1*	*0.867*	*0.0878*	*0.711*	*1.000*
4.5	*13*	*1*	*0.800*	*0.1033*	*0.621*	*1.000*
5.5	*12*	*1*	*0.733*	*0.1142*	*0.540*	*0.995*
6.3	*11*	*1*	*0.667*	*0.1217*	*0.466*	*0.953*
6.4	*10*	*1*	*0.600*	*0.1265*	*0.397*	*0.907*
6.8	*9*	*1*	*0.533*	*0.1288*	*0.332*	*0.856*
9.2	*8*	*2*	*0.400*	*0.1265*	*0.215*	*0.743*
15.9	*6*	*1*	*0.333*	*0.1217*	*0.163*	*0.682*

sex=2

time	*n.risk*	*n.event*	*survival*	*std.err*	*lower 95% CI*	*upper 95% CI*
1.8	*15*	*1*	*0.933*	*0.0644*	*0.815*	*1.000*
2.3	*14*	*1*	*0.867*	*0.0878*	*0.711*	*1.000*
2.8	*13*	*1*	*0.800*	*0.1033*	*0.621*	*1.000*
3.8	*12*	*1*	*0.733*	*0.1142*	*0.540*	*0.995*
5.8	*11*	*1*	*0.667*	*0.1217*	*0.466*	*0.953*
22.1	*4*	*1*	*0.500*	*0.1708*	*0.256*	*0.977*

>plot(fit3, xlab="time in months",ylab="Survival Probability")

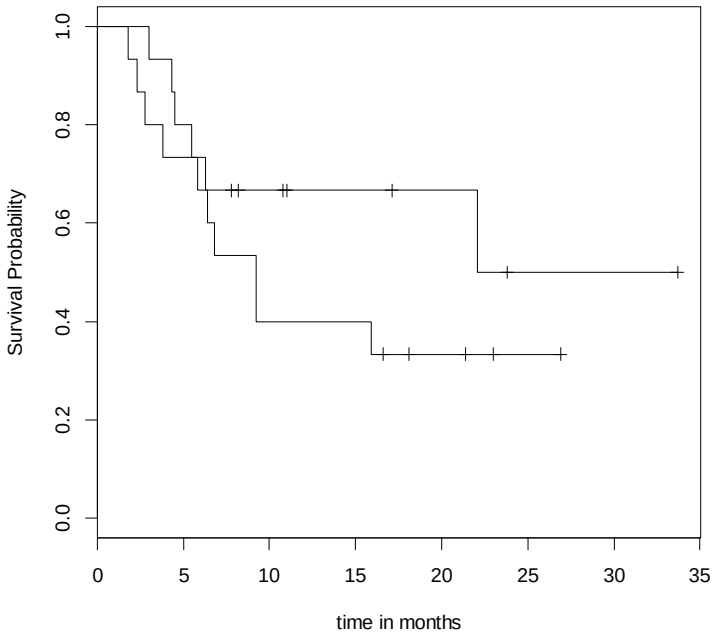

Figure 2. KM survival curves for remission times of patients compared with sex

Testing survival distributions related with sex using Log-rank test

> fit5=survdiff(Surv(rtime,revent)~sex,data=data,rho=0)
> fit5
Call:
survdiff(formula = Surv(rtime, revent) ~ sex, data = data, rho = 0)

	N	Observed	Expected	(O-E)^2/E
sex=1	15	10	8.21	0.389
sex=2	15	6	7.79	0.410

Chisq= 0.8 on 1 degrees of freedom, p= 0.367
> fit7=survfit(Surv(rtime,revent==1)~treat,data=data)

> summary(fit7)
Call: survfit(formula = Surv(rtime, revent == 1) ~ treat, data = data)

Table 3: Remission Probabilities of melanoma patients obtained through K-M estimate related with treatment type.

treat=1

time	*n.risk*	*n.event*	*survival*	*std.err*	*lower 95% CI*	*upper 95% CI*
1.8	*11*	*1*	*0.909*	*0.0867*	*0.754*	*1.000*
2.3	*10*	*1*	*0.818*	*0.1163*	*0.619*	*1.000*
3.8	*9*	*1*	*0.727*	*0.1343*	*0.506*	*1.000*
5.5	*8*	*1*	*0.636*	*0.1450*	*0.407*	*0.995*
6.3	*7*	*1*	*0.545*	*0.1501*	*0.318*	*0.936*
6.4	*6*	*1*	*0.455*	*0.1501*	*0.238*	*0.868*

treat=2

time	*n.risk*	*n.event*	*survival*	*std.err*	*lower 95% CI*	*upper 95% CI*
2.8	*19*	*1*	*0.947*	*0.0512*	*0.852*	*1.000*
3.0	*18*	*1*	*0.895*	*0.0704*	*0.767*	*1.000*
4.3	*17*	*1*	*0.842*	*0.0837*	*0.693*	*1.000*
4.5	*16*	*1*	*0.789*	*0.0935*	*0.626*	*0.996*
5.8	*15*	*1*	*0.737*	*0.1010*	*0.563*	*0.964*
6.8	*14*	*1*	*0.684*	*0.1066*	*0.504*	*0.929*
9.2	*10*	*2*	*0.547*	*0.1215*	*0.354*	*0.846*
15.9	*6*	*1*	*0.456*	*0.1311*	*0.260*	*0.801*
22.1	*3*	*1*	*0.304*	*0.1518*	*0.114*	*0.809*

> plot(fit7, xlab="time in months",ylab="Survival Probability")

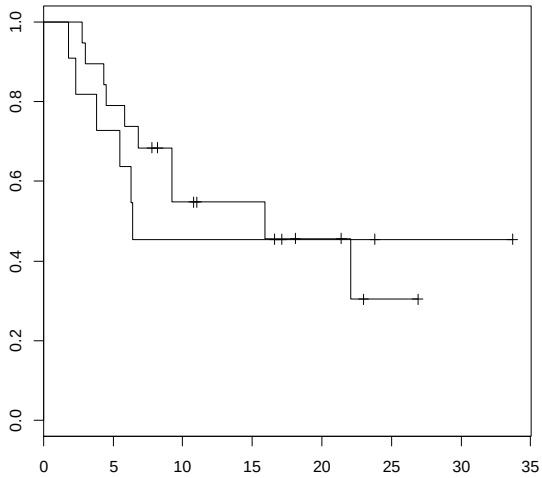

Figure 3. KM survival curves for remission times related with treatment type

```
>  fit7=survdiff(Surv(rtime,revent)~treat,data=data,rho=0)
> fit7
Call:
survdiff(formula = Surv(rtime, revent) ~ treat, data = data,
    rho = 0)

          N  Observed  Expected  (O-E)^2/E
treat=1  11         6      5.55     0.0366
treat=2  19        10     10.45     0.0194

Chisq= 0.1  on 1 degrees of freedom, p= 0.811
```

IV. DISCUSSION

Figure 1 Shows survival probabilities for remission time of all patients in survival curve form. It shows the relation between survival (remission) probabilities and time for each observation. Figure 2 shows the survival curves of male and female patients. This figure also gives the idea about survival probabilities of male and female patients. Survival curve for male is differing from the survival curve for female. In early period two curves are overlapped.

Figure 3 shows the survival curve of patient receiving BCG and C. Parvum treatment. It also shows some difference in survival curves. But these significantly difference may be check using log-rank test.

When using log-rank test regarding sex of the patients, the value of p is 0.367>0.05.So accept null hypothesis is that the survival distribution of male patient is same as that of female patients.

When using log-rank test regarding treatment type of the patients, the value of p is 0.811>0.05.So accept null hypothesis is that the survival distribution of BCG treatment receiving patients is same as that of C.Parvum treatment receiving patients.

V. CONCLUSION

We obtained Survival Probabilities, Survival curve and test value using log-rank test are obtained by the use of R Statistical software which is freely available. Kaplan-Meier method of estimating survival curves only gives pictorial representation of the survival distributions but it does not take whole follow-up period into account. Survival curves related to sex and related to type of treatment shows slight difference in the survival curves of corresponding survival distributions but log-rank test determines there is no significant difference between these curves for both related to sex and related to treatment type. So Log-rank test is used to test the null hypothesis that there is no significant difference between the two survival distributions in the probability of an event at any time point.

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Cystic Lymphangioma of Mesentery - A Case Report

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Abstract- Lymphangiomasare malformations that occur most often in the head and neck or axilla of young children but are detected occasionally in adults at other sites such as mesentery , omentum and mesocolon . We present a case of cystic lymphangioma occurring in the splenic mesentery in a 28 year old male.

Index Terms- cystic lymphangioma, mesentery

I. INTRODUCTION

Lymphangiomas are malformations that arise from sequestration of lymphatic tissue that fail to communicate normally with the lymphatic system.[1]They become markedly dilated under the pressure effect of accumulating lymph. Lymphangiomas can be acquired due to obstruction following surgery , irradiation or infection.Intra abdominallymphangiomas are rare and occur in the mesentery ,omentum and mesocolon.[2]We report a case of cystic lymphangiomaoccuring in the splenic mesentery in an adult.

II. CASE REPORT

A 28 year old male presented with complaints of abdominal pain of 5 years duration. USG abdomen showed a large multilocular cystic lesion in the splenic hilum. Exploratory laprotomy revealed a lobular cystic lesion measuring 22 x22 x 4 cm in the mesentery at the splenic hilum. The cyst along with the spleen was excised and sent for histopathological examination.Postoperative period was uneventful.Gross examination revealed a multiloculated thin walled cyst measuring 22 x 12 x 4 cm filled with 600 ml of serosanguinous straw coloured fluid.

Fig 1 .Multiloculated thin walled cyst

On microscopic examination the cyst wall was lined by flattened endothelial cells with proteinaceous material in the lumen.Wall showed attenuated smooth muscle ,lymphoid aggregates, foamy histiocytes and cholesterol clefts. A diagnosis of cystic lymphangioma of mesentery was made.

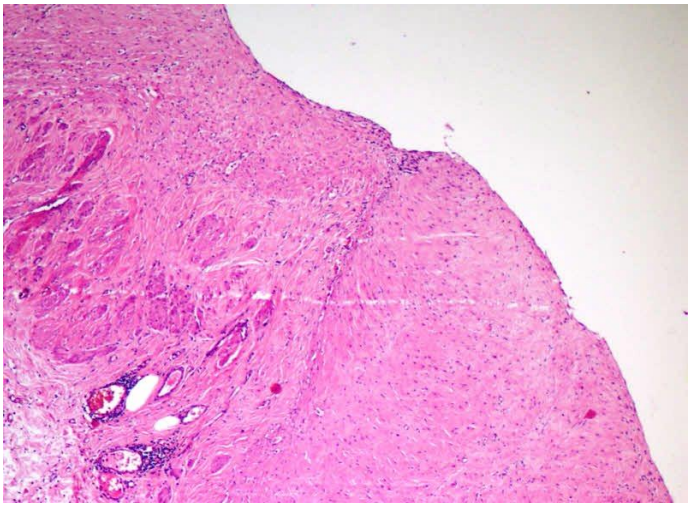

Fig 2.Cyst wall lined by fattened endothelial cells .Wall shows smooth muscle and aggregates of lymphocytes H&E x100

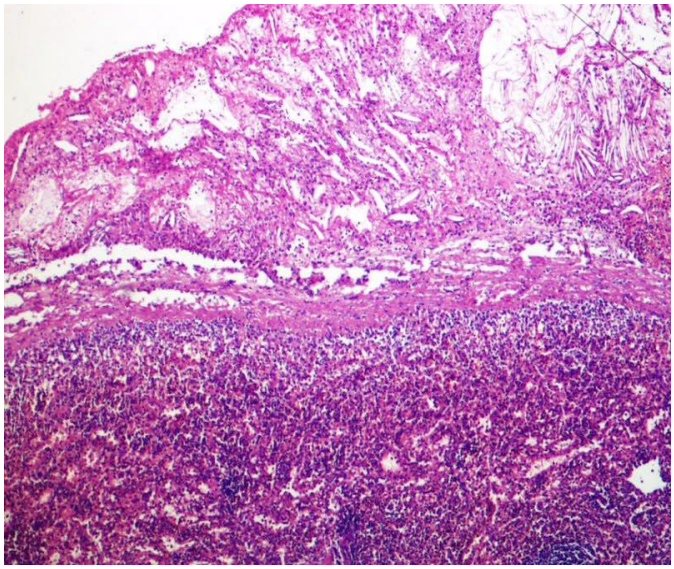

Fig 3.Cyst wall seen closely adherent to spleen but not infiltrating into the parenchyma H&E x 200

III. DISCUSSION

Lymphangiomasusually occur in children and are commonly(95%) found in head, neck and axillary region. The remaining 5% are located in the mesentery ,retroperitoneum,abdominal viscera ,lung and mediastinum. Lymphangioma of abdomen is rare especially in adults.[2,3,4]

Lymphangiomas are classified histologically into capillary, cavernous and cystic.[4,5,6]The capillary lymphangiomas are usually situated superficially in the skin and consists of small thin walled lymphatics . The cavernous lymphangioma consists of larger lymphatic spaces having a connection with normal adjacent lymphatics. Cystic lymphangiomas are usually multilocular although they may be unilocular. They contain serous or chylous fluidand show dilated lymphatic spaces of various sizes associated with collagen and smooth muscle bundles in the stroma but lacks connection to the adjacent normal lymphatic spaces.[5]

Clinical signs and symptoms are related to the size of the lesion.Abdominal pain and distension are thecommon presenting features of cystic lymphangioma of abdomen.Most of these patients develop symptoms of acute abdomen caused by intestinal obstruction , volvulus or infarction.[3,7,8] A preoperative diagnosis is possible in most of these cases by USG or CTscan.USG shows multiple homogenous nonenhancingareas with variable attenuation .

Cystic lymphangioma has to be differentiated from cystic mesothelioma, enteric duplication cyst and pseudocyst of pancreas.[2]Cystic mesothelioma shows glandular spaces of varying sizes lined by mesothelial cells. Presence of an enhancing thick wall is helpful in characterizing a lesion as either a pseudocyst or an enteric duplication cyst. Pseudocyst of pancreas is characterized by thick irregular wall with a ragged inner surface and absence of a lining epithelium.Treatment of cystic lymphangioma is complete surgical excision and sometimes may require resection of the adjacent bowel or other closely associated structures because incomplete excision may lead to recurrence.

IV. CONCLUSION

Lymphangiomas are uncommon benign lesions that may rarely present as cystic lesions in the abdomen. Awareness of this entity is very essential while evaluating intra abdominal cystic masses.

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Cognitive Behavioral Therapy on Frontal EEG Asymmetry among Students with Higher Levels of Test Anxiety

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Abstract- Individuals with more anxiety reveal more right frontal activity and they respond towards negative emotions and the individuals with left frontal activity exhibit more positive emotions. The present study aims to investigate the impact of CBT on frontal asymmetry on the students having more levels of test anxiety and EEG activity was measured in 16 higher secondary students with higher levels of test anxiety before and after cognitive behavioral therapy (CBT).After CBT students shifted towards greater relative left frontal brain activity. Students with greater left frontal activity were having lower levels of test anxiety and the results indicated that CBT reduced test anxiety and brought out the changes in frontal asymmetry. It is concluded that CBT has an impact on frontal asymmetry and test anxiety levels of students.

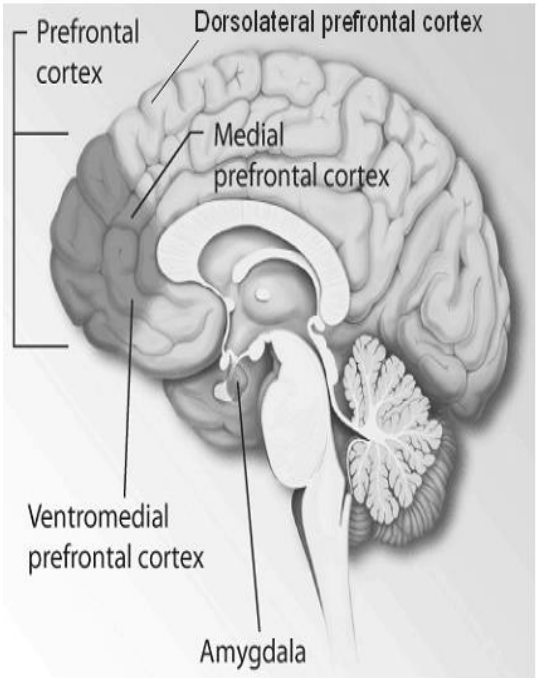
Index Terms- cognitive behavioral therapy, electroencephalogram, frontal asymmetry, test anxiety

I. INTRODUCTION

Motivated by a range of possible applications in the field of human-computer interaction, research on emotion recognition from facial expressions, speech and physiological signals receives increasing attention. Many application areas can benefit from emotion recognition systems, ranging from applications that track the user's affective states and give corresponding feedback (e.g., automatic tutoring applications) to personalized photo or music selection applications. Furthermore, they can be useful in exploring reactions to advertisements, for monitoring emotional states in the healthcare area or in detecting which product aspects cause frustration. The correlates of emotion in human EEG have been discovered more than two decades ago. In particular, the phenomenon of"frontal EEG asymmetry" has played a prominent role in the emotion research. According to Davidson's influential approach/withdraw motivational model of emotion [2003] left frontal activity indicates a positive or approach-related emotion, whereas higher right frontal activity indicates a negative or withdrawal-related emotion. The degree of activation is inferred from the spectral power in the alpha band (8-12 Hz), with lower values in alpha power being associated with a higher degree of activity.

The role of hemispheric laterality in emotion has been a source of debate for decades. Contrary to more simplistic notions of right hemisphere superiority across all domains of emotion, recent conceptualizations of hemispheric asymmetries have taken into account the variety of psychological experiences that fall under the broad rubric of emotion. For example, Davidson (1993) has proposed that while right posterior regions specialize in the perception of affective stimuli of both positive and negative valence, both Cognition and emotion. The left frontal regions may be more active during the experience of approach-related (i.e. positive) emotions and the right frontal regions may be more active during the experience of withdrawal-related (i.e. negative) emotions. Indeed, these Findings suggest that frontal EEG asymmetry may represent a predisposition that underlies individual differences in reactivity to valence stimuli. More specifically, individuals who exhibit hypo activation of the left frontal region and who, therefore also demonstrate elevated responsively to negative stimuli, may be at increased risk for experiencing episodes of depression. This formulation is consistent with the results of a wide range of studies examining biobehavioural aspects of depression (Ian Gotlib, 1998). For example, Robinson and his colleagues (e.g. Robinson, Kubos, Starr, Rao, & Price, 1984; Robinson & Price, 1982) have presented data supporting an association between left frontal lesions and subsequent depression (for a review, see Robinson & Downhill, 1995). Furthermore, several investigations utilizing positron emission tomography (PET) have similarly found left frontal hypo activation in depressed subjects (e.g. Bench et al., 1993; Martinot et al., 1990; but see also Drevets et al., 1992). in reviewing the literature, one will variously encounter references to frontal EEG activity and frontal EEG activation. At first blush, insistence upon this distinction may appear pedantic, but in fact, strict attention here can significantly enhance understanding of research in this area, and is relevant to making inferences concerning the distinction between frontal EEG asymmetry as a moderator or a mediator. For the purposes of this article, and as a general recommendation, activity will refer to a tonic recording of cortical processes as measured by EEG, while activation will refer to the change in EEG activity in response to a provocation, such as the presentation of an emotional stimulus. For example, one may be interested in measuring an individual's asymmetry in frontal EEG activity at rest as well as that same individual's asymmetrical frontal EEG activation in response to an experimental manipulation, such as the presentation of an image of a venomous snake. While this implies that activity refers only to resting or baseline measures, this is not necessarily the case. Indeed, one could measure an individual's resting activity both at baseline and following a stimulus presentation. The difference between those post-stimulus and baseline activity measures, however, would represent that individual's inferred activation in

response to the stimulus. This distinction is important for maintaining conceptual clarity when reviewing the literature on frontal EEG asymmetry. Thus, in reviewing the literature, one of the functions of this article will be to make clear this distinction in past literature as well as in the arguments and data presented here.



II. OBJECTIVES

- to study the impact of CBT on frontal EEG activity among the students having higher levels of test anxiety
- to improve the quality of academic life of the students

Frontal Asymmetry: ln (right) - ln(left)

1.2 Research question

Does Cognitive Behavior Therapy bring out changes in the frontal asymmetry scores of the brain waves among students having higher levels of test anxiety?

III. DATA ACQUISITION

200 students (age range 17–19 yr.) were selected and they were given speilberger's Test Anxiety questionnaire and scores were tabulated to select 16 students with higher levels of Test anxiety and they were divided into two groups and in each group 8 girls and 8 boys were present. Subjects were free from medical, psychiatric, and sleep disorders as determined by history, physical examination, biochemical screening tests, electrocardiograms, and psychological screening questionnaires. Their normal EEG pattern was recorded. Experimental group was given cognitive behavior therapy for about three weeks. Then their post test scores of EEG activity, and test anxiety were recorded. Pre test and post test scores of experimental group were statistically analyzed to tabulate the data. It is concluded that CBT has an impact on frontal asymmetry and test anxiety among students.

IV. DATA ANALYSIS

Absolute Power: The actual power (voltage) in the patient's EEG database. (Power is microvolt's squared.) Squaring the microvolt's we get the power of that brain region

Interhemispheric and Intrahemispheric Coherence: Interhemispheric (between left and right hemisphere sites) and intrahemispheric coherence (between sites in the same hemisphere) measures the similarity or correlation of the EEG signal between regions

Table-1 Mean scores of Different electrodes of FP1,F3, F8, FP2, F4, Fz, and F7 of Boys and Girls of the total sample

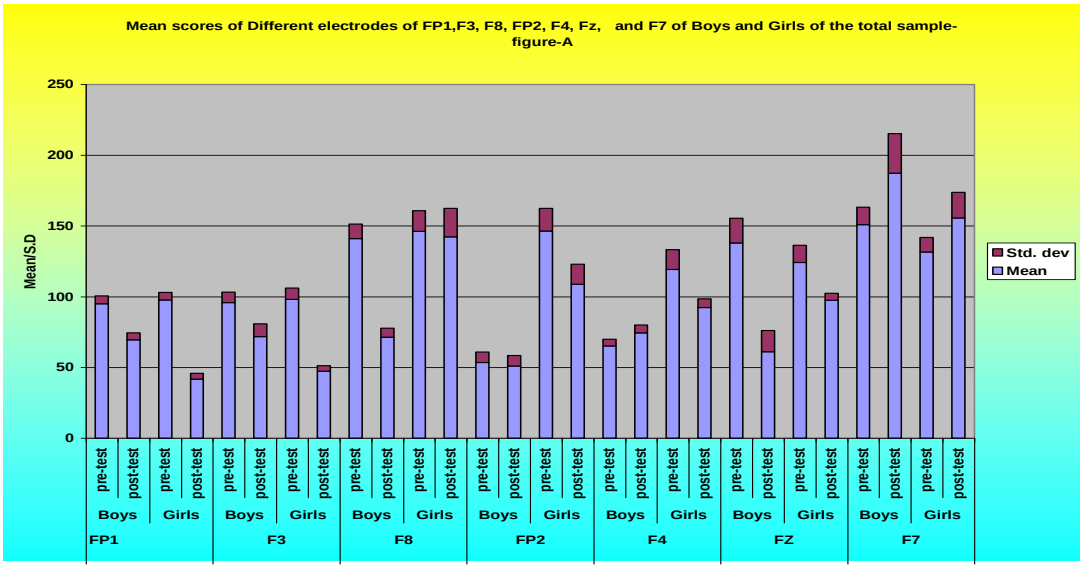
brain regions	electrode	Boys				Girls			
		pre-test		post-test		pre-test		post-test	
		mean	S.D	mean	S.D	mean	S.D	mean	S.D
frontal	FP1	94.99	5.46	69.50***	4.89	97.58	5.522	41.69***	4.06
	F3	95.83	7.40	71.82***	8.91	98.16	7.90	47.35***	3.93
	F8	140.99	10.4	71.34***	6.37	146.2	14.63	142.20*	20.23
	FP2	53.36	7.52	50.91**	7.52	146.36	15.99	108.85***	14.06
	F4	65.18	4.67	74.46***	5.41	119.17	14.07	92.36***	6.22
	FZ	137.96	17.5	61.14***	14.9	124.1	12.13	97.48***	4.93
	F7	150.86	12.3	187.36***	27.81	131.58	10.32	155.68***	18.00

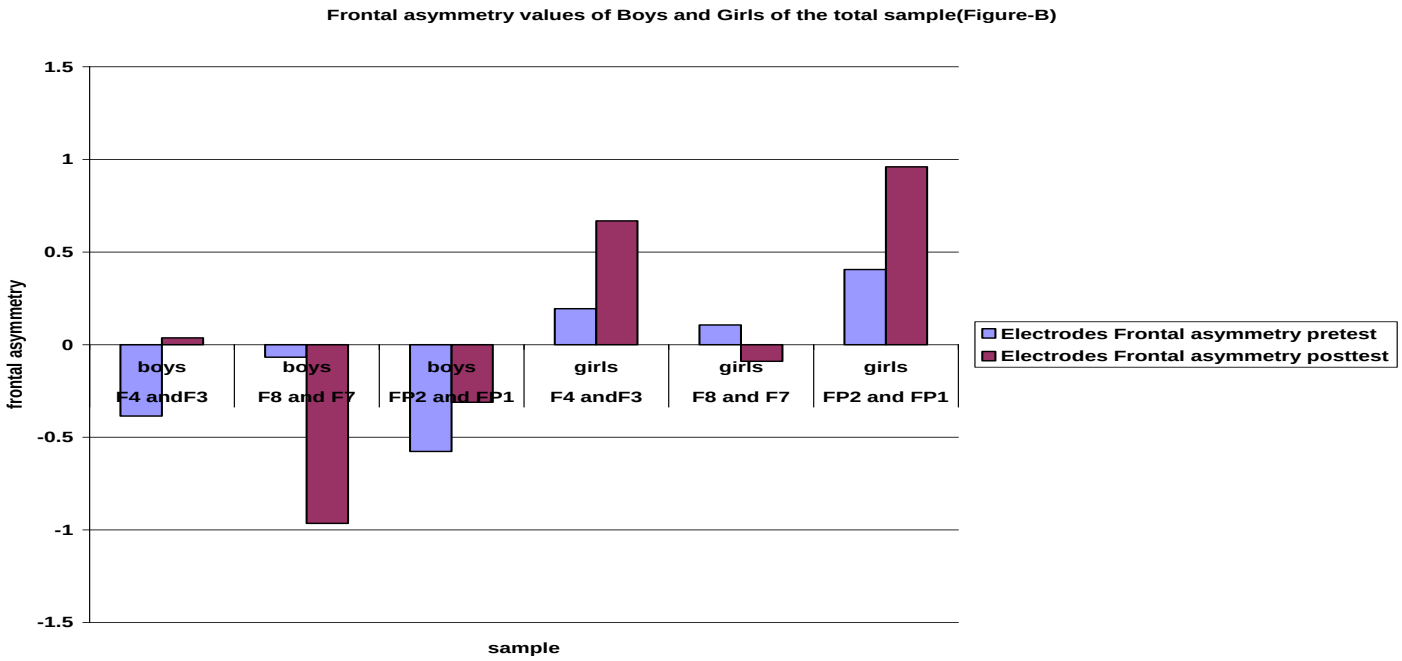
***P< 0.05, **P < 0.01, ***P< 0.001, N.S= Not Significant**

Table-2 Frontal asymmetry values of Boys and Girls of the total sample

Variables	Gender	Electrodes	No	Frontal asymmetry	
				pretest	posttest
absolute powers of Right frontal and left frontal brain regions	Boys	F4 andF3	8	-0.3854	0.0360
		F8 and F7	8	-0.0676	-0.9655
		FP2 and FP1	8	-0.5767	-0.3112
	Girls	F4 andF3	8	0.1939	0.6681
		F8 and F7	8	0.1053	-0.09
		FP2 and FP1	8	0.4050	0.9597

***P< 0.05, **P < 0.01, ***P< 0.001, N.S= Not Significant**





It is clear from the table 1 and 2 also from figures A and B that there are significant differences between pretest 'mean' scores and post test ' mean ' scores of right and left sides of the brain from the boys and girls sample. Significant changes are due to the application of Cognitive behavior therapy. This study is in line with the following studies.

. Davidson et al. (2003) examined changes in brain electrical asymmetry and immune function after a meditation program. They found that the meditation group showed significant larger increases in left-sided anterior activation (during baseline and positive, negative emotion induction) compared with a Wait-list group. EEG alpha power is inversely related to scalp-recorded cortical activity, positive asymmetry scores are thought to reflect greater relative left EEG cortical activity (Allen et al., 2004a; Davidson, 1988). Additionally, Craig (2005) has described a likely anatomical basis for autonomic nervous system lateralization. The right anteriorinsula (located deep to T4) receives ascending projections originating from afferent nerves that mediate sympathetic nervous system functions, whereas the left anterior insula (located deep to T3) receives Projections from afferent nerves that mediate parasympathetic nervous system functions. Our results are consistent with a report of functional brain changes due to CBT in patients with spider phobia (*Paquette V,* et al, 2003). Frontal EEG activity and ocular parameters may be used to monitor and predict changes in neurobehavioral performance associated with sleep loss and circadian misalignment. This global increase in slow EEG activity was less pronounced in the occipital derivation (Pz-Oz,), and the late evening decrease was more pronounced compared with the frontal derivation Brainwave Optimization. We propose that a wider range of human behaviors may be fundamentally motivated by the brain's intention to regulate itself. Indeed, we believe that the brain's impetus to self-regulate may eventually explain pathological behaviors more convincingly than can conventional psychological, sociological or genetic theories. Greater self-regulation by the human brain is necessary because the brain is our central command center for our biology at a global level. The brain is the central control center for all human experience and functioning. So to improve any aspect of our well-being, we should ultimately aim to facilitate better functioning of the brain. Peter Sterling's model of allostasis (2004) provides a theoretical elaboration of the centrality of the brain for human self-regulation. In the past, the physiological and medical sciences have been based on the model of homeostasis, or stability through constancy. Homeostasis considers various systems in terms of their requirement to maintain various set points at constant values. Deviations from these set-points are treated as disease states. The cause of these deviations is understood to bodys functionality of local mechanisms in the system. That is to say, dysfunctional local mechanisms are seen to interfere with preservation of the set-points. The aim of medical therapy is to correct the local mechanisms which are associated with set-point deviations, e.g., disease. In contrast, the allostasis model emphasizes that systems maintain stability through change. They shift their set-points based on the changing demands their environments. With regard to models of brain asymmetry and emotion increased alpha positive predictor for treatment outcome (Ulrich et al., 1984; Bruder et al 2001) in a refinement of their theory Heller, Nitschke, Etienne, and Miller (1997) proposed a distinction between subtypes of anxiety related to different patterns of brain activation. Anxious arousal (e.g. panic, state anxiety, and sympathetic nervous system hyper reactivity) should be associated with greater right parietal-temporal activation, while anxious apprehension (e.g. worry, rumination, anticipation of future threat) should be associated with greater left anterior activity since it involves linguistic processing (Heller et al., 2003; Heller et al., 2002; Nitschke et al., 2000). Partial support for their assumptions comes from studies examining regional EEG asymmetry in self-reported anxious arousal (Nitschke et al., 1999) and during experimentally manipulated anxious-arousal (Heller et al., 1997) the different models of brain asymmetry and emotion each are strong in

explaining different components of emotion (e.g. perception, experience, expression, and arousal). The right-hemisphere model emphasizes emotional perception and expression. The valence model, which is mainly focusing on experience of emotion, was subsumed by the AAE model of emotion which has probably the most empirical support. The AAE model is probably the most elaborated allowing clear predictions (e.g. about trait asymmetry as diathesis variable) and thus enables good empirical evaluation. The integrative model of Heller and colleagues confines the valence hypothesis to anterior brain regions and in part the 50 right-hemisphere hypothesis to posterior regions. Its strength is the integration of seemingly discrepant findings regarding different types of anxiety and comorbidity of anxiety and depression. Relative left frontal hemisphere activation has been proposed to be associated with a self enhancing regulatory style inhibiting negative affective responses (Tomarken & Davidson, 1994). Since CBT leads to a decrease of PTSD symptoms and anxiety we hypothesized that participants receiving CBT would exhibit a greater decrease in right anterior and posterior activation during exposure to a trauma-related accident. Specifically, there was an increase in the activation of the left medial temporal gyrus in response to the paradigm of empathy. The same process occurred with the posterior cingulate gyrus, which had its activation increased in response to the condition of forgiveness after the treatment. From this study Farrow et al. 2005 concluded that CBT can promote changes in the brain area. Schwartz et al, 1996 replicated the previous findings of significant decrease of the activity of the right caudate nucleus in those who responded to therapy. CBT favors the restructuring of thought, modification of feelings and behaviors, and promotes new learning. Consequently it involves synaptic changes (2006) and this may be responsible for significant changes of mean scores of absolute powers of Different electrodes of EEG of the sample.

Table-3 Inter Hemispheric Coherence of FP1 and FP2 of the brain in the total sample

Variables	Gender	Test	No	'r'	't'
FP1 vs FP2	Boys	Pre Test	8	0.2317 N.S	1.14
		Post Test	8	0.2870 N.S	1.43
	Girls	Pre Test	8	0.3175 N.S	1.6
		Post Test	8	0.7212***	4.99

***P< 0.05, **P < 0.01, ***P< 0.001, N.S= Not Significant**

It is clear from the table 3that the calculated correlation coefficient values were significantly lesser than that of table correlation coefficient values in the case of Boys (p< 0.01, N=8,) in pre-test and in the post test. Hence the hypothesis was accepted and proved that there are no significant Inter Hemispheric Coherence between Pre Test scores of FP1 vs. FP2 and Post Test scores of FP1 vs. FP2 before and after applying Cognitive Behavior Therapy.

In case of Girls (p< 0.01, N=8) the calculated correlation coefficient values were significantly lesser than the table correlation coefficient values in pre-test and in the post test it is greater than the table correlation coefficient values. Hence the hypothesis was rejected and proved that there are significant Inter Hemispheric Coherence between Pre Test and post test scores of FP1 vs. FP2 before and after applying Cognitive Behavior Therapy.

A model that might explain some inconsistencies in the study of brain asymmetries in relation to emotion and psychopathology has been suggested by Heller and colleagues (Heller, 1990, 1993; Heller & Nitschke, 1997), and subsequently refined (Heller et al., 2003; Heller & Nitschke, 1998; Heller et al., 1997; Heller, Schmidtke, Nitschke, Koven, & Miller, 2002; 48 Nitschke, Heller, & Miller, 2000; Nitschke, Heller, Palmieri, & Miller, 1999). This neuropsychological model of emotion integrates in part the competing right hemisphere hypothesis (Gainotti et al., 1993) and the valence hypothesis (Davidson, 1992). Based on and psycho physiological data the model incorporates psychological theories of emotion decomposing emotional states into two components, valence and arousal. The model proposes that the valence dimension (pleasant, unpleasant) is dependent on functions of anterior regions of the cortex. When the left frontal region is active relative to the right, affective valence is pleasant, whereas relative right-frontal activity is associated with negative affect. Furthermore, the model posits that the arousal dimension depends on right posterior regions of the brain. Nitschke et al. (2000) proposed that the right posterior anxious arousal system promotes sympathetic nervous system activity, spatial attention, visual scanning of the environment, and sensitivity to meaningful nonverbal cues. According to the model, anxiety and depression are associated with different patterns of brain activity in right posterior regions, with anxiety to be associated with increased and depression with decreased activity. The neuropsychological model of emotion by Heller and colleagues. This model explained some inconsistencies of prior EEG asymmetry research in depression and anxiety (Bruder et al., 1997; Kentgen et al., 2000; Metzger et al., 2004). Empirical evidence for the model comes from two studies investigating depressed patients (Keller et al., 2000) and students with high and low levels of depression (Heller, Etienne and Miller, 1995; Keller et al., 2000) using a neuropsychological task measuring a hemispheric correlation.

Table-4 Inter Hemispheric Coherence of F3 and F4 of the brain in the total sample

variables	Gender	Test	No	'r'	't'
F3 vs F4	Boys	Pre Test	8	0.347 N.S	1.77
		Post Test	8	0.4258**	2.25
	Girls	Pre Test	8	0.285 N.S	1.42
		Post Test	8	0.7505***	5.44

***P< 0.05, **P < 0.01, ***P< 0.001, N.S= Not Significant**

It is clear from the table 4 that the calculated correlation coefficient values were significantly lesser than that of table correlation coefficient values in the case of Boys ($p< 0.01$, N=8,) in pre-test and in the post test. Hence the hypothesis was rejected and proved that there are no significant Inter Hemispheric Coherence between Pre Test scores of F3 vs. F4 and Post Test scores of F3 vs. F4 before and after applying Cognitive Behavior Therapy.

In case of Girls ($p< 0.01$, N=8) the calculated correlation coefficient values were significantly greater than the table correlation coefficient values in pre-test and in the post test it is lesser than the table correlation coefficient values. Hence the hypothesis was rejected and proved that there are significant Inter Hemispheric Coherence between Pre Test and post test scores of F3 vs. F4 before and after applying Cognitive Behavior Therapy.

Cognitive restructuring taken place during CBT differs from student to student and at the same time CBT keeps on refining day by day that will cause changes differently in students' brain and it has been proved in the present investigation through changes in their frontal asymmetric values, which is not uniform in all the cases. Moscovitch et al (2011) noted that higher activity in the left frontal brain pretreatment correlated with lower social anxiety and depression scores post treatment. These findings suggest that the effects of CBT can be measure biologically. The absence of activation in the dorsolateral prefrontal cortex (BA 10) and parahippocampal gyrus, after CBT,provides strong support to the view that CBT reduces phobic avoidance by deconditioning contextual fear learned at the level of the hippocampus/parahippocampal region, and by decreasing cognitive misattributions and catastrophic thinking at the level of the prefrontal cortex (Gorman et al., 2000). This deconditioning process would prevent the reactivation of the traumatic memories by allowing the phobic subjects to modify their perception of the fear-evoking stimuli. Once this perception has been reframed, the phobogenic stimuli would not constitute a threat anymore. Such cognitive restructuring would render obsolete the activation of the brain regions previously associated with the phobic reaction. The lack of an association between EEG asymmetry and attention bias is also consistent with results obtained recently by Sutton, Davidson, and colleagues (1997), who suggest that attention bias on the emotion Stroop task may be related to alpha asymmetry in the anterior temporal region, rather than at the midfrontal sites .Considered collectively, these findings suggest that frontal EEG asymmetry is not related to cognitive bias in depression. Although it is possible that our sample size did not afford sufficient statistical power to detect a relation between frontal asymmetry and cognitive functioning, the fact that neither self-report nor behavioral measures of cognitive bias were found to be related to asymmetry raises questions concerning the association between frontal EEG asymmetry and cognitive vulnerability to depression. Moreover, given the finding in the present study that vulnerability to test anxiety may be associated with a loss of frontal EEG asymmetry, investigators in this area might profitably begin to examine the relation between frontal asymmetry and positive illusions (e.g. Taylor & Brown, 1988; Tomarken & Davidson, 1994).

V. EDUCATIONAL IMPLICATIONS

Overall, students who received the CBT treatment programme reported lower levels of anxiety at post treatment, although data analysis also confirmed this conclusion. The support for this explanation on the discrepancy between students and parent measures can be found in research findings suggesting that a low level of agreement between children' student's self-report and other informants, such as parents, is fairly typical.(Meehan BT et al ,2003, and Youngstrom,2000) This current review focuses specifically on the role of interventions based on the techniques of cognitive behavioral therapy (CBT) in secondary schools for preventing and reducing suicidality, depression and anxiety in students. A central aim of these interventions is to improve the social skills and problem-solving-abilities students require to cope with life and its many challenges, and thereby decreasing the likelihood of depressive symptoms developing when the student is faced with biological or environmental stressors. As these interventions can be delivered and implemented at different levels we chose to evaluate the relative effectiveness of the following approaches. Students with conduct disorders are more likely to be living in lone parent families, with parents who have no educational qualifications, in families where neither parents are employed, in low-income households or in social sector housing Meltzer, H et al, (2000). Obviously, this does not mean that students from these backgrounds are all conduct disordered, even if the proportion is larger than in other groups. As the name implies, this therapy addresses both cognition - thoughts, feelings, emotions - and behaviors, attempting to change dysfunctional ways of thinking or misguided thought patterns that often lead to dysfunctional and sometimes harmful behaviors. The therapy focuses on the present, current thought patterns, identifying distortions, and applying interventions that specifically target those errant thoughts. These interventions and techniques are problem-solving solutions, first guiding individuals in how to evaluate and modify beliefs, and then, how to change correspondingly unhealthy behaviors and interactions. Metaphorically, CBT interventions resemble a step-by-step "how-to" manual, giving instructions, and then empowering the students to follow the steps, observe how they feel, and report back to the therapist the successes or challenges encountered. Typically, this type of therapy is short-term and goal-oriented, with occasional "checkups" to gauge progress and help in correcting any missteps. Clinical psychology research is as important to the nation's health and well being as medical research. In the same way that medical scientists work to understand the prevention, genesis, and spread of various genetic and infectious diseases, scientists conduct rigorous psychological research studies to understand, prevent, and treat the human condition as it applies psychologically to individuals, couples, families, cultures, and diverse communities. However, cognitive-behavioral therapy is focused on "getting better" rather than "feeling better". So while we are encouraged that students improve their symptoms with cognitive-behavioral therapy, we are more interested in helping them with the underlying thoughts and core beliefs that caused their emotional distress, helping them rid themselves of problematic, inaccurate thoughts, and replacing them with thoughts that are healthy and accurate. An approach to therapy that emphasizes the pursuit of evidence on

which to base its theory and techniques, as well as encourages its students or clients to consider evidence before taking action; or an approach to therapy is supported by research findings, and those findings provide evidence that it is effective. Because cognitive behavioral therapists base their applications on evidence-based research, the amount of CBT research surpasses the amount of psychotherapeutic research in nearly all other areas. Beyond its efficacy in treating acute depression, it has prophylactic effects and is acceptable to various populations in a range of settings. Good theoretical accounts of the emergence of depression in adolescence are forthcoming; to date, however, attempts at primary prevention are unconvincing. Our understanding of factors contributing to positive outcomes is growing, allowing CBT to be tailored to individual client needs.

VI. CONCLUSIONS

In conclusion, the present results provide support for the association between abnormal frontal EEG asymmetry and vulnerability to anxiety and depression. Results of recent studies suggest that attachment models of anxiety and depression may provide an explanatory mechanism for the association between left frontal hypo activation and depression (cf. Dawson, 1994). Future investigations might continue to explore the relation between frontal EEG asymmetry and attachment, as well as the specificity of frontal EEG abnormality to anxiety and depression. Further examination of these areas will no doubt force a reconceptualisation of the nature of the associations among cognition, emotion, and psychopathology.

VII. FUTURE DIRECTIONS

As a sub-field of emotion and motivation, the study of frontal EEG asymmetry holds substantial promise. In terms of basic science, this sub-field promises to inform us regarding the fundamental properties of emotion, both in terms of how emotions occur and what properties they entail. In more applied settings, the possibility yet remains that frontal EEG asymmetry may serve as a useful liability marker for test anxiety. Regardless of the application of this measure, it is increasingly important that theoretical predictions surrounding frontal EEG asymmetry are put to more rigorous tests, especially those provided by testing explicit meditational and mode rational models. In the absence of such explicit tests, the field will remain a collection of studies merely suggestive of moderating and mediating influences, around which much exciting and potentially important speculation and theory is generated. It is possible to more explicitly test that speculation and theory with models that more adequately can support or refute such inferences. The results of these explicit tests will guide the field in investigating whether and how frontal EEG asymmetry may serve as a risk marker for psychopathology, and in investigating what underlying physiological systems influence and are influenced by emotion. Gaining a deeper understanding of the fundamental properties of emotion will require clear thinking in terms of how the various components of emotion and emotional experience are related. It is in this spirit that the present remarks are offered.

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Indepth Interview of the Status of Widows in Rural Areas of Pondicherry

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***Abstract*- INTRODUCTION:** In India, National Family Health Survey: (NFHS - 2) data has given an estimate that there are more than 33 million widows comprising of 8% of the total female population.1 The widows are surrounded by cultural and social practices that seldom she can come out from them. **OBJECTIVES:** To determine the social, cultural deprivation faced by them after widowhood and to describe their attitudes towards widow's remarriage. **RESULTS:** Widows opined that younger widows without children can remarry. Widows said their relatives treated them badly and were blamed for the death of the husbands. The widows felt shocked and angered towards the last rites. Majority of widows did not have bank account. Majority of the older widows opined that getting rice for food itself was difficult, but for the younger widows educating their children was difficult to be met. Widows who had attained widowhood after 50 years opined that they expected the government to supply rice. Younger widows emphasized the government should give some loan for their survival. Conclusion: Majority of the widows were socially and culturally deprived.

***Index Terms*-** Widows, In-depth interview, Remarriage, Pondicherry

I. INTRODUCTION

In India, National Family Health Survey: (NFHS - 2) data has given an estimate that there are more than 33 million widows comprising of 8% of the total female population. According to Census 1991, more than 65% of widows in the age group of 60 and above were in the states of Goa, Assam, Karnataka, West Bengal, Andra Pradesh and Tamil Nadu. [1]

II. DESCRIPTION OF THE PROBLEM

The death of a woman's husband marks her transition from a wife to a widow. In the Hindu society of India and Nepal, a widow was physically alive but socially dead. The widows are marginalized in terms of property rights, inheritance and custody of her children. The situations are similar in almost all the South Asian countries. [2]

The widows are surrounded by cultural and social practices that seldom she can come out from them. The widows are tolerated but never welcomed in the society. Upon widowhood, most widows in rural India are subject to economic decline, social isolation and related deprivation. In spite, of the financial assistance that is available for widows who remarry, the society is not positive even today towards widow's remarriage. Due to growing globalization, the social changes which occur are fast but few areas which remain unchanged or where the change is slow is marriage and related customs, especially widow's remarriage.

OBJECTIVES:

1. To determine the social and cultural deprivation faced by them after widowhood.
2. To describe the attitudes and awareness to social issues like widow's remarriage.
3. To determine the awareness of benefits provided to them by the government and suggestions for improvement.

III. MATERIALS AND METHODS

The study was carried out in the service area of Jawaharlal Institute Rural Health Center (JIRHC) which the rural field practice area of Department of Preventive and Social Medicine, Jawaharlal Institute of Postgraduate Medical Education and Research Center (JIPMER), Pondicherry.

Parameters studied:

1. **Attitude towards remarriage**
2. **Reasons for unwillingness to remarry**
3. **Attitude of relatives towards widows**
4. **Reaction of society towards widowhood**
5. **Children's Negative Behaviour/Instances Of Change/Reasons**
6. **Attitude Towards Last Rites**
7. **Bank Account/Ration Card**
8. **Awareness Of The Benefits And Its Satisfaction/ Problems In Getting The Benefits**
9. **Expenditure Difficult To Meet**
10. **Expectation From The Government For Their Welfare**
11. **Disturbed By Other Males**

Brief Procedure:

The study was conducted in the two villages of Jawaharlal Institute Rural Health Center (JIRHC) viz., Ramanathapuram and Pilliarkuppam. In-depth interview was conducted among ten of the widows regarding widow's remarriage, social deprivation in detail and their satisfaction of the incentives provided to them by the government. The interview schedule was pretested other village outside the service area. After making a few modifications based on the responses obtained, the interview

schedule was finalized. The information collected was recorded in their own language using a recorder. All the data was collected by the chief investigator under the supervision of co-investigator.

IV. RESULTS

INDEPTH-INTERVIEW

ATTITUDE TOWARD REMARRIAGE

OVERALL

Overall irrespective of the age of attainment of widowhood the widows had similar views regarding remarriage only younger widows without children who would fall within the preview of the societal norms. They were of the opinion that younger widows could think of remarriage, of course, with the support of their parents. Otherwise living without remarriage was the societal norm.

REASONS FOR UNWILLINGNESS TO REMARRY

OVERALL

Overall, all the widows gave the similar reason for not getting married again. But still the burden was greater with the younger widows who had to bring their children by educating them and getting them married. Two of the respondents also added that since they faced lot of difficulties in their first marriage they didn't want to marry again.

ATTITUDE OF THE RELATIVES TOWARDS WIDOWS

OVERALL

Overall, majority of the widows who had attained widowhood after 50 years said that although their relatives treated them badly they did not consider it as it was within the societal norms. However, younger widows opined that they felt sad and depressed the way the relatives treated them badly.

REACTION OF THE SOCIETY TOWARDS WIDOWHOOD

OVERALL

Overall, irrespective of age the widows opined that they were treated badly by the society but the emphasis was more with the younger widows whom some, were blamed for the death of the husbands.

CHILDREN'S NEGATIVE BEHAVIOUR/ INSTANCES OF CHANGE/REASONS

OVERALL

Overall, majority of widows did not find any change in their children's behaviour. But one of the younger widows said that her son got married on his own, as there was no male member to check his activities.

ATTITUDE TOWARDS LAST RITES

OVERALL

Overall, the widows irrespective of the age felt shocked and angered about the last rites but they felt that it was necessary as it was ongoing in a society. But one of the younger widows said that it was not necessary.

BANK ACCOUNT/RATION CARD

OVERALL

Overall, majority of widows did not have bank account. But all of them had a ration card.

AWARENESS OF THE BENEFITS AND ITS SATISFACTION/ PROBLEMS IN GETTING THE BENEFITS

OVERALL

Overall, all the widows perceived that the government did not do anything specific for the widows. Except for a few respondents, they did not get Rs. 10,000/-. Two of the respondents opined that they did not get pension also in spite of the submission of the application long ago.

EXPENDITURE DIFFICULT TO MEET

OVERALL

Overall, there was a difference between the older and younger widows as majority of the older widows opined that getting rice for food itself was difficult, but for the younger widows educating their children and to meet other expenses was difficult.

EXPECTATION FROM THE GOVERNMENT FOR THEIR WELFARE

OVERALL

Overall, there were differences in the expectations from the government towards the welfare of widows. Widows who had attained widowhood after 50 years opined that they expected the government to supply rice similar to handicapped persons. But younger emphasized the government should give some loan for their survival.

DISTURBED BY OTHER MALES

OVERALL

Majority of the widows did not face problem from other males, however few of the younger widows had faced such a problem. They need some protection.

V. DISCUSSION

In-depth interview with the widows found overall irrespective of the age of attainment of widowhood the widows had similar views regarding remarriage giving emphasis that only younger widows without children can remarry which would fall within the preview of the societal norms. Mala Bhandari in his study from Vrindavan and Varanasi observed that 88.75%. 92.5% and 78.75% categorically stated that widows should not remarry. In Varanasi the corresponding figures are 100%, 89.29% and 100% respectively. Very few of them quoted emotional reasons for this (i.e. unable to forget the first husband) 11.25%, 81.25%, 55% in Vrindavan and 53.57%, 42.86%, 92.86% in Varanasi stated that they hate the idea of remarriage.12.5%, 8.75%, 22.5% in Vrindavan and 28.58%, 42.86%, 92.86% in Varanasi cite religion as opposed to the idea of remarriage. [3]

The widows were not respected and visited by their family members was lower after widowhood. Overall, majority of the widows who had attained widowhood after 50 years said that although their relatives treated them badly they did not consider it as it was within the societal norms. However younger widows

opined that they felt sad and depressed the way the relatives treated them. One would think that sympathy for a widowed mother would bring a positive change in the behaviour. But the study shows that widows fall even lower on the social scale. This reiterates the fact that widows suffer doubly, as widows and as women. The community based study from Vrindavan and Varanasi found that 8.75% indicating bad behavior towards mother, it rose to 11.25% when the mother becomes a widow. Similarly indifferent behavior rise from 61.25% to 66.25%. Interestingly though 30% claimed to have been treated well by their children before widowhood, this figure fell to 22.5% on widowhood. The same trend was seen among the respondents living in boarding houses and on the streets. Of the respondents in the rehabilitation homes, 15% claimed that no respect had ever been shown to them as wives. This figure increases to 21.25% on widowhood. Similarly 38.75% claimed that they had evoked respect as wives. But the number falls to 32.75% on widowhood. This trend is seen among the widows living in boarding houses and on the streets. Their social isolation were indicated by the fact that 36.25%, 78.75% and 75% of the respondents of Vrindavan in the three sections claimed that relatives and friends never visited them. The corresponding figures for Varanasi were 17.86%, 32.14% and 67.86%. [3] Also the other studies consistently showed that the frequency of interaction with friends and emotional support from friends had a positive contribution to the well-being of older adults. [4]

The high figures are a good indication of the influence of social conditioning. It is also interesting to note that changes in life style was not a voluntary choice of most of the respondents, but something that becomes an involuntary decision, motivated more often than not by economic reasons. Social insecurity is a consequence of economic disadvantages; from childhood she is conditioned to think that her very purpose of wearing ornaments or dressing up is for the husband. Along with the economic reasons is also the social conditioning. Not only is her social status decimated but she is also considered inauspicious. On the other hand since she has no social status, she is not welcomed at social functions. The widows are conditioned by society to think that they are inauspicious and will bring ill luck to the newly married couple. Studies from Vrindavan and Varanasi found that women who changed to simple food were 66.25%, 83.75% and 92.5% in all the three sections of Vrindavan. In Varanasi 100% in all the sections stated the same. A majority 96.25% in rehabilitation homes, 98.75% in boarding houses and 97.5% of those on the streets indicated economic reasons for this change in diet. Every respondent in Varanasi reiterated this. Only 1.25% respondents in rehabilitation homes of Vrindavan stated that they started wearing simple clothes because of lack of interest. The corresponding figures in boarding houses and on the streets are 3.75% and 5% respectively. But as high as 96.25%, 96.25% and 92.5% in all the three sections cited economic disadvantages for the change in dressing. But in Varanasi all 100% in all the three sections attributed economic reasons for change in dressing habits. 85.5%, 96.25% and 80% in all three sections of Vrindavan did not wear bangles, flowers or the mangalsutra. In Varanasi, the corresponding figures are 100%, 89.29% and 96.43%. In Vrindavan while 98.75% in rehabilitation homes attended religious functions, the number fell to 85% in the case of social functions. The corresponding figures for those on the street are 85% to 76.25%. In contrast in Varanasi 100% of the respondents attended both religious and social functions. 85% of those living in rehabilitation homes of Vrindavan attended marriage ceremonies, but only 58.75% went near the bridal couple. The corresponding figures for those in boarding houses are 12.5% only attended marriage ceremonies, but not one went near the bridal couple. In Varanasi of the 85.71% of respondents in rehabilitation homes who attended marriage ceremonies not one went near the bridal couple. Of the 96.43% of respondents on the streets who attended marriage ceremonies, only 11.35% went near the bridal couple. 38.75%, 100% and 86.25% in all the three sections of Vrindavan were not allowed to enter the marriage hall. The corresponding figures for Varanasi are 85.71%, 100% and 92.86% in all three sections. [3]

Younger widows opined they were blamed for the death of their husband. Mala Bhandari in his study from Vrindavan and Varanasi observed that 3.75% of widows from rehabilitation homes, 1.25% from boarding homes and 2.5% of widows from the streets were blamed by their in-laws. The relatives blamed them in 7.14%, 3.57% in other two sectors of Varanasi. However 14.29% of the widows from Varanasi said that they were blamed by their parents itself. [3]

Despite the fact the women had been left to fend for themselves, there was no change in their children behaviour after the death of their husband. Studies from Vrindavan and Varanasi found that 50% in boarding homes and 40% on the streets in Vrindavan stated that their children had not turned negative towards them. Widows in Varanasi who had suffered negative behavior from their children were 17.86%, 10.71% and 28.57% in all three sections. [3]

According to a study from Vrindavan and Varanasi found that 26.25% in boarding homes and 6.25% on the streets in Vrindavan stated that the reason was not able to check their activities without the male member. About 17.86% of the widows in rehabilitation home in Varanasi who had suffered negative behavior from their children attributed it to bad company of their children. 3.57% of the widows in boarding home said that the reason of change in their children behaviour was without the male member they were unable to check their activities. [3] According to Chen et al in his study among elders in Beijing found that the children had generally positive contribution to the well-being of older parents. [5]

In the current study it was found that all the widows in the study said that felt insecure when last rites were performed. In in-depth interview also it was found that overall, the widows irrespective of the age felt shocked and angered about the last rites but they felt that it was necessary as it is ongoing in a society. According to a study from Vrindavan and Varanasi found that most of the respondents interviewed were particular on their last rites being conducted in a befitting manner. About 76.25% of those in rehabilitation homes were confident that the last rites were conducted properly. In contrast, 97.55% of those in boarding houses and 92.55% of those on the streets are faced with a sense of insecurity. In Varanasi almost all the respondents interviewed were not sure whether their last rites were conducted properly. [3]

Majority of widows had no savings left by their husband. Studies from Vrindavan and Varanasi found that 66.25% and 73.75% of those in rehabilitation homes and boarding houses had

been left with some savings by their late husbands. In sharp contrast, 73.75% of those on the streets had no savings at all. In Varanasi the figures were 85.71%, 100% and 100% in all three sections had no savings left for them by their spouses. [3]

Majority of them had no bank account. That means the widows have not made an attempt to provide some security for themselves. This indicates a level of enterprise that can be utilized for empowering them further by introducing the banking concept to the widows. A Study from Vrindavan and Varanasi found that 66.25% and 73.75 % of those residing in rehabilitation homes and in boarding houses of Vrindavan had bank account. The corresponding figures for Varanasi were 35.71% and 60.71%. On the other hand 73.75% of those on the streets had no bank accounts. But in Varanasi only 14.28% were without bank accounts. Also it should be noted that while only 33.75% of the women in the rehabilitation homes of Vrindavan had no bank accounts, the corresponding figure for Varanasi are 64.29%. [3]

It was found that all the widows had a ration card. But none of them said that they receive old age pension. Among the respondents in rehabilitation homes of Vrindavan, 62.5% were holding ration cards. Again in boarding houses of Vrindavan 81.25% held ration cards. In Varanasi of the respondents in boarding houses, 50% were holding ration card. [3] According to a case study done in Delhi 48.24% of the widows were knowledgeable about Sharia Laws that provided them the right to inherit parent's and husband's property. But in reality women did not claim share from such property. [5]

In-depth interview it was observed that overall, there was a difference between the older and younger widows as majority of the older widows opined that getting rice for food itself was difficult, but for the younger widows educating their children and to meet other expenses was difficult. According to a study from Vrindavan and Varanasi found that 50%, 31.25% of the widows in rehabilitation homes and boarding homes felt that their children marriage was most difficult to meet. About 23.75% of the widows in the streets said that food itself was difficult to manage. The most difficult expense for the widows of Varanasi was their children marriage in 57.14%, 46.43% and 60.17% in three sectors respectively. [3]

In in-depth interview, it was found that overall all the widows perceived that the government did not do anything specific for the widows. Except for a few respondents, others did not get Rs. 10,000/-. Two of the respondents opined that they did not get widow pension also in spite of the submission of the application long ago.

In in-depth interview it was observed that there were differences in the expectations from the government towards the welfare of widows. Widows who had attained widowhood after 50 years opined that they expected the government to supply rice. But younger emphasized that the government should give them some loan for their survival. Older widows seemed to worry over their next meal. These responses are an eye opener as it is need based. But they should also think in terms of larger issues, like advocating a change in society's attitude towards widows. Certainly an attempt will have to be made to stop them from being mere beneficiaries. Efforts will have to be made to make them independent and work for the dignity of a decent living and existence. The community based study from Vrindavan and Varanasi found that 58.75% of those in rehabilitation homes in Vrindavan felt that social mindset should change. In sharp contrast 95% of those in boarding homes and 65% of those on the streets felt that eligibility of widows for pension was of paramount importance. For them money was urgent and immediate need, since they were always leading a hand to mouth rehabilitation homes would like jobs to made available to them. But 93.75% of those in boarding homes and 81.25% of those on the streets emphasized on the need for pension benefits. [3]

VI. CONCLUSION

The widows are socially, economically and culturally deprived. As per the societal norms they are against widow's remarriage. They need more help from the government to support their living

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Comprehensive Study on the Profession and the Prevalence of Antisperm Antibodies among Infertile Male Population

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Abstract- A total number of 402 infertile males aged between 25 and 50 were studied for the presence of antisperm antibodies (ASA) in their blood. They were grouped together basing on their occupation and the nature of their work, which included drivers, business men, gold smith Industrial workers and teachers. Prevalence of ASA among these groups was studied and the results were analyzed. It was observed that the formation of ASA was associated with the temperament of the individual and the environment of occupation. Stress, tension pollution reduced physical activity and occupational temperature ware directly proportional. Stress associated personal habits like smoking and alcoholism were found to be related factors. Further 85% of the infertile men with obesity showed ASA positivity.

Index Terms- Antisperm antibodies, Infertility

I. Introduction

Infertility was observed from ancient days1 the intensity is raising steadily day by day2. In recent years it reaches to an alarming level3. Various etiological agents such as microbial infections, immunological interference, physical and physiological problems may contribute for this4 family destructive process.

Though there were many causes for this condition one important factor being the immunological involvement5.The known mechanism noted was the sequestrated nature of sperm antigens when contacted with the immunocompetent cells of the individual began to produce the antisperm antibodies. This autoimmunity leads to the destruction of the sperm cells and its functions.

The process of formation of ASA was favored in certain clinical and surgical procedures6. Apart from this it was suspected that there were other factors such as the tentionable jobs, and the environment in which the individual work may also influenced the ASA formation7. Hence this study was designed to explore the relationship of occupation and environmental influence in the formation of antisperm antibodies.

II. Materials and Methods

Materials

A total number of 402 infertile male patients who were attending infertility clinic of Rajah Muthiah Medical College Hospital Annamalai Nagar Chidambaram were used as study population. Blood samples were collected from patients of 25-50 years of age groups. The study was conducted during the period between December 2008 and March 2012. Clinical and familial history along with other personal details like profession, habits, smoking and consuming alcohol were also recorded.

Methods

Sera were separated from the blood samples and were tested by Enzyme immuno assay for the presence of antisperm antibodies. The results are tabulated. Profession wise distribution of ASA positivity was calculated. Personal data collected from the infertile men were analyzed the results were recorded as against the ASA positivity.

III. Results

Age wise distribution of infertile cases showed more numbers (175/402) in the age group of 36-40 years. Next higher number was seen (88/402) in the age group of 31-35 years. Least numbers of infertile cases were observed in the age group of 25-30 years shown in Table-1.

Among the 402 infertile males 56 were positive for anti sperm antibody (13.7%) shown in diagram-1. Analysis of ASA in respect to the profession of the infertile male cases showed a highest incidence in gold smiths 20/56 (35.7%) this was followed by the driver profession 18/56 (32.1%). Results in other groups such as business men 9/56 (16%) and industrial workers 4/56 (7.1%). School teachers were also showed ASA positivity but to a fairly low level of 8.9%.

Due to stress and strain in their work spot some people use to smoke use Jardha or Bettle Nut and some people consume alcohol regularly. Among the 56 positive cases 26 of them were smokers, 22 of them used Jardha and 8 of them were alcoholics. In case of obesity among the 56 ASA positive cases 48 of them were obese only 8 had normal weight.

Table: 1 Age wise distribution of ASA in infertile men

Age groups in years	No. of patients	Percentage
25-30	8	1.99
31-35	88	21.89
36-45	175	43.53
41-45	74	18.40
46-50	57	14.19
Total	402	100

Fig. 1Prevalence of ASA positivity

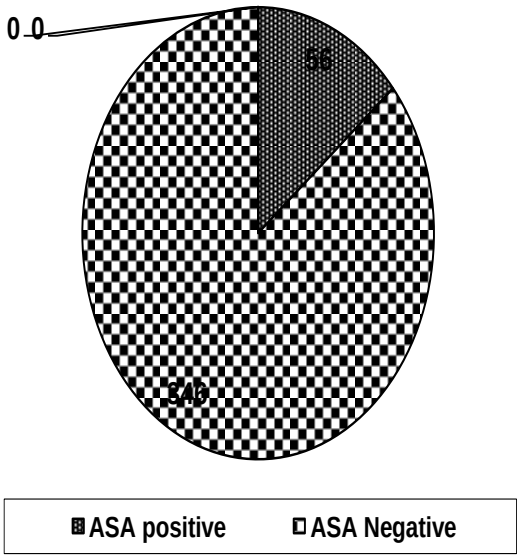
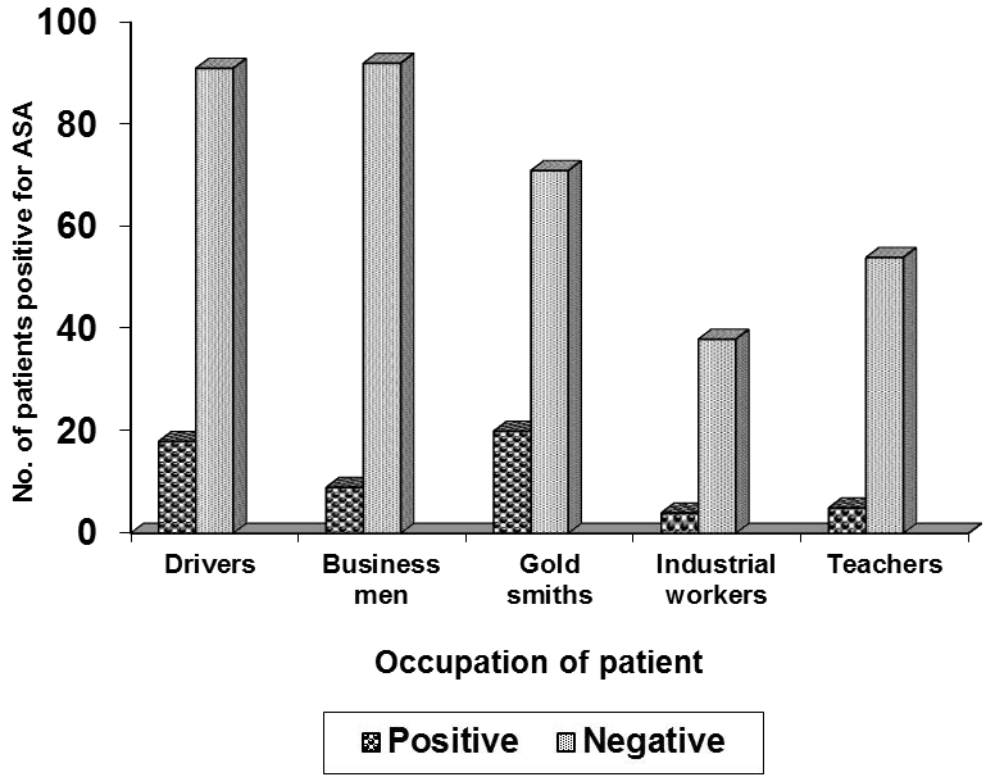


Table: 2 Occupation and ASA positivity

Occupation	ASA positive	ASA negative	Total
Drivers	18	91	109
Business men	9	92	101
Gold smiths	20	71	91
Industrial workers	4	38	42
Teachers	5	54	59
Total	56	346	402

Fig. 2



IV. Discussion

Twenty present of all cases where the male is the only contributing factor to infertility can be corrected positively by life style8. Sperm cells are protected from immune system by a natural protective mechanism called the blood-testes barrier9. If by any means breaches this barrier then the immune system has access to sperm the sperm behave as an antigen and antibodies are formed10.ASA positivity in fertile male was seen in 56 cases out of 402 tested. Among the various occupational groups tested gold smiths were top in the list with 35.7% ASA positivity. This could be explained that the gold smith who sat all the time near the flame while at work. Goldsmiths referred in the present study are exposed not only to radiant heat but also to potentially toxic fumes11.

Similarity the driver community was also closer to the hot engine during their long duty time. Exposure to heat may lead to low sperm count, reduced motility and variant morphology12. Heat intense occupations such as bakers, cooks welders and drivers are potentially exposed to elevated radiant heat and this may impair spermatogenesis or possibly epididymal function13.

When speaking about business men (shop keepers) they had no active work and were held at one spot without having much of physical movements showed 16% of ASA positivity. ASA positivity was also observed among the industrial worker (7%) though they were active workers the industrial pollution may account for the infertility. Protection of the workers through provisions of personal protective will go a long way in reducing the exposure to industrial pollution of the workers and preserve their reproductive ability.

Occupational environment may also affect significantly to a greater number of people. Toxicant encountered in specific occupational settings as they come in to contact with chemicals in a variety of work including direct exposure, industrial emissions of pesticides and their residues, ingested foods or contaminated water14.

There was significant of delay in conception following male exposure to textile dyes, plastic manufacturing and welding. Similarly men who involved in unpacking or handling of antibiotics had a significant association with delayed pregnancy15. Other groups in the community among the educated population affected was the school teachers particularly teaching 10th and-12th Standards showed ASA positivity of (8.9%).The stress and strain they undergo on every year due to public examination may account for their mounting tension in life. Further the processing of personal data collected by questioner of their habits revealed that smoking and other forms of nerve stimulating agents, and the use of alcohol seem to be the aggravating factors in the production of ASA.

Smokers sperm count are an average of 13-17% lower than non smokers. A study on smokers who were followed for 5 to 16 months after stopping reported that their sperm counts rose 50 to 800% suggesting that toxic chemicals in the smoke are responsible and any reduction in the sperm count is reversible16. Male smokers have an increase in sperm abnormalities there by suggesting a mutagenic effect17. Cigarette smoking may alter hormone levels in males18.

Interestingly limited studies suggested that to be mutagenic to human spermatozoa and lead to cancer, birth defect and genetic disease in offspring19. Excessive alcohol intake is associated with direct testicular toxicity20. A large 50% reduction in conception was found in experiments. Test animals were given intoxication doses of alcohol 24 hours prior to mating. Alcohol reduces fertilization success21.

Different studies reported that obesity was significantly associated with erectile dysfunction. It was found that heavier a man was lesser the testosterone he produced and the lack of their essential male hormone intern it caused erectile dysfunction. The low level of testosterone is also associated with abnormalities in blood flow to the penis22.

Further obesity in males was recently proposed to constitute an important role, approximately 30-40%infertility cases can be attributed to problems with male partners23. Present study results also confirm the previous findings. A maximum number of infertile cases were obese. In men of erectile dysfunction 79 % of the patients are found to be more weight or obese. This can be explained in part by elevated levels of several pro inflammatory cytokines in obese individuals24.

Report by Hammound revealed that obesity is associated with altered spermatogenesis and erectile dysfunctions25. The present study report showed among the ASA positive men 85.7% were obese. Therefore the obesity was found as an important risk factor in infertility.

V. Conclusion

The study report revealed that the level of ASA was highest in gold smith population. It was followed by the driver community. It clearly showed that the environment may play a role in infertility and also the production of ASA. Pollution free surroundings, healthy physical activity and peaceful working atmosphere are essential for normal fertility. Obesity was the most important risk factor observed in the infertile population.

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Estimation of confidence intervals for Multinomial proportions of sparse contingency tables using Bayesian methods

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Abstract **-** Multinomial distribution, widely used in applications with discrete data, witnessed varieties of competing intervals from frequentist to Bayesian methods, still prove to be interesting in the case of zero counts or sparse contingency tables. The methods commonly recommended in both approaches are considered based on its influence of zero counts, polarizing cell counts, and aberrations. The inference based on comparative study shows that Bayesian approach, with an appropriate prior could be a good choice in dealing with a sparse data set without any imputation for zero values.

Index Terms**:** Bayesian inference, Coverage probabilities, Dirichlet distributions, Multinomial distributions, sparse data.

I. Introduction

The cell of contingency table contains frequency of outcomes of categorical response variables and its number denotes the dimension and size is determined by number of categories related to each of the variable. Generally, inferential methods for categorical data assume multinomial or Poisson sampling models. The observed counts {ni; i=1,2,…,k} could be considered as k levels of a single categorical variable or for k=IJ cells of a two way categorical variables with levels I and J. Agresti (1992) has explained the different sampling k models and in particular, the present work is based on the multinomial distribution $(N,\{\pi_1,\pi_2,\ldots,\pi_k\})$. Maximum likelihood estimates (MLE) of cell probabilities can be derived easily as sample cell proportions but interval estimation of multinomial probabilities too has drawn then active attention.

The impact of sparseness provides an ample scope to have a comparative study among these methods as well as Bayesian procedures. Agresti and Yana (1987) have stated that the asymptotic approximations may be quite poor for sparse table, even for a large N. Further Szyda et al (2008) observed that sparseness could occur even when k is relatively large. Subbiah and Srinivasan (2008) have studied the nature of sparseness in a 2x2 table based on a summary measure. Also, recent developments have favored Bayesian approaches as more suitable methods to handle sparseness as compared to three standard recommendations while handling sparse or zero counts (Agresti, 1992, Subbiah et al 2008).

The objective of this paper is to draw comparisons that include Bayesian approach with non-informative priors for underlying parameters. Study envisages use of typical 2x2 data sets in the literature and a large contingency table (Szyda etal, 2008). Frequentist property of coverage probabilities for Bayesian approach have also been studied and compared with the available results of classical approaches using recent computational tools. The following section provides a comprehensive list of active methods in the literature considered for comparison of confidence intervals for multinomial proportions.

II. Confidence intervals for Multinomial Proportions

In the case of Bayesian inference, Dirichlet distribution is the widely used and recommended conjugate prior distribution for the multinomial probability parameters (Gelman et al, 2000). However, to obtain posterior distribution, a relationship between Gamma distribution and Dirichlet distribution has been used and presented as

$$Y_i \sim \text{Gamma}(\alpha_i,1)$$
$$V=\sum_{i=1}^{k} Y_i \sim \text{Gamma}(\alpha_0,1) \text{ where } \alpha_0=\sum_{i=1}^{k}\alpha_i$$
$$\text{Then } \left(\frac{Y_1}{V},\frac{Y_2}{V},\ldots,\frac{Y_k}{V}\right) \sim \text{Dirichlet}(\alpha_1,\alpha_2,\ldots,\alpha_k).$$

With a proper choice of hyper parameters $\{\alpha_i\}$ a complete Bayesian scheme can be implemented. However, recent advances in the Monte Carlo simulations, posterior summaries can directly be obtained from simulating Dirichlet distribution. The typical scheme (MD) is

$$n_i \sim \text{Multinomial}(N,\{\pi_1,\pi_2,\ldots,\pi_k\})$$

$$\pi \sim \text{Dirichlet}(\theta_1,\theta_2,\ldots,\theta_k) \text{ so that}$$
$$\pi|\theta \sim \text{Dirichlet}(\theta_1+n_1,\ldots,\theta_k+n_k)$$

Setting $\theta_j = 1\ (j = 1,2,\ldots,k)$ will yield a uniform density and Tuyl et al (2009) have favoured this choice as a better non informative prior for $\{\theta_j\}_{j=1}^{k}$

Further, a simulation study has been carried out to compare the performance of the intervals in terms of repeated experiments. Bayesian estimates obtained from incorporating objective priors might require such a test based on frequentist approach. Agresti

and Min (2005) have attempted this in evaluating the Bayesian confidence intervals for binomial proportions. The corresponding procedure for multinomial proportions includes following steps

1) Consider any data set with cell count $\{n_1,n_2,\ldots\ldots n_k\}$
2) Compute its MLE p = {ni/N} and assume p as population parameter
3) Simulate Multinomial(N, p) for L times
4) Obtain confidence interval using the required methods
5) Coverage Probability = (Number of intervals in (iv) that include p) / L

Similar attempts have been made for classical approaches or Bootstrap intervals in literature that are cited earlier in this paper. This work includes Bayesian methods by considering contingency tables with non-zero but low counts and has an appreciable distance between the counts. However for comparison purpose other standard procedures such as QH-Quesenberry and Hurst (1964), GM-Goodman (1965), FS-Fitzpatrick and Scott (1987), SG- Sison and Glaz (1995) and methods due to central limit theorem (CLT) and its continuity corrected version (CLT-CC) have also been considered.

III. Motivating Data Sets

If X and Y denote two categorical response variables, X with I categories and Y with J categories leading to k = IJ possible combinations that can be represented in a contingency or cross-classification table with cells contain frequency counts of outcomes for a sample. As a case of a hypothetical example, suppose that a clinical trial is undertaken to compare the effect of a new drug or other therapy with the current standard drug or therapy. Ignoring side effects and other complications, the response for each patient is assumed to be simply "success" or "failure." For a single stand-alone experiment, the observed data can be shown in the following table:

Table 1: Hypothetical responses in one segment of a clinical trial

	Response Success	Failure	Total
Treatment	a	b	m
Control	b	d	n
Total	r	s	N

Sparse tables often contain cells having zero counts and such cells are called empty cells. Contingency tables are referred to as sparse when many cells have small frequencies besides some of them being zeros too. It is extremely important to describe the location of zero cells in the 2 x 2 table, as the same is also crucial in studying the nature of sparsity and could affect the analysis. Sparsity is not restricted to the tables with smaller sample sizes alone but could also occur with large sample size due to high concentration of frequencies in certain cells and poor or none in other cells. The impact of sparsity is felt in estimation of summary measures like odds ratio, computational complexity and asymptotic approximations. Even for large contingency tables, due to the small sample size and the resulting sparseness of the data table, the asymptotic distributions of the tests may not be relied in hypothesis testing (Szyda et al 2008).

The characteristics of the data sets (referred to as I to X) collected from the published literature with 2 x 2 tables are summarized to provide the length and breadth of the sparsity in the data sets. Table 2 provides the details of source and distribution of zero cells. Apart from zero cells, proportion of non-zero cell counts with frequency less than six is also described, so that the sparse nature of the data sets are completely described. Also, to understand the spread of counts in individual tables minimum and maximum of range calculated for each table in a data set is presented. This value provides a quick view of polarization of counts; for example data set V shows a very high range so that cell counts are extremely different in their sizes. Zero minimum indicates equal cell counts in a data set (Kishore, 2007), whereas Efron (1996) has a table with zero in all the four cells. Also, based on Subbiah and Srinivasan (2008) nature of sparseness of each of these data sets has been classified to indicate the typical real time data variability among the collected literature and the results are summarized in the same table

Table 2: General description of the ten illustrative data sets

Data No	Source of data sets	Zero entries		Positive entries < 6		Range of table totals		No of tables with nature of Sparseness		
		No	%	No	%	Min	Max	Mild	Moderate	Severe
I	Kishore (2007)	5	18	4	14	0	17	3		
II	Agresti (1990)	7	35	9	45	5	6	5		
III	Smith et al (1995)	2	2	10	11	12	158	2		
IV	Sweeting et al (2004)	37	40	9	10	15	1128	19	2	
V	Sweeting et al (2004)	2	7	6	21	688	66153		1	1
VI	Efron (1996)	16	10	43	26	0	48	3	2	1
VII	Tian et al (2007)	48	25	45	23	25	2852	27	11	3
VIII	Tian et al (2007)	67	35	27	14	25	2852	30	9	3
IX	Cochran (1954)	4	25	3	19	17	40	1	3	
X	Warn et al (2002)	17	9	18	10	7	177	15	2	

Apart from these ten 2 x 2 tables, another contingency table (Szyda et al 2008) has been considered whose size is 4 x 5 of which 12 cells (60%) are zero where as minimum and maximum among remaining non-zero cells are 5 and 66 respectively. This data illustrates the presence of more zeros and extreme non-zero counts with high range even in a large size tables. These observations among many such real time studies provide a notion for comparative study using relevant characteristics which are prevalent in data sets summarized in contingency tables.

IV. Results

Bayesian data analysis can be referred to posterior inference given a fixed model and data and computation has been carried out in WinBUGS and R. However, sufficient search indicates non availability of classical methods in open sources and these methods are implemented using Macros in EXCEL except SG which is obtained through SAS.

Results from the computations include lower and upper limits of 95% confidence intervals calculated from the closed form classical methods. 2.5 and 97.5 percentiles from posterior samples are used to obtain lower and upper limits of Bayesian confidence intervals after a run of 50000 single MCMC chain with burn-in of initial 50% and convergence has also been monitored using kernel density. However, Table 3 provides results from one data set as an illustrative case and subsequently observations from the comparative analysis have been presented. This data set has as many characteristic as desired in explaining the performance of these procedures; especially, under sparseness, low non-zero counts and the impact on corresponding results.

The comparisons are based mainly on length of intervals (shorter or wider), aberrations; many studies have considered coverage probability as a tool for comparing performance of intervals. However, very limited or no studies have included Bayesian method in this comparison and this study has considered Bayesian MD procedure and compare with existing results. The data characteristics such as sparseness in terms of presence of zeros and low cell counts range of cell counts in a table and size of the table. Though computation tools become abundant in the present scenario, these procedures require a keen attention in the availability to the user community.

Table 3: Comparison of seven simultaneous confidence interval procedures with α = 0.05 for five different 2 x 2 tables

QH							
LL	UL	LL	UL	LL	UL	LL	UL
0.000	0.415	0.205	0.848	0.000	0.415	0.152	0.795
0.059	0.638	0.059	0.638	0.000	0.394	0.186	0.814
0.186	0.814	0.000	0.394	0.030	0.567	0.096	0.702
0.138	0.761	0.009	0.487	0.186	0.814	0.000	0.394
0.052	0.746	0.000	0.527	0.254	0.948	0.000	0.527
GM							
LL	UL	LL	UL	LL	UL	LL	UL
0.000	0.362	0.229	0.829	0.000	0.362	0.171	0.771
0.068	0.603	0.068	0.603	0.000	0.342	0.208	0.792
0.208	0.792	0.000	0.342	0.035	0.527	0.109	0.672
0.155	0.735	0.010	0.441	0.208	0.792	0.000	0.342
0.061	0.713	0.000	0.471	0.287	0.939	0.000	0.471

CLT							
LL	UL	LL	UL	LL	UL	LL	UL
0.000	0.000	0.170	0.920	0.000	0.000	0.080	0.830
0[a]	0.562	0[a]	0.562	0.000	0.000	0.139	0.861
0.139	0.861	0.000	0.000	0[a]	0.435	0[a]	0.673
0.061	0.772	0[a]	0.283	0.139	0.861	0.000	0.000
0.104	0.580	0.000	0.000	0.288	1.141	0.000	0.000
CLT-CC							
LL	UL	LL	UL	LL	UL	LL	UL
0[a]	0[a]	0.125	0.875	0[a]	0[a]	0.034	0.784
0[a]	0.521	0[a]	0.521	0[a]	0[a]	0.098	0.819
0.098	0.819	0[a]	0[a]	0[a]	0.394	0[a]	0.632
0.020	0.730	0[a]	0.241	0.098	0.819	0[a]	0[a]
0[a]	0.641	0[a]	0[a]	0.216	1[b]	0[a]	0[a]
SG							
LL	UL	LL	UL	LL	UL	LL	UL
0.000	0.346	0.364	0.892	0.000	0.346	0.273	0.801
0.000	0.527	0.000	0.527	0.000	0.277	0.250	0.777
0.250	0.783	0.000	0.283	0.000	0.450	0.083	0.616
0.167	0.692	0.000	0.359	0.250	0.776	0.000	0.276
0.143	0.684	0.000	0.398	0.571	1.000	0.000	0.398
FS							
LL	UL	LL	UL	LL	UL	LL	UL
0[a]	0.089	0.456	0.635	0[a]	0.089	0.365	0.544
0.168	0.332	0.168	0.332	0[a]	0.082	0.418	0.582
0.418	0.582	0[a]	0.082	0.085	0.248	0.252	0.415
0.335	0.498	0.002	0.165	0.418	0.582	0[a]	0.082
0.146	0.426	0[a]	0.140	0.574	0.854	0[a]	0.140
MD							
LL	UL	LL	UL	LL	UL	LL	UL
0.002	0.226	0.228	0.710	0.002	0.235	0.179	0.645
0.078	0.487	0.077	0.487	0.002	0.218	0.212	0.679
0.209	0.676	0.002	0.221	0.044	0.402	0.119	0.550
0.165	0.619	0.017	0.320	0.211	0.677	0.002	0.219
0.067	0.562	0.003	0.303	0.267	0.813	0.003	0.307

[a] Lower limit is less than zero [b] Upper limit is greater than one

In terms of length of intervals for the data sets (I to X), SG (63%) and QH (31%) yield wider intervals compared to other methods. SG has the maximum length in most of the cases where range of cell counts are markedly as high as 6821. Even in such polarized tables, only small count cells have this property and QH produces long intervals for other cells of corresponding tables. Data set IV, VI, VII, VIII, and few tables of X can be considered as illustrative cases which exhibit this observation where the distance among the individual counts are notably high and more tables are available with the presence of zeros in different position of four cells. This property is apparent in the data set III in which all cell counts are non-zero counts except in two cells among the total of 88 cells (22 tables).

In the data set V, possibly a rare table with an extreme characteristic in that cell counts are too wider (10, 45870, 40, 66163) is available. MD provides a wider interval only in this case for the low counts and QH has shared this for other larger values and except this case, MD has not shown this property in any of the other tables considered for the comparisons that are presented here and other data sets with which this study has made extended comparisons.

While considering other methods, no case has an interval with maximum length due to FS. However, two methods based on CLT share this property in almost all cases in similar cases though CLT-CC yield wider interval in slightly more cases. However, these methods possess a feature in that for zero cells they provide intervals of zero length which is due to the presence of sample proportion in their mathematical form. But it has been observed that for tables with all low counts so that total is also marginally low, wider intervals could be due to CLT methods; a single table in data I and data set II that has uniformly low counts and at least one zero cell in all the tables illustrate this observation.

In the case of shorter intervals, FS dominates uniformly in all the four cells of all data sets considered for the study; 72%, 95%, 60% and 94% of occasions are the supportive numerals for this property. In each case, CLT methods immediately succeed FS in this property but this may be due to its feature already mentioned and hence could be avoided from comparison. Surprisingly SG yield shorter intervals in two tables of the data set V where counts are extremely varying in nature (range 8632 and 66153). No other methods exhibit this property in any of cases considered for the comparative study.

It is observed that aberrations exist in three procedures due to CLT based methods and FS. But those cells cannot be identified with any particular characteristic of a cell like zero count. In the case of zero counts these methods will yield a degenerate case with lower and upper limits are same value. This feature is an obvious outcome of their mathematical forms. Also, a closer look of CLT indicates that the procedure will be resulted with a smoothing by the chi square value whenever cell counts are zero. This kind of smoothing would encourage the recommendation of Bayesian procedure as observed in Agresti (1992). Also, from Table 3 it can be observed that upper limit can also have estimates that are not possible for a proportion; in limits of CLT based intervals exceed one where as SG yields exactly one as upper limit where the observed proportion is quite nearer to one and as low as five.

Further nature of sparseness has been considered in understanding the performance of these methods in term of extreme lengths; three classifications of sparseness also demonstrate this behavior. QH and SG perform uniformly across these classifications and CLT based intervals provide wider intervals even in the case of mild sparse as well as all the four cells are with low counts. However, because FS dominates uniformly while comparing shorter intervals, nature of sparseness has not been considered in those cases.

The analysis schemes have been extended to a data set that has a 4 x 5 contingency table (Syzdaetal, 2008) with many zeros. Results have shown that no major changes in terms of longest interval are visible when compared to 2 x 2 tables. QH has dominated uniformly over all non zero cells in the table followed by SG and GM. However, unlike the case 2 x 2 table, this behavior does not distinguish between low or high non zero counts. Also, FS yields smaller intervals in all cases that may not be a required feature for an interval estimator. Bayesian method yields a better compromise estimates when compared to these methods with extreme values. The inevitable 0.0 as estimates for zero counts in the case of CLT methods are apparent for this data set too. But CLT-CC yields a negative lower limit for a case where the count is five. Hence when table size (k) or total counts (N) become large, the negative lower limit could appear in the case of cell counts over and above five.

Also, the outcome of simulation studies indicate a consistent behavior of Bayesian confidence intervals when compared to classical approach though MD intervals are uniformly narrower than other counterparts and achieves coverage probability less than 0.95. Agresti and Coull (1998) have pointed out that such property can also be preferred in certain cases and very wider intervals which may tend to provide very high coverage probability in most of the cases. This attempt includes another data of size 1 x 7 (Quesenberry and Hurst,1964) that has been used almost in all similar studies that is beyond the data sets considered in this Section. Figure 1 presents the illustrative details of the consistent behavior of MD and the extreme performance of QH, GM, and SG; CLT methods and FS are not considered for this comparison based on their performance that is observed earlier.

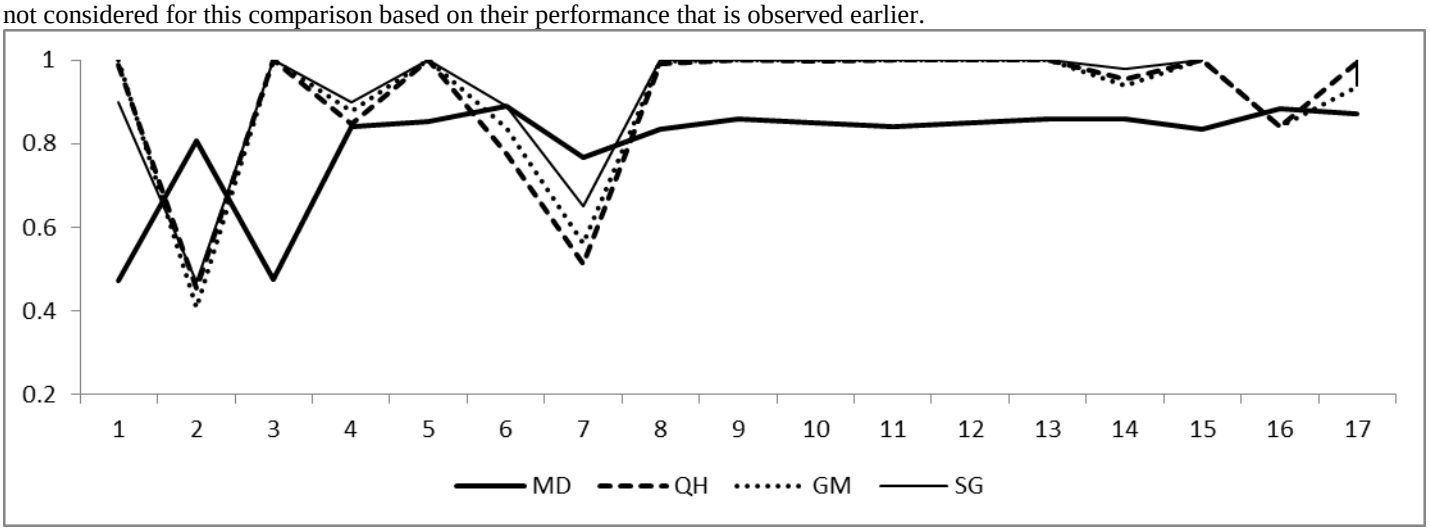

Figure 1: A comparison of coverage probabilities for the nominal 95% QH, GM, SG and MD intervals for multinomial proportions

All these data sets represent the varied feature of contingency tables so that methods can be compared for the performance of the methods. Extreme range of cell counts where classical methods are consistently closer to one whereas MD provides around 85% as its coverage probability in all such cases. From the figure it can be observed that there is a reversal tendency in the case of second set where cell counts are low and classical procedures have as low as 45%. Similar effect is also the case when cell counts are high but of notably apart from each other. Hence, in the absence of a perfect definition of sparseness in a general I x J tables, MD has a consistent behavior in terms of coverage probability though the numerical value is below the nominal value.

V. Conclusions

Multinomial proportions have found many applications and Burda et al (2008) have provided illustrative cases for multinomial discrete models. Lee et al (2011) have applied the methods for constructing confidence intervals for multinomial proportions to the design process of grain tracing and recall system. Kern (2006) has used a pig data to illustrate the Bayesian inference on multinomial probabilities. The present study has emphasized the need to understand, implement and compare the procedures in obtaining confidence intervals for multinomial proportions, especially when the data is sparse in nature.

In general, comparison and subsequent recommendation of any statistical procedures are based on their performance, wider availability to the users, computational issues, and aberrations. Further sparseness also plays an important role in deciding the procedure to be adopted. Agresti (1990) has stated that for sampling zeros, it is not sensible to use 0.0 as the best estimate of a probability. In view of this, classical methods, which yield 0.0 as estimator either because of its form or through auto corrections, need not be recommended if the data sets do have more zeros. Hence zero counts irrespective of other cell counts need a careful investigation in using an estimation method. Bayesian procedures even with a non informative prior yield estimates in such situations similar to its performance uniformly over 2 x 2 tables considered as a basic form in many categorical data studies.

Such similarity is consistently visible when the size of the table increases and range of the cell counts is relatively higher. However a classical method needs careful choice when k or N changes also when the cell counts are comparatively different. Further it may be easier in the present scenario to obtain a computation tool or mechanism, but still the availability of these classical methods is restricted to Wald type intervals. However, CLT based methods are not recommended when data is sparse irrespective of zero or non zero counts and size of the table.

Bayesian procedure has distinct advantages in obtaining the confidence limits without any aberrations; possessing acceptable frequentist coverage probabilities and practically important in computational flexibility and availability. It has been observed that exact inference plays important role in statistical inference of discrete data; however, for sparse data large sample chi-square statistics are often unrealistic (Agresti and Coull, 1998). More importantly, all of these conclusions are drawn for sparse data which is more realistic even for large contingency tables. Hence, this comparative study has emphasized the need to apply Bayesian methods with an objective prior for estimating multinomial proportions in categorical data with presence of zero or low cell counts and has an appreciable difference between the cell counts; in some cases the number of categories also plays a role to choose between methods. Bayesian methods have been identified as unified approach to handle varied situations of cell frequencies that would generally arise in the analyses of contingency tables and its applications.

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Cosmo Super Star! (A Science Fiction article)

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I. INTRODUCTION

What is the shape of universe?... What was the first life originated?... What is the Chromosome number of first organism?... The first organism evolved from matter or antimatter?... oh… still there is a mystery even in 21st Century about the exact shape of Universe, the Origin of first life. There are so many super stars on the earth now. If so who was the **FIRST SUPER STAR?**

It is hypothesized by the author that the prehistoric human who was spontaneously evolved by electro magnetic radiation from the dark flame (Cosmo) shall be called the first super star having **THREE CHROMOSOME** and he was much senior to **AMOEBA** and lived in **MARS.** It is speculated that the prehistoric superstar formulated the relative position of Sun, Earth, Moon in the Universe in prehistoric code with **THREE TINY DOTS**. It is hypothesized by the author that the three tiny dots shall not be a fiction and in fact author considers as a **DISCOVERY** about the real relative position of Sun, Earth, Moon in the Cosmo universe. It is focused that the Sun, Earth, Moon shall not be considered as independent space object but shall be considered as an integral part of **THREE-IN-ONE** Star of the whole cosmo universe.

Respected Sir

The main focus of this article / book is related to **Origin of Life** due to Electro-magnetic radiation. It is speculated that the electro magnetic radiation evolved from neutrinos and dark matters might be responsible for evolution of all matters like planet, asteroid, and other organisms. It is focused that the Evolution of Various Organisms is due to variation and increase in natural frequency of oscillation of Earth (Heart beat). It is hypothesised that the lineage of Evolution of Organism Shall be identified according to the Chromosome level of organism and it is speculated that the increase in Chromosome level is closely associated with increase in natural frequency of oscillation of Earth. It is focused that the variation in natural frequency of oscillation of earth shall be very well understood from the theory of **SCHUMANN RESONANCE.** It is hypothesized by the author that the earth heart beat is closely associated with **Human Heart Beat.**

It is hypothesized that the whole cosmo universe has only one Sun, one Earth, one Moon. It is hypothesized that in the existing solar system Sun, Earth, Moon shall be considered as integral part and forming only one **STAR** in the whole Universe, which was already formulated in **PREHISTORIC TAMIL CODE IN THREE TINY DOTS.** Three tiny dots shall means the relative position of Sun, Earth and Moon in the cosmo universe.

In the expanding Universe it is speculated that Moon is going away from Earth, which results in increase in natural frequency of oscillation of Earth due to reduction in intensity of magnetic flux axis connecting Moon and Earth. As per prehistoric law of Tamil code the Earth and Moon shall be considered as magnet bar in which moon shall be considered as North pole, Earth shall be considered as South pole and Sun is considered as located at mid point and vertical to the magnetic bar.

It hypothesized that the present day alarming of **Global warming** may be due to **asymmetry** in relative position of SUN, EARTH, MOON compared to prehistoric time which was **symmetric** in those period rather than increase in **carbon di oxide** level in the atmosphere.

Further it is hypothesized that from the law of prehistoric Tamil code, the Sun, Earth, Moon are not really **REVOLVING IN THE ORBIT**, but they are rotating only **ON ITS AXIS** in stationary position. Various cycles of climatic and environmental condition may be due to relative position of outer core of Sun, Earth, Moon. It is also hypothesized that the **INNER CORE OF THE SUN** also has microbial rotation on its axis which makes variation in unpredicted environment condition such as unexpected cyclones, earth quake, unexpected volcanic activity, unexpected new disease such as dengue fever, bird fever and other fast growth of cancer in the modern time. Further it is speculated that the inner core of Sun may be completing one cycle by 3000 A.D.

It is focused that the first life originated was THREE **CHROMOSOME PREHISTORIC HUMAN** and whose population might be living in MARS due to different genetic structure and subsisting with **OZONE** breath. During the course of evolution of expanding Universe the **MODERN HUMAN** with 46 chromosome might have been evolved and subsisting with **OXYGEN** breath. It is speculated that the oxygen might be evolved from the prehistoric ozone environment in much later period and in fact oxygen shall be considered as **SPECIES** to ozone. Further the prehistoric human might be having **GREEN** or **BLUE** colour blood due to ozone breath and subsequently the colour of the blood might be changed to **RED** colour due to oxygen breath in the modern time. It is focused by the Author that the first life originated might be prehistoric human having only three chromosome and much senior to **AMOEBA** having 12 chromosome. The prehistoric human might have derived three chromosome from the star i.e. Sun, Earth, Moon.

It is focused by author that prehistoric human (Akkie, Akkilan population) while living in **MARS** region might be well versed in advancement in scientific technology like astronomy, mathematical science, evolutionary science etc and formulated the prehistoric scientific laws in code form. It s hypothesized that during that period they might have formulated the relative position of Sun, Earth, Moon in **three dot code form.** Further they might have estimated that total chromosome of prehistoric human as three numbers and formulated in code form in three tiny dots.

It is speculated by the author that in modern communication the method of communication such as Braille code, Morse Code, Binary Code, American standard computer code etc might have been derived from the philosophy of prehistoric **Tamil code form.**

In modern science (Quantum Physics) still there is mystery exists about what is the shape of Universe?. Further scientists not able to conclude where from the huge radiant energy comes? It is hypothesized that the prehistoric human population already formulated that entire cosmo universe contains **THREE VACUUM** region of different level and formulated the Universe in code form with **three tiny dots**.

It is hypothesized by the author that the whole cosmo universe could derive huge radiant energy from three fundamental **TAMIL NEUTRINOS** (so called **Aether** energy in modern physics). Author hypothetically call these three fundamental Tamil Neutrinos as (i) **A** – Neutrino (ii) **K** - Neutrino (iii) **J** – Neutrino. A **-** Stands for **photon**, K – Stands **for Electron**, J – Stands for **Proton**.

It is speculated that the MARS planet cannot support the life of modern human with 46 chromosome and present attempt being made by NASA for making tour to MARS may be a unuseful attempt.

Further it is focused by the author that not only MARS Planet, all other planets, Asteriods, Astmosperic gases, Earth resources like Ocean, Soil, Rock, Coke, Coal, natural gas, crude oil, etc shall be considered as the **byproduct** of electromagnetic equilibrium in the evolution process of early universe. It is hypothesized by the author that even **HUMAN BLOOD** itself shall be considered as the byproduct of electromagnetic equilibrium takes place every day in humanbody.

Further it is hypothesized by the author that during the course of evolution of organisms the **48** chromosome organism might be evolved at much later period after evolution of 46 chromosome organism. The **Apes** family may be having close **genetic** characteristics due to close proximity of chromosome level with 46 chromosome (modern human). It is hypothesized by the author that **Darwin Sir** evolution theory is related to very far below in the **lineage** region of chromosome build up process of organisms.

Adam has three chromosome?.... Adam lived in MARS?....
If so HEAVENLY FATHER is Tamil based?....

Whether Prehistoric human population still exists? Or Died? New scientific theory (Biocentrism) suggests that **death** is not the terminal event. Further it is the **axioms** of science that **Energy Never Dies.** It can neither be created nor destroyed. But does this energy transcend from one world to another. If so where the energy of Prehistoric human gone?

The Prehistoric tamil law already formulated that the whole cosmo universe has only **three regions of vacuum** and it was formulated with **three tiny dots.** The bottom most region shall be considered as **Material Universe** where all matters including planets exist. The other two upper regions shall be considered as containing Dark energy consists full of Neutrinos and anti-neutrinos derived from **Dark Flame** (Cosmo).

It is hypothesized by the author that all material form of organisms such as human, animals; plants, after death may rise upto the upper region of vacuum and still exists with **Dark Flame** of Universe.

It is further focused by the author that the existence of whole cosmo universe shall be considered as a **Global Tamil Family** and it was formulated by Prehistoric Tamil Code with three **tiny dots.** The three tiny dot shall mean **Father, Mother** and **Child.**

It is speculated that in the global family the child is considered going away from mother and hence the mother is **heated** up. In **Astronomical** term, it shall be understood that Moon is going away from Earth and hence **Earth gets warmed** and it is understood as **GLOBAL WARMING.** The **Mayan Calender** predicted the **ZERO POINT** of earth occurs by **2012 AD.** But it is speculated that the Prehistoric Tamil Populations already estimated that the **Zero** point shall occur by **3000 AD.**

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Efficient Flash Translation layer for Flash Memory

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Abstract- Flash is a type of electronically erasable programmable read- only memory (EEPROM). Flash memory is important as nonvolatile storage for mobile consumer electronics due to its low power consumption and shock resistance.

NAND flash memory has many advantageous features as a storage medium, such as superior performance, shock resistance, and low power consumption. However, the erase-before-write nature and the limited number of write/erase cycles are obstacles to the promising future of NAND flash memory. An intermediate software layer called Flash Translation Layer (FTL) is used to overcome these obstacles.

Many efforts for optimizing the working of address mapping schemes have been done by different research workers. Though various schemes are designed and proposed but there is no literature available providing mathematical computations comparing the performance of the various mapping schemes in the form of time complexity. In this paper we have tried to find out the comparative cost of block merge operation required during garbage collection for some representative mapping schemes like BAST and FAST .

I. INTRODUCTION

1.1 Characteristics of NAND flash memory

NAND flash memory has several limitations:

1) The previous data should be erased before a new data can be written in the same place. This is usually called *erase-before-write* characteristic.

2) Normal read/write operations are performed on a per-*page* basis, while erase operations on a per-*block* basis. The erase block size is larger than the page size by 64~128 times. In MLC (Multi-Level Cell) NAND flash memory, the typical page size is 4KB and each block consists of 128 pages.

3) Flash memory has limited lifetime; MLC NAND flash memory wears out after 5K to 10K write/erase cycles.

There are three basic operations in NAND flash memory: read, write (or program), and erase. The read operation fetches data from a target page, while the write operation writes data to a page. The erase operation resets all values of a target block to 1. NAND flash memory does not support in-place update.

A NAND flash memory chip is composed of a fixed number of *blocks*, where each block typically has 32 *pages*. Each page in turn consists of 512 bytes of the main data area and 16 bytes of the spare area. The page is the basic unit of read and write operations in NAND flash memory.

NAND flash memory is usually used as a storage medium in place of Hard Disk Drive (HDD) because of its non- volatility and large I/O unit. It is not straightforward to replace HDDs with NAND flash memory due to its erase before- write nature. The erase unit (called block) is larger than the read and write unit (called page) by 32 – 128 times.

Other semiconductor devices such as SRAMs and DRAMs, the write operation compared to the read operation. As the write operation usually accompanies the erase operation.

Another limitation of NAND flash memory is that the number of program/erase cycles for a block is limited to about 100,000 – 1,000,000 times. Thus, the number of erase operations should be minimized to improve the overall performance and the lifetime of NAND flash memory.

1.2 Flash Translation Layer (FTL)

The FTL is one of the core engines in flash-based SSDs that maintain a mapping table of virtual addresses from upper layers to physical addresses on the flash.

The main goal of FTL is to emulate the functionality of a normal block device with flash memory, hiding the presence of erase operation and erase-before-write characteristics. Two important functions of FTL are address translation and garbage collection.

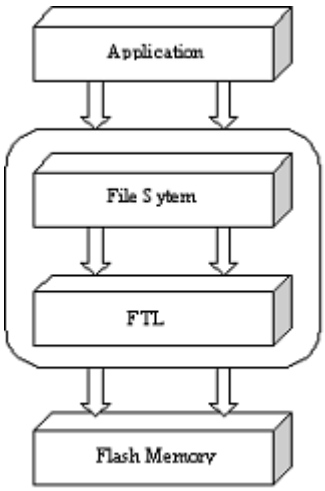

Figure 1 Architecture of Flash File System

Fig. 1 shows the software architecture of the flash file system. This section focuses on the FTL layer shown in Fig. 1. The file system layer issues a series of read or write commands each with a logical sector number, to read data from, or write data to, specific addresses in flash memory. The logical sector number is converted to a real physical sector number of flash memory by some mapping algorithm in the FTL layer.

FTLs can be categorized into two classes according to their mapping granularities: Sector-mapped FTLs and block-mapped FTLs. A Sector-mapped FTL literally maps a logical address into a physical address in a page unit. It is highly flexible as a logical page can be written to any physical page in NAND flash memory. On the other hand, the mapping unit of a block-mapped FTL is a block. And later on a hybrid mapping scheme proposed by D. Park came into existence.

II. Types of Mapping Schemes

2.1 Sector – Mapping FTL

For example, Figure 2 shows an example of sector mapping. In the example, we assume that a block is composed of four pages and so there are totally 16 physical pages, where each page is organized into the sector and spare areas. If we also assume that there are 16 logical sectors, the row size of the mapping table is 16. When the file system issues a command - "write some data to LSN (Logical Sector Number) 9", the FTL algorithm writes the data to PSN (Physical Sector Number) 3 according to the mapping table in case the PSN 3 has not been written before.

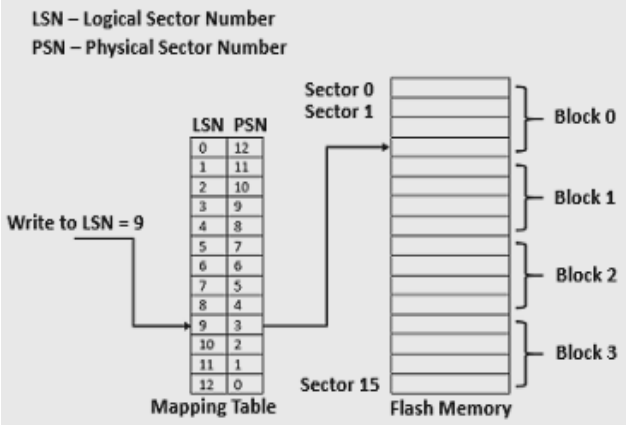

Figure 2 Sector Mapping

But, in other case, the FTL algorithm looks for an empty physical sector, writes data to it, and adjusts the mapping table. If there is no empty sector, the FTL algorithm will select a victim block from flash memory, copy back the valid data to the spare free block, and update the mapping table. Finally, it will erase the victim block, which will become the spare block.

In order to rebuild the mapping table after power outage, the FTL algorithm either stores the mapping table to flash memory or records the logical sector number in the spare area on each writes to the sector area.

Table 1. Measures of Sector mapping scheme

Garbage collection cost	Block Erase is done when a block is completely utilized.
RAM requirement	Proportional to flash size
Search time	Not required
Usefulness	Useful in case strict time requirement

2.2 Block-mapping FTL

The pure block-mapping FTL is another classic FTL scheme Block-mapping table is used to store and manage the mapping information between LBN and Physical Block Number (PBN). If there are m pages in a block, the size of the block-mapping table is m times smaller than its page-mapping counterpart. In a block-mapping FTL, one LPN must be mapped to a fixed page offset in any physical block (i.e., direct mapping). If this page offset has been written before, the LPN cannot be written to any other page in this block even if there are free pages in the same physical block. In this case, all existing valid data in the block as well as the data to be written must be copied to a new clean block, and the old block is marked for erase, incurring one erase and a number of read/write operations. Compared with the page-mapping FTL the block-mapping FTL requires extra operations to serve a request, adversely affecting the performance. Since both the block-mapping and page-mapping FTLs have their aforementioned disadvantages, they are rarely used in SSD commercial products in their pure forms.

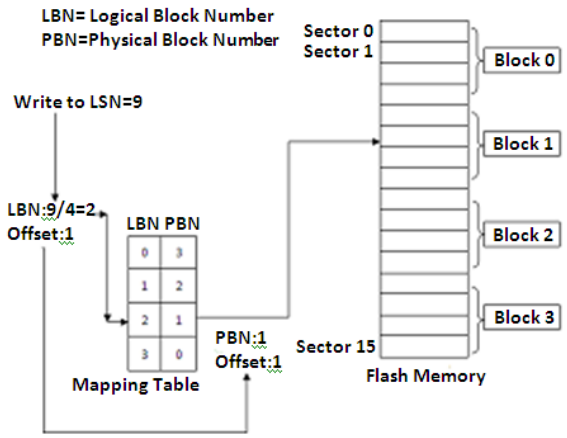

Figure 3. Block mapping

figure 3 shows an example of the block mapping algorithm. Assuming that there are four logical blocks, the row size of the mapping table is four. If the file system issues the command "write some data to LSN 9", the FTL algorithm calculates the logical block number 2(=9/4) and the sector offset 1(=9%4), and then finds physical block number 1 using the mapping table. Since in the block mapping algorithm, the physical sector offset equals the logical sector offset, the physical sector location can be easily determined.

Table 2. Measures of block mapping scheme

Garbage collection cost	Block Erase is done hen a block is completely utilized.
RAM requirement	Proportional to flash size
Search time	Not required
Usefulness	Useful in case of strict time requirement

2.3 Hybrid FTL

A family of the hybrid mapping schemes is introduced to address the shortcomings of the Sector-mapping and block-mapping FTLs. In a typical hybrid FTL, physical blocks are logically partitioned into two groups: data blocks and log blocks. When a write request arrives, the hybrid FTL first writes the new data in a log block and invalidates the data in the corresponding target data block.

Block-mapping information for data blocks and page-mapping information for log blocks are kept in a small RAM for performance purposes. When all the log blocks are full, their data are flushed into the data blocks immediately and they are then erased to generate new free log blocks. More specifically, the valid data in data blocks and the valid data in the corresponding log block must be merged and written to a new clean data block. This process is called a merge operation. Further, merge operations can be classified into three types depending on their overhead. Full merge occurs, when the log block is selected as a victim block and not written sequentially from the first page to the last page, and all the valid data in it and in its corresponding data block are copied to a new clean block.

This process requires m read operations, m write operations and two erase operations, where m is the number of pages in a block. When the log block is written sequentially from the first page to the last page of a logical block, this log block can replace the corresponding data block, a merge operation called switch merge. This type of merge requires only one erase operation. Partial merge takes place when the log block is written sequentially from the first page to a middle page in a block, and the last part of data will be copied from the corresponding data block. Partial merge requires several read and write operations and one erase operation. A number of variations of the hybrid FTL schemes have been proposed recently, including BAST, FAST, LAST, Superblock Reconfigurable FTL. More recently, Demand-based FTL (DFTL) was proposed to address the RAM-capacity problem of the page-mapping FTL by storing only the "hot" mapping information in RAM based on temporal locality of workloads.

DFTL is shown to significantly outperform hybrid FTLs.

A hybrid technique, as its name suggests, first uses a block mapping technique to get the corresponding physical block , and then, uses a sector mapping technique to find an available empty sector within the physical block.

Figure 4 shows an example of the hybrid technique. When the file system issues the command "write some data to LSN 9", the FTL algorithm calculates the logical block number 2(=9/4) for the LSN, and then, finds the physical block number 1 from the mapping table. After getting the physical block number, the FTL algorithm allocates an empty sector for the update. In the example, since the first sector of the physical block 1 is empty, the data is written to the first sector location. In this case, since the two logical and physical sector offsets (i.e., 2 and 1, respectively) differ from each other, the logical sector number 9 should be written to the spare area in page 1 of the physical block 1. For rebuilding the mapping table, not only this information but also the logical block numbers have to be recorded in the spare areas of the physical blocks.

When reading data from flash memory, the FTL algorithm first finds the physical block number from the mapping table using the given LSN, and then, by reading the logical sector numbers from the spare areas of the physical block, it can get the most recent value for requested data.

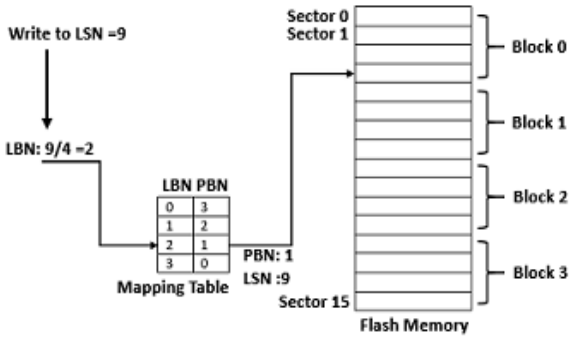

Figure 4. Hybrid Mapping

III. Comparative Scheme

3.1) Block Associative Sector Translation (BAST) exclusively associates a log block with a data block. In presence of small random writes, this scheme suffers from ***log block thrashing*** that results in increased full merge cost due to inefficiently utilized log blocks.

Table 3. Measures of BAST scheme

Garbage collection cost (Worst case: considering number of random requests = number of log blocks	((3T/100)*N)Read + ((3T/100)*N) Write + (2*3T/100) Erase T: total blocks in flash N: number of pages/block Log block: 3% of T
RAM requirement	Less, Proportional to number of log block
Search time (worst case	Time to search the page map table of a single log block
Usefulness	In case of sequential read write and update pattern

3.2) Fully Associative Sector Translation (FAST) allows log blocks to be shared by all data blocks. This improves the utilization of log blocks as compared to BAST. FAST keeps a

single sequential log block dedicated for sequential updates while other log blocks are used for performing random writes. Thus, it cannot accommodate multiple sequential streams and does not provide any special mechanism to handle temporal locality in random streams.

Table 4. Measures of FAST mapping scheme

Garbage collection cost (worst and best case : considering number of random requests = K* number of log Blocks)	((3T/100)*K*N) Read + ((3T/100) * K*N) Write +(2*3T/100) Erase T: total blocks in flash N: number of ages/block Log block: 3% of T
RAM requirement	Less, Proportional to num, of random log blocks
Search time (worst case)	Time to search the page map table of all log blocks.
Usefulness	

3.3) Super Block FTL scheme utilizes existence of *block level* spatial locality in workloads by combining consecutive logical blocks into a super block. It maintains page- level mappings within the superblock to exploit temporal locality in the request streams by separating hot and cold data within the superblock. However, the three-level address translation mechanism employed by this scheme causes multiple OOB area reads and writes for servicing the requests. More importantly, it utilizes a fixed superblock size which needs to be explicitly tuned to adapt to changing workload requirements.

3.4) Locality-Aware Sector Translation (LAST)

Scheme tries to alleviate the shortcomings of FAST by providing multiple sequential log blocks to exploit spatial locality in workloads. It further separates random log blocks into hot and cold regions to reduce full merge cost. In order to provide this dynamic separation, LAST depends on an external locality detection mechanism. However, Lee et al. themselves realize that the proposed locality detector cannot efficiently identify sequential writes when the small-sized write has sequential locality. Moreover, maintaining sequential log blocks using a block-based mapping table requires the sequential streams to be aligned with the starting page offset of the log block in order to perform switch-merge. Dynamically changing request streams may impose severe restrictions on the utility of this scheme to efficiently adapt to the workload patterns.

Table 5. Measures of LAST mapping scheme

Garbage collection cost	Better than FAST
RAM requirement	Same as that of FAST
Search time	Time to search a page map tables of all log block
Usefulness	Useful in case of random Read write & update pattern.

IV. DFTL Architecture

DFTL makes use of the presence of temporal locality in workloads to judiciously utilize the small on-flash SRAM. Instead of the traditional approach of storing all the address translation entries in the SRAM, it dynamically loads and unloads the page-level mappings depending on the work- load access patterns. Furthermore, it maintains the complete image of the page-based mapping table on the flash device itself. There are two options for storing the image: (i) The OOB area or (ii) the data area of the physical pages. We choose to store the mappings in the data area instead of OOB area because it enables us to group a larger number of mappings into a single page as compared to storing in the OOB area. For example, if 4 Bytes are needed to represent the physical page address in flash, then we can group 512 logically consecutive mappings in the data area of a single page whereas only 16 such mappings would fit an OOB area. Moreover, the additional space overhead incurred is negligible as compared to the total flash size. A 1GB flash device re- quires only about 2MB (approximately 0.2% of 1GB) space for storing all the mappings.

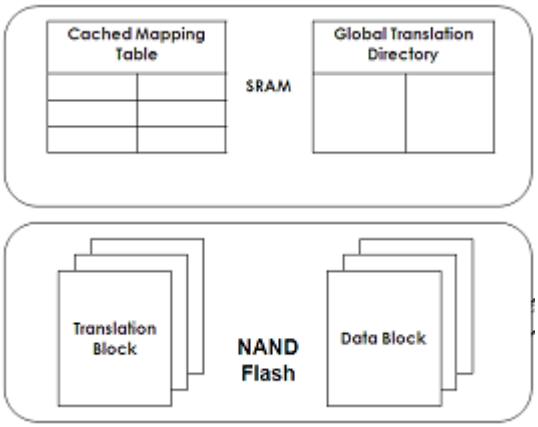

Figure 5 DFTL Architecture

5) Comparison of Existing State-of-the-art FTLs with DFTL

Table 2 shows some of the salient features of different FTL schemes. The DFTL architecture provides some intrinsic advantages over existing state-of-the-art FTLs as follows:

5.1) ***Full Merge*** - Existing hybrid FTL schemes try to reduce the number of full merge operations to improve their performance. DFTL, on the other hand, completely does away with full merges. This is made possible by page- level mappings which enable relocation of any logical page to any physical page on flash while other hybrid FTLs have to merge page-mapped log blocks with block- mapped data blocks.

5.2) Partial Merge - DFTL utilizes page-level temporal locality to store pages which are accessed together within same physical blocks. This implicitly separates hot and cold blocks as compared to LAST and Superblock schemes [13, 20] which require special external mechanisms to achieve the segregation. Thus, DFTL adapts more efficiently to changing workload environment as compared with existing hybrid FTL schemes.

5.3) Random Write Performance **-** As is clearly evident, it is not necessarily the random writes which cause poor flash device performance but the intrinsic shortcomings in the design of hybrid FTLs which cause costly merges (full) on log blocks during garbage collection. Since DFTL does not require these expensive full-merges, it is able to improve random write performance.

5.4) Block Utilization **-** In hybrid FTLs, only log blocks are available for servicing update requests. This can lead to low block utilization for workloads whose working-set size is smaller than the flash size. Many data blocks will remain un-utilized (hybrid FTLs have block-based map- pings for data blocks) and unnecessary garbage collection will be performed. DFTL solves this problem since up- dates can be performed on any of the data blocks.

V. Conclusion

We argued that existing hybrid FTL schemes exhibit poor performance for enterprise-scale workloads with significant random write patterns. We proposed a complete paradigm shift in the design of the FTL with our Demand-based Flash Translation Layer (DFTL) that selectively caches page- level address mappings. Our experimental evaluation using Disk Sim with realistic enterprise-scale workloads endorsed DFTL's efficacy for enterprise systems by demonstrating that DFTL offered (i) Improved performance, (ii) reduced garbage collection overhead, (iii) Improved overload behavior and (iv) Most importantly unlike existing hybrid FTLs is free from any tunable parameters. As a representative example, a predominantly random write-dominant I/O trace from an OLTP application running at a large financial institution showed a 78% improvement in average response time due to a 3-fold reduction in garbage collection induced operations as compared to a state-of-the-art FTL scheme.

Table 6. Comparative analysis of various FTL schemes

FTL Scheme	Merge cost of block during GC	Lookup performance	RAM req.	Mapping granularity
Pure page level	N/A	No lookup cost	Much more	Page
Pure block level	Much more	No lookup cost	Very less	Block
BAST	More	Less	Less	Page for log blocks. Blocks for data blocks
FAST	Lesser than BAST	Much more	Less	Page for log blocks. Blocks for data blocks
LAST	Lesser than FAST	Much more	Less	Page & block for log blocks. Block for data blocks
Demand paged	N/A	Lesser	Lesser	Page

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Use of Carbon Capture Storage Technique for Clean Power Generation

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Abstract- Global demand for energy shows no signs of slowing, carbon dioxide emissions keep surging to new records. More than ever, the need for a fundamental shift to a cleaner and more reliable energy system is clear. The technologies with the greatest potential for energy and carbon dioxide (CO2) emission reduction, however, are making the slowest progress. Carbon capture and storage (CCS) is not seeing the necessary rates of investment into full-scale demonstration projects and nearly one-half of new coal-fired power plants are still being built with inefficient technology.

The power generation sector is expected to contribute more than one-third of potential CO2 emissions reductions worldwide by 2020 under the 2DS, and almost 40% of 2050 emission reduction. Enhanced power generation efficiency, a switch to lower-carbon fossil fuels, increased use of renewable and nuclear power, and the introduction of CCS are all required to achieve cleaner power generation.

I. INTRODUCTION

The Indian power sector is growing at a rapid pace and coal is, and will continue to be, the primary energy source for the Indian power sector that underpins the country's economic development. About 68% of India's current power generation is from coal and that share is expected to remain in the same range for many years. A significant percentage of India's existing coal-fired power plants are aging and are no longer very efficient and reliable, especially those operated by state utilities. In recent years, India power generators have focused on increasing plant load factors (PLF, or capacity factors) to increase electricity output, but often at the expense of plant efficiency. Heat rates (i.e., fuel consumption per unit of electricity generated) higher than 3,000 kcal/kWh (i.e., efficiencies lower than about 28%) are common among the older, especially smaller (100-200 megawatt, MW) units operated by many Indian state utilities. Plant reliability and efficiency suffer as a result of:

1. The high-ash content of most Indian coals, which often exceeds 45%, and overall poor coal quality.
2. Limited maintenance of equipment.
3. Absence of incentives to maintain or improve efficiency and reliability.
4. Inadequate operating budgets: this includes the cost involved in the maintenance of the plant machineries and purchase of new technological instruments, the cost involved in the reduction of CO_2 emissions to the atmosphere etc.
5. Limited knowledge and experience on state-of-the art operating and maintenance (O&M) practices, which adversely impacts both plant efficiency and reliability.

Several Energy Efficient R&M (Renovation and Modernization) projects have already been implemented under cooperative programs with Germany, Japan, and the World Bank. Since 1982, USAID/India has sought to promote and demonstrate better coal utilization technologies and practices in India.

II. CARBON CAPTURE AND STORAGE

Carbon Capture and Storage refers to technology attempting to prevent the release of large quantities of CO_2 into the atmosphere from fossil fuel use in power generation and other industries by capturing CO_2, transporting it and ultimately, pumping it into underground geologic formations to securely store it away from the atmosphere. The value of CCS as an important CO_2 abatement tool is already recognized by multiple authoritative organizations, including the UK Committee on Climate Change (CCC), the Intergovernmental Panel on Climate Change (IPCC) and the International Energy Agency (IEA). The IEA asserts that the cost of reducing emissions to 2005 levels by 2050 will increase by 70% if CCS is not deployed. For many large industrial emitters, CCS is seen as the only viable long term solution to mitigating the future costs of CO_2 emissions.

There are three main approaches to CO2 capture, for industrial and power plant applications.

1. *Post-combustion* systems: separate CO_2 from the flue gases produced by combustion of a primary fuel (coal, natural gas, oil or biomass) in air.
2. *Oxy-fuel combustion* systems: uses oxygen instead of air for combustion, producing a flue gas that is mainly H_2O and CO_2 and which is readily captured. This is an option still under development.
3. *Pre-combustion* systems: process the primary fuel in a reactor to produce separate streams of CO_2 for storage and H_2 which is used as a fuel.

The basis for CO_2 capture:

The main application of CO_2 capture is likely to be at large point sources: fossil fuel power plants fuel processing plants and other industrial plants, particularly for the manufacture of iron, steel, cement and bulk chemicals. Capturing CO_2 directly from small and mobile sources in the transportation and residential &

commercial building sectors is expected to be more difficult and expensive than from large point sources. An alternative way of avoiding emissions of CO_2 from these sources would be by use of energy carriers such as hydrogen or electricity produced in large fossil fuel-based plants with CO_2 capture or by using renewable energy sources. In this report we concentrate our discussion on Post-Combustion capture technology since CO_2 is primarily obtained from burning the fossil fuels.

III. CARBON CAPTURE

Post-Combustion capture

Capture of CO_2 from flue gases produced by combustion of fossil fuels and biomass in air is referred to as post-combustion capture. Instead of being discharged directly to the atmosphere, flue gas is passed through equipment which separates most of the CO_2. The CO_2 is fed to a storage reservoir and the remaining flue gas is discharged to the atmosphere. Current anthropogenic CO_2 emissions from stationary sources come mostly from combustion systems such as power plants, cement kilns, furnaces in industries and iron and steel production plants. All the CO_2 capture systems described in this section are aimed at the separation of CO_2 from the flue gases generated in a large-scale combustion process fired with fossil fuels. Flue gases or stack gases found in combustion systems are usually at atmospheric pressure. Because of the low pressure, the large presence of nitrogen from air and the large scale of the units, huge flows of gases are generated, the largest example of which may be the stack emissions coming from a natural gas combined cycle power plant.

In principle post-combustion capture systems can be applied to flue gases produced from the combustion of any type of fuel. However, the impurities in the fuel are very important for the design and costing of the complete plant. Flue gases coming from coal combustion will contain not only CO2, N2, O2 and H2O, but also air pollutants such as SOx, NOx, particulates, HCl, HF, mercury, other metals and other trace organic and inorganic contaminants. There are several commercially available process technologies which can in principle be used for CO2 capture from flue gases. However, comparative assessment studies have shown that absorption processes based on chemical solvents are currently the preferred option for post-combustion CO2 capture. At this point in time, they offer high capture efficiency and selectivity, and the lowest energy use and costs when compared with other existing post-combustion capture processes.

IV. TRANSPORTATION OF CO_2

Transport is that stage of carbon capture and storage that links sources and storage sites. In the context of long-distance movement of large quantities of carbon dioxide, pipeline transport is part of current practice. Except when plants are located directly above a geological storage site, captured CO_2 must be transported from the point of capture to a storage site.
Pipelines today operate as a mature market technology and are the most common method for transporting CO_2. Gaseous CO_2 is typically compressed to a pressure above 8 MPa in order to avoid two-phase flow regimes and increase the density of the CO_2, thereby making it easier and less costly to transport.CO_2 also can be transported as a liquid in ships, road or rail tankers that carry CO_2 in insulated tanks at a temperature well below ambient, and at much lower pressures. The first long-distance CO_2 pipeline came into operation in the early 1970s. In the United States, over 2,500 km of pipeline transports more than 50 $MtCO_2$ per year from natural and anthropogenic sources, mainly to sites in Texas, where the CO2 is used for EOR. These pipelines operate in the 'dense phase' mode (in which there is a continuous progression from gas to liquid, without a distinct phase change), and at ambient temperature and high pressure. In most of these pipelines, the flow is driven by compressors at the upstream end, although some pipelines have intermediate (booster) compressor stations.

In pipeline transportation, the volume is reduced by transporting at a high pressure, this is routinely done in gas pipelines, where operating pressures are between 10 and 80 MPa. A transportation infrastructure that carries carbon dioxide in large enough quantities to make a significant contribution to climate change mitigation will require a large network of pipelines. The carbon dioxide stream ought preferably to be dry and free of hydrogen sulphide, because corrosion is then minimal, and it would be desirable to establish a minimum specification for pipeline quality. However, it would be possible to design a corrosion resistant pipeline that would operate safely with a gas that contained water, hydrogen sulphide and other contaminants. Pipeline transport of carbon dioxide through populated areas requires attention be paid to design factors, to overpressure protection, and to leak detection. Liquefied natural gas and petroleum gases such as propane and butane are routinely transported by marine tankers; this trade already takes place on a very large scale. Carbon dioxide is transported in the same way, but on a small scale because of limited demand. The properties of liquefied carbon dioxide are not greatly different from those of liquefied petroleum gases, and the technology can be scaled up to large carbon dioxide carriers.

Transport of CO_2 through pipeline requires that CO_2 be compressed and cooled to the liquid state. The properties of CO_2 are considerably different from other fluids commonly transported by pipeline, such as natural gas. Thus, it is necessary to use accurate representations of the phase behaviour, density, and viscosity of CO_2 in the design of the pipeline. Transport of CO_2 at lower density results in high pressure drop per unit length which results in inefficient transport of CO_2. To reduce difficulties in design and operation, it is generally recommended that a CO_2 pipeline operate at pressures greater than 8.6 MPa by which any sharp changes in the compressibility of CO_2 can be avoided across a range of temperatures that may be encountered in the pipeline system.

V. STORAGE

This section examines types of geological formations that have received extensive consideration for the geological storage of CO_2: oil and gas reservoirs, deep saline formations and unminable coal beds.

CO_2 storage in hydrocarbon reservoirs or deep saline formations is generally expected to take place at depths below 800 m, where the ambient pressures and temperatures will

usually result in CO_2 being in a liquid or supercritical state. Under these conditions, the density of CO_2 will range from 50 to 80% of the density of water. This is close to the density of some crude oils, resulting in buoyant forces that tend to drive CO_2 upwards. Consequently, a well-sealed cap rock over the selected storage reservoir is important to ensure that CO_2 remains trapped underground. When injected underground, the CO_2 compresses and fills the pore space by partially displacing the fluids that are already present (the 'in situ fluids'). In oil and gas reservoirs, the displacement of in situ fluids by injected CO_2 can result in most of the pore volume being available for CO_2 storage. In saline formations, estimates of potential storage volume are lower, ranging from as low as a few percent to over 30% of the total rock volume.

Once injected into the storage formation, the fraction retained depends on a combination of physical and geochemical trapping mechanisms. Physical trapping to block upward migration of CO_2 is provided by a layer of shale and clay rock above the storage formation. This impermeable layer is known as the "cap rock". Additional physical trapping can be provided by capillary forces that retain CO_2 in the pore spaces of the formation. In many cases, however, one or more sides of the formation remain open, allowing for lateral migration of CO_2 beneath the cap rock. In these cases, additional mechanisms are important for the long-term entrapment of the injected CO_2. The mechanism known as geochemical trapping occurs as the CO_2 reacts with the in situ fluids and host rock. First, CO_2 dissolves in the in situ water. Once this occurs (over time scales of hundreds of years to thousands of years), the CO_2- laden water becomes more dense and therefore sinks down into the formation (rather than rising toward the surface).

Next, chemical reactions between the dissolved CO_2 and rock minerals form ionic species, so that a fraction of the injected CO_2 will be converted to solid carbonate minerals over millions of years. Yet another type of trapping occurs when CO_2 is preferentially adsorbed onto coal or organic-rich shales replacing gases such as methane. In these cases, CO_2 will remain trapped as long as pressures and temperatures remain stable. These processes would normally take place at shallower depths than CO_2 storage in hydrocarbon reservoirs and saline formations.

VI. CO2 STORAGE MECHANISMS IN GEOLOGICAL FORMATIONS

The effectiveness of geological storage depends on a combination of physical and geochemical trapping mechanisms. The most effective storage sites are those where CO_2 is immobile because it is trapped permanently under a thick, low-permeability seal or is converted to solid minerals or is adsorbed on the surfaces of coal micro pores or through a combination of physical and chemical trapping mechanisms.

Physical trapping: stratigraphic and structural

Initially, physical trapping of CO_2 below low-permeability seals (cap rocks), such as very-low-permeability shale or salt beds, is the principal means to store CO_2 in geological formations. In some high latitude areas, shallow gas hydrates may conceivably act as a seal. Sedimentary basins have such closed, physically bound traps or structures, which are occupied mainly by saline water, oil and gas. Structural traps include those formed by folded or fractured rocks. Faults can act as permeability barriers in some circumstances and as preferential pathways for fluid flow in other circumstances. Stratigraphic traps are formed by changes in rock type caused by variation in the setting where the rocks were deposited. Both of these types of traps are suitable for CO_2 storage. Care must be taken not to exceed the allowable overpressure to avoid fracturing the cap rock or re-activating faults.

Physical trapping: hydrodynamic

Hydrodynamic trapping can occur in saline formations that do not have a closed trap, but where fluids migrate very slowly over long distances. When CO_2 is injected into a formation, it displaces saline formation water and then migrates buoyantly upwards, because it is less dense than the water. When it reaches the top of the formation, it continues to migrate as a separate phase until it is trapped as residual CO_2 saturation or in local structural or stratigraphic traps within the sealing formation. In the longer term, significant quantities of CO_2 dissolve in the formation water and then migrate with the groundwater. Where the distance from the deep injection site to the end of the overlying impermeable formation is hundreds of kilometres, the time scale for fluid to reach the surface from the deep basin can be millions of years.

Geochemical trapping

Carbon dioxide in the subsurface can undergo a sequence of geochemical interactions with the rock and formation water that will further increase storage capacity and effectiveness. First, when CO_2 dissolves in formation water, a process commonly called solubility trapping occurs. The primary benefit of solubility trapping is that once CO_2 is dissolved, it no longer exists as a separate phase, thereby eliminating the buoyant forces that drive it upwards. Next, it will form ionic species as the rock dissolves, accompanied by a rise in the pH. Finally, some fraction may be converted to stable carbonate minerals (mineral trapping), the most permanent form of geological storage (Gunter *et al.*, 1993). Mineral trapping is believed to be comparatively slow, potentially taking a thousand years or longer. Nevertheless, the permanence of mineral storage, combined with the potentially large storage capacity present in some geological settings, makes this a desirable feature of long term storage.

VII. COST OF CO2 CAPTURE AND STOREGE OPERATIONS

CCS applied to a modern conventional power plant could reduce CO_2 emissions to the atmosphere by approximately 80-90% compared to a plant without CCS. Capturing and compressing CO_2 requires much energy and would increase the fuel needs of a coal-fired plant with CCS by about 25%. These and other system costs are estimated to increase the cost of energy from a new power plant with CCS by 21-91%.

	Natural gas combined cycle	Pulverized coal	Integrated gasification combined cycle
Without capture (reference plant)	0.03 - 0.05	0.04 - 0.05	0.04 - 0.06
With capture and geological storage	0.04 - 0.08	0.06 - 0.10	0.06 - 0.09
With capture and Enhanced oil recovery	0.04 - 0.07	0.05 - 0.08	0.04 - 0.08

Table 1: Costs of energy with and without CCS (2002 US$ per kWh)

VIII. CONCLUSION

Large reductions in emissions of CO_2 to the atmosphere are likely to be needed to avoid major climate change. Capture and storage ofCO_2, in combination with other CO_2 abatement techniques, could enable these large reductions to be achieved with least impact on the global energy infrastructure and the economy. Capture and storage is particularly well suited to use in central power generation and many energy-intensive industrial processes. CO_2 capture and storage technology also provides a means of introducing hydrogen as an energy carrier for distributed and mobile energy users.

For power stations, the cost of capture and storage is about $50/t of$CO_2$ avoided. This compares favorably with the cost of many other options considered for achieving large reductions in emissions. Use of this technique would allow continued provision of large-scale energy supplies using the established energy infrastructure. There is considerable scope for new ideas to reduce energy consumption and costs of CO_2 capture and storage which would accelerate the development and introduction of this technology.

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Phytochemical Analysis of Methanolic Extracts of Leaves of Some Medicinal Plants

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Abstract- The present investigation deals with the phytochemical studies of leaves of different medicinal plants like Andrographis paniculata (Burm. f.) Wall ex Nees of the family Acanthaceae, Bauhinia acuminata Linn. of the family Caesalpiniaceae, Clerodendrum indicum (Linn.) O. Kuntze syn. C. siphonanthus R. Br. of the family Verbenaceae, Nerium odorum Soland. of the family Apocynaceae and Sida humilis Cav. Syn. S. veronicaefolia Lam., syn. S. cordata (Burm. f.) Borss. of the family Malvaceae. Methanolic (90%) extracts of leaf powders have been screened for qualitative determination of different secondary metabolites like starch, alkaloids, flavonoids, tannins, reducing sugars, amino acids and lignins by specific chemical color reaction tests.

Index Terms- Medicinal plants, Methanolic extracts Phytochemical study, Secondary metabolites.

I. INTRODUCTION

Nature has provided a complete store house of remedies to cure all ailment of mankind. Use of plants as a source of medicine has been inherited from the onset of human civilization and is an important component of the healthcare system. The aims of this paper are to evaluate the preliminary phytochemical characters such as determination of phermacognostic principals of some medicinal plants of different families. In recent years, chemical analysis and biological assays have begun to play an important role in ethnobotanical studies (Jana et al., 2009). In several cases, such analyses have led to the discovery of novel bioactive phytochemicals.

In the present investigation, an attempt was made by some microchemical tests to study the phytochemicals of some medicinally important plants like *Andrographis paniculata* of the family Acanthaceae which is used to treat infections and some diseases (Anand et al., 2011); *Bauhinia acuminata* of the family Caesalpiniaceae is reported as medicinally important in traditional system of medicine and are used extensively for the treatment of inflammation, headache, fever, tumors, skin infections etc. (Kumar et al., 2010); *Clerodendrum indicum* of the family Verbenaceae, which is reported to use as medicine for the treatment of asthma, pyreticosis, cataract etc. (Chandrashekar and Rao, 2012) ; *Nerium odorum* of the family Apocynaceae is also used in manifests as nausea, vomiting, colic etc. (Patel et al., 2010) and *Sida cordata* of the family Malvaceae can be effectively used in curing the bacterial diseases, bronchial asthma, cold etc. (Kalaiarasan and John, 2010).

II. MATERIAL AND METHODS

The leaf samples of *Andrographis paniculata*, *Bauhinia acuminata*, *Clerodendrum indicum*, *Nerium odorum* and *Sida humilis* (Figure-1) were collected from the medicinal garden of Rampurhat College, Rampurhat, Birbhum, located in the lateritic belt of West Bengal. The selected plant species have carefully identified with the help of different floras (Paria, 2005; Maheswari, 2000; Panigrahi and Murthy, 1989; Varma, 1981).

The leaf samples of selected plant species were carefully separated, cleaned, shade dried, mechanically grinded and coarsely powdered. Finally, the leaf powders were extracted (Soxhlet extraction) with 90% methanol and those extracts were used for different chemical color reaction tests for identification of different phytochemical groups. Phytochemical screening was carried out to assess the qualitative chemical composition of crude extracts using commonly employed precipitation and coloration reaction to identify the major natural chemical groups such as starch, alkaloids, flavonoids, tannins, reducing sugars, amino acids and lignins. General reactions in these analysis revealed the presence or absence of these compounds in crude extract tested (Brindha and Saraswathy, 1981).

III. RESULTS

The preliminary phytochemical screening carried out on methanolic extracts of *Andrographis paniculata*, *Bauhinia acuminata*, *Clerodendrum indicum*, *Nerium odorum* and *Sida humilis* leaves revealed the presence of phytoconstituents such as starch, alkaloids, flavonoids, tannins, reducing sugars, amino acids and lignins (Table-1). The chemical color reaction tests for chemical constituents of leaf extracts of the investigated plants are shown in Figure-2. Methanolic extracts of the leaf of five above mentioned plants have shown positive results for starch, alkaloids, amino acids and lignins. But flavonoids are absent in *Andrographis paniculata* and *Sida humilis*. Tannins are absent in *Andrographis paniculata*, *Bauhinia acuminata* and *Sida humilis*. Reducing sugars are absent in *Sida humilis* only. Leaf extracts of *Nerium odorum* gives faint coloration for starch, alkaloids, flavonoids, tannins and reducing sugars which indicates presence of these in very low quantity (Table-1, Figure-2). On the other hand *Sida humilis* gives light coloration for alkaloids, amino acids and lignins. From Table-1 it is observed that tannins are present in low quantity in the plant *Clerodendrum indicum*.

IV. DISCUSSION AND CONCLUSION

Andrographis paniculata, *Bauhinia acuminata*, *Clerodendrum indicum*, *Nerium odorum* and *Sida humilis* plants were traditional plants, used in herbal medicine. Chemical analysis and biochemical assays are very important aspects in pharmacognostic evaluation of medicinal plants (Choudhury et al., 2009; Harborne, 1973). Through the chemical tests in the methanolic extracts of leaves of the five investigated plants, it is found that the important phytochemical groups (starch, alkaloids, amino acids and lignins etc.) are present in all the cases which actually confirms their medicinal properties (Chandrashekar and Rao, 2012; Anand et al., 2011; Kalaiarasan and John, 2010; Kumar et al., 2010; Patel et al., 2010; Sugumaran and Vetrichelvan, 2008).

It was also found that *Andrographis paniculata* have antimicrobial activity (Anand et al., 2011). It is also observed by Sugumaran and Vetrichelvan (2008) that tannins are absent in ethyl acetate extract of *Bouhinia purpurea* Linn. leaves. In the study prototype results were observed in *B. acuminata*. The leaves of *Clerodendrum viscosum* possess flavonoids, alkaloids, carbohydrates, tannins (Chandrashekar and Rao, 2012), which is similar to the results of chemical analysis in *Clerodendrum infortunatum*. In the present findings, the methanolic extract of leaves of *Sida humilis* revealed that the absence of flavonoids, tannins and reducing sugars in contrast with the result of Kalaiarasan and John (2010), in *Sida cordifolia*.

However, further studies are required in this direction for its comprehensive analysis including qualitative or semi qualitative analysis, characterize its chemical structure and assess its biological activities.

It may be concluded that the above five plants are very useful plant. These plants may be used to cure some common and other various diseases. It is necessary of exploration of maximum potential of these plants in medicinal field and pharmaceutical sciences for their appropriate application.

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Table-1: Phytochemical analysis of leaves of some selected plant species

Sl. No.	Microchemical Tests	Coloration	Selected Plant species				
			Andrographis paniculata	***Bauhinia acuminata***	***Clerodendrum indicum***	***Nerium odorum***	***Sida humilis***
a.	**Starch by weak Iodine solution**	Blue black	+ + +	+ + +	+ +	+	+ +
b.	**Alkaloids by Wagner's reagent**	Dark brown	+ +	+ + +	+ +	+	+
c.	**Flavonoids by 10% NaOH solution**	Yellowish brown	_	+ +	+ + +	+	_
d.	**Tannins by 10% aqueous Lead acetate solution**	Light yellow ppt.	_	_	+	+	_
e.	**Reducing sugars by Benedict's reagent**	Brick red	+ + +	+ +	+ + +	+	_

f.	**Amino acids by Ninhydrin reagent**	Lemon yellow	+ +	+ +	+ + +	+ +	+
g.	**Lignins by Phloroglucinol reagent**	Yellowish orange	+ + +	+ + +	+ + +	+ + +	+

(+, ++, +++ represent degree of intensity of color change i.e. presence of phytochemical groups and

– represents no change of color i.e. absence of phytochemical groups)

Figure-1: Photographs of investigated medicinal plants

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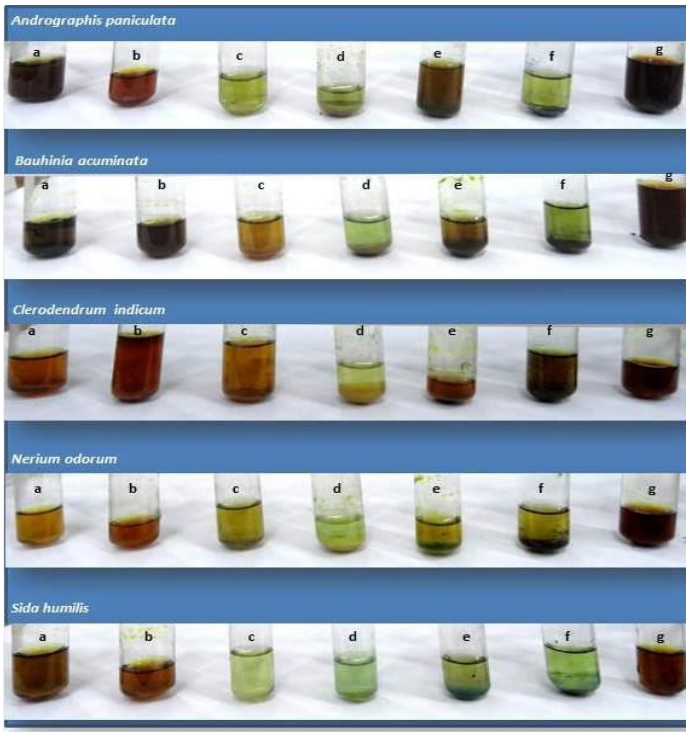

Figure-2: Chemical color reaction tests of leaf extracts of some medicinal Plants.

a: Starch (Blue black), b: Alkaloids (Dark brown), c: Flavonoids (Yellowish brown), d: Tannins (Light yellow ppt.), e: Reducing sugars (Brick red), f: Amino acids (Lemon yellow), g: Lignins (Yellowish orange)

Relative w-projective and dimension

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Abstract- In this paper we generalize the notion of projective, injective and flat modules and dimension. Hence, we introduce and study the notion of Quasi –w projective modules and dimension.

Index Terms- projective, injective and flat module with dimension, Quasi –w projective

I. INTRODUCTION

Throughout this paper , all rings are associative and all modules – if not specified otherwise are left and unitary. Let R is a ring and M be an R-module as usual we use pdR(M) , idR(M) and fdR(M) to denote respectively. The classical projective dimension, injective dimension and flat dimension of M. we use also gldim(R) and wdim(R) to denote respectively , the classical global and Quasi w dimension of R. The character module Homz(M,Q/Z) is denoted by M1.

Recall that a ring R is called left perfect if every flat module is projective.

Example of perfect rings includes Quasi-Frobenius rings or left Artinian. It is shown that a ring is Quasi-Frobenius if and only if every projective module is injective if and only if every injective module is projective. We introduce and study a new generalization of projective and injective modules and dimension. The relation between the Quasi-w projective dimension and other dimension are discussed.

Definition 1.1:- For an R-module M the Quasi w-projective dimension of M qwpdR(M) is defined to be the smallest integer $n \geq 0$ such that $Ext_R^{n+1}(M,M) = 0$ for all flat modules M. If no such integer exist set qwpdR (M) = ∞. If qwpdR(M) = 0 then M will be called a Quasi w-projective module.

Example 1.2:- Consider the local Quasi-Frobenius ring $R = \frac{m[Z]}{Z^2}$ where m is a field and denote by $\bar{Z}$ the residue class in R of Z, then $\bar{Z}$ is a Quasi –w projective R-module which is not projective.

Proof :- Since R is Quasi –Frobenius, every projective and every flat since R is perfect modules M is injective then $Ext_R^i((\bar{Z}, M)) = 0$ for all I >0. Thus $\bar{Z}$ is a Quasi –w-projective R-module Now, if we suppose that $\bar{Z}$ is projective then it must be free since R is local a contradiction since $\bar{Z}^2 = 0$. So we conclude that $\bar{Z}$ is not projective as desired.

In [4] the authors defined and studied a refinement of flat Modules which they called the IF modules. Recall that an R-module if $Tor_R^i(I,M) = 0$ for all right injective R-module I and all I > 0.

Proposition 1.3 :- Let R be a right coherent ring then every Quasi-w projective R-module is an IF R-module.

Proof :- Let M be a Quasi –w projective R-module. Let E be an injective right R-module then $\bar{E}$ is flat then $Ext_R^i(M, \bar{E}) = 0$ for all i > 0. While $Ext_R^i(M, \bar{E}) = (Tor_R^i(\bar{E}, M))$. Hence Tor_R^i (E, M) = 0 thus M is an IF- module.

Proposition 1.4 :- Let M be a Quasi –w projective R-module then

i) $Ext_R^i(M, M^*) = 0$ for all i > 0 and all M* with finite flat dimension.

ii) $Either$ M is Projective or $fd_R(M) = \infty$

Proof :-

i) Since $Ext_R^i(M, M^*) = 0$ for all flat modules M* and all i >0, the proof is immediate by dimension shifting.

ii) Suppose that $fd_R(M) < \infty$ and pick a short exact sequence $0 \to M^* \to P \to M \to \infty$ WHERE P is projective. Clearly $fd_R(M^*) < \infty$, then $Ext_R^1(M, M^*) = 0$ thus the short exact sequence splits and so M is isomorphic to a direct summand of P and then projective.

Corollary 1.5 :- A module M is Quasi –projective if and only if it is flat and Quasi –w projective.

Proof :- Let M be an R-module. The cotorsion dimension of M $cd_R(M)$ is smallest integer n such that $Ext_R^{n+1}(\bar{M}, M) = 0$ for all flat module $\bar{M}$. The left cotorsion dimension of the R , Cot . D(R) is the supremum of cotorsion dimension of R module. It is show in [5, corollary 7.26] that l .cot. D(R) = sup{ qP / $\bar{M}$- flat}.

Proposition 1.6 :- Let M be an R-module and consider the following condition-

1) M is a quasi –w projective module
2) $Ext_R^i(M, P) = 0$ for all i >0 and the projective modules P.
3) $Ext_R^i(M, P) = 0$ for all i >0 and all module P with finite projective dimension.

$Proof$:- (1)$\to$(2) It is trivial

(2)$\leftrightarrow$(3) Result by dimension shifting.

Let $\bar{M}$ be a flat module. By Lazard's Theorem [2 section 1N.6 The. 1] there is a direct system $(L_i)i \in I$ of finitely generated free R-modules such that $\overrightarrow{lim} L_i \cong \bar{M}$

If M is finitely presented from [2, Exercise 3, P-187] we have $Ext^i_R(M,\bar{M}) \cong \overrightarrow{\lim} Ext^i_R(M,L_i)$. Thus in this case the implication (3) →(1) holds.

If l .cot . D (R) < ∞ then $qpd_R(\bar{M}) < \infty$. Hence, in this case also the implication (3)→(1) holds.

Proposition 1.7 :- The following statements are equivalent-
1) R is left perfect
2) Every flat module is quasi w-projective.

In particular if the class of all quasi-w projective modules are closed under direct limits then R is left Perfect

Proof :- If R is left perfect it is clear then every flat module is quasi-w projective. As t o the converse let $\bar{M}$ be a flat module. By (1) it is quasi-w projective and so projective by proposition 2.4. then R is left perfect. If the class of all quasi –w projective module is closed under direct limits then any direct limit of projective modules is both flat and quasi-w projective since every projective module is both flat and quasi-w-projective then by corollary 1.5 every direct limit of projective modules is projective. Thus R is left perfect.

Proposition 1.8 :- The following are equivalent –
(1) Every R-module is quasi-w projective
(2) R is quasi- Frobenius

Proof :- This follows from the fact that a ring is quasi-Frobenius if and only if every projective module is injective and that is quasi- Frobenius rings are Perfect.

A left (right) R-module M is said FP- injective if Ext^1_R (M,M) = 0 for every finitely presented left (right) R-module M.

A ring R is said to be FC if it is left and right coherent and left right self FP- injective.

Proposition 1.9 :- The following are equivalent-
1) R is FC
2) Every finitely presented (left and right) module is quasi w-projective.

Proof :- Let M be a finitely presented right or left module and $\bar{M}$ be a flat right or left module then $\bar{M}$ is FP-injective by [8 , lemma 4.1], So , $Ext^i_R(M,\bar{M})$ = 0 for i >0. Thus M is quasi w-projective.
As to converse for any finitely presented right or left module M we have $Ext^i_R(M,R)$= 0 for all i >0 by 2 thus R is self right and left FP- injective.

Proposition 1.10 :- For any R-module M and any positive integer n the following assertions are equivalent –
1) qwpdR(M)$\leq$ n
2) $Ext^i_R(M,\bar{M})$ = 0 for all i>n and a;; R-module $\bar{M}$ with finite flat dimension.
3) $If\ 0 \rightarrow G_n \rightarrow G_{n-1} \rightarrow \ldots\ldots\ldots\ldots \rightarrow G_0 \rightarrow M \rightarrow 0$

Is an exact sequence of modules withG_0……G_{n-1} are quasi-w projective modules then Gn is a quasi w-projective module.

Proof :- (1) ↔ The Proof of this equivalence is standard homological algebra.
(2)→(1) obvious

(1) →(2) Set P=qwpdR(M). By induction on m=$fd_R(\bar{M})$ we prove that $Ext^i_R(M,\bar{M})$ =0 for all i>P. The induction start is given by (1). If m>0 pick the short exact sequence $0 \rightarrow \bar{M} \rightarrow P \rightarrow \bar{M} \rightarrow 0$ where P is a projective module. Clearly fdR ($\bar{M'}$) =m-1. Thus $Ext^i_R(M,\bar{M})$ =0 for all i>n. from the long exact sequence…… $\rightarrow Ext^i_R(M,P) \rightarrow Ext^i_R(M,\bar{M}) \rightarrow Ext^{i+1}_R(M,\bar{M'}) \rightarrow \ldots\ldots\ldots\ldots$

It is clear that $Ext^i_R(M,\bar{M}) = 0$ for all i >n.

Proposition 1.11 :- For any R module M qwpdR (M)≤ pd_R(M) with equally if fd_R(M) is finite.

Proof :- The first inequality follows from the fact that every projective module is quasi-w projective. Now set $qwpd_R(M) = n < \infty$ and consider an n-step projective resolution of M as follows $0 \rightarrow M' \rightarrow Pn-1 \rightarrow \ldots\ldots\ldots\ldots \rightarrow P1 \rightarrow P0 \rightarrow M \rightarrow 0$ where all Pi are projective. Clearly M' is quasi w-projective. If $fd_R(M) < \infty$ then $fd_R(M') < \infty$ and then it is projective by proposition 1.4. Hewnce $Pd_R(M) \leq n$, and so the equality holds.

[2] Quasi w- projective dimension of rings :

Definition 2.1 :- The left quasi –w projective dimension of a ring R , l .qwpd(R) is defined by setting l .qwpd(R)= sup { qwpd(M) / M is a left R-module.

Theorem 2.2 :- Let R be a ring and n be a positive integer then the following are equivalent –
(1) l.qwpd(R) $\leq n$.
(2) Qwpd(R/I) $\leq n$ For every left ideal I of R
(3) idR($\bar{M}$) $\leq n$ for all projective module P.

Proof :- (1)→(2) and (3)→(4) are obvious.
(2)→(3) Let $\bar{M}$ be a flat module. Since qwpdR(R/I) $\leq n$. For every ideal I of R we have $Ext^i_R(R/I,\bar{M})$ = 0 for all i>n. Thus using the Baer Criterion $id_R(\bar{M}) \leq n$.

(4)→(1) Let M be an arbitrary module. Since idR(P) $\leq n$ for each projective module P we have $Ext_R^i(M,P)$= 0 for all >n and projective module P. By dimension shifting we get that $Ext_R^i(M,P)$= 0 for all i>n and all module P with finite projective dimension. l.cot. D(R) $\leq$ sup.{id_R(M)/ P projective} $\leq n$. Thus given a flat module $\bar{M}$ we have $qpd_R(\bar{M}) < \infty$. Hence $Ext_R^i(M,\bar{M})$= 0 for all i>n. consequently qqwpd$_R$ (M) $\leq n$.

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Relative w-projective and dimension

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iii) $Ext_R^i(M,M^*)$ = 0 for all i > 0 and all M* with finite flat dimension.

iv) $Either$ M is Projective or $fd_R(M)=\infty$

Proof :-

i) Since $Ext_R^i(M,M^*)$=0 for all flat modules M* and all i >0, the proof is immediate by dimension shifting.

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4) Every flat module is quasi w-projective.

In particular if the class of all quasi-w projective modules are closed under direct limits then R is left Perfect

Proof :- If R is left perfect it is clear then every flat module is quasi-w projective. As t o the converse let $\bar{M}$ be a flat module. By (1) it is quasi-w projective and so projective by proposition 2.4. then R is left perfect. If the class of all quasi –w projective module is closed under direct limits then any direct limit of projective modules is both flat and quasi-w projective since every projective module is both flat and quasi-w-projective then by corollary 1.5 every direct limit of projective modules is projective. Thus R is left perfect.

Proposition 1.8 :- The following are equivalent –

(3) Every R-module is quasi-w projective
(4) R is quasi- Frobenius

Proof :- This follows from the fact that a ring is quasi-Frobenius if and only if every projective module is injective and that is quasi- Frobenius rings are Perfect.

A left (right) R-module M is said FP- injective if Ext_R^1 (M,M) = 0 for every finitely presented left (right) R-module M.

A ring R is said to be FC if it is left and right coherent and left right self FP- injective.

Proposition 1.9 :- The following are equivalent-

3) R is FC
4) Every finitely presented (left and right) module is quasi w-projective.

Proof :- Let M be a finitely presented right or left module and $\bar{M}$ be a flat right or left module then $\bar{M}$ is FP-injective by [8 , lemma 4.1], So , $Ext_R^i(M,\bar{M})$ = 0 for i >0. Thus M is quasi w-projective.

As to converse for any finitely presented right or left module M we have $Ext_R^i(M,R)$= 0 for all i >0 by 2 thus R is self right and left FP- injective.

Proposition 1.10 :- For any R-module M and any positive integer n the following assertions are equivalent –

4) $qwpdR(M)\leq$ n
5) $Ext_R^i(M,\bar{M})$ = 0 for all i>n and a;; R-module $\bar{M}$ with finite flat dimension.
6) $If\ 0 \rightarrow G_n\rightarrow G_{n-1}\rightarrow\ldots\ldots\ldots\ldots\rightarrow G_0\rightarrow M\rightarrow 0$

Is an exact sequence of modules with$G_0\ldots\ldots\ldots\ldots G_{n-1}$ are quasi-w projective modules then Gn is a quasi w-projective module.

Proof :- (1) ↔ The Proof of this equivalence is standard homological algebra.

(2)→(1) obvious

(1) →(2) Set P=qwpdR(M). By induction on $m=fd_R(\bar{M})$ we prove that $Ext_R^i(M,\bar{M})$ =0 for all i>P. The induction start is given by (1). If m>0 pick the short exact sequence $0\rightarrow\bar{M}\rightarrow P\rightarrow\bar{M}\rightarrow 0$ where P is a projective module. Clearly $fdR(\bar{M}')$ =m-1. Thus $Ext_R^i(M,\bar{M})$ =0 for all i>n. from the long exact sequence…… $\rightarrow Ext_R^i(M,P)\rightarrow Ext_R^i(M,\bar{M})\rightarrow Ext_R^{i+1}(M,\bar{M}')\rightarrow\ldots\ldots\ldots\ldots$

It is clear that $Ext_R^i(M,\bar{M}) = 0$ for all i >n.

Proposition 1.11 :- For any R module M qwpdR (M)≤ $pd_R(M)$ with equally if $fd_R(M)$ is finite.

Proof :- The first inequality follows from the fact that every projective module is quasi-w projective. Now set $qwpd_R(M) = n < \infty$ and consider an n-step projective resolution of M as follows $0 \to M' \to Pn\text{-}1 \to \ldots\ldots\ldots\ldots \to P1 \to P0 \to M \to 0$ where all Pi are projective. Clearly M' is quasi w-projective. If $fd_R(M) < \infty$ then $fd_R(M') < \infty$ and then it is projective by proposition 1.4. Hewnce $Pd_R(M) \leq n$, and so the equality holds.

[2] Quasi w- projective dimension of rings :

Definition 2.1 :- The left quasi –w projective dimension of a ring R , l .qwpd(R) is defined by setting l .qwpd(R)= sup { qwpd(M) / M is a left R-module.

Theorem 2.2 :- Let R be a ring and n be a positive integer then the following are equivalent –

(4) l.qwpd(R) $\leq n$.
(5) Qwpd(R/I) $\leq n$ For every left ideal I of R
(6) idR($\bar{M}$) $\leq n$ for all projective module P.

Proof :- (1)→(2) and (3)→(4) are obvious.
(2)→(3) Let $\bar{M}$ be a flat module. Since qwpdR(R/I) $\leq n$. For every ideal I of R we have $Ext_R^i(R/I, \bar{M})$ = 0 for all i>n. Thus using the Baer Criterion $id_R(\bar{M}) \leq n$.
(4)→(1) Let M be an arbitrary module. Since idR(P) $\leq n$ for each projective module P we have $Ext_R^i(M,P)$= 0 for all >n and projective module P. By dimension shifting we get that $Ext_R^i(M,P)$= 0 for all i>n and all module P with finite projective dimension. l.cot. D(R) $\leq$ sup.{id_R(M)/ P projective} $\leq n$. Thus given a flat module $\bar{M}$ we have $qpd_R(\bar{M}) < \infty$. Hence $Ext_R^i(M,\bar{M})$= 0 for all i>n. consequently $qqwpd_R(M) \leq n$.

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Regional Variation in Elastic Fibers on Different Regions of the Skin in Neonatal Goat (Capra hircus)

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Abstract- The present study was conducted in 10 neonates (birth to 3 months of age) indigenous goat. With orcein the elastic fibers stained dark brown in the reticular layer of dermis. These were sparsely distributed in the papillary layer. They are rare and finely branched sparsely distributed in the papillary layer and arranged perpendicular to skin surface, abundantly present in between and around hair follicles but sparse around sweat glands. The fibers are found intermingled in arrector pili muscles. Mostly present superficial to the level of the sweat glands. Regional difference reported is common in all samples that elastic assimilation was intensely found in neck and ventral region of the skin. The fibers were found moderate to weak in other body regions such as dorsal, thigh and flank.

Index Terms- Neonates, Dermis, Elastic, Papillary & Reticular layer.

I. INTRODUCTION

Goat fiber production is affected by genetic and environmental influences. Environmental influences include bio–geophysical factors photoperiod, climate–herbage system and soil–plant trace nutrient composition, nutrition factors and management factors. Nutrition and management directly influences rate of stocking, supplementary feeding of energy and protein, live weight change, parturition and management during rearing along with health.

Elastin, a major fibrillar protein component of various connective tissues, is critical for the normal physiological function of organs such as skin, lungs, and the blood vessels (Sandberg, Ryhanen & Foster 1981-82). Elastic fibers impart elasticity to tissues such as skin, lungs, ligaments and arterial walls.

In the present study skin samples were processed in the laboratory to see the regional changes in elastic fibers. The mammalian skin fibers represent an interesting biological material and also are in use in various industries.

II. MATERIALS AND METHODS

The present study was conducted in 10 neonates (birth to 3 months of age) indigenous goat. The skin samples were collected from different regions of the body of neonatal goat- dorsal, ventral, thigh, flank and neck region were taken with the help of razor blade, scissors and forceps. The tissues were fixed in 10% neutral buffered formalin solution for 36 to 48 hours. (Lillie and Fullmer, 1976; and Drury and Wallington, 1980). The tissues were then processed in laboratory by adopting standard methods (Drury and Wallington, 1980) of dehydration, clearing and embedding. They were processed and sectioned using routine histological procedures. The paraffin tissue sections of 3-5 um thickness were stained with orcein stain to study elastic fibers configuration.

III. RESULTS AND DISCUSSION

Elastic fibers as reported in the study are found in high intensity in the neck and ventral region as manifested with the working caliber of elasticity.

Elastic fibers are the 'rubberband' fibers responsible for the 'snap back' quality of young skin. As described by Dr. Peter Pugliese M.D., elastic fibers are important 'youth Protein' fibers of skin. They give young, smooth, firm skin its resiliency - its retraction and 'snap'. Elastic fibers are considered as the skin's very important "rubberbands."

With orceine the elastic fibers were stained dark brown in the reticular layer of dermis.

Regional differences reported are general in all samples that elastic assimilation was highest (intense) found is neck and ventral region of the skin. The fibers were found moderate to weak in other body regions such as dorsal, thigh and flank.

The elastic was abundantly present in between and around hair follicles but sparse around sweat glands. They intermingled in the arrector pili muscles. The elastic fibers were mostly presented superficial to the level of the sweat glands. As reported in other structural aspects the dermis did not show much structural variation in different regions of the body.

The arrangement and size of the hair follicles varied in different regions of the body in neonatal goat. In neck region the follicles were oval to elongate in shape, arranged in two rows superficial and middle, mostly in group of three, which were oriented parallel to the surface and well clinched with elastic fibers (fig 4). These were nearly of equal size at places shows an oblique radial arrangement. In dorsal and flank region the linear arrangement was more conspicuous well supported by elastic fibrils.

The sebaceous glands in all the skin regions were simple branched alveolar types. These glands were always associated with hair follicles and located just above the sweat glands, between hair follicles, arrector pili muscle and skin fibers, presented only in upper part of papillary layer of dermis in all the skin regions.

The sweat glands as reported were sparsely layered with elastic fibers, tubular, deeply located in the dermis below the sebaceous gland.

Table 1. Histological analysis of cutaneous elastic in neonatal goat

Neonates	Neck	Dorsal	Thigh	Ventral	Flank
1	+++	++	+++	+++	++
2	+++	+++	++	++	++
3	++	+++	+++	+++	++
4	+++	++	++	++ +	+++
5	+++	++	++	+++	++
6	+++	++	++	++	++
7	+++	++	++	+++	++
8	+++	++	++	+++	+++
9	++	+++	++	+++	++
10	+++	++	++	+++	+++

+++ Intense, ++ Moderate

PHOTOGRAPHS

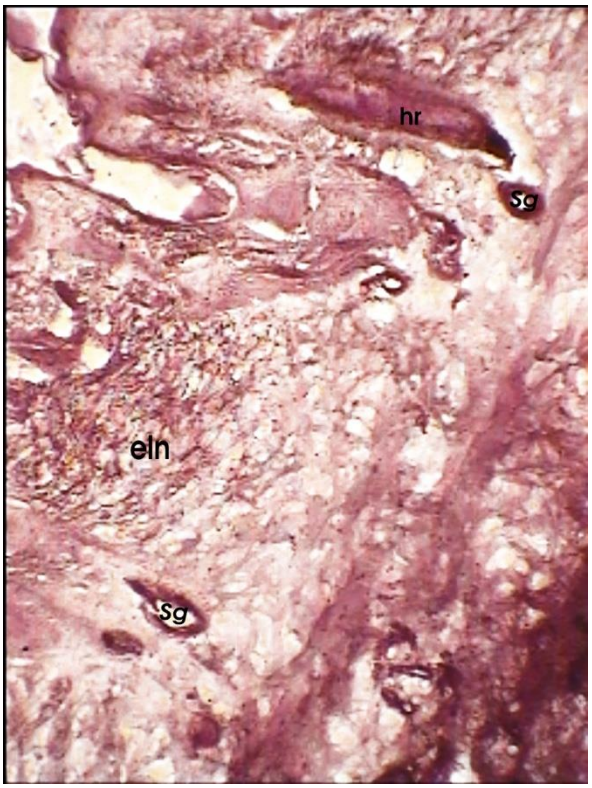

Fig 1. Photomicrograph – cross section of skin of Neonatal goat, dorsal, orcein stain Showing hr-hair, sg-sweat gland, eln- elastic fibres.

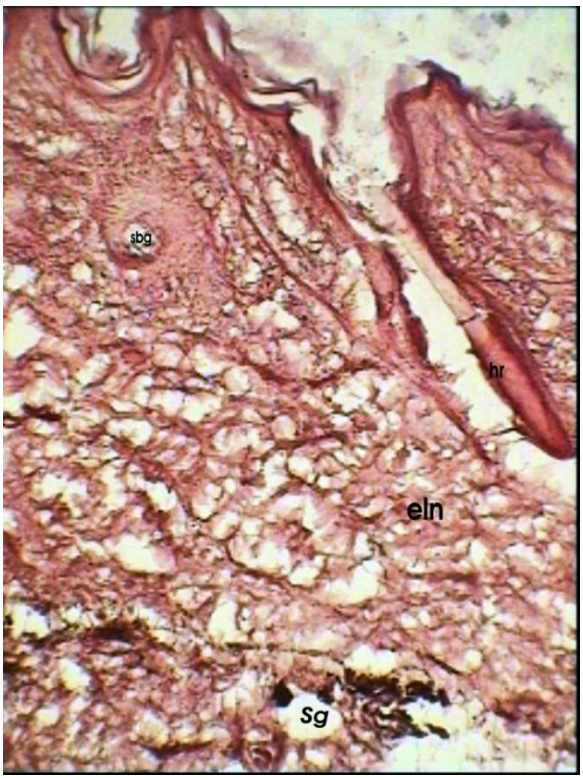

Fig 2. Photomicrograph – cross section of skin of Neonatal goat, ventral, orcein stain Showing hr-hair, sg-sweat gland, eln- elastic fibres & sbg-sebaceous gland.

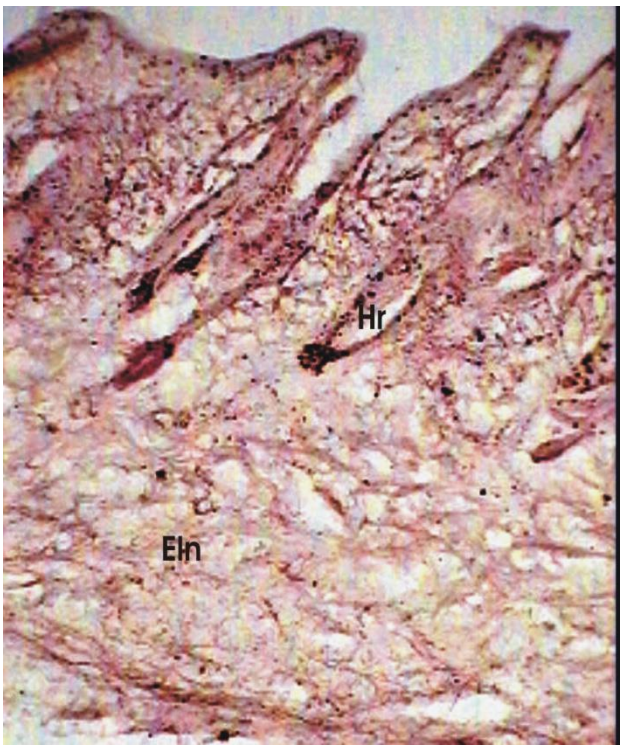

Fig 3. Photomicrograph – cross section of skin of Neonatal goat, Thigh, orcein stain Showing hr-hair, sg-sweat gland, eln-elastic fibres.

Fig 4. Photomicrograph – cross section of skin of Neonatal goat, Neck, orcein stain Showing hf-hair follicle, sg-sweat gland, eln-elastic fibres & sbg-sebaceous gland.

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Supplementation of Concentrate/Legume Hay with Maize Silage on Nutrient Utilization and Nitrogen Balance in Nellore Ram Lambs

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Abstract- An on-farm experiment was conducted with Nellore ram lambs by feeding intensively for five months period with sole maize silage (R-I), silage + concentrate at 0.5 per cent body weight (R-II), silage + concentrate at 1.0 per cent body weight (R-III), silage + concentrate at 1.5 per cent body weight (R-IV), silage + lucerne hay (R-V) and silage + GN haulms (R-VI) to develop a feeding system based on maize (Zea mays) silage with supplementation of concentrate mixture and or legume hay/straw at appropriate level to study the nutrient utilization and nitrogen metabolism and compared with the performance of ram lambs fed sweet sorghum bagasse based complete diet (R-VII). The DMI (g/d) was significantly (P<0.01) higher by 25.19, 29.74, 34.82, 31.04, 27.6 and 26.56 per cent respectively with R-II, R-III, R-IV, R-V, R-VI and R-VII rations in comparison to R-I ration (sole silage). The DMI (g/d) was almost similar in lambs fed rations R-III, R-V and R-VI. The lowest DMI (g/d) was observed in ram lambs fed R-I ration. The DMI (g/kg w0.75) in ram lambs was also significantly (P<0.01) different among the seven rations and was highest in ration R-III and lowest in ration R-I in comparison to the other rations. Almost similar DMI/kg w0.75 was observed in ram lambs fed rations R-II, R-IV, R-V and R-VI. Mean digestibility coefficients of DM, OM, CP, EE, CF and NFE in ram lambs fed R-I, R-II, R-III, R-IV, R-V, R-VI and R-VII rations were non significantly different among the seven rations. Average NDF, ADF, hemicellulose and cellulose digestibilities were comparable and insignificantly increased as the level of concentrate increased in the ration. The N intake (g/d) was significantly (P<0.01) higher by 41.14, 52.82, 61.87, 51.93, 37.72 and 64.94 per cent, respectively with R-II, R-III, R-IV, R-V, R-VI and R-VII rations in comparison to R-I ration. The N out go through faeces and urine in ram lambs was significantly (P<0.01) different among the seven rations. The N balance (g/d) in lambs was either significantly or non significantly higher by 41.24 (>0.05), 62.38 (P>0.05), 76.59 (P<0.01), 73.24 (P<0.01), 56.49 (P>0.05) and 72.99 (P<0.01) per cent, respectively with R-II, R-III, R-IV, R-V, R-VI and R-VII rations in comparison to R-I ration. Based on the results of this study it is concluded that, supplementation of concentrate and or legume hay did not affected the digestibilities of nutrients. However affected the nitrogen utilization in Nellore ram lambs.

Index Terms- concentrate, legume hay, maize silage, nutrient utilization, nitrogen balance

I. INTRODUCTION

Maize (*Zea mays*) is the nutritious feed for small and large ruminants have high protein efficiency ratio (PER), relatively high digestible energy (DE) and total digestible nutrients and thus maize fodder can play an important role in supplying animal feed throught the year if we cultivate them. (Desai, and Deore, 1984). Feeding of silage based rations is becoming popular among the farmers rearing sheep on commercial basis in India particularly in Andhra Pradesh and Karnataka. However, a feeding system based on silage needs to be developed for rearing of ram lambs on commercial basis since literature on silage feeding in ram lambs is limited. Silage, which is anaerobically fermented green fodder, is valued throughout the world as a source of animal feed during lean months (Ragothaman Venkataramanan *et al.,* 2010). Maize is the third most important cereal crop of the world. It is used as food, feed and forage. Maize fodder can safely be fed at all stages of growth without any danger of oxalic acid, prussic acid as in case of sorghum or other fodders. Therefore, green maize fodder is referred as 'king of crops' suitable for good silage making (Muhammad *et al.,* 1990). Very limited numbers of sheep farmers are feeding their ruminants with silage in India. Farmers in Andhra Pradesh are showing interest in preparing and feeding of silage to their ruminant animals particularly small ruminants like growing sheep to obtain optimum body weight.

In view of the farmers' awareness on feeding of silage to small ruminants for meat production, an attempt was made to feed the maize silage by supplementing concentrate and legume hay at certain levels to study the nutrient digestibilities and nitrogen metabolism in growing Nellore ram lambs.

II. MATERIALS AND METHODS

This on-farm experiment was carried out with maize silage at Indugula village in Tipparthy mandal of Nalgonda district, Andhra Pradesh which is 140 km away from Hyderabad. The main source of irrigation for food crops or forage crops here is by rains, bore wells and small tanks. Average rainfall was 50-60 mm per annum and occurs chiefly due to southwest monsoons every year from June to September. Soil is of mostly red (chalka) type. The experimental animals faced the maximum environmental

temperature of about 44^0C in the month of May and the minimum was about 23^0C during the entire experimental period.

Silage making

The farmer who reared the experimental ram lambs was having 20 acres of land with bore well as well as sprinkler irrigation facility. The farmer was already rearing 100 sheep which were being maintained by grazing on open fields. The land used to grow the maize crop for silage making was ploughed thoroughly for 3 times with a tractor and about 2 tons of farm yard manure per acre was applied as a basal fertilizer. Nitrogen, phosphorus and potassium fertilizers were purchased from local market and applied at the rate of 60, 24 and 10 kg per acre, respectively. A 36V92 variety of maize seed was selected and purchased from Pioneer seeds Pvt. Ltd., Hyderabad for growing the maize fodder for silage making. Physical or chemical treatment of seed was not done prior to sowing.

Sowing, Fertilizer application and Irrigation

Seed rate was 8.5 kg/acre. While sowing with hand, the distance maintained between the rows was 1.5 feet and between the seeds was 7 inches. About 5 g of urea fertilizer was placed along with a maize seed while sowing. Corbofuron granules were placed in the rows after sowing to prevent the infestation by insects after germination. Maize seeds were sowed in the winter season (15^{th} November, 2009). Irrigation was done by sprinklers for first 60 days followed by bore well once in 15 days till harvest. Herbicide (Atrazin @ 1.5 litres in 200 litres of water per acre) was used on the next day of sowing to prevent the growth of weeds in the field. Monocrotophos an insecticide was sprayed over the growing fodder on 40^{th} day. Booster dose of urea was applied on 15^{th}, 30^{th} and 40^{th} day.

Harvesting of Maize Fodder

Maize fodder was harvested on 87^{th} day after sowing when the cob containing one fourth to half milky grains. Harvested green fodder was allowed to wilt in the field for 4-6 h to reduce the moisture content to around 65 per cent.

Construction of Silo

Two silo pits were constructed near the experimental animal shed with the dimensions of 9'L x 9'W x 8'H so as to accommodate about 10 tons of silage in each pit. All the inside walls and bottom of silo was cemented to prevent seepage of ground water if any. All the sides of the silo were covered with HDPE plastic cover before filling the pit with chopped maize fodder.

Chopping

Harvested and wilted green fodder was brought to the site of silage pit from the field by using a tractor. The 10 HP motor capacity chop cutter was arranged at one edge of silo in order to allow the chopped green fodder directly to fall into the pit. About 5-6 whole maize plants were kept in the chop cutter at a time so as to cut the fodder to a size of ½ to ¾ inches.

Silage Additives

Sugarcane molasses, urea (fertilizer grade) and common salt were added at 1, 0.5 and 0.5 per cent, respectively while making the silage. They were mixed in water (50 litres/ton of fodder) in a plastic drum thoroughly with a stick and were sprinkled uniformly all over the maize fodder while chopping it in a chop cutter.

Compacting and sealing

Chopped green maize fodder was trampled (compacted) with wooden planks by two persons for every one foot level in the pit Great care was taken while trampling chopped silage fodder to prevent trapping of air in the pit so as to maintain strict anaerobic environment in the silo. After filling the chopped fodder to about 2 feet's above the ground level it was tightly covered with HDPE plastic covers and heavy weight bags filled with sand were kept over the pit to prevent entry of air and water into silo. Pit slope to one side was maintained to drain water quickly if rain occurs. Silo was opened on 39th day after sealing for the feeding of experimental ram lambs.

Preparation of Concentrate Mixture

Concentrate feed ingredients were procured from the local Hyderabad market. Concentrate mixture with 17% CP and 70% TDN (Table 3) was prepared in the feed mill located at Department of Animal Nutrition, College of Veterinary Science, Rajendranagar, Hyderabad and transported to Indugula village for feeding of ram lambs.

Lucerne (*Medicago sativa*) crop was grown at the farmer's fields at Indugula village and was harvested at 50-60% flowering stage. Harvested green lucerne fodder was dried under shade for 3-4 days in order to contain 12-14% moisture. While drying, the fodder was turned upside down three times a day to prevent growth of fungus and to hasten the process of drying. Prepared hay was filled in the gunny bags and stored for feeding of experimental ram lambs. Groundnut (*Arachis hypogea*) haulms were purchased (Rs. 3.00 per kg) from another farmer in Indugula village for the feeding of experimental ram lambs. Sweet sorghum bagasse (SSB) was procured from ICRISAT, Patancheru, Hyderabad to prepare complete ration. Concentrate ingredients were purchased from local market in Hyderabad. Complete ration with 50 per cent level SSB was processed into mash according to the formula (Table 3) using hammer mill through 8 mm sieve at the feed mill of Department of Animal Nutrition, College of Veterinary science, Rajendranagar, Hyderabad and was transported to Indugula village to feed the experimental ram lambs.

Selection and Grouping of Ram Lambs

Forty nine 3-4 months old growing Nellore ram lambs with an average body weight of 14.26±0.24 kg were purchased from Karimnagar, Karimnagar district of Andhra Pradesh and were randomly distributed into seven groups of seven animals each. The average body weight (kg) of ram lambs in seven treatments was 14.33±0.85 (T_1), 14.33±0.47 (T_2), 14.20±0.56 (T_3), 14.23±0.87 (T_4), 14.30±0.46 (T_5), 14.32±0.62 (T_6) and 14.10±0.86 (T_7), respectively.

Experimental Silage Rations

Seven experimental groups were fed with respective rations as mentioned below for a period of five months.

The first group (T_1) of growing Nellore ram lambs was fed sole maize silage *ad libitum.*

The second group (T_2) animals were fed concentrate mixture @ 0.5 per cent of body weight + maize silage *ad libitum.*

The third group (T_3) animals were fed concentrate mixture @ 1.0 per cent of body weight + maize silage *ad libitum.*

The fourth group (T_4) animals were fed concentrate mixture @ 1.5 per cent of body weight + maize silage *ad libitum.*

The fifth group (T_5) of ram lambs was fed lucerne hay to meet 25 per cent of dry matter requirement and maize silage *ad libitum*.

The sixth group (T_6) of ram lambs was fed groundnut haulms (straw) to meet 25 per cent of dry matter requirement and maize silage *ad libitum*.

The seventh group (T_7) ram lambs were fed solely on SSB based complete ration (50:50).

Housing, Feeding Watering and Management

The ram lambs were housed according to groups in well ventilated, clean pens (24'LX10'W) with an open area (24'LX10'W) for movement during the day time. Ordinary flooring (soil) was maintained in the pens. All the experimental ram lambs were offered their respective feeds at 9.00 and 15.00 h by weighing on an electronic digital balance and residue if any was weighed after 24 h. The growth trial was conducted for a period of 150 days. All the experimental animals were offered clean, fresh drinking water round the clock. Hygienic surroundings were maintained throughout the experimental period. All the animals were treated for external and internal parasites with Ivermectin and Fenbendazole drugs, respectively, in the beginning as well as after three months of experimental period. Animals were vaccinated against PPR disease after seven days of first deworming.

Metabolism study

Digestion cum metabolic studies were conducted at the end of the growth trial in ram lambs to assess the nutrient utilization and nitrogen and energy balance of the experimental rations. The animals were kept in clean, well ventilated individual metabolic cages (40'' length, 26'' width) with feeding and watering arrangement during the metabolic trial. Animals were shifted to metabolic cages 3 days prior to collection period to acclimatize them to metabolic cage environment.

The collection period lasted for seven days. During the collection period the daily feed consumption, leftover as well as faeces and urine voided were recorded at 9.00 h before feeding. During the period of metabolism trial 24 h collection of faeces was made using faecal bags harnessed to the ram lambs. The daily urine output of each lamb was measured by collecting urine in glass bottles kept at the bottom of the metabolic cages. Few drops of toluene was added to the urine collection bottles daily to avoid nitrogen loss.

Collection of Samples

Representative samples of each feed offered and residues were collected and pooled for 7 days. Daily DM was estimated from respective samples and were pooled and preserved for estimation of other nutrients. The samples of all the experimental feeds and leftover after drying were ground separately in a laboratory Wiley mill through a 1 mm screen and preserved in air tight bottles for subsequent analysis.

Faeces from each animal was collected in separate containers, weighed, mixed thoroughly and aliquoted for dry matter and nitrogen estimation..

For dry matter, aliquots of $1/10^{th}$ of daily faeces voided by each animal was taken in previously weighed petri dishes and dried overnight in hot air oven at 100 ± 5^{o}C. The daily sample from each animal for seven day collection period was pooled, ground in Wiley mill through a 1 mm screen and stored in polythene bags for further analysis.

For faecal nitrogen estimation $1/100^{th}$ part of faeces voided each day by individual animal was weighed and frozen in refrigerator for further analysis.

For nitrogen estimation, $1/20^{th}$ part of total urine voided daily by individual animal, after thorough mixing, was pipetted out in duplicate into Kjeldahl flasks containing 30 ml of concentrate sulfuric acid. The aliquots, thus pooled in the flasks, were maintained separately for each animal.

Proximate Analysis

Dry matter, crude protein, crude fiber, either extract or total ash were estimated according to procedure nos. 4.1.03, 4.2.02, 4.6.01, 4.5.01 and 4.1.02, respectively as described by AOAC (1997).

Crude protein estimation

A known quantity of the ground sample (appropriate aliquots, in case of wet faeces and urine) was digested with suitable quantity of concentrated H_2SO_4 in the presence of catalytic digestion mixture ($CuSO_4$ and K_2SO_4 in 1:10 ratio) by using Turbotherm (Gerhardt, Germany). An acid blank was also run along with the samples for correction of any N contribution by the acid itself. The digested sample was then quantitatively transferred in to a volumetric flask with repeated washing with distilled water. The N content of the sample was estimated by distilling a suitable aliquot into an auto analyzer (Vapodest, Gerhardt, Germany). The N content multiplied by the factor 6.25 gave the CP content of the sample, which was expressed as percentage on DMB. Fibre fractions in feeds, faeces and residues were performed as per the method described by Van Soest *et al.* (1991).

Statistical analysis of the data was carried out according to the procedures suggested by Snedecor and Cochran (1994). Analysis of variance was utilized to test the significance of various treatments and the difference between treatment means was tested for significance by Duncan's Multiple Range and F Test (Duncan, 1955).

III. RESULTS AND DISCUSSION

Chemical Composition

The chemical composition of experimental rations of maize silage fed to growing Nellore ram lambs is presented in the Table 1. The per cent DM, OM, CP, EE, CF, NFE, TA, NDF, ADF, hemicellulose, cellulose and lignin values were 29.86, 93.07, 5.49, 2.42, 23.5, 61.66, 6.93, 58.72, 41.9, 16.82, 22.15 and 7.58 for maize silage; 93.51, 89.11, 17.27, 3.59, 8.23, 60.02,

10.89, 32.05, 13.32, 18.73, 7.09 and 3.11 for concentrate mixture (17 per cent CP and 70 per cent TDN); 89.33, 90.28, 15.06, 2.14, 32.49, 40.59, 9.72, 57.66, 39.72, 17.94, 30.17 and 4.95 for lucerne hay; 91.46, 88.09, 9.23, 1.20, 36.72, 40.94, 11.91, 69.54, 46.70, 22.84, 30.52 and 6.57 for groundnut haulms; 92.67, 89.23, 11.51, 1.75, 28.64, 47.33, 10.77, 58.28, 31.29, 26.99, 24.08 and 4.35 for sweet sorghum bagasse (SSB) based complete ration (50R:50C), respectively on DM basis. The DM, OM, and CP content of whole crop maize silage prepared at two third milk line stage used in the present study was similar as observed by Shaver *et al.* (1984) whose values were 33.6, 95.9 and 6.8 per cent, respectively, for DM, OM and CP in the corn silage and the similar values were also recorded by Filya (2004) for CP (6.50%), Hemicellulose (18.30%) and cellulose (24.50%) in whole crop maize silage that the crop was harvested at two third milk line stage and the ash content (4.10) was higher in the present study maize silage. This higher ash content in the present maize silage might be due to tropical climatic effect and soil condition. Reddy and Reddy (1988) reported the chemical composition of maize silage for feeding of adult sheep as DM 30.67, OM 89.34, CP 6.94, CF 29.43, EE 0.96, NFE 52.01 and TA 10.66 per cent on DMB.

The chemical composition of maize silage reported by Marina *et al.* (2007) for sheep was DM (26.4), OM (95.5), CP (6.2), NDF (58.2), ADF (32.1) per cent, on DMB and these results were almost similar to the present study maize silage fed to Nellore ram lambs. Almost the same proximate composition of maize silage was observed by Rowghani *et al.* (2008) and Sohail *et al*. (2010).

Rama Prasad *et al*. (1999) reported the proximate composition of groundnut haulms as 86.2 (OM), 10.6 (CP), 30.70 (CF), 2.6 (EE) and 32.3 (NFE) per cent on DMB. In another study Mandal *et al*. (1999) reported the chemical composition of groundnut haulms as 27.44, 11.81, 1.5, 20.53, 46.7 and 19.46 per cent, respectively for DM, CP, EE, CF, NFE and TA. Vara Prasad *et al*. (2000) reported the chemical composition of groundnut haulms as 13.1, 21.2, 40.2, 44.9, 38.8 and 29.8 for CP, CF, NFE, NDF, ADF and cellulose, respectively on DMB. Chemical composition of GN haulms was reported by Murthy *et al*. (2001) for CP, EE, CF, NFE and TA as 13.12, 3.22, 31.86, 38.95 and 12.85 per cent, respectively on DMB. Jatinder Singh *et al.* (2009) reported the CP, CF, EE, total ash and NFE as 5.30, 33.15, 1.33, 6.37 and 53.85 per cent, respectively on per cent DM in groundnut haulms.

Ranjhan (1998) reported the chemical composition (%) of lucerne hay as DM (85), CP (21.3), CF (29.4), NFE (35.2), EE (1.4), ash (12.7), NDF (43.6) and ADF (35.8). Chemical composition of lucerne hay as reported by Sanjivkumar and Bhatt (2000). contained 85.60, 19.10, 21.60, 2.80, 42.10 and 14.40 per cent of OM, CP, CF, EE, NFE and TA, respectively on DMB. Wildeus *et al*. (2007) reported the chemical composition of alfalfa hay in terms of CP, NDF and ADF as 15.2, 70.2 and 41.5 per cent, respectively on DMB. The same author in same publication he was also reported 16.8, 64.3 and 47.3 per cent of CP, NDF and ADF, respectively on DMB in another experiment.

Metabolism Study

Dry matter intake

The dry matter intake (DMI) by ram lambs fed maize silage and SSB based rations is presented in Table 2. The DMI expressed as g/d or g/kg $w^{0.75}$ was 608.40±26.61, 68.91±3.53; 813.24±36.07, 79.38±4.51; 865.89±24.80, 86.44±1.27; 933.36±38.43, 81.62±1.01; 882.26±19.20, 80.99±3.65; 840.38±20.81, 84.59±1.61 and 828.47±57.90, 76.51±3.06 g/d in growing Nellore ram lambs fed rations R-I, R-II, R-III, R-IV, R-V, R-VI and R-VII, respectively. The DMI (g/d) was significantly ($P<0.01$) higher by 25.19, 29.74, 34.82, 31.04, 27.6 and 26.56 per cent respectively with R-II, R-III, R-IV, R-V, R-VI and R-VII rations in comparison to R-I ration (sole silage). DMI (g/d) was almost similar in lambs fed rations R-III, R-V and R-VI. Lowest DMI (g/d) was observed in ram lambs fed R-I ration. DMI when expressed as g/kg metabolic body weight in growing Nellore ram lambs was also significantly ($P<0.01$) different among the seven rations and was highest in ration R-III and lowest in ration R-I in comparison to the other rations. Comparable DMI per $kgw^{0.75}$ was observed in ram lambs fed rations R-II, R-IV, R-V, R-VI.

High DMI in lambs fed R-IV ration in the present experiment might be due to higher in take of concentrate along with maize silage *ad lib* in comparison to other rations. Likewise, lower intake in R-I ration might be due to sole maize silage feeding. Silage is a bulky feed (high moisture – about 70%) which fills the gut with little amount of silage intake. Hence, the animals received less dry matter with sole silage feeding. Supplementation of forage based rations plus 75-100 g each of maize plus GNC, sunflower cake or palm kernel cake (PKC) has markedly improved the DMI in growing sheep (Pratap Reddy *et al*., 1989; Devasena and Krishna, 1996). Maximum DMI occurred when grass silage was supplemented with high level of concentrate (Keane *et al*., 2006). Similarly supplementation of legume fodder to guinea grass hay based diet has significantly increased the total DMI (Upadhyay, 1987; Ash, 1990) in goats and in sheep by Devasena and Krishna (1996). DMI per kg metabolic body weight in lambs fed sole silage ration (R-I) in the present study was in contrast to the value given by Marina *et al.* (2007) in sheep fed maize silage alone. This increase in DMI with increase in concentrate supplementation is in agreement with the results observed by Das (2008) in Sikkim local male kids supplemented with concentrate @ 0.5, 1.0 and 1.5% of body weight. Similar increase in dry matte intake with increase in level of concentrate supplementation was observed in Black Bengal kids by Das and Ghosh (2001). Previous reports suggest that DMI increased due to supplementation in low quality forage diets (Dixon and Egan, 2000; Rafiq *et al.* 2002 and Dixon *et al*. 2003). DMI when expressed as per cent body weight was not significantly different among the ram lambs fed different experimental silage based and SSB based rations.

Dry matter digestibility

Dry matter digestibility coefficients determined was 65.67±3.22, 66.94±2.30, 69.54±2.67, 70.94±1.99, 70.83±0.65, 68.46±2.78 and 64.29±1.03 per cent, respectively (Table 3, Fig. 1) for the silage rations R-I (sole silage), R-II (silage + concentrate mixture @ 0.5% body weight), R-III (silage + concentrate mixture @ 1.0% body weight), R-IV (silage + concentrate mixture @ 1.5% body weight), R-V (silage + lucerne hay to meet 25% DM requirement), R-VI (silage + groundnut

haulms to meet 25% DM requirement) and R-VII (sweet sorghum bagasse based complete ration). DM digestibility was increased non significantly by 1.9, 5.57, 7.43, 7.29 and 4.08 per cent with R-II, R-III, R-IV, R-V and R-VI rations, respectively in comparison to R-I ration and DM digestibility of SSB based ration was 2.15 per cent lower in comparison to R-I ration. Chauhan and Brar (1989) reported increased DM digestibility with supplementation of concentrates to maize silage based rations in buffalo calves. Devasena and Krishna, (1996) reported increased DM digestibility with supplementation of legume forages in sheep. Veereswara Rao *et al*. (1993) also reported increased DM digestibility with supplementation of legume forages to basal forage of NB_{21}. Singh and Samantha (1998) reported increased DM digestibility with supplementation of legume forages to basal forages.

5.3.4.3 Organic matter digestibility

Organic matter digestibility of experimental rations R-I, R-II, R-III, R-IV, R-V, R-VI and R-VII was 68.96±2.82, 70.64±2.41, 72.15±0.52, 73.01±1.52, 71.66±3.74, 69.95±3.57 and 65.06±2.06 per cent, respectively (Table 3, Fig. 1). Non significantly increase in OM digestibility by 2.38, 4.42, 5.55, 3.77 and 1.42 per cent was observed with R-II, R-III, R-IV, R-V and R-VI rations, respectively in comparison to R-I ration. OM digestibility of SSB based ration was 5.93 per cent lower in comparison to R-I ration. Chauhan and Brar (1989) reported increased OM digestibility with supplementation of concentrate to maize silage based rations in calves. Singh and Samantha (1998) reported increased OM digestibility with supplementation of legume forages to basal non legume forages. Insignificant increase in OM digestibility was observed by Veereswara Rao *et al*. (1993) in lambs by supplementing NB_{21} green forage with legume fodder.

Crude protein digestibility

Crude protein digestibility coefficient (%) was 62.60±3.54, 63.82±2.50, 65.19±4.85, 70.63±1.48, 71.15±6.03, 69.49±5.03 and 67.25±1.28 for the silage rations R-I, R-II, R-III, R-IV, R-V, R-VI and R-VII, respectively (Table 3, Fig. 1). CP digestibility was increased insignificantly by 1.91, 3.97, 11.37, 12.02, 9.92 and 6.91 per cent in R-II, R-III, R-IV, R-V, R-VI and R-VII rations, respectively in comparison to R-I ration. Pratap Reddy *et al*. (1989) reported insignificant increase in CP digestibility when concentrate was supplemented with basal forage rations. Increased CP digestibility with supplementation of concentrates at different levels with maize silage based rations in calves was reported by Chauhan and Brar (1989. Varaprasad *et al*. (1995) reported increase in CP digestibility in lambs fed Co-1 grass supplemented with concentrate. This increase in CP digestibility in the present experiment might be due to gradual increase in dietary CP concentration which might have satisfied adequate N concentration for rumen microbes (Russel *et al*., 1992). These CP digestibilities were almost similar with the results of Das (2010) in Sikkim local male kids fed mixed grass supplemented with concentrate @ 0.5, 1.0 and 1.5% of body weight.

Ether extract digestibility

Ether extract digestibility coefficient determined was 62.43±1.93, 62.94±2.29, 63.15±1.78, 63.72±3.01, 66.68±2.80, 64.01±2.67 and 60.08±2.67 per cent for the rations R-I, R-II, R-III, R-IV, R-V, R-VI and R-VII, respectively (Table 3, Fig. 1). Numerical increase (P>0.05) in EE digestibility by 0.81, 1.14, 2.02, 6.37 and 2.47 per cent with R-II, R-III, R-IV, R-V and R-VI rations, respectively in comparison to R-I ration and EE digestibility of SSB based ration was 3.91 per cent lower than R-I ration. Chauhan and Brar (1989) reported non significantly increased EE digestibility with supplementation of concentrates to maize silage based rations in calves. Pratap Reddy *et al*. (1989) reported insignificant increase in EE digestibility when concentrate was supplemented to h basal forage rations.

Crude fibre digestibility

Crude fibre digestibility of the silage rations R-I, R-II, R-III, R-IV, R-V, R-VI and R-VII was 60.26±4.80, 60.95±3.93, 61.57±4.40, 62.45±4.88, 65.08±4.05, 62.75±5.28 and 57.41±5.00 per cent, respectively (Table 3, Fig. 1). Non significant increase in CF digestibility by 1.13, 2.13, 3.51, 7.41 and 3.97 per cent with R-II, R-III, R-IV, R-V and R-VI rations, respectively in comparison to R-I ration and CF digestibility of SSB based ration was 4.94 per cent lower than R-I ration. Pratap Reddy *et al*. (1989) reported insignificant increase in CF digestibility when concentrate was supplemented to basal forage rations. Varaprasad *et al*. (1995) reported increase in CF digestibility in lambs fed Co-1 grass supplemented with concentrate. Devasena and Krishna, (1996) reported increased CF digestibility with supplementation of legume forage to basal ration in sheep.

Nitrogen free extract digestibility

Nitrogen free extract digestibility coefficient of the rations R-I, R-II, R-III, R-IV, R-V, R-VI and R-VII were 64.08±2.27, 65.77±2.95, 66.39±3.16, 66.78±2.04, 65.39±1.86, 64.12±2.69 and 61.10±1.97 per cent, respectively (Table 3, Fig. 1). Numerical increase in NFE digestibility by 2.57, 3.48, 4.04, 2.00 and 0.06 per cent in lambs fed R-II, R-III, R-IV, R-V and R-VI rations, respectively in comparison to those fed R-I ration and NFE digestibility of SSB based ration was 4.88 per cent lower than R-I ration. Pratap Reddy *et al*. (1989) reported insignificant increase in NFE digestibility when concentrate was supplemented to basal forage rations. Varaprasad *et al*. (1995) also reported increase in NFE digestibility in lambs fed Co-1 grass supplemented with concentrate. Similar findings were noticed by Devasena and Krishna (1996) in sheep fed colonial guinea grass supplemented with groundnut cake plus maize premix.

Digestibilities of cell wall constituents

The mean NDF, ADF, hemicellulose and cellulose digestibility coefficients of R-I, R-II, R-III, R-IV, R-V, R-VI and R-VII rations were 62.32±2.42, 59.43±3.37, 60.94±0.48, 57.56±4.24; 64.51±2.68, 60.44±0.88, 62.49±2.00, 59.63±2.35; 65.65±2.37, 62.28±2.27, 65.41±2.21 and 62.07±3.11; 65.81±3.10, 62.29±2.98, 66.63±3.55, 64.00±2.51; 62.58±1.53, 62.14±1.05, 64.23±2.95, 61.61±2.15; 62.09±2.50, 59.79±2.97, 62.83±1.39, 58.37±1.24 and 60.86±2.81, 58.61±1.06, 60.31±1.90 and 57.17±2.42, respectively (Table 3 and Fig. 2). The digestibility coefficients of fibre fractions were comparable among the experimental maize silage based and SSB based rations. However, insignificantly numerical increase in digestibility of fibre fraction in the silage rations supplemented

with concentrate at 0.5%, 1.0% and 1.5% of body weight as well as silage supplemented with legume hay and legume straw was observed. Cell wall constituent digestibility of SSB based ration (R-VII) was comparable with sole silage ration (R-I). Singh and Samantha (1998) reported increased NDF digestibility with supplementation of legume forages to basal non legume forages. Marina *et al.* (2007) reported almost similar results in sheep fed maize silage alone as noticed in the R-I ration (sole silage) of present study.

Increased digestibility of forage based rations supplemented with concentrate mixture is due to improved fermentation facilitated by improved availability of higher digestible nutrients to the microbes (Sehgal *et al.* 1999). Supplements which provide critical nutrients enhance the rumen ecosystem so as to increase the microbial growth, rate of fibre digestion and propionate production (Lindsay, 1970). Digestibility of total diet generally increased with increased proportions of concentrates in the diet (Xu *et al*., 2008). Numerically lower digestibility of nutrients in R-I ration in comparison to other R-II, R-III and R-IV rations indicated that without concentrate supplementation it was not possible to supply sufficient amount of rumen degradable N and other nutrients required by rumen microbes for optimum rumen microbial activity (Leng, 1990).

Nutritive Value

The data on nutritive value of seven experimental silage rations in terms of DCP, TDN, DE and ME are presented in Table 4. The DCP and TDN values were 3.44±0.19, 60.51±1.70; 4.47±0.23, 61.66±1.43; 5.33±0.23, 62.15±1.62; 6.62±0.20, 63.02±1.02; 5.62±0.52, 62.78±1.15; 4.44±0.32, 60.59±1.21 and 7.74±0.15 and 55.47±2.22 per cent, respectively for the R-I, R-II, R-III, R-IV, R-V, R-VI and R-VII rations. The DCP content was significantly (P<0.01) higher by 23.04, 35.46, 48.04, 38.79, 22.52 and 55.56 per cent in rations R-II, R-III, R-IV, R-V, R-VI and R-VII rations, respectively in comparison to ration R-I. Highest DCP was found with R-VII ration followed by R-IV, R-III, R-V, R-II, R-VI ration and least DCP content was observed in ration R-I. Pratap Reddy *et al*. (1989) reported increase in DCP value when concentrate was supplemented to basal forage rations. Varaprasad *et al*. (1995) in lambs fed Co-1 forage, Devasena and Krishna (1996) in lambs fed colonial guinea grass observed that, supplementation of concentrate or legume forage to basal diet increased the DCP content of the ration.

The TDN content was non significantly (P>0.05) higher by 1.87, 2.64, 3.98, 3.61 and 0.13 per cent in R-II, R-III, R-IV, R-V and R-VI rations in comparison to R-I ration but lower by 8.33 per cent in R-VII ration in comparison to R-I ration. The TDN value was highest in ration R-IV and lowest in ration R-VII. The DCP and TDN values were increased in the rations as the concentrate proportion increased. This might be due to high energy and protein content of concentrate feed than the maize silage feed alone. Pratap Reddy *et al*. (1989) reported increase in TDN value when concentrate was supplemented to basal forage rations. These results were also corroborating with the findings of Varaprasad *et al*. (1995) in lambs fed Co-1 forage and Devasena and Krishna (1996) in lambs fed colonial guinea grass supplemented with concentrate and/or legume fodder.

The DE and ME values determined for different experimental maize silage and SSB based rations for growing Nellore ram lambs was 11.16±0.31, 9.15±0.26; 11.38±0.26, 9.32±0.22; 11.47±0.30, 9.40±0.25; 11.63±0.19, 9.53±0.15; 11.58±0.21, 9.50±0.17; 11.18±0.22, 9.17±0.18 and 10.23±0.41 and 8.39±0.34 MJ per kg dry matter, respectively (Table 4) for R-I, R-II, R-III, R-IV, R-V, R-VI and R-VII rations. There was a significant (P<0.05) difference in the DE and ME values of seven rations. The DE and ME values of ration R-I to R-VI were almost same and was lower in R-VII than the other rations. In contrast, Chauhan and Brar (1989) reported increased TDN, DE and ME values due to supplementation of concentrates to maize silage based rations in calves.

The DE and ME values (MJ/kg DM) of experimental rations fed to ram lambs was almost same in R-I to R-VI rations but it was lower in R-VII ration in comparison to other rations. This might be due to similar intake of energy by animals among the rations R-I to R-VI. The ME content (MJ/kg DM) of maize silage used in the present study is similar as was reported by Sohail *et al.* (2010). They reported ME value as 8.95 MJ/kg DM of maize silage for sheep.

Nitrogen Balance

Mean N intake, N loss through faeces, urine and also total loss (g/d) is shown in Table 5. Nitrogen intake (g/d) was 5.35±0.23, 9.09±0.31, 11.34±0.38, 14.03±0.86, 11.13±0.16, 8.59±0.22 and 15.26±1.07 in lambs fed R-I, R-II, R-III, R-IV, R-V, R-VI and R-VII rations, respectively. The N intake (g/d) was significantly (P<0.01) higher by 41.14, 52.82, 61.87, 51.93, 37.72 and 64.94 per cent in lambs fed R-II, R-III, R-IV, R-V, R-VI and R-VII rations, respectively in comparison to R-I ration. Nitrogen intake in the R-IV and R-VII rations of present study was higher than the other experimental rations. This might be due to greater protein content in these rations. Mtenga and Kitlay (1990) reported higher N intake in goats when basal forage ration was supplemented with concentrate.

Nitrogen out go (g/d) through faeces and urine in ram lambs fed R-I, R-II, R-III, R-IV, R-V, R-VI and R-VII rations was 2.01±0.24, 2.19±0.30; 3.31±0.34, 3.84±0.19; 3.92±0.16, 4.39±0.71; 4.12±0.36, 5.03±0.50; 3.24±0.73, 3.63±0.58; 2.63±0.46, 3.33±0.41 and 5.02±0.52 and 6.01±0.88 g/d, respectively. Nitrogen out go through faeces and urine in ram lambs was significantly (P<0.01) different among the seven experimental rations. The total N loss (g/d) in lambs fed rations R-I, R-II, R-III, R-IV, R-V, R-VI and R-VII was 4.20±0.48, 7.15±0.37, 8.31±0.86, 9.15±0.85, 6.87±1.05, 5.96±0.75 and 11.03±1.19, respectively. The N loss through faeces and urine as well as total N loss (g/d) were also significantly (P<0.01) different among the ram lambs fed different experimental rations. The loss being highest in lambs fed ration R-VII and lowest in rams fed R-I ration probably due to variation in N intake in lambs fed different experimental rations. The loss being highest in lambs fed R-VII ration and lowest in rams fed R-I ration.

Nitrogen balance in experimental ram lambs was 1.14±0.42, 1.94±0.23, 3.03±1.13, 4.87±0.45, 4.26±0.95, 2.62±0.56 and 4.22±0.47 g/d and was either significantly or non-significantly higher by 41.24 (P>0.05), 62.38 (P>0.05), 76.59 (P<0.01), 73.24 (P<0.01), 56.49 (P>0.05) and 72.99 (P<0.01) per cent in lambs fed R-II, R-III, R-IV, R-V, R-VI and R-VII rations,

respectively in comparison to R-I ration and the nitrogen retention expressed as per cent intake or per cent absorbed was 21.48±7.76, 32.94±11.24; 21.44±2.60, 33.54±3.73; 26.00±8.95, 38.72±11.71; 34.95±3.24, 49.28±3.63; 38.51±8.89, 52.67±8.92; 31.00±7.05, 43.39±7.72 and 28.07±3.66 and 41.47±4.70, respectively for R-I, R-II, R-III, R-IV, R-V, R-VI and R-VII rations. Pratap Reddy *et al*. (1989) reported increase in N retention when concentrate was supplemented at various levels with basal forage rations. Veereswara Rao *et al*. (1993) observed higher N retention in lambs by feeding NB21 green forage supplemented with legume fodder.

Nitrogen balance (g/d) was significantly ($P<0.01$) different among the seven experimental rations being significantly higher in rations R-IV, R-VII and R-V and lower in rations R-I and R-II rations. Higher nitrogen retention in R-IV, R-VII and R-V than other rations fed animals in the present study might be due to more nitrogen intake from respective rations. Pratap Reddy *et al*. (1989) reported increase in N retention when concentrate was supplemented to basal forage rations. Veereswara Rao *et al*. (1993) observed higher N retention in lambs by supplementing NB_{21} green forage to legume fodder. Devasena and Krishna (1996) observed higher N balance in lambs fed colonial guinea grass (*Panicum maxium*) supplemented with groundnut cake plus maize premix.

The results of this study demonstrate that feeding of sole maize silage and supplementation with concentrate or legume hay did not affected the digestibilities of nutrients. However, Nitrogen balance in the silage fed ram lambs was affected with supplementation of concentrate or legume hay.

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Effects of Select Instructional Practices on Improving the Achievement of Low Achievers

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Abstract- The major goal of science education today is fostering students' intellectual competencies, such as independent learning, problem-solving, decision-making and critical thinking. In order to achieve this, science teaching must be shifted from traditional schooling to more constructivist-oriented instruction. The aim of this study was to find out the effects of various instructional practices such as problem based learning, 5 E model instruction and portfolio writing practices on enhancing the achievement of low achievers. Pre test- post test parallel group design was adopted for the present study. The sample consisted of 120 low achievers, where 90 students were categorised in three experimental groups and 30 students in control group. Instructional materials were prepared based on the topic from 8th standard science text book according to the nature of instructional strategies selected. The major finding of the study is that the select instructional strategies are more effective than the traditional method of teaching science. Among the instructional practices, the problem based learning practice is more effective in enhancing the achievement of low achievers.

Index Terms- Problem based learning, 5-E model instruction, portfolio writing, low achievers.

I. INTRODUCTION

This article guides a stepwise walkthrough by Experts for writing a successful journal or a research paper starting from inception of ideas till their publications. Research papers are highly recognized in scholar fraternity and form a core part of PhD curriculum. Research scholars publish their research work in leading journals to complete their grades. In addition, the published research work also provides a big weight-age to get admissions in reputed varsity. Now, here we enlist the proven steps to publish the research paper in a journal.

Rapid changes in the world including technological advancements influence the science and as a consequence the interaction between the science and the society has been changing as well. It has a large impact in developing the ethical and moral values of the society we live in. The research in this area finds wide scope as the scientific progress made the technological developments possible. In the present society, scientific and technological innovations increasingly permeate almost every aspects of daily life. In a world based on science and technology, it is education that determines the level of prosperity, welfare and security of the people.

It is essential to make science education effective and relevant for a large and necessarily more diverse fraction of the population. To do so, we need to transform how students think so that they can understand and use science like scientists do. Effective teaching facilitates that engaging students in thinking deeply about the subject at an appropriate level and then monitoring that thinking guiding it to more expert like. The achieving of scientific literacy should be considered to be almost as important as achieving basic literacy. Conventional science teaching relied heavily on lectures, reading, and teacher led demonstrations.

II. RATIONAL FOR THE STUDY

The major goal of science education today is fostering students' intellectual competencies, such as independent learning, problem-solving, decision-making and critical thinking (American Association for the Advancement of Science (AAAS), 1994; National Research Council (NRC), 1996). In order to achieve this, science teaching must be shifted from traditional schooling to more constructivist-oriented instruction. Teaching has to be restricted to posing questions for the student to answer and to creating situations to facilitate discovery by the student. The purpose of science education is not simply to produce the next generation of Scientist, but it can spread basic scientific literacy throughout our population so that wise decisions can be reached about how to address the problems-global as well as local- we are facing today. Traditional science instruction concentrates on teaching factual knowledge, but it lacks the mental organisational structure that facilitates the retrieval and effective application of that knowledge. A significant community of science education researchers is experimenting various approaches to the development and testing of new pedagogical methods. Here the investigator attempts to investigate the effect of select instructional strategies on improving the understanding of science concepts. And the target population is low achievers. Low achievers are neglected minority owing to their poor academic achievement in school (Coren, 1992). They often face discrimination from peers. In general, low achievers' self perception of academic ability is significantly lower than peers. Under the existing school system, school authorities did not arrange remedial programmes for low achievers to promote their academic excellence. Thus, teacher interaction with low achievers is less motivating and less supportive. Hence the investigator proposed to conduct the experiment on low achievers.

The significance of this study lies in its potential to contribute to the literature and to educational practice related to science teaching, with special focus on instruction aimed at promoting the cognitive processes of low achievers in the classroom.

Review of related literature

Ridlon (2009) used two different instructional approaches for teaching mathematics through problem centred approach. Year 1 involved low achievers, whereas Year 2 was mixed ability students. The experimental treatment was a problem-centered approach (PCL) where potentially meaningful tasks were posed to the class and solved in collaborative groups. The groups presented and defended their solution strategies to their peers. Regardless of perceived ability level, the PCL approach appeared to significantly enhance achievement and improve attitude towards mathematics. Low achievers seemed to gain the most, narrowing the gap between them and their mixed ability peers. PCL appeared useful with any curriculum if the teacher understood and properly implemented the components of the approach.

Wu and Tsai (2005) conducted a study to explore the effects of long-term constructivist-oriented science instruction on elementary school students' process of constructing cognitive structures. Furthermore, such effects on different science achievers were also investigated. The findings showed that the students in the constructivist-oriented instruction group attained significantly better learning outcomes in terms of the extent and integration of their cognitive structures, metacognition engagement and the usage of information processing strategies. Moreover, it was also revealed that both high achievers and low achievers benefited from the constructivist-oriented instructional activities, but in different ways.

Wolf (1989) pointed out that students need time to study their works and to select the pieces that best exemplify what they have learned; thus the writing portfolio represents what learning has taken place during the duration of the class. Murphy (1998) claims that the time students spend selecting writings to be included in portfolios leads to reflection, a necessary part of portfolios. Further, portfolios give a more accurate representation of the writing accomplishments or shortcomings of a student and a teacher.

Hawkins (1988) studied the changing practices in classrooms to improve the behavior of low achievers. The effects of a package of instructional methods on the academic achievement, behaviour and social bonding of 77 experimental and 83 control students in grade seven who were low achievers in math are described. The potential for preventing discipline problems through promoting sound teaching practices in mainstream classrooms is reviewed in the study.

III. STATEMENT OF THE PROBLEM

Review of related literature revealed that there are limited studies associated with the academic performance of low achievers. Further, the investigator, as a teacher educator experienced many behavioural and academic problems faced by the slow learners during the school visits. Hence the present study is an attempt to analyse the influence of various instructional strategies for improving the academic performance of low achievers.

The problem for the present study can be entitled as 'Effects of select instructional practices on improving the achievement of low achievers'.

Here the select instructional practices are problem based learning, 5-E model instruction and portfolio writing practices.

Theoretical Framework

Problem-based learning *(PBL)*: It is an instructional method in which students learn through facilitated problem solving. In problem based learning, student learning centers on a complex problem that does not have a single correct answer. Students work in collaborative groups to identify what they need to learn in order to solve a problem. They engage in self-directed learning and then apply their new knowledge to the problem and reflect on what they learned and the effectiveness of the strategies employed. The teacher acts to facilitate the learning process rather than to provide knowledge. Research studies suggest that problem based learning is an instructional approach that offers the potential to help students develop flexible understanding and lifelong learning skills.

5-E Model Instruction: It helps the students to use the process skills of science to develop an understanding of the scientific concepts. It involves engage, explore, explain, elaborate and evaluate.

Engage: Brings the student's mind into the frame to learn something new

Explore: Provides a common experience for all students and helps the teacher identify prior knowledge

Explain: Here the student constructs an explanation. The teacher provides information to increase the accuracy of the explanation.

Elaborate: Builds on current understanding to increase the depth and breadth of understanding

Evaluate: Provides an opportunity for students to assess their own understanding and be able to demonstrate the depth and breadth of that understanding to others.

Portfolio Writing: Portfolios invite 'introspection'. They are both receptacles and vehicles for individual reflection. But reflection in learning is much more than introspection. Portfolio reflection provides a way to 'make learning visible'. This 'visible learning' becomes the basis for conversations and other interactions among students and teachers and parents. Portfolios have an audience, so does reflection. Reflection involves more than just 'looking back'; it occurs before, during, and after reading and writing or learning process.

IV. RESEARCH QUESTIONS

1. How do the instructional practices affect the achievement of low achievers?
2. Which instructional strategy has the most significant effect on the achievement?
3. Whether the instructional practices have any effect on gender?

V. RESEARCH METHODOLOGY

Design*:* In the present study the investigator adopted experimental method- pre test post test parallel group design. Here the investigator used three experimental groups, receiving

the treatment based on three instructional practices problem based learning, 5-E model instruction, portfolio writing practices respectively. And the control group had the traditional way of teaching.

Sample: At the initial stage of the sample selection 300 8th standard students were selected from two schools from Kollam district. They were exposed to the pre test and 120 low achievers were identified using 'sigma distances' (M-σ). Out of the 120 low achievers, 90 students were randomly selected and allotted in three experimental groups and 30 in control group, giving due representation to gender.

Instruments: Personal data sheet, achievement test based on the selected topic from science, teaching-learning materials based on select instructional practices.

Topic: The investigator has selected the topic 'Force-internal force, external force, balanced force, unbalanced force, Newton's First law, Inertia' for preparing instructional materials for the present study (8th standard science text book).

Procedure*:* The achievement test in science was administered (pre test) on the students under study and low achievers were identified. The identified low achievers were grouped into three experimental groups and one control group. The experimental groups were taught using the select instructional practices, while the control group was taught using the conventional method.

Problem based learning: Students are at the center of learning when teachers implement problem based learning. First, a problem or a topic of discussion is presented to stimulate student interest. Then students were asked to work in small groups to investigate the problem. As the process progresses, the students were given opportunity to challenge their ideas by other group members or by the teacher if necessary. The process is cyclical and repeated several times as new information is learned and ideas have been modified to generate new learning needs. Thereby the scientific facts and concepts were taught indirectly by integrated within the scientific process.

5-E model instruction: The 5 E learning cycle model requires instruction to include the following discrete elements: engage, explore, explain, elaborate and evaluate. The *engage* component in the model is intended to capture students' attention. Here the teacher engages students by creating surprise or doubt through a demonstration.

The *explore* phase of the learning cycle provides an opportunity for students to observe, record data, isolate variables, design and plan experiments, create graphs, interpret results, develop hypotheses, and organize their findings. Here the teacher frame questions, suggest approaches, provide feedback and assess understandings.

Students are introduced to models, laws, and theories during the *explain* phase of the learning cycle. Then the teacher guides students toward coherent and consistent generalizations, helps students with distinct scientific vocabulary and provides questions that help students use this vocabulary to explain the results of their explorations.

The *elaborate* phase of the learning cycle provides an opportunity for students to apply their knowledge to new domains, which include raising new questions and hypotheses to explore. Here the teacher includes related numerical problems for students to solve.

The *evaluate* phase of the learning cycle continues to include both formative and summative evaluations of student learning. Here teacher use tests to evaluate the concepts discussed. Tests include questions from the lab and ask students questions about the laboratory activities. Students were asked to interpret data from a lab similar to the one they completed and also to design experiments as part of their assessment.

Portfolio writing practices: Portfolios are known as the thinking journals, because it reflects the whole thinking process of a student in the class. It helps the student to think upon their activities while learning. Once the instruction was over, the investigator asked the students think on what they have learned in the class. Then the students were trained to write their understandings, suggestions and doubts they came across while learning .There by the students were able to recapitulate the content. In order to practice the students in writing portfolios, the teacher asked them to write it after each class. Thus the students were aware of the importance of writing portfolios. Also by reviewing the portfolios the teacher understood that students were tuned in such a way that portfolios enhanced their understanding of the concepts.

The same tests were administered again and the pre test-post test scores were compared to test the effectiveness of the select instructional practices implemented. The data thus obtained were analysed using appropriate statistical techniques and interpreted accordingly.

VI. RESULTS

Table 1- Significance of difference between mean pre test scores of the control and experimental group 1 (Problem Based Learning)

Group	N	Mean	S.D	t-test	Level of significance
Control group	30	4.93	1.01	1.08	NS
Experimental group 1	30	5.23	1.13		

NS- Not Significant

Table 2- Significance of difference between mean pre test scores of the control and experimental group 2 (5-E model instruction)

Group	N	Mean	S.D	t-test	Level of significance
Control group	30	4.93	1.01	0.38	NS
Experimental group 2	30	5.03	0.99		

NS- Not Significant

Table 3- Significance of difference between mean pre test scores of the control and experimental group 3 (Portfolio writing practices)

Group	N	Mean	S.D	t-test	Level of significance
Control group	30	4.93	1.01	0.64	NS
Experimental group 3	30	5.10	0.99		

NS- Not Significant

From tables 1, 2 and 3, the obtained t-values 1.08, 0.38 and 0.042 are less than the critical value 1.96 at 0.05 level of significance. It implies that the difference in the achievement test scores is not significant. This reveals that there is no difference in the achievement scores of control group and experimental groups before giving treatments.

Table 4- Significance of difference between mean post test scores of the control and experimental group 1 (Problem Based Learning)

Group	N	Mean	S.D	t-test	Level of significance
Control group	30	12.3	1.74	11.8	0.01
Experimental group 1	30	17.9	1.92		

The calculated t-value 11.83 is very much greater than the critical value 2.58 at 0.01 level of significance. It implies that the difference between mean is statistically significant. This reveals that experimental group 1 is superior to the control group in post test achievement scores.

Table 5- Significance of difference between mean post test scores of the control and experimental group 2 (5-E model instruction)

Group	N	Mean	S.D	t-test	Level of significance
Control group	30	12.3	1.74	7.11	0.01
Experimental group 2	30	15.1	1.27		

The calculated t-value 7.11 is very much greater than the critical value 2.58 at 0.01 level of significance. It implies that the difference between mean is statistically significant. This reveals that experimental group 2 is superior to the control group in post test achievement scores.

Table 6- Significance of difference between mean post test scores of the control and experimental group 3 (Portfolio writing practices)

Group	N	Mean	S.D	t-test	Level of significance
Control group	30	12.3	1.74	6.21	0.01
Experimental group 3	30	14.6	1.04		

The calculated t-value 6.21 is very much greater than the critical value 2.58 at 0.01 level of significance. It implies that the difference between mean is statistically significant. This reveals that experimental group 3 is superior to the control group in post test achievement scores.

From tables 4,5 and 6 it is also clear that highest mean value (17.9) obtained for the treatment, problem based learning. Hence it can be considered as superior to other two treatments in improving the achievement of low achievers.

Table 7- Significance of difference between the post test scores of experimental groups with respect to Gender

Experimental Groups	Variable	Category	N	Mean	S.D	t-test	Level of significance
Group 1	Gender	Boys	16	17.5	1.86	1.23	NS
		Girls	14	18.3	1.94		
Group 2	Gender	Boys	18	15.1	1.47	0.23	NS
		Girls	12	15.2	0.94		
Group 3	Gender	Boys	16	14.5	1.26	0.56	NS
		Girls	14	14.7	0.73		

NS- Not Significant

The calculated t-values 1.23, 0.23 and 0.56 are less than the critical value 1.96 at 0.05 level of significance. This indicates that the difference in the post test achievement scores the three experimental groups is not significant. That is the treatments have no effect on gender.

Summary of Findings

The major finding of the study is that the select instructional strategies are more effective than the traditional method of teaching science. Among the instructional practices, the problem based learning practice is more effective in enhancing the achievement of low achievers. Also the treatments have no effect on the variable, gender.

VII. EDUCATIONAL IMPLICATIONS

The present study revealed that problem based learning can be considered as one of the most viable technique to individualise instruction in improving the academic performance of low achievers. In the light of research findings, it is felt that the results may contribute to alleviation of difficulties of the students in approaching science concepts. Students start to see how the knowledge they learn helps them to solve problems in life, therefore giving them a love for learning and turning them into lifelong learners. As education learns to embrace this new type of teaching, teachers will have to learn to give the control of the problem and the classroom to the students. The challenge in the present day science is to develop a mindset to boost up the standards of scientific research. And it is the sole responsibility of the stakeholders and researchers in all fields to make it happen by implementing more effective pedagogical practices.

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Technical Vocational Education (TVE) Institutions and Industries Partnership: Necessity for Graduates Skills Acquisition

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Abstract- The quality of TVE graduates from technological institutions has been a subject of major concern for most industries in Nigeria. Most industries in Nigeria complaint on TVE graduates as possessing low level skills required for employment in industries and lack of confidence in carrying out their duties and responsibilities. This paper discussed extensively on TVE institutions and industry partnership as a necessity for graduates' skills acquisition. It outlined the concept of TVE and also discussed on the present status of TVE in Nigeria. The causes of a skill gap between the industries and TVE institutions were further explained. The paper also highlights on how to bridge the gap between the industries and TVE institutions. Conclusions were drawn and recommendations made in order to establish linkage between the industries and TVE institutions amongst which are: that TVE institutions and local industries should collaborate to organize seminars and workshops where they will share information on the changing trends in industrial practices and how these changes can be incorporated into the curriculum of the school and also TVE institutions should be charged with oversight responsibilities by putting up structures and policies that will facilitate industrial- institutional links. This will go a long way in ensuring that TVE graduates acquire the right skills that make them be employable in the industries or other related organizations.

Index Terms- Technical Vocational Education, Institutions, Industries, Graduates Skills Acquisition.

I. Introduction

The existing technical skills gap between the graduates of TVE and the industry has become a major concern of parents, business leaders and educators in Nigeria. Employers of labor have continued to express their concern and worry over the quality of the current graduates of TVE programs in their short of relevant skills required for employment (Idris & Rajuddin, 2012). Most industries and employers of labor in Nigeria complaint, stem from inadequate skill requirement of TVE graduates for most cutting edge technology, low practical knowledge and lack of confidence in carrying out their duties and responsibilities. Most Nigerian TVE graduates have to be subjected to several re -training program since most of the graduates are considered non-employable going by the quality of training acquired from their various institutions (Olorunfemi & Ashaolu, 2008). Various studies have queried the relevance of graduates and research results to the industry considering the low academic status and skill acquired byproduct from various TVE institutions. The training program is not addressing the growing needs of the industry and the society. A change of direction is required to close up the widening gap between the TVE graduates and the industry in terms of the requisite skills required for employment in the industries.

II. Concept of Technical Vocational Education (TVE)

The Federal Republic of Nigeria (FRN) in the National Policy on Education has placed a premium on TVE in view of its important role in technological and industrial development of Nigeria. It is recognized as that aspect of education which leads to the acquisition of practical and applied skills as well as basic scientific knowledge (FRN, 2004). Olawepo (Osifeso, 20011) defines TVE as a type of education designed for preparing the individual and learner to earn a living (to be self-reliant) or increase his earning in an occupation where technical information and understanding of the laws of science and technology application to modern design production, distribution and services as essentials for positive change. According to Dike (2009) TVE is that aspect of education which leads to the acquisition of skills as well as basic scientific knowledge. It is a planned program of courses and learning experiences that begins with exploration of career options, supports basic academic and life skills, and enables achievement of high academic standards, leadership, preparation for industry-defined work, and advanced and continuing education (Maclean & Wilson, 2009). Oni (2007) described TVE as that type of education which fits the individual for gainful employment in recognized career as semi-skilled workers or technicians or sub-professionals. One of the most significant aspects of TVE is its inclination towards the world of work and the emphasis of the curriculum on the acquisition of employable skills. The need to link training in TVE to employment either self or paid employment is at the base of all the best practices and approaches observed throughout the world.

III. Present Status of TVE in Nigeria

UNESCO (2004) identified the two main objectives of TVE as; the need to train the workforce for self-employment and the need to raise the productivity of the informal sector of the economy. They pointed out that dearth of resources has led to

cuts in the volume of training that is provided in public institutions. These reductions are a hindrance to pursuing the main objectives of providing training and raising production. Considering the expensive nature of TVE as a type of education, it is important that an expanded system with needed and adequate facilities and equipment will lead to the success of the system. In the same vein Islam and Mia (2007) expressed that both formal and non-formal TVE lacked an effective link between training and the world of work. They further noted that as a result of its lack of coherent mode, practical skills training does not produce the required skills for the labor market. Additionally, the trainees also lacked training experience, initiative and motivation to carry out their duties effectively.

According Netherland Organization for International Cooperation in Higher Education (NICHE, 2010) there are various challenges that TVE graduates are facing in terms of practical skills acquisition, in most developing countries, especially in Nigeria, TVE is narrow in scale, scope, quality and relevance. The programs are not relevant to the requirements of the local labor market, the curricula and syllabi are out-of-date and the institutions lack the tools and equipment essential for practical skills acquisition. Where present, the equipment in workshops and laboratories is often out-of-date, bearing little resemblance to the technologies presently used by industry (NICHE, 2010). Insufficient training equipment leads to trainee overcrowding during practical lessons, with most of the students only observing the demonstration and not having the chance to get some hands-on practice. Due to the fact that the institutions are poorly resourced, the education and training remain theoretical and the graduates are not considered more skilled than their academic counterparts by the labor market. The institutions thereby acquire a poor image, and produce graduates with lower employability (NICHE, 2010).

However, public TVE institutions, according to Atchoarena & Esquieu (Dasmani, 2011) continued to attract a great deal of criticism. First, they were unable to train skilled workers to meet the requirements of industries and were unaware of the need for continuing education. Second, they were extremely costly. Often, the graduates of these institutions joined the ranks of the unemployed, an indication that the training provided did not match the jobs available. In most developing countries, including Nigeria, Ghana, Cameroon to mention but a few. Public TVE institutions have not been able to adapt to the new structure of the labor market and the new skill requirements of industries in both the formal and informal sectors. To buttress this fact, in a recent survey research carried out over 40 million graduates are unemployable into the industries because of lack of technical vocational and skills required for jobs available. Indeed there are vacancies for jobs in Nigeria but right now, there are little or no manpower with appropriate skills to fill the vacancies for the available jobs (Alfred, 2010).

Olaitan (2009) stated that TVE is still viewed as second class type of education in Nigeria by youths and the vulnerable unemployed graduates. No matter the campaign organised and the number of papers written by those knowledgeable about the usefulness of TVE, preference for academic qualification is on the increase among Nigeria citizens. It is in vogue among the private institutions in Nigeria only to realize in many years later that most graduates are out seeking for jobs that don't exist or not relevant to the academic training. In general, the quality of training is low, with too much emphasis on theory and certification rather than on skills acquisition and proficiency testing. Inadequate instructor training, obsolete tools, machines and equipment, and lack of instructional materials are some of the factors that combine to reduce the effectiveness of training in meeting the required knowledge and skills required for gainful employment. High quality skills training requires appropriate workshop tools, machines and equipment, adequate supply of training materials, and practice by learners (African Union, 2007).

According to Omotayo, *eta'l.,* (2008) other challenges that hinder the skill acquisition of graduates of TVE institutions include: inadequate funding, poor planning (including inaccurate research data), inadequate facilities and infrastructure, teacher incompetence, and inconsistent monitoring and evaluation, poor materials and resources. The problem with education funding in Nigeria is not just its inadequacy, it is also the fact that the system is fraught with corruption and fund misappropriation.

IV. Causes of Skill Gap Between Industry and Institutions

The teaching methods in TVE institutions have not succeeded in impacting positively on practical skills acquisition of graduates of TVE institutions. These methods of instruction include: the traditional lecture method based on sound theoretical background, group discussions, field trips and Students Industrial Work Experience Scheme (SIWES). However, modifications of these approaches are very imperative to meet the current industry challenges Oduola (Olorufemi & Ashaolu, 2008). Most of the multi-nationals organizations in Nigeria source their major and scientific research skills and facilities from their home countries with little or no input from the local scientific research. The main institution's training facilities are outdated and inadequate making practical training difficult to meet the modern trends in the industries.

According to Olorufemi & Ashaolu (2008) the industries in Nigeria do not have much involvement in training, funding and curriculum development of our technological institutions, which are the trademark of industrialization in the developed nations. There is no direct link and interaction between the industries and the technological institutions in research activities and manufacturing. The identified poor basic foundation in science subjects affects the students learning approach to science, engineering and technology principles and applications. Many studies have also revealed the relationship between learning approaches and workload (Kolani, *eta'l.*, 2006). The incessant hostility in various campuses has been connected to so much idle time resulting in lower quality of training since the workload does not reflect in-depth practical skills training. The practical class period on the timetable are expended in unfruitful exercise. Student learning becomes unproductive when the basic principles are not understood thereby leading to poor motivation.

V. Industry and TVE Institutions Partnership

While Nigerian industries are fast growing in terms of the improvement in science and technology, unprecedented demand for better graduates has been set up. However, the industry often complains that the existing technological institution's curricula fall short to tackle the practical problems in the industry. For instance, the industry expects the technological institutions to train their future employees with the latest technology. This is because technological institutions lack proper curricula that are suitable for the industries (Ayofe*, eta'l.,* 2009). This causes a gap between the technological institutions and industry that needs to be bridged so that the graduates can fit into the industries upon graduation. The industry is constantly broadening and the knowledge is increasingly becoming complex. The role of developing better curriculum in technological institution's program is important in bridging the gap between the changing technology and the industrial needs of employers. Technological institutions should make provision for conducive learning environment and industry oriented curriculum that the business community perceived as meeting their needs. Curricula are expected to be developed with the objective of producing skilled and employable workers that can easily fit into the industry without being re-trained (Ayofe, *eta'l.,* 2009).

However, the quality of TVE graduates from technological institutions particularly at tertiary level has been a subject of major concern for most industries in Nigeria. Most industries complaint results from inadequate skill requirement for most cutting edge technology, low practical skill and confidence. Most Nigerian engineering graduates are exposed to several re-training programs since most of the graduates are regarded as unemployable going by the quality of instruction acquired from their various institutes (Olorunfemi & Ashaolu, 2008). The economic development of any nation depends on its level of human resources development, especially in the area of science and technology for industrialization. Nigeria is far from undergoing any landmark in technological development towards industrialization due to poor infrastructures, despite the large number of graduates from various technological institutions that have failed to impact positively on the development of industries for industrialization of its economic sector (Olorunfemi & Ashaolu, 2008).

VI. Bridging the Skill Gap Between Industry and TVE Institutions

The persistent skills gap between the industry and technological institutions has made dependence on importation of skilled workers, what then can be done to bridge this skill gap? In bridging the gap, many countries have encountered and tackled this issue by introducing a strong technological component to the curriculum of their institutions of learning (Ayofe, *eta'l.,* 2009). This usually comes in many different ways; prevailing among them is offering students courses in technological education, work attitude and work principle, followed by a subsequent engagement in industrial and commercial firms, where they get actual experience in a real work environment.

Productive programs have been carried out in countries like Australia, Canada, Germany, United States, and Britain. The success story of such programs in these countries is guaranteed by the existence of a huge industrial sector, which works in collaboration with schools. Other countries have chosen to establish training centers, which have workshops that impart students with real work experience. These training centres' are normally established, financed, and ran by the private sector and schools pay fees for their students to make use of these centres. Developing skills in TVE, are normally presented as a stand-alone program or generally in the subjects taught. The skills are strengthened, self-learning, lifelong learning, research skills, time management skills, critical thinking skills etc. These mechanisms have been seen to be most effective when they are woven into technological institution's curriculum, rather than tackling them as stand-alone subjects. The program guarantees that the students acquire adequate knowledge and has the skills needed to use it in his job. Knowledge and skills are two important components that have been found to be necessary for both groups of students; the one that joins the labor force and the one that go for training in technological institutions.

Government and the established private sector should also put in place preparations for specialized students of TVE institutions to go through short-term practical training in their chosen careers through Student Industrial Work Experience Scheme (SIWES) in order to enhance their knowledge in their chosen occupation**.** Appropriate educational curricula must be planned and implemented by our institutions of learning, particularly the technical colleges, polytechnics, mono-technics, universities and other specialized training institutions. The curricula must be relevant to the peculiarities of our situation must address most essentially the current industrial demands with the aim of making our technological institutions graduates relevant to the needs of the industry (Ayofe, *eta'l.*, 2009). In an attempt to address these challenges and establish the link between TVE institutions and the industries Reddan & Harrison (2010) argued that TVE institutions need to restructure their programs to be responsive to the needs of the job market, especially the industry. To achieve this goal, TVET curricula must focus on outcomes in terms of the skills, knowledge and attitudes required by industry. That is, TVE provision should be responsive to the demands of industry.

VII. Conclusion and Recommendations

Technical vocational education is considered as the bedrock of industrialization in developed countries the world over because of its impact on productivity, social and economic development. In a developing country like Nigeria, most training programs in TVE institutions are not in line with the needs and aspirations of the labor market and the industries. The institutions are poorly funded and lack of adequate infrastructural and workshop facilities. The linkage between TVE institutions and the industries is very weak. However, the establishment of the linkage program between the TVE institutions and the industries will have profound benefits on the students, teachers and the industries. It will avail the industries with the opportunities to evaluate the performance of the highly motivated TVE graduates,

which will eventually serve as a basis from which the industries can seek for future full-time employees (Kati, *eta'l.,* 2007).

In order to establish effective partnership between the TVE institutions and the industries, the following recommendations are made:

- TVE institutions and local industries should collaborate to organize seminars and workshops where they will share information on the changing trends in industrial practices and how these changes can be incorporated into the curriculum of the schools.
- TVE institutions should be charged with oversight responsibilities by putting up structures and policies that will facilitate industrial- institutional links
- Seminars, field trips and excursions should be incorporated into the academic curriculum, so that it will help expose students to real working environments while still undergoing academic training.
- Government, communities, industries, religious organizations, labor unions Non-Governmental Organizations (NGOs), wealthy individuals in the society should be encouraged to funding of TVE for the smooth running of TVE in terms provision of infrastructural and workshop facilities such as tools, equipment and machines for the acquisition of skills of the graduates of TVE to enable them fit into the labor or be gainfully employed in industries or related organizations or to become self-employed.

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Analysis of Smoke of Diesel Engine by Using Biodiesel as Fuel

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Abstract- This study represents the analysis of smoke of biodiesel by using smoke tester. In this article biodiesel is taken as a fuel instead of diesel and quantity of emitted pollutants HC and CO is evaluated by taking different quantity of biodiesel at different load. This work shows how use of biodiesel will affect the emission of pollutants. Diesel Engine is compression ignition engine and use diesel as fuel, in this engine alternative fuel can be used. One alternate fuel is biodiesel. Biodiesel can be used in pure form or may be blended with petroleum diesel at any concentration in most injection pump diesel engines and also can be used in Vehicle, Railway, and Aircraft as heating oil.

Index Terms- Biodiesel, Diesel Engine, CO, HC, Smoke, Smoke Tester.

I. INTRODUCTION

Bio diesel is the ester based variety oxygenated fuels derived from natural renewable biological sources such as vegetable oils. Biodiesel fuel can be made from new or used vegetable oil and animal fats. Biodiesel is high quality fuel made through a chemical process called Trans esterification. Needed ingredients are vegetable oil, Methanol and Lye (NaOH).

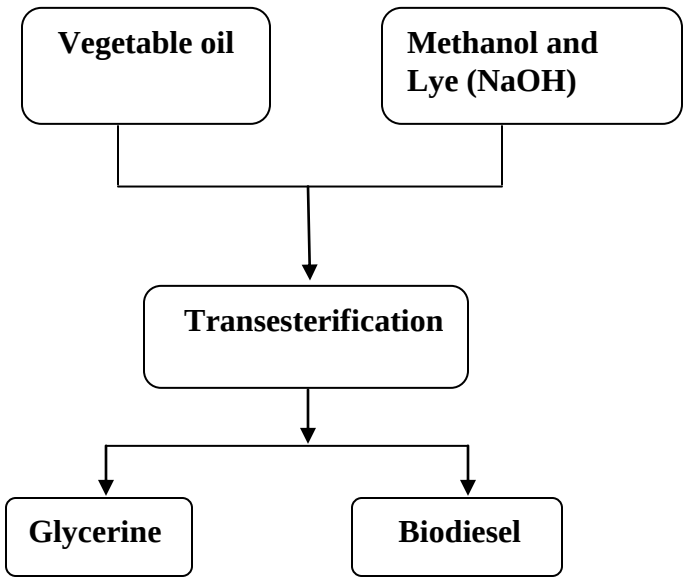

Fig. 1 Process of making Biodiesel

The chemical reaction that occurs through this process breaks down the oil into a layer of biodiesel which rises to the top of the reactor, and a layer of Glycerine which falls to the bottom. The Glycerine is drained is drained off and used for other purposes, composted or otherwise disposed off. The biodiesel is then washed, dried and filtered to remove any extra impurities and its ready to be used as a fuel in diesel engines without any modifications to the engine.

Biodiesel is just twice as viscous as diesel having better lubricating properties and much higher cetane ratings than today's lower sulphur diesel fuels. Biodiesel addition reduces fuel system wear and in low levels in high pressure systems increases the life of the fuel injection equipment that relies on the fuel for its lubrication. Depending on the engine, this might include high pressure injection pumps, pump injectors (also called unit injectors) and fuel injectors.

II. ANALYSIS & FINDINGS

In this study ***castor oil*** is taken as vegetable oil mixed with and forms biodiesel used in different quantity at different load to take observations about emitted pollutants HC and CO.

TABLE I : Quantity of HC in ppm

Load (lbs)	**Biodiesel**	**Fuel consumption (ml)**	**Time of consumption**	**HC (ppm)**
20	20%	40	4 min 30 sec	1700
	50%	60	4 min 30 sec	1500
	80%	40	4 min	1400
	100%	50	4 min 15 sec	1340
40	20%	30	3 min 15 sec	1800
	50%	60	4 min 30 sec	1580
	80%	40	3 min 45 sec	1450
	100%	40	3 min 45 sec	1400
60	20%	40	3 min 30 sec	1800
	50%	40	4 min 30 sec	1650
	80%	60	3 min 45 sec	1560
	100%	60	4 min	1500

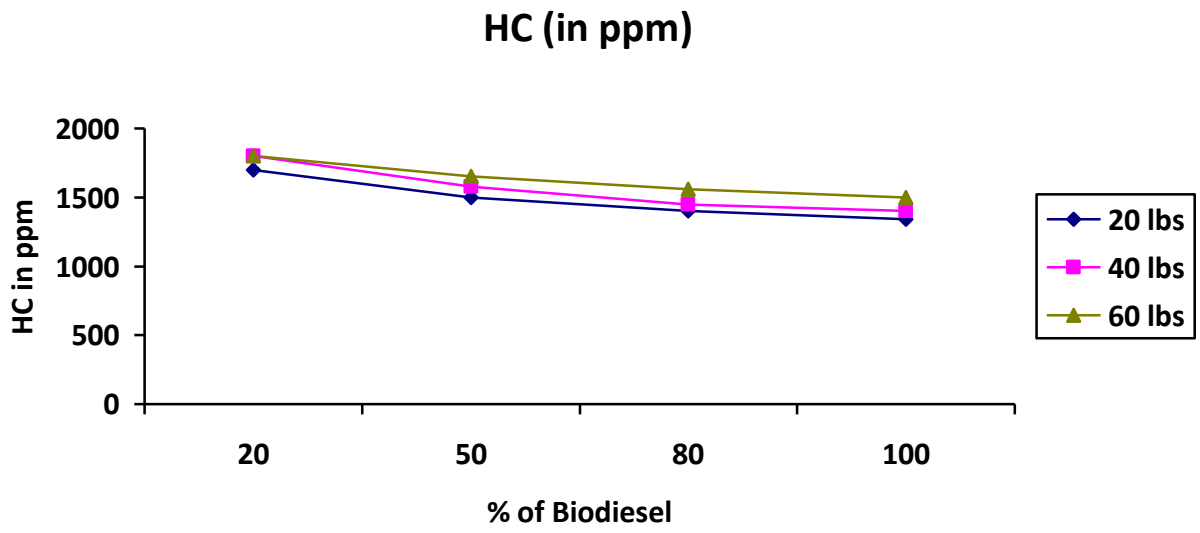

Fig. 2 Graph between quantity of HC & Biodiesel %

TABLE II : Quantity of CO in %

Load (lbs)	Biodiesel	Fuel consumption (ml)	Time of consumption	CO (%)
20	20%	40	4 min 30 sec	2.7
	50%	60	4 min 30 sec	2.5
	80%	40	4 min	2.3
	100%	50	4 min 15 sec	2.1
40	20%	30	3 min 15 sec	2.8
	50%	60	4 min 30 sec	2.6
	80%	40	3 min 45 sec	2.4
	100%	40	3 min 45 sec	2.1
60	20%	40	3 min 30 sec	2.8
	50%	40	4 min 30 sec	2.7
	80%	60	3 min 45 sec	2.5
	100%	60	4 min	2.3

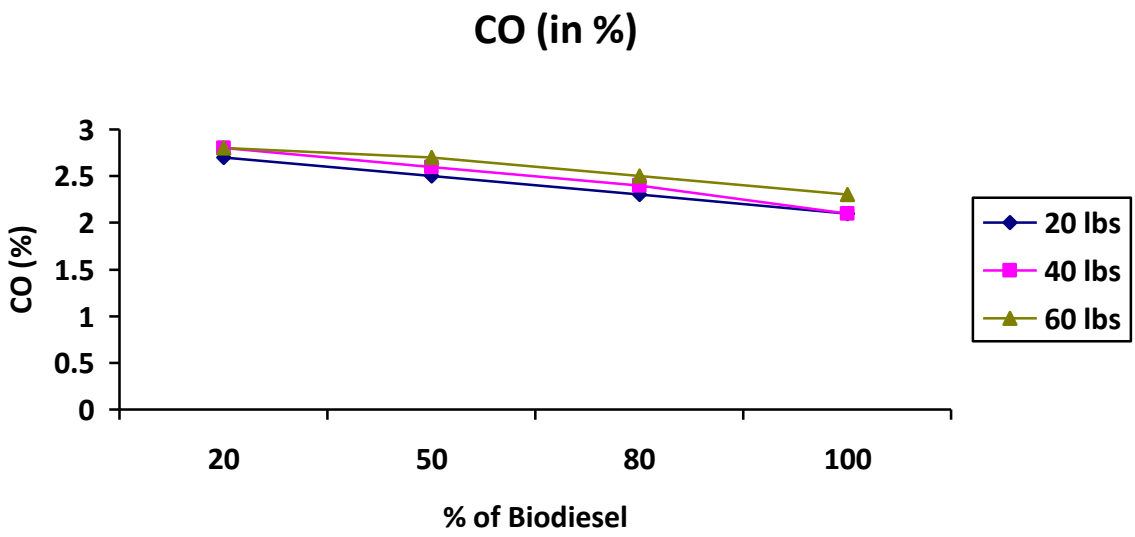

Fig. 3 Graph between quantity of CO % & Biodiesel %

III. RESULTS AND DISCUSSIONS

After experiment we find that hydrocarbons (HC) in ppm decreases with increasing of biodiesel percentage as shown in graph in fig 2.

Order of impact by HC –

Pure diesel > 20% biodiesel mixed diesel > 50% biodiesel mixed Diesel > 80% biodiesel mixed diesel > 100% pure biodiesel

At 20 lbs load –

20 % biodiesel emitted 1700 ppm HC
50 % biodiesel emitted 1500 ppm HC
80 % biodiesel emitted 1400 ppm HC
100 % biodiesel emitted 1340 ppm HC

At 40 lbs load –

20 % biodiesel emitted 1800 ppm HC
50 % biodiesel emitted 1580 ppm HC
80 % biodiesel emitted 1450 ppm HC
100 % biodiesel emitted 1400 ppm HC

At 60 lbs load –

20 % biodiesel emitted 1800 ppm HC
50 % biodiesel emitted 1650 ppm HC
80 % biodiesel emitted 1560 ppm HC
100 % biodiesel emitted 1500 ppm HC

Hence, load increasing also affects the pollutants emission,% of HC increases with load increasing.

And , carbonmonooxide (CO) percentage also decreases with increasing biodiesel percentage as shown in graph in fig.3

Order of impact by CO –

Pure diesel > 20% biodiesel mixed diesel > 50% biodiesel mixed Diesel > 80% biodiesel mixed diesel > 100% pure biodiesel

At 20 lbs load –

20 % biodiesel emitted 2.7 % CO
50 % biodiesel emitted 2.5 % CO
80 % biodiesel emitted 2.3 % CO
100 % biodiesel emitted 2.1 % CO

At 40 lbs load –

20 % biodiesel emitted 2.8 % CO
50 % biodiesel emitted 2.6 % CO
80 % biodiesel emitted 2.4 % CO
100 % biodiesel emitted 2.1 % CO

At 60 lbs load –

20 % biodiesel emitted 2.8 % CO
50 % biodiesel emitted 2.7 % CO
80 % biodiesel emitted 2.5 % CO
100 % biodiesel emitted 2.3 % CO

IV. CONCLUSION

During this study we find that use of Biodiesel in diesel engine reduces the percentage of emitted pollutants, hence with increasing quantity of biodiesel Emission of HC and CO decreases. In this experiment castor oil is taken as vegetable oil and mixed with methanol makes Biodiesel, and this biodiesel used in diesel engine instead of diesel to get the results about emission of HC (hydrocarbons) & CO (carbon monooxide). So we finds quantity of HC & CO reduced with increasing quantity of biodiesel. But this quantity of pollutants increases with load increasing.

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Internet of Things: A New Paradigm

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Abstract- Like other milestones in technology, the Internet of Things enables us to measure what could not be measured before. Until today it remains difficult to come up with a clear definition of the Internet of Things. For companies this means additional information for high resolution management of industry and business processes. For citizens the possible implications are manifold, ranging from consumer empowerment to rethinking society. Different infrastructures and networks will compete and interact in the future Internet of Things. The Internet of Things is not synonymous with ubiquitous / pervasive computing, the Internet Protocol (IP), communication technology, embedded devices, the Internet of People or the Intranet / Extranet of Things, yet it combines aspects and technologies of all of these approaches. This paper elaborates the phrase Internet of things, discusses and analyses various definitions of IoT in use and try to distinguish it from other related technologies/concepts.

Index Terms- : IoT, RFID, Things, ubiquitous

I. INTRODUCTION (THE EVOLUTION OF IOT)

Communication between computers started with the EDI (Electronic Data Interchange) that made possible direct dialogue between two PCs. With the Internet, all the computers connected to the Internet can talk to each other and with the connection of mobile phones, the connection has become mobile. The evolution that we have been witnessing in the last few years is an extension of this Internet to all the things that surround us. Up until now the Internet has been the first place for uniting people by means of different types of social media (Email, Blog, Facebook, Twitter, Flickr etc), now it is being transformed into the tool that will allow all objects to interact and in certain cases to gain access to the collective knowledge that they will generate. The growth of the Internet is an ongoing process: only twenty-seven years ago it was connecting about a thousand hosts and has grown ever since to link billions people through computers and mobile devices. One major next step in this development is to progressively evolve from a network of interconnected computers to a network of interconnected objects, from books to cars, from electrical appliances to food, and thus create an 'Internet of things' (IoT) [1]. The concept of the Internet of Things first became popular through the Auto-ID center and related market analysts publications [2]. If all objects of daily life were equipped with radio tags, they could be identified and inventoried by computers [3][4]. However, unique identification of things may be achieved through other means such as barcodes or 2D-codes as well. With all objects in the world equipped with minuscule identifying devices, daily life on Earth would undergo a transformation [5][6]. Companies would not run out of stock or waste products, as involved parties would know which products are required and consumed[6]. Mislaid and stolen items would be easily tracked and located, as would the people who use them. Our ability to interact with objects could be altered remotely based on your current status and existing user agreements [3]. These objects will sometimes have their own Internet Protocol addresses, be embedded in complex systems and use sensors to obtain information from their environment (e.g. food products that record the temperature along the supply chain) and/or use actuators to interact with it (e.g. air conditioning valves that react to the presence of people). The scope of IoT applications is expected to greatly contribute to addressing today's societal challenges: health monitoring systems will help meet the challenges of an ageing society [7]; connected trees will help fight deforestation [8]; connected cars will help reduce traffic congestion and improve their recyclability, thus reducing their carbon footprint. This interconnection of physical objects is expected to amplify the profound effects that large-scale networked communications are having on our society, gradually resulting in a genuine paradigm shift. To complement this overview, it is worth noting three points that highlight the complex nature of IoT. First, it should not be seen as a mere extension of today's Internet but rather as a number of new independent systems that operate with their own infrastructures (and partly rely on existing Internet infrastructures). Second, as detailed in a recent ISTAG report [9], IoT will be implemented in symbiosis with new services. Third, IoT covers different modes of communication: things-to-person communication and thing-to-thing communications, including Machine-to-Machine (M2M) communication that potentially concerns 50-70 billion 'machines', of which only 1 % are connected today[10]. These connections can be established in restricted areas ('intranet of things') or made publicly accessible ('Internet of things'). The advent of IoT is taking place in an ICT environment affected by several major trends [11]. 'Scale' is one of them: the number of connected devices is increasing, while their size is reduced below the threshold of visibility to the human eye. 'Mobility' is another: objects are ever more wirelessly connected, carried permanently by individuals and geo-localisable. 'Heterogeneity and complexity' is a third trend: IoT will be deployed in an environment already crowded with applications that generate a growing number of challenges in terms of interoperability.

II. VARIOUS TECHNOLOGY COMPONENTS IMPLEMENTING IOT

Services:

Systems integrators and services organizations provide integration and solution implementation services for IoT projects.

Software:

Middleware and application infrastructure vendors provide information for analytical engines regarding IoT endpoints and enable vertical market solutions.

Hardware:

GPS chips, RFID sensors, actuators, and embedded and external hardware devices capture location and status information.

Network:

Network access, satellite, and transport infrastructure vendors provide the network connectivity that underlies IoT solutions.

Analytics Solutions:

Business intelligence and analytical software solutions such as data mining and predictive analytics, video image analysis, pattern recognition, and artificial intelligence algorithms determine whether to act on or ignore a pattern.

III. THINGS IN IoT

In the context of "Internet of Things" a "thing" could be defined as a real/physical or digital/virtual entity that exists and move in space and time and is capable of being identified. [Internet of Things – Strategic Research Roadmap 2009]

Things are commonly identified either by assigned identification numbers, names and/or location addresses. In the future, 'things' as boring as household appliances could be networked to the Internet. Hypothetically speaking, a fridge could be aware of what is stored inside and order another litre of milk when it runs out or even make people aware that it is close to its expiry date. Things are expected to become active participants in business, information and social processes where they are enabled to interact and communicate among themselves and with the environment by exchanging data and information "sensed" about the environment, while reacting autonomously to the "real/physical world" events and influencing it by running processes that trigger actions and create services with or without direct human intervention. The Internet of Things allows people and things to be connected anytime, anyplace, with anything and anyone, ideally using any path/network and any service. This implies addressing elements such as convergence, content, collections (repositories), computing, communication, and connectivity in the context where there is seamless interconnection between people and things and/or between things and things. The applications are endless. 'It's an exciting time for anyone involved with RFID and its associated technologies. The international move towards the development of an "Internet of Things" continues to raise the profile of RFID as a technology with enormous potential to revolutionise global markets, bringing both economic and societal benefits.

IV. DIFFERENT VIEWS AND CONCEPTS ON IoT DEFINITIONS

The term Internet of Things is not well defined and has been used and misused as a buzzword in scientific research as well as marketing and sales strategies. Until today it remains difficult to come up with a clear definition of the Internet of Things. Followings are some popular definitions.

(i) According to Cluster of European Research Projects on the Internet of Things (CERP-IoT 2009): "Internet of Things (IoT) is an integrated part of Future Internet and could be defined as a dynamic global network infrastructure with self-configuring capabilities based on standard and interoperable communication protocols where physical and virtual 'things' have identities, physical attributes, and virtual personalities and use intelligent interfaces, and are seamlessly integrated into the information network. In the IoT, 'things' are expected to become active participants in business, information and social processes where they are enabled to interact and communicate among themselves and with the environment by exchanging data and information 'sensed' about the environment, while reacting autonomously to the 'real/physical world' events and influencing it by running processes that trigger actions and create services with or without direct human intervention. Interfaces in the form of services facilitate interactions with these 'smart things' over the Internet, query and change their state and any information associated with them, taking into account security and privacy issues."

(ii) The Internet of Things is a concept in which the virtual world of information technology integrates seamlessly with the real world of things. The real world becomes more accessible through computers and networked devices in business as well as everyday scenarios. With access to fine-grained information, management can start to move freely from macro to micro levels and will be able to measure, plan and act accordingly. However, the Internet of Things is more than a business tool for managing business processes more efficiently and more effectively – it will also enable a more convenient way of life.

(iii) CASAGRAS project coordinator Ian G Smith said. The Internet of Things is "A world where physical objects are seamlessly integrated into the information network, and where the physical objects can become active participants in business processes. Services are available to interact with these 'smart objects' over the Internet, query and change their state and any information associated with them, taking into account security and privacy issues."

(iv) Another approach towards a definition of the Internet of Things can be derived from logistics where it is common to ask for the *right product* in the *right quantity* at the *right time* at the *right place* in the *right condition* and at the *right price*. In this analogy the *right product* relates to accurate and appropriate information about a uniquely identifiable physical object as well as its form, fit and function. This includes the usage of Auto-ID and appropriate sensor information or any other kind of linked information to the object that can be accessed through the Internet of Things. The *right quantity* can be achieved through high granularity of information combined with filtering and intelligent processing. The *right time* does not necessarily mean anytime, but more precisely 'when needed'. It may be sufficient

to receive information about an object only once a day or only in the case of a status change. Consequently, right-time does not equal real-time, a term that is mentioned quite often in relation to the Internet of Things.

(v) The future Internet of Things links uniquely identifiable things to their virtual representations in the Internet containing or linking to additional information on their identity, status, location or any other business, social or privately relevant information at a financial or non-financial pay-off that exceeds the efforts of information provisioning and offers information access to non-predefined participants. The provided accurate and appropriate information may be accessed in the right quantity and condition, at the right time and place at the right price.

(vi) While there's been no single standard definition of IoT solutions since the possibilities of Internet of Things was first discussed, a variety of technologies are used in implementation (e.g., RFID tags, barcodes, and GPS technologies). These technologies are used to monitor and transfer the status of physical assets to remediate business problems like supply chain inefficiencies in industries such as manufacturing, healthcare, transportation, and retail, as well as to inspire innovation in organizations. For example, information embedded in RFID-tagged containers is collected by readers along the supply chain at the supplier's shipping department, throughout the transportation process, and during storage, collection, and installation along the assembly line. The availability of accurate, current data about supply chain events enables firms to deliver superior customer service while reducing inventory held as a buffer against uncertainty. Other IoT applications include food traceability, patient monitoring, and medical device and medicine tracking. A variety of forces are aligning to drive increased enterprise demand for IoT solutions, including declining device costs, widely deployed wireline and wireless IP networks, and regulatory mandates.

In the year 2012, various survey were conducted in Europe to understand: i) how well the term ''Internet of Things'' resonates with global enterprises; ii) what timelines corporations have for implementing IoT applications and solutions; iii) the benefits and challenges associated with IoT solution deployment; iv) and which third-party partners companies seek assistance from as they deploy IoT solutions. The study methodology included an online survey of hundreds of global enterprise IT decision-makers in specific industries including manufacturing; healthcare; oil, gas, and petroleum; retail and hospitality; transportation and logistics; and government. These decision-makers participated in or were aware of their corporate initiatives related to IoT solutions and applications.

Key results from this online study include:

• *Enterprise decision-makers have a positive perception of IoT solutions***.** 64% of survey respondents were familiar with the term ''Internet of Things.'' More than 70% of the respondents who were familiar with the term have a positive perception of it. In addition, 85% of the respondents who were familiar with the term strongly or completely agree with the following definition of ''Internet of Things solutions''.

''Smart interconnected devices that businesses use to get more visibility into the identification, location, and condition of products, assets, transactions, or people to drive more effective and timely business decisions or to improve customer interactions.''

• *Barcodes and RFID devices are valuable components of IoT solution implementations.* Firms consider many different devices valuable to implementing IoT solutions. Specifically, 72% of enterprises state that barcodes are valuable to enable IoT solutions; 71% say the same thing about real-time location tracking technologies like active RFID and 58% of firms consider passive RFID devices as valuable IoT solution technologies.

• *Many firms are planning to deploy IoT solutions and applications in the future.* Companies are in the early stages of deploying IoT solutions, as only 15% of surveyed enterprises already have an IoT solution in place. However, future momentum for these solutions appears strong: 53% of surveyed enterprises are planning to implement IoT solutions within the next 24 months.

V. ANALYSIS OF DEFINITIONS

While these definitions list the possible technical components of the Internet of Things, they still have some major shortcomings. *Firstly*, they list components that have been mentioned before in relation to other visions such as pervasive or ubiquitous computing and therefore it is difficult to distinguish IOT from these concepts. *Secondly*, it misses wider consideration of current developments and user interactions in the Internet commonly referred to as Web 2.0/Web 4.0. Similar to the relationship between the World Wide Web (WWW) and the Internet, the addition of Web x.0 functionality may be seen as a user-centric extension to the Internet of Things rather than an integral part of it. However, whereas the development of the Internet began more than thirty years before the realisation of the WWW in the early 1990s, the Internet of Things is already being influenced by Web 2.0 functionality right from the beginning. Both technology developments have been happening in parallel rather than consecutively. *Thirdly*, they do not provide a reason why or how the Internet of Things will be a self-sustainable and successful concept for the future. Self-sustainability encompasses viability, including a dynamic global network infrastructure with self-configuring capabilities based on standards and interoperable communication protocols as well as openness for future extensions, ideas, and technologies. Economic success may never have been a part of a definition for the Internet or other technical network infrastructures. Nevertheless, we consider it a valid consideration within a holistic definition approach as economic success and adoption is just as important as technical sustainability in a forward-looking statement.

For the purposes of differentiation it may be best to consider what the Internet of Things is not – or at least not exclusively. A corresponding blog discussion has been started by Tomas Sánchez López (Sánchez López 2010). He considers that the Internet of Things is not only:

• *ubiquitous / pervasive computing*, which does not imply the usage of objects nor does it require a global Internet infrastructure

• the *Internet Protocol (IP)*, as many objects in the Internet of Things will not be able to run an Internet Protocol

• a *communication technology*, as this represents only a partial functional requirement in the Internet of Things similar to the role of communication technology in the Internet
• an *embedded device*, as RFID tags or Wireless Sensor Networks (WSN) may be part of the Internet of Things, but stand-alone they miss the back-end information infrastructures and in the case of WSN the standards to relate to 'things'
• the *application*, just as Google or Facebook could not be used in the early 90's to describe the possibilities offered by Internet or WWW

With these negations in mind it is easier to differentiate the Internet of Things. However, all fields of research that has been mentioned above overlap partially with IoT.

VI. CONCLUSION AND OUTLOOK

Internet of things can help to improve citizens' quality of life, delivering new and better jobs for workers, business opportunities and growth for the industry. The Internet of Things will make it possible for objects to get information about their position in the world, to interact with other objects and to have access to comparative information for data gathered in their vicinity. Future developments in the Internet of Things will optimise the information flow in industrial and social scenarios and revolutionise business and private communication. Thanks to the recent advances of miniaturisation and the falling costs for RFID, sensor networks, NFC, wireless communication, technologies and applications, the Internet of Things suddenly became relevant for industry and end-users. Detection of the physical status of things through sensors, together with collection and processing of detailed data, allows immediate response to changes in the real world. This fully interactive and responsive network yields immense potential for citizens, consumers and business.

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Cloud Computing Server Operating System Availability Enhancement by Utilizing the USB Flash and SD Card Devices

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***Abstract*-** Cloud computing is rapid growing revolution in the way of information technology. As it grows rapidly providing high availability is a major task. This paper examines the cloud computing in the context of how it provides high availability by utilizing USB devices in the servers and explores the revolutionary transformations and challenges it brings to IT management.

***Index Terms*-** SaaS, PaaS, IaaS, VMs.

I. INTRODUCTION

Cloud computing enlargement has taken all the attention of various communities like researches, student, business, consumer and government organizations. Cloud Computing is a marketing term which is also known as utility, computing, deliver the service as software, platform and infrastructure as a service in pay-as-you-go model to consumers. Berkeley report says on this services as "Cloud computing, the long held dream of computing as a utility, has the potential to transform a large part of the IT industry, making software even more attractive as a service" [4]. However, the basic idea behind the cloud model is that anything that could be done in computing whether on an individual PC or in a corporate data center from storing data to collaborating on documents or crunching numbers on large data sets can be shifted to the cloud. Certainly, cloud computing enables a new platform and location-independent perspective on how we communicate, collaborate and work. So long as you can access the Web, you are able to work when and where you wish. With fast, reliable Internet connectivity and computer power, it does not matter where the document, the e-mail or the data the user sees on the screen comes from.

Cloud computing enables providers to use distant data centers for cloud computing. Still, while some have predicted the end of the PC era with the rise of the cloud computing model, many believe that most organizations and even individuals will continue to make use of traditional PCs and laptops, even if more and more of their use will be to access the cloud [1].

Many organizations embrace both public and private cloud computing by integrating the two models into hybrid clouds. These hybrids are designed to meet specific business and technology requirements, so that activities and tasks are allocated to traditional IT, external or internal clouds, as appropriate, helping to optimize security and privacy with a minimum investment in fixed IT costs.

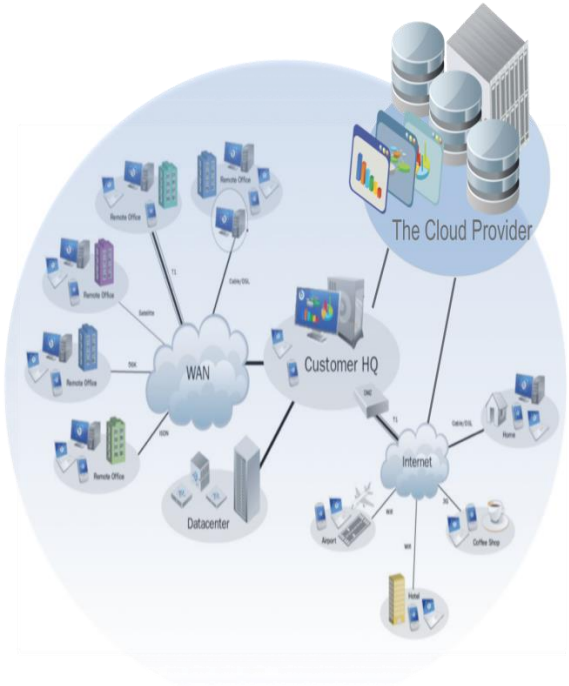

Figure: Cloud computing and Cloud storage

In addition to the different cloud computing models, there are distinctions among the most common cloud service models as shown in Figure 1. Available to anyone with Internet access, cloud service models include:

●Software as a Service (SaaS) cloud model Enables software to be delivered from a host source over a network as opposed to installations or implementations.

●Platform as a Service (PaaS) cloud model Enables operating systems and middleware services to be delivered from a managed source over a network.

●Infrastructure as a Service (IaaS) cloud model Enables the entire infrastructure to be delivered as a service over a network, including storage, routers, virtual systems, hardware and servers.

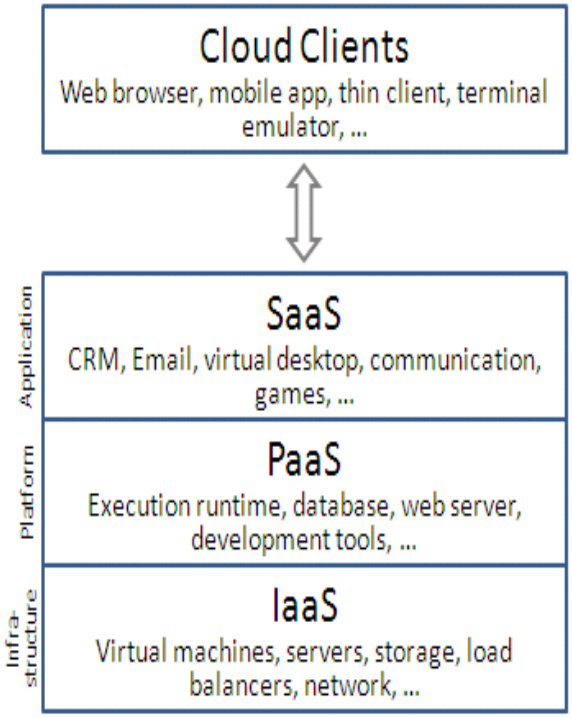

Figure 1: Cloud computing service models

II. Cloud Infrastructure as a Service (IaaS).

The capability provided to the consumer is to provision processing, storage, networks, and other fundamental computing resources where the consumer is able to deploy and run arbitrary software, which can include operating systems and applications. The consumer does not manage or control the underlying cloud infrastructure but has control over operating systems, storage, deployed applications, and possibly limited control of select networking components [7].

2.1Virtualization

One of the main cost-saving, hardware-reducing, and energy-saving techniques used by cloud providers is virtualization. Virtualization is done with software-based computers that share the underlying physical machine resources among different virtual machines (VMs). With OS virtualization each VM can use a different operating system (OS), and each OS is isolated from the others. Many companies use VMs to consolidate servers, enabling different services to run in separate VMs on the same physical machine. VMs allow time-sharing of a single computer among several single-tasking operating systems. Utilizing VMs requires the guest operating systems to use memory virtualization to share the memory of the one physical host. Memory Virtualization removes volatile random access memory (RAM) resources from individual systems and aggregates those resources into a virtualized memory pool available to any computer in the cluster. Memory virtualization leverages large amount of memory which improves overall performance, system utilization, and increased efficiency. Allowing applications on multiple servers to share data without replication also reduce the total amount of memory needed [6].

2.2 Computing and Storage

Data centers have emerged in the past few years as a new paradigm for interconnecting computing and storage on a massive scale. There are several viewpoints from which to approach the development of data centers: as the outgrowth of large web server farms (for web-hosting), as the convergence of computing and networking (high-performance computing as typified by the Cloud Computing paradigm) and as a convergence of local area networks and storage networks. Several technological innovations have spurred the rapid deployment of data centers; notably, 10 Gbps Ethernet technology, the specification of Fiber Channel over Ethernet (FCoE) standards, server virtualization, and the development of high-performance Network Interface Cards (NICs) [5].

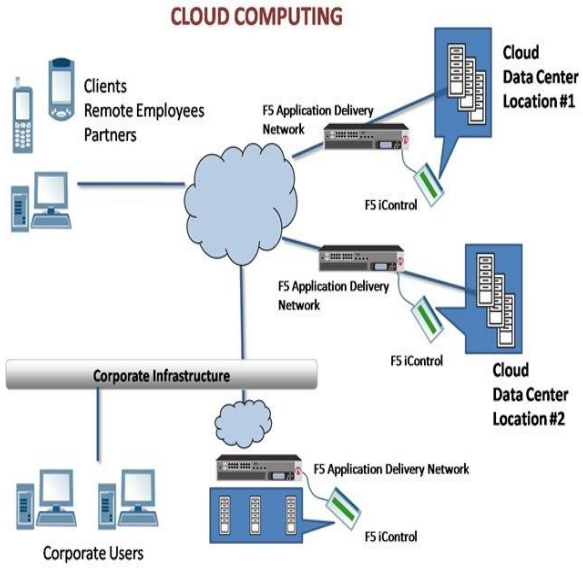

Figure: Data center managed services

While cloud computing models are attractive because of their flexibility and cost effectiveness, certain challenges must be addressed in order to provide a viable option to traditional data services.

The externalized aspect of outsourcing can make it harder to maintain data integrity and privacy, support data and service availability, demonstrate compliance, and secure highly available access to applications and information. In short, cloud computing can present an added level of risk.

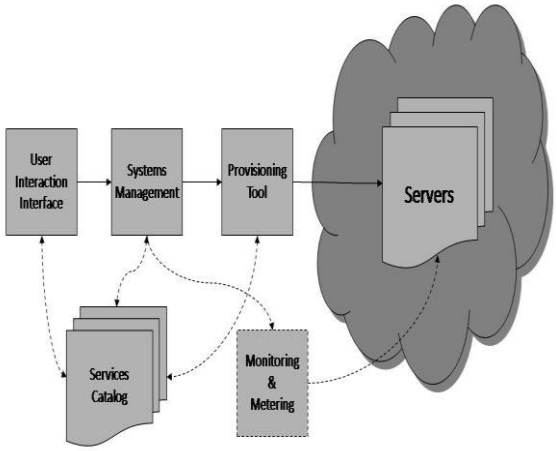

Figure: Server Management Architecture

In contrast, high availability can be achieved through software that runs on top of an operating system, through the chosen server architecture or even as a feature within the application itself. Applications that have high-availability properties built in eliminate the need to run additional software. High-availability systems are often designed to quickly recover from a component failure within the architecture, sometimes with a very limited amount of application downtime or with a period of degraded capacity, while the failed components are quickly resolved.

Since availability of servers is the major issue in the cloud computing. Failure or crashing of any operating system leads to increase the downtime. For this USB Flash and SD cards are used as recovery. Server operating system is installed on USB devices which are need to be connected to the server hardware. After installation of operating system set the boot sequence order of server with hard disk(From where the current Server operating system is booted), USB flash and SD card. If there is failure or crash occurs in the present operating system then the operating system will recover from the USB devices. This will increases the performance rate of the system by reducing the down time of server.

III. Conclusion

This paper described the cloud computing can be used to address strategic problems with which IT continually deals like availability and reliability. Computing in the context of how it provides high availability by utilizing USB devices in the Servers. The advantages are clear with the main ones being resiliency, efficiency, scalability, flexibility and easier outsourcing. Cloud computing genuinely does have the potential to radically change the way organizations purchase, manage and provide computing resources to their employees.

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Digital Light Processing and its Future Applications

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Abstract- Electronic projection display technology had its origin in the Oil film projector system. Developed in the early 1940s oil film projectors have been the workhorse for applications that require projection displays of the highest brightness. But due to various limitations a number of solid state technologies have challenged the oil firm projector, including CRT-LCD light valves and active-matrix- LCD panels. In response to various limitations of the LCD technologies and to provide superior image quality, high brightness projection display systems have been developed based on Digital Light Processing technology. DLP is based on the Digital Micro Mirror Device (DMD), a Micro Electro Mechanical System, invented in 1987 by Larry Hornbeck at Texas Instruments. DMD is a fast, reflective digital light switch that precisely controls a light source using a binary pulsewidth modulation technique. It can be combined with image processing, memory, a light source, and optics to form a DLP system capable of projecting large, bright, high contrast colour images with better colour fidelity and consistency than current displays.

This paper describes the structure, working, operation, advantages and future use of the DLP system. It also compares the DLP system with other display technologies and presents the DLP technology based products in the market.

Index Terms- Digital Micromirror Device, DMD, Digital Light Processing, DLP, micro electromechanical systems, MEMS, DMD Modulator, digital imaging, projection displays.

I. INTRODUCTION

Digital Light Processing is a revolutionary new way to project and display information. Based on the Digital Micro mirror Device developed by Texas Instruments, DLP creates the final link to display digital visual information. DLP technology is being provided as subsystems or "engines" to market leaders in the consumer, business, and professional segments of the projection display industry. In the same way the compact disc revolutionized the audio industry, DLP will revolutionize video projection.

II. DLP TECHNOLOGY

DLP technology is based on an optical semiconductor, called a Digital Micromirror Device (DMD), which uses mirrors made of aluminum to reflect light to make the picture. The DMD is often referred to as the DLP chip. The chip can be held in the palm of your hand, yet it can contain more than 2 million mirrors each, measuring less than one-fifth the width of a human hair. The mirrors are laid out in a matrix, much like a photo mosaic, with each mirror representing one pixel.

When you look closely at a photo mosaic, each piece of it holds a tiny, square photograph. As you step away from the mosaic, the images blend together to create one large image. The same concept applies to DMDs. If you look closely at a DMD, you would see tiny square mirrors that reflect light, instead of the tiny photographs. From far away (or when the light is projected on the screen), you would see a picture.

The number of mirrors corresponds to the resolution of the screen. DLP 1080p technology delivers more than 2 million pixels for true 1920x1080p resolution, the highest available.

III. A BRIEF HISTORY OF INNOVATION

1977

• Texas Instruments scientist Dr. Larry Hornbeck begins exploring how the principles of reflection can be used to manipulate light.

1987

• Dr. Hornbeck develops the Digital Micromirror Device: an optical semiconductor capable of steering photons with unparalleled accuracy.

1992

• Texas Instruments forms the Digital Imaging Venture Project to explore the commercial viability of the Digital Micromirror Device.

1993

• Digital Light Processing technology is named; the Digital Imaging division (later to become the DLP® Products division) is established to unlock its potential for commercial projection display applications.

1994

• Prototype projectors are used to publicly demonstrate Digital Light Processing™ technology for the first time.

1995

• The DLP® Products division of TI announces its first customer agreements.

1996

• The first commercial DLP® systems are shipped to InFocus, nView and Proxima. Digital Projection signs on to manufacture DLP® projectors.

1997

• February: The Motion Picture Academy of Arts and Sciences chooses DLP® technology to project the Oscars®; DLP® technology has been used at the Academy Awards® ever since.

1998

• June: DLP® Products receives an Emmy Award for Outstanding Achievement in Engineering Development from the Academy of Arts and Sciences. Dr. Larry Hornbeck also received an Emmy for inventing DLP® technology.

1999

• January: The first DLP® projector specifically engineered for home theatre is shown by Dream Vision.

• June: DLP Cinema® projector technology is publicly demonstrated for the first time on two screens in Los Angeles and New York.

2000

• March: TI announces agreements with Christie and Barco to manufacture digital cinema projectors with DLP Cinema® technology.

• May: The world's first sub-3-lb. DLP® projector is introduced by PLUS Corporation, demonstrating DLP® technology's capability to lead the market in portable projectors.

2001

• January: Sharp announces the first 16:9 projector, greatly anticipated by home theatre enthusiasts.

• March: Mitsubishi launches a flagship DLP® HDTV based on HD1 chipset.

• June: The first sub-2-lb.projector is announced by InFocus with its ground-breaking LP120 DLP® projector.

2002

• January: Samsung announces their first DLP® HDTV, priced at $3,999.

• April: HP enters the projector market with DLP® technology.

• May: Dell enters the projector market with DLP® technology.

• June: NEC is named the third DLP Cinema® manufacturer partner.

2003

• March: DLP Cinema® 2K resolution chip is introduced at ShoWest.

2004

• DLP® becomes number one supplier of microdisplay technology, according to TSR.

• InFocus becomes first TI customer to ship 1 million DLP® projectors.

2005

• January: HP, Optoma and Radio Shack introduce the first "Instant Theatre" projectors, incorporating sounds system and DVD player with DLP® projection into one, consumer-friendly unit.

2006

• January: DLP® technology achieves greater than 50% market share in the worldwide front projection market for first time.

• January: TI introduces DLP® HDTVs with LED technology.

• March: DLP Cinema® projectors surpass 1,000 deployed milestones; 1,200 projectors deployed worldwide.

IV. HOW DLP WORKS?

DLP display technology digitally manipulates (or processes) light to produce film-like, all-digital images. DLP integrates a projection lamp and an electronic video signal from a source such as a VCR or computer, and the processed light produces an all-digital picture.

The key to this complete digital process is the Digital Micro mirror Device (DMD), a thumbnail-size semiconductor light switch. The DMD consists of an array of thousands of microscopic-size mirrors, each mounted on a hinge structure so that it can be individually tilted back and forth. When a lamp and a projection lens are positioned in the right places in the system, DLP processes the input video signal and tilts the mirrors to generate a digital image.

Imagine that you are in a football stadium at night and it's half-time. All the lights in the stadium are turned off; there is a blimp floating a few hundred feet above the field.

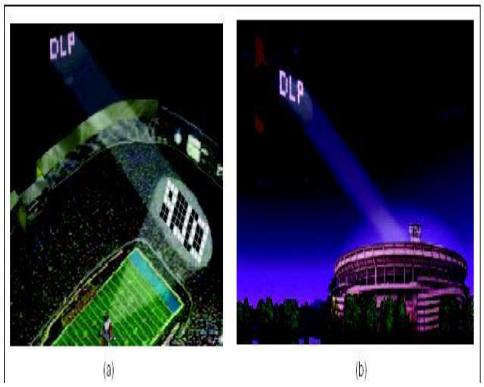

Figure 1. Fans in a stadium reflecting light toward a blimp

In figure 1(a) Light is projected from a spotlight toward fans in the stadium. When cued by a numbered signal, these fans hold up their reflective seat cushions and tilt them toward or away from a blimp. By doing so, the fans in the stadium are reflecting pixels of light toward the blimp. The result is that the light pattern created by the seat cushions projects an image onto the surface of the blimp. (b) A distant viewer sees the image on the blimp.

As part of the half-time show, a powerful spotlight placed at the 50 yard line blasts light into the stands. All the fans in one area of the stadium are asked by the announcer to hold up the reflective seat cushions that were in their seats before the game. Each reflective cushion has a different number on the back. The announcer asks you to tilt your seat cushion so that the light from

the spotlight reflects directly up to the blimp each time your number appears on the scoreboard screen. If your number is not displayed, you are to tilt the cushion away from the blimp and direct light down to the field. Each fan in the stands now controls a pixel of light. You can imagine that if some fans reflect the light toward the blimp, an image of some sort will appear on its side [Figure 1(a)].

Now imagine a viewer looking toward the stadium and the blimp from a remote distance [Figure 1(b)]. When looking at the blimp, he or she will see an image on the side of the blimp that is generated by the sports fans tilting their reflective seat cushions and reflecting light onto the side of the blimp.

DLP technology accomplishes this same task, but it does so by processing light that is focused onto the DMD (Figure 2). At speeds greater than 1,000 times per second, the mirrors are electronically
tilted. Light from a lamp is digitally reflected from the DMD, through a projection lens, and onto a screen. Colour is added through a colour wheel filter system.

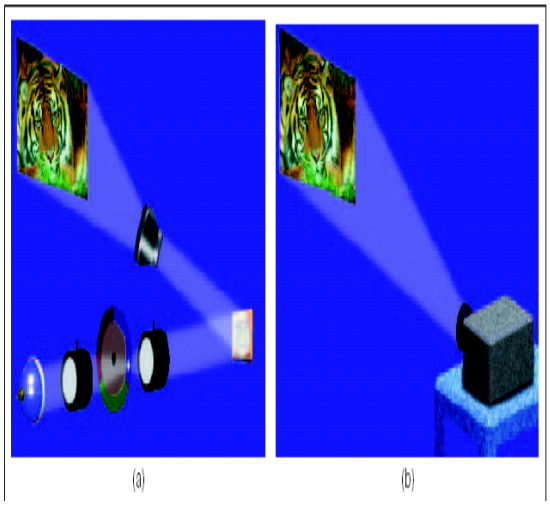

Figure 2. DLP system operation

In much the same way the image was formed on the side of the blimp (figure 2(a)). DLP forms images for video. Shining light on the DMD and tilting the mirrors creates a digital image. Colour is added by placing a red, green, and blue colour wheel filter system in the optical path. As the wheel spins, the mirrors are tilted on for the exact amount of time required for each colour. At any given instant, only one of the primary light colour is hitting the DMD, but when the filter system spins fast enough, the colour blend to create a full-colour digital image. In figure 2(b), a viewer sees an image on the screen from a DLP projector.

V. DLP STRUCTURE

A Digital Micro Mirror Device chip is the heart of Digital Light Processing projector, DMD can be described simply as a semiconductor light switch. The micro mirrors are mounted on the DMD chip and it tilts in response to an electrical signal. The tilt directs light toward the screen, or into a "light trap" that eliminates unwanted light when reproducing blacks and shadows. Other elements of a DLP projector include a light source, a colour filter system, a cooling system, illumination and projection optics.

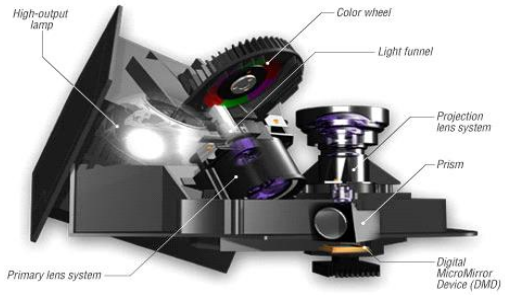

Figure 3. DLP Projector internal structure

A DLP based projector system includes memory and signal processing to support a fully digital approach. Depending on the application, a DLP system will accept either a digital or analog signal. Analog signals are converted into digital in the DLPs front –end processing. Any interlaced video signal is converted into an entire picture frame video signal through interpolative processing. The signal goes through DLP video processing and becomes progressive Red (R), Green (G) and Blue (B) data. The progressive RGB data is then formatted into entire binary bit planes of data.

A. DMD Basic Principle

A DMD chip has on its surface several hundred thousand microscopic mirrors arranged in a rectangular array which correspond to the pixels in the image to be displayed. The mirrors can be individually rotated ±10-12°, to an on or off state. In the on state, light from the projector bulb is reflected into the lens making the pixel appear bright on the screen. In the off state, the light is directed elsewhere, making the pixel appear dark.

To produce greyscales, the mirror is toggled on and off very quickly, and the ratio of on time to off time determines the shade produced. Contemporary DMD chips can produce up to 1024 shades of gray.

The mirrors themselves are made out of aluminium and are around 16 micrometres across. Each one is mounted on a yoke which in turn is connected to two support posts by compliant torsion hinges. In this type of hinge, the axle is fixed at both ends and literally twists in the middle. Tests have also shown that the hinges cannot be damaged by normal shock and vibration, since it is absorbed by the DMD superstructure.

Two pairs of electrodes control the position of the mirror by electrostatic attraction. Each pair has one electrode on each side of the hinge, with one of the pairs positioned to act on the yoke and the other acting directly on the mirror. The majority of the time, equal bias charges are applied to both sides simultaneously.

To move the mirrors, the required state is first loaded into an SRAM cell located beneath each pixel, which is also connected to the electrodes. Once all the SRAM cells have been loaded, the bias voltage is removed, allowing the charges from the SRAM cell to prevail, moving the mirror. When the bias is restored, the mirror is once again held in position, and the next required movement can be loaded into the memory cell.

The bias system is used because it reduces the voltage levels required to address the pixels such that they can be driven directly from the SRAM cell, and also because the bias voltage can be removed at the same time for the whole chip, so every mirror moves at the same instant.

B. DMD Modulator

The DMD chip is comprised of over one million mirrors. The size of each mirror is less than 1/5" the width of a human hair. The DMD is monolithically fabricated by Complementary Metal Oxide Semiconductor-like processes over a CMOS memory.

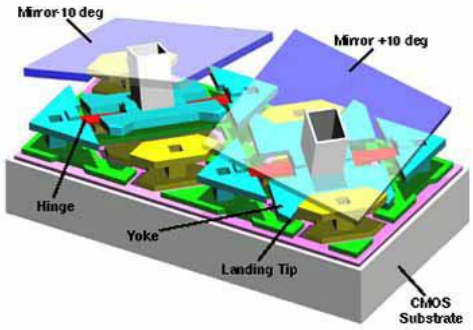

Figure 4. Isometric schematic of DMD pixels

Each light switch has an aluminum mirror, 16 µm² that can reflect light in one of two directions depending on the state of the underlying memory cell. Rotation of the mirror is accomplished through electrostatic attraction produced by voltage differences developed between the mirror and the underlying memory cell. With the memory cell in the on (1) state, the mirror rotates to +10°, with the memory cell in the off (0) state, the mirror rotates to -10°.

The mechanical portion of each pixel consists of a three layer structure. The centre layer, called beam layer, is suspended over the bottom electrode layer by thin torsion hinges. The top mirror layer is attached to the beam layer with a via post. The yoke may rotate about the torsion hinge axis to either side, landing on the electrode layer at specific tilt angles of +/- 10 degrees. Manipulation of the mirrors is accomplished electro statically utilizing the address electrodes on either side of the torsion hinge. These address electrodes are tied to the SRAM cell residing in the silicon backplane beneath each mirror structure.

After passing through condensing optics and a colour filter system, the light from the projection lamp is directed at the DMD. When the mirrors are in 'on' position, they reflect light through the projection lens and onto the screen to form a digital, square-pixel projected image. Each mirror on the DMD array is electro statically tilted to the 'on' or 'off' positions. The technique that determines how long each mirror tilts in either direction is called pulse width modulation. The mirrors are capable of switching on and off more than 1000 times a second this rapid speed allows digital gray-scale and colour reproduction.

C. DMD Cell Structure

The DMD pixel is a monolithically integrated MEMS superstructure cell fabricated over a CMOS SRAM cell. Figure 5 shows the pixel structure of in an exploded view illustrating how the various layers interrelate, including the underlying static random access memory (SRAM) cell that is used to address the pixel. An organic sacrificial layer is removed by plasma etching to produce air gaps between the metal layers of the superstructure. The air gaps free the structure to rotate about two compliant torsion hinges. The mirror is rigidly connected to an underlying yoke. The yoke, in turn, is connected by two thin, mechanically compliant torsion hinges to support posts that are attached to the underlying substrate. The address electrodes for the mirror and yoke are connected to the complementary sides of the underlying SRAM cell.

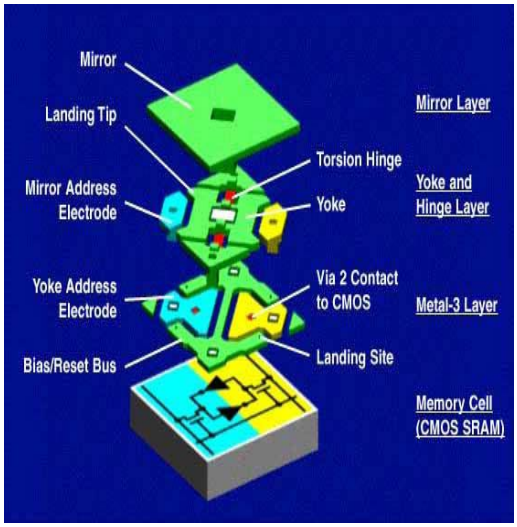

Figure 5. Pixel structure of a DMD cell

The yoke and mirror are connected to a bias bus fabricated at the metal-3 layer. The bias bus interconnects the yoke and mirrors of each pixel to a bond pad at the chip perimeter. The DMD mirrors are 16 µm² and made of aluminum for maximum reflectivity. They are arrayed on 17 µm centres to form a matrix having a high fill factor (~90%). The high fill factor produces high efficiency for light use at the pixel level and a seamless (pixilation-free) projected image

Electrostatic fields are developed between the mirror and its address electrode and the yoke and its address electrode, creating an efficient electrostatic torque. This torque works against the restoring torque of the hinges to produce mirror and yoke rotation in the positive or negative direction. The mirror and yoke rotate until the yoke comes to rest (or lands) against mechanical stops that are at the same potential as the yoke. Because geometry determines the rotation angle, as opposed to a balance of electrostatic torques employed in earlier analogy devices, the rotation angle is precisely determined.

The fabrication of the DMD superstructure begins with a completed CMOS memory circuit. A thick oxide is deposited over metal-2 of the CMOS and then planarized using a chemical mechanical polish (CMP) technique. The CMP step provides a completely flat substrate for DMD superstructure fabrication,

ensuring that the projector's brightness uniformity and contrast ratio are not degraded.

Through the use of six photomask layers, the superstructure is formed with layers of aluminum for the address electrode (metal-3), hinge, yoke and mirror layers and hardened photo-resist for the sacrificial layers (spacer-1 and spacer-2) that form the two air gaps. The aluminum is sputter-deposited and plasma-etched using plasma-deposited SiO2 as the etch mask. Later in the packaging flow, the sacrificial layers are plasma-ashed to form the air gaps. The packaging flow begins with the wafers partially sawed along the chip scribe lines to a depth that will allow the chips to be easily broken apart later.

The partially sawed and cleaned wafers then proceed to a plasma etcher that is used to selectively strip the organic sacrificial layers from under the DMD mirror, yoke, and hinges.

Following this process, a thin lubrication layer is deposited to prevent the landing tips of the yoke from adhering to the landing pads during operation. Before separating the chips from one another, each chip is tested for full electrical and optical functionality by a high-speed automated wafer tester.

D. Electronic Operation in DMD Cell

The DMD pixel is inherently digital because of the way it is electronically driven. It is operated in an electrostatically bistable mode by the application of a bias voltage to the mirror to minimize the address voltage requirements. Thus, large rotation angles can be achieved with a conventional 5-volt CMOS address circuit. The organization of the DMD chip is shown in fig. Underlying each DMD mirror and mechanical superstructure cell is a six-transistor SRAM. Multiple data inputs and demultiplexers (1:16) are provided to match the frequency capability of the on-chip CMOS with the required video data rates. The pulse width modulation scheme for the DMD requires that the video field time be divided into binary time intervals or bit times. During each bit time, while the mirrors of the array are modulating light, the underlying memory array is refreshed or updated for the next bit time. Once the memory array has been updated, all the mirrors in the array are released simultaneously and allowed to move to their new address states.

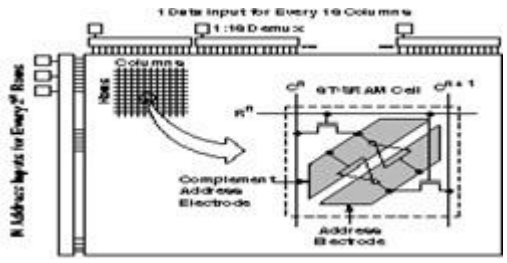

Figure 6. Organisation of the DMD chip

This simultaneous update of all mirrors, when coupled with the PWM bit-splitting algorithm, produces an inherently low-flicker display. Flicker is the visual artifact that can be produced in CRTs as a result of brightness decay with time of the phosphor. Because CRTs are refreshed in an interlaced scan-line format, there is both a line-to-line temporal phase shift in brightness as well as an overall decay in brightness. DLP-based displays have inherently low flicker because all pixels are updated at the same time (there is no line-to-line temporal phase shift) and because the PWM bit-splitting algorithm produces short-duration light pulses that are uniformly distributed throughout the video field time (no temporal decay in brightness).

VI. DIFFERENT CHIP MODELS OF DLP

Like digital video camcorders, DLP devices come in either one or three chip models. One chip DLP systems use a projection lamp to pass white light through a colour wheel that sends red-green-blue colours to the DMD chip in a sequential order to create an image on-screen. Only one DMD chip is used to process the primary red, green and blue colours.

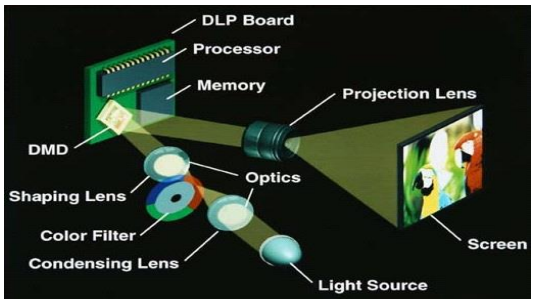

Figure 7. Single Chip Model

In three chip DLP systems use a projection lamp to send white light through a prism, which creates separate red, green and blue light beams. Each beam is send to their respective red, green and blue DMD chip to process the image for display on-screen. One chip models are said to produce a display of over 16-million colours. Three chip models can produce a display of over 35-trillion colours.

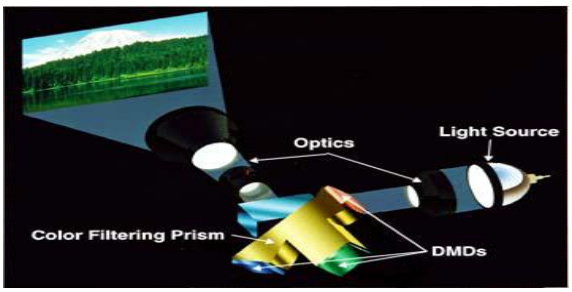

Figure 8. Three Chip Model

VII. LCD PROJECTORS (LIQUID CRYSTAL DISPLAY)

Liquid crystal Display Projectors operate by shining light through transparent LCD cells. Most LCD projectors as advanced polysillicon LCDs, which use three separate colour panels (red, green, and blue) to produce the desired colour. LCDs have excellent colour saturation, usually have adjustable brightness and contrast, are typically brighter than DLPs at the same luminance output, and have a broader range of connectivity.

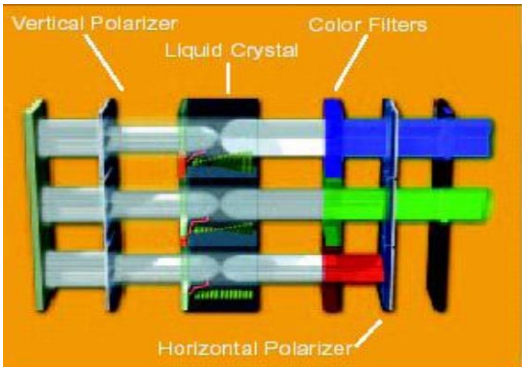

Figure 9. Three closely spaced red, green, and blue LCD sub-pixels

LCDs are polarization-dependent, so one of the polarized light components in not used. This means that 50% of the lamp light never even gets to the LCD because it is filtered out by a polarizer. Other light is blocked by the transistors, gate and source lines in the LCD cell. In addition to light losses, the liquid crystal material itself absorbs a portion of the light. The result in that only a small amount of the incident light is transmitted through the LCD panel and on to the screen.

A. Amorphous Silicon LCD

Am-Si LCDs are built by depositing transistors on a large glass substrate. A transistor is located in the corner of each pixel while a thin conductive grid connects to each pixel on the panel. Pixels are made up of three individually controlled sub-pixel strips (red, green, and blue) to create a pixel capable of producing many colour combinations. Am-Si panels are used to create single-panel projectors, but these projectors suffer from poor image quality due to the side-by-side sub-pixel colour scheme.

B. Polycrystalline Silicon LCD

Polycrystalline silicon LCD, more commonly referred to as poly-Si, is a very popular LCD technology for projection display. These LCDs are fabricated at high temperatures on quartz substrates. Poly-Si LCD panels are much smaller than am-Si panels. They have smaller transistors and greater fill factors as well, but to date, poly-Si panels are monochromatic (meaning they don't have the colour stripping found in am-Si panels). Colour is created in poly-Si projectors by using three separate LCD panels, beam-splitting mirrors, and a prism system. White light is split into red, green, and blue components. Each component of light is directed to its own LCD panel, where the light modulation occurs. The modulated light is then recombined by a prism so that the pixels from each panel are overlaid on each other to produce a colour image. The challenge for these three-panel poly-Si projectors is the precision alignment that is necessary to make the separate red, green, and blue image planes converge to produce a uniform, aligned picture.

C. Cathode Ray Tube Technology

Cathode ray tube technology is used in nearly all of today's computer monitors and televisions. Electron beams are scanned back and forth and directed at a phosphor-emitting surface. When electrons hit this surface, light is emitted. By scanning the beams at rates faster than the eye can detect, a full image can be created.

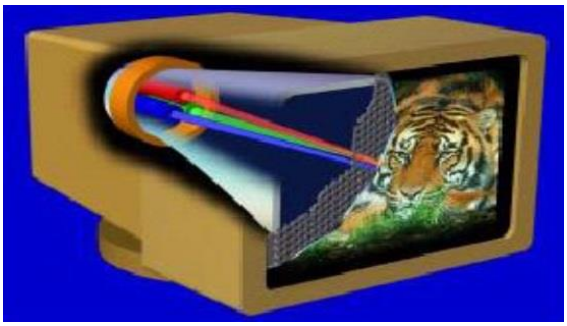

Figure 10. The CRT Monitor

The problem with CRTs is that they are not digital but analogy displays. CRT technology probably will be replaced in the future by emerging LCD and DLP technology. In addition to being based on an old, analogy technology, CRTs also lack the brightness necessary for many larger screen applications. Further brightness limitations arise when CRT projection systems attempt to drive higher resolution video signals. The CRT's brightness decreases as resolution increases, limiting the CRT's potential as an optimal solution for HDTV. Because a CRT display system usually relies on three electron guns (Figure 10), one for each primary colour (red, green, blue), it also requires constant alignment and tweaking for optimum picture quality.

VIII. LCD VS DLP PROJECTOR

The gap between the mirrors in a DMD pixel (left) is smaller than the gap in an LCD display (right), resulting in a sharper display.

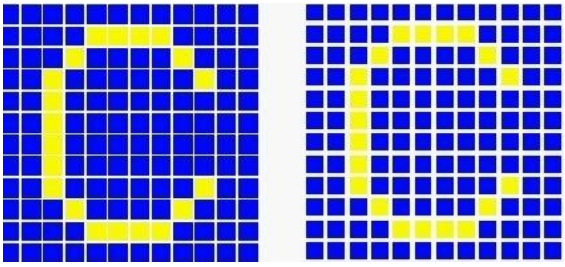

Figure 11. DMD Vs LCD

LCD projector tends to produce more saturated colour and sharper images. Depending on the resolution and size images may become pixilated. DLP Projectors typically offer deeper blacks and higher contrast. They may not project very detailed images well.

LCD Projectors operate by shining light through transparent LCD cells. Most LCD projectors use advanced polysillicon LCDs, which use three separate colour panels (red, green, and blue) to produce the desired colour.

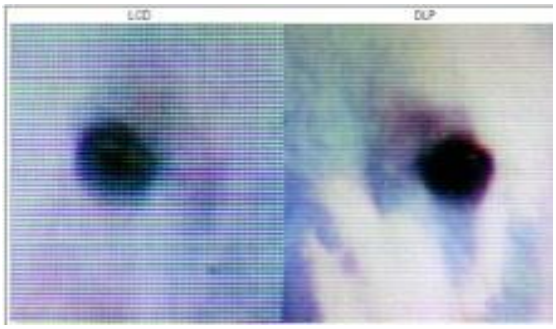

Figure 12. Actual close-up photographs of both (a), an LCD-projected image, and (b), a DLP-projected image.

Projected images are produced by the combination of light shining through the LCD cells. LCDs have excellent colour saturation, usually have adjustable brightness and contrast, are typically brighter than DLPs at the same lumen output and have broader range of connectivity. Brighter, sharper, versatile, more reliable, consistent picture quality are the advantages of DLP.

IX. DMD TELEVISION

DLP (Digital Light Processing Television) is a technology that uses one, or sometimes three, electronic chips called Digital Micromirror Devices, or DMDs, to produce a vivid picture with a high contrast ratio on a high-definition large-screen TV. More than a million micromirrors mounted on the chip — five micromirrors, laid side by side, would fit across a human hair — respond to electric signals to focus the light from a white lamp either on, or away from, the TV screen. A one-chip DMD can produce more than 16 million colours when the light of each micromirror passes through a colour wheel; a three-chip DMD produces several trillion colours.

Figure 11. DLP Television

DLP was developed by Texas Instruments in the late 1980s, and at one time was used in thousands of theaters around the world. At home, DLP could strut its stuff best on large screens — from 43" to around 84"— which made it ideal for home theatres.

DLPTs are frequently referred to as "projection" TVs, with rear-screen and front-screen projection models available. A 1080p DLPT has a resolution of 1,920x1, 080 pixels, and an aspect ratio of 16:9. A 65" screen can be viewed easily from a distance of 5 yards.

DLPTs had several advantages over big-screen plasma and LCD TVs of the time:

1. They were more economical, inch per inch.
2. They were more svelte, being 13" to 15" slim and of a significantly lighter weight.
3. They were more reliable, simply because they had fewer parts vulnerable to failure. Lamp bulbs did need to be changed every 15 years or so.
4. They were immune to burn-in, which was caused by excessive gaming or by an ever-present logo in the corner of the screen.

Like its brethren the PC and the digital camera, each successive generation of DLPTs brought about an improved viewing experience by eliminating common causes for complaints. For example, speeding up the colour wheel banished the annoying "rainbows" on the screen, and the too-narrow viewing angle — 30 or 40 degrees, similar to that of a PC screen — was expanded to up to 180 degrees. That said, however, the Society of Motion Picture and Television Engineers did still recommend that viewers be within 30 degrees of the picture "cone" for optimal viewing.DLP televisions were eventually discontinued in 2012.

X. FUTURE USES OF DLP

DLP has a number of potential uses beyond home theatre, television and film projection. DLP image projectors are becoming more common in business environments. Other applications that could incorporate its high-definition image creation are photo finishing, three-dimensional visual displays, holographic storage, microscopes, spectroscopes and medical imaging. A number of these technologies are already in development. Scientists and developers are likely to discover even more uses for DMDs and DLP technology in the future.

XI. ADVANTAGES OF DLP

1. Brighter.

DLP projectors are among the brightest available because DLP technology brings more light from lamp to screen, resulting in more effective presentation even when ambient light is difficult to control.

2. Sharper.

DLP projection's unique technology comes closest to producing the exact mirror image of an incoming video or graphic signal, resulting in projection that's seamless at any resolution.

3. Versatile.

DLP technology allows projectors to be small and light, often weighing as little as 1kg- making them versatile enough for use in conference rooms, living rooms and classrooms.

4. More reliable.

Display system using DLP technologies are able to recreate their incoming source material with each projection experience that will not fade over time.

5. *Consistent picture quality.*

A data projector based on DLP technology delivers knockout picture quality again and again because, being all-digital, recreates its image source every time of use. Unlike competing analogue technologies such as LCD, the semiconductor that makes DLP projection possible is virtually immune to heat, humidity vibration and other factors.

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Training and Empowerment of Rural Women in Kerala: The Level of Awareness Created in Terms of Knowledge, Attitude and Practice

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***Abstract*-** The study focuses on training and empowerment of rural women in Kerala. The aim of the study is to analyse the level of awareness created among rural women about women's rights and women's issues. The level of awareness is measured in terms of knowledge, attitude and practice. A training package for the empowerment of rural women known as 'Training for Women's Groups in Rural Areas' was selected for the study. The training was implemented for the members of Mahila Samajams (women's organisations) and imparted through ten voluntary organisations working in ten districts of Kerala. The long term objective of the training was the empowerment and self reliance of women. A descriptive study was conducted selecting 175 respondents by simple random sampling method from the universe of 350 women who had attended the women's training programme. Similarly, another 175 respondents who had not attended the training programme were selected randomly for comparative purpose. Thus a total number of 350 respondents were included in the study. Personal interview with the help of interview schedule was the main tool used for collecting information from both the categories of respondents. Result of the study highlights that trained women have a higher level of awareness, more positive attitude and greater will be their responding capacity on women's rights and issues.

***Index Terms*-** Attitude, Empowerment, Knowledge, Practice

I. INTRODUCTION

Women in the Indian society have a degraded status because of the prevalence of multiple factors such as illiteracy, exploitation, unemployment, female infanticide, child marriage, dowry, prostitution, rape, widowhood, wife beating and purdah system. It must be remembered that women's status is not a just a matter of cultural and social history of traditions but it is basically rooted in the political and economic structure of our society which needs to be changed. Women are denied the right to own or inherit property and they become dependent on men, which renders them vulnerable to exploitation. This unfortunate state of affairs is also seen in the state of Kerala, which claims a higher literacy rate and a better health care system as compared to other states. Also the sex ratio of 1084 females for every 1000 males is favourable to women in Kerala when compared to other states; unfortunately the same is not the case with their status in real life. In order to pave the way for a healthier and more progressive future of the nation it is imperative to help women rise from their shackles and to empower them so that they may be able to contribute to society constructively and significantly.

Empowerment of Women

Women's empowerment can be viewed as a continuum of several interrelated and mutually reinforcing components. Empowerment is an active, multidimensional process, which enables women to realize their full identity and potential in all spheres of life. Women's empowerment allows women to be appreciated and acknowledged for who they are and what they do. It is not particularly the ideology of feminism that empowers women, but rather their capacities to face bravely the individual and social facts of their actual situations.

Most poor women have never been allowed to think for themselves or to make their own choices except in unusual circumstances when a male decision maker has been absent of has abdicated his role. Women must be convinced of their innate right to equality, dignity and justice. An empowerment process is one which tackles the condition and position of women, a process which questions about the power structures and gender subordination within these are continually revised and explored. This empowerment process however may most effectively be instigated by means of implementing appropriate training programmes for the selected section of women.

Training and Empowerment of Rural Women

The purpose of training is to achieve a change in the behaviour of those being trained. The general overall objective of the training for rural women is to equip them with the basic knowledge, attitude and skills to play effective roles in promoting the process of development. Training of women functionaries in rural development has become an important issue with special concern for women in development. While the basic concepts of training viz. transfer of knowledge, skill, change of attitudes would remain the same for any training. The identification of the training needs of women and monitoring and evaluating such training would require greater attention. Training has to bring about definite changes in the trainees, like promoting achievement, particularly in the areas of knowledge, skills, attitudes and practices. In training the focus is on learning by an individual of the new ways of doing things. Training structural and organised efforts through which an atmosphere of learning, sharing and synthesizing of information, knowledge and skills are transmitted to the trainees with the help of trainers.

Training can be used as an agent of basic change in the status of women. Training brings about a change in the self image of women, awareness of their inner strength, helps them in making valuable contributions to society and enables them to take on new roles, and to develop the use of questioning and enhances their decision making skills. Training for empowerment places great stress in the creation of an atmosphere of learning. The basic rationale for training women leaders is to promote leadership among the members of Mahila Samajams (women's organisations) and the womenfolk in the village. Training helps women to empower women's organisations to act as catalysts at the local level and as pressure groups with other agencies securing social and economic justice for women. It helps them to plan out their objectives and action programmes and also to identify the areas in which they need to bring a change. Training has become a need for women since they have to enhance their self esteem learn new behaviour for managing the situations and develop leadership and learn building skills. Thus training becomes the most vibrant component of human resource development programmes.

II. RESEARCH ELABORATIONS

Training for Women's Groups in Rural Areas

A training package for the empowerment of rural women known as 'Training for Women's Groups in Rural Areas' was selected for the study. The training was implemented by a Dutch Charitable organisation known as CEBEMO for the members of Mahila Samajams (women's organisations) and imparted through ten voluntary organisations working in ten districts of Kerala. The long term objective of the training was the empowerment and self reliance of women. The specific objectives of the training were:

1. To improve the level of awareness of women on women's rights and women's issues.
2. To change the socio economic status of women in family and society.
3. To strengthen the functioning of Mahila Samajams.

The training was given for forty days in different phases over a period of two years. The methodology for the training was lectures, group discussions, debates, workshops, role plays, audio-visual programmes and paper presentations. As part of the training the participants were given individual assignments like preparation of essays on social problems and women issues, home visits, conducting awareness programmes in the local areas, conducting social survey of the villages and making a study of the people's organisations.

Objective of the Study

The aim of the study is to analyse the level of awareness created among rural women about women's rights and women's issues.

The level of awareness is measured in terms of knowledge, attitude and practice.

The hypothesis formulated to study this objective is that trained women will have a higher level of awareness, more positive attitude and greater will be their responding capacity towards women's rights and issues.

Universe

The universe of the study consists of 350 Mahila Samajam (women's organisation) members who have attended the women's training programmes known as "Training for Women's Groups in Rural Areas" implemented through ten voluntary organisations working in ten districts of Kerala.

Sampling

One hundred and seventy five respondents were selected by simple random sampling method from the universe of 350 women who had attended the women's training programme. Similarly, another 175 respondents who had not attended the training programme were selected randomly for comparative purpose. They were selected from the same Mahila Samajams from which some women attended the training programmes and having similar socio economic profile. Thus a total number of 350 respondents were included in the study.

Design

The design of the study was descriptive and diagnostic in nature. While the descriptive design helps to study in depth the characteristics of individuals, group and situation, the diagnostic design helps to explain the association between variables related to the study.

Sources of Data

The source consisted of primary and secondary data. The primary data source was the respondents. The secondary data for the study were books and journals, reports and records related to the topic.

Tool of Data Collection

Personal interview with the help of interview schedule was the main tool used for collecting information from both the categories of respondents. The same interviews schedule was administered for both the categories of respondents who have attended the training and who have not. Methods like observations, focused group discussions and informal discussions were also used for eliciting data. The interview schedule was very helpful in collecting information directly from the respondents and besides it was the ideal tool since the respondents were not highly educated.

III. MAJOR FINDINGS

The study had been conducted with the purpose of gaining insight into the empowerment of women through training programme.

Among the 350 respondents, 121 (34.6%) belonged to the age group of 36-40 years followed by 87 (24.9%) in the age group of 31-35 years. In the group wise distribution of both categories of respondents i.e. those who attended training and those who did not attend the training programme, the highest percentage belonged to the age group of 36-40 years.

With respect to education it was found that among the respondents, 31.7 per cent are educated up to the high school level and 27.1 per cent of the respondents have completed their SSLC. The highest level of education among respondents in both categories is Pre-degree level. However 4.3 per cent of the respondents can only read and write. This further indicates that all the respondents are literate.

The study indicates that 320 respondents hail from nuclear families and 30 respondents live in joint families. Among the 350 respondents, it is seen that 333 (95.1%) are married and the remaining 17 (4.9%) are widowed.

The occupational profile of the respondents shows that the highest percentage (36.8%) is engaged in domestic work and 28 per cent in wage labour. A few respondents (3.4%) have salaried jobs and others are engaged either in cottage industries or in cultivation.

It was also seen that the majority of the respondents in both categories hold some positions in the Mahila Samajam.

Knowledge

The researcher made an attempt to study the level of knowledge of women regarding different Acts relating to women's rights and various welfare measures for women and children. It was found that the respondents who attended the training have higher mean score (mean=77.18, SD=6.50) when compared to the respondents who did not attend the training (mean=49.05, SD=6.85). The observed difference is statistically significant since the 't' value (39.41) is significant at 0.05 level.

For instance, the knowledge level of both categories of respondents on Hindu Succession Act varies. Significantly among the attendees of training, half (49.7%) of the respondents have good knowledge followed by 48.6 per cent have very good knowledge. While for non attendees more than half (54.9%) of the respondents have only a little knowledge and 41.7 per cent have very little knowledge whereas only 3.4 per cent have a good knowledge on the topic.

Hence it can be stated that the respondents who attended the training have more knowledge on women's rights and women's issues. It may be pointed out that one of the main reasons for the prevalence of dowry system, disparity in wages, inequality in the distribution of land, sexual abuse, existence of child marriage is due to lack of adequate knowledge on women's rights.

The two categories of respondents do differ statistically with regards to knowledge of Acts, source of Acts, government welfare programmes and problem solving. Hence it can be pointed out that training is a powerful tool and key to greater knowledge for the women.

Attitude

It is found that there is a significant difference in the attitude of women who attended training and who did not attended training regarding women's status and education. The mean score for the attendees of training is 37.01, SD= 2.57 and for non attendees 27.83, SD =2.78. The observed difference is statistically significant, as the't' value (32.03) is significant at 0.05 level. Thus it shows that it is the training which makes difference in the attitude level of women.

This indicates that training definitely accounts for a better attitude and it may be stated that the women who have undergone training have a better understanding of the concept of equality between men and women. Only if one has a positive attitude can this concept and equality be fully understood and accepted. In order to attain this social change is required and for this women need to come into the main streams of society. It must be remembered that in rural areas women are not very receptive to change as they are bound by age old customs and traditions. Therefore it is only through constant education and training can new perception and changes be brought about.

Practice

From the analysis it is found that the respondents who attended training have a greater questioning and responding capacity than those who did not. The mean score for the attendees is 18.83, SD=3.94 and for non attendees is 13.62, SD =3.32. The difference is statistically significant as the't' value (13.36) is significant at 0.05 level. Training helps women to have more responding and questioning capacity.

It is observed that the respondents who attended the training have a better capacity in questioning and responding to social issues especially those pertaining to women. They take initiative in protesting against issues like eve teasing, dowry harassment, rape, wife beating and disparity at both places of work and worship. The issues are brought before government authorities by way of oral and written complaints, public meetings, rally etc. The training has helped them to be more aware of the seriousness of problems especially those concerning women. In the case of those who have not attended the training it was observed that there is a lack of initiative in responding to issues which affect them.

It may be concluded that the training programme prompts women to think for themselves and to consider the importance of issues and to execute appropriate action in this direction. Thus training helps in enhancing the bargaining power and reacting mentality of women giving them a stronger hold in society.

The analysis of the study reveals that the hypothesis is proved correct i.e. trained women have a higher level of awareness, more positive attitude and greater will be their responding capacity on women's rights and women's issues.

IV. CONCLUSION

The process of women's empowerment begins in the mind, by changing women's consciousness. The process of empowerment involves not just an improvement in physical and social conditions, but also ensuring equal participation in decision making process, control over resources and mechanisms for sustaining these gains. The findings of the study show that there were remarkable changes in women in the level of awareness as a result of the training programme. It is strongly suggested that other voluntary organisations and nongovernmental organisations could implement similar training programmes for the empowerment of women. It can be concluded that training helps women to increase their level of awareness regarding women's rights and women's issues.

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Autonomic Communication - A New Wave

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Abstract- Next generation networks will certainly face requesting access from different parts of the network. The heterogeneity of communication and application software's changing situations in the environment, from the users, the operators, the business requirements as well as the technologies. Users will be more and more mobile, protocols, etc. will increase and render the network more complex to manage. Autonomic networking aims to design networks that are capable to self-manage, while optimizing their configurations and interactions to the changing needs of the users and the environment. The present paper deals with overview of autonomic communication and its role in digital global world.

Index Terms- Adaptability, autonomic communication, configurable, technology, wireless Communication.

I. Introduction

Autonomic communication studies the individual network element as it is affected by and affects elements and the often numerous groups to which it belongs as well as network in general. Autonomic is a way, being able to self-configure, self-monitor, self-adapt, and self-heal. Autonomic communication studies the individual network element as it is affected by and affects elements and the often numerous groups to which it belongs as well as network in general. The goals are to understand how desired element's behaviors are learned, influenced or changed, and how, in turn, these affect other elements, groups and network [1].

The Autonomic Communication (AC) is to allow self-adaptable, self-configurable, technology independent, robust, secure, scalable, and easily deployed services and infrastructures. Autonomic Communication is a paradigm in which the applications and the services need not ported onto a pre-configured network, but where the network itself grows out of the services that end users desire.

II. Advantages of Autonomic Communication

The following are the main benefits of this advanced technology:

1. Reduce network complexity
2. Cope with proliferation of WLANs and numerous ad hoc, peer to peer networking paradigms which may occupy same frequency space.
3. Share the cost of managing the networks (as IBM's Autonomic Computing)[1].
4. Support mobile, personalized Services
5.Improve efficiencies through context awareness, etc.
6. End-user is becoming a resource-powerful engine.
7. Contributes its resources to the network
8.The autonomic, powerful end-user is becoming the network.
9.The network will be made up of zillions of autonomic nodes.

III. Issues In Autonomic Communication

The increasing density of the global network operators, developers and users both dramatic advantages and significant challenges:

3.1 Industry

The need to maintain diverse and complex networks is often a significant (and increasing) cost of doing business. A more autonomic infrastructure would reduce these costs and facilitate new opportunities, but only if made sufficiently flexible, robust and secure for use across the spectrum of corporate communications.

3.2 Operators

Increasing interconnectivity potentially allows improved robustness capabilities and closer integration with the and bandwidth, but also increases the complexity of management and the fragility of protocols in coping with a highly dynamic and largely scale free environment composed of diverse networks and technologies. Fine-grained mobility and roaming require that the relationships between operators, and between operators and users, be extensively re-thought.

3.3 Developers

Mobile and pervasive networks allow applications and services to extend into the environment, both providing and benefiting from sensing personal and social goals of users, but at the cost of massively increased programming and configuration complexity [4].

3.4 Users

Mobility and ubiquity tilt the balance of communications systems in the users' direction, placing individually- and socially-focused adaptations at the core of the systems architecture, but with the danger that the increased potential for surveillance and complexity will erode the privacy of individual and further disenfranchise entire social groups.

IV. Vision of Autonomic Communication

1. It is to allow self-adaptable, self- configurable, technology independent, robust, secure, scalable, and easily deployed services and infrastructures.

2. ACs is a paradigm in which the applications and the services need not ported onto a pre- configured network, but

where the network itself grows out of the services that end users desire.

The following desirable properties are envisioned

1. Zero- effort deployment
2. Autonomously Controlled Network Entities – self-organized, self- managed, self-configured, self- healed, self-learning system [2].
3. Morphing capabilities - changing (add, delete, modify) functionality.

V. Features of Autonomic Communication

The main features of autonomic communication are the ability to adapt to an evolving situation where new services can become available.

5.1 Network Architectures

Network Architectures that are heterogeneous, resilient, evolvable, distributed and highly dynamic. Multi- mission/ reprogrammable functionality and communications Services [4]. As Network Architectures consists of traditional communication and coordination models which are heterogeneous in nature means no autonomic adaptability.

5.2 Self- aware Communications Protocols

Self-aware Communications Protocols that enable self-adaptive anytime/ anywhere operations in a heterogeneous mix of networks.

5.3 Security and Protection

Since autonomic nodes could self program their networking behavior out of cooperative mechanisms, this calls for novel techniques for trusted software.

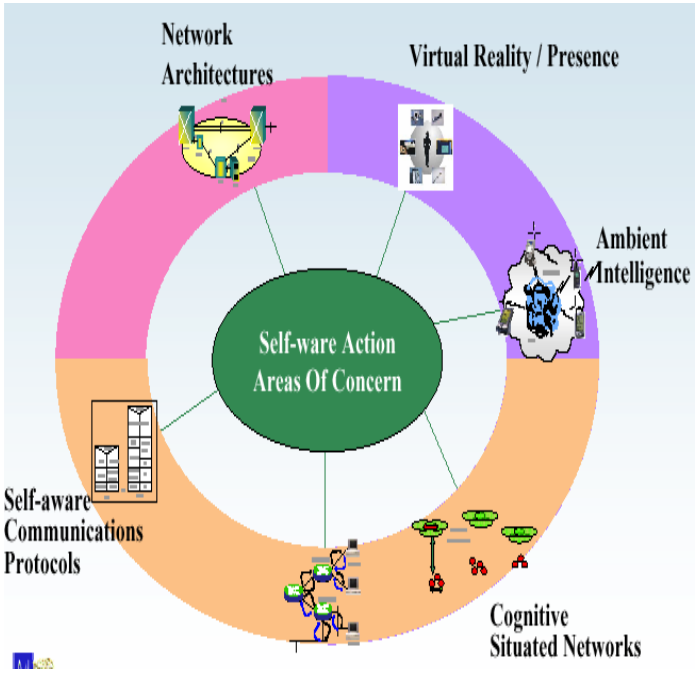

Fig 1: Autonomic communication system self-ware areas

5.4 Cognitive Situated Networks

Cognitive Situated Networks no longer act as a means to simply propagate information from one machine to the other but become a living partner enabling "on- the- fly" interactions with the environment. Nano-sensor technology to gather, filter, aggregate, and map information that live in multiple and dynamic contexts.

5.5 Virtual Reality & Virtual Presence

Virtual Reality & Virtual Presence natural interaction by Video tracking, Eye tracking, Gesture and Emotion Recognition.

5.6 Ambient Intelligence

It is an emerging interface paradigm in which the computer intelligence is embedded in a digital environment that is aware of the presence of the users and is sensitive to their needs, habits, gestures and emotions. In this interaction through all senses - devices- to- device, devices- to- person, devices- to- environment and devices- to- cyber. Figure 1 shows different components of this technology.

VI. Functionalities Of An Autonomic System

In order to grant an autonomic behavior to autonomic system, the following functionalities will be foreseen and enabled both at the autonomic node or component as a single point of decision and the whole system.

6.1 Self-locating

With this feature the autonomic node establishes, and dynamically updates, a reference system to identify neighbor nodes and locate the resources required by its coordination scheme. The reference model will help in the behavior correlation process.

6.2 Self-configuring

With the ability to dynamically configure itself on the fly, an information pervaded environment can adapt immediately and with minimal human intervention to the deployment of new components or changes in the information-pervaded environment.

6.3 Self-healing

Through self-healing, systems state can be evaluated and corrective actions can be initiated without disrupting system operation. The corrective actions may lead the system/subsystem to alter its own state and/or influence changes in other elements of the environment. The information pervaded environment as a whole becomes more resilient as changes are made to reduce or help to eliminate the impact of failing components.

6.4 Self-optimizing

This feature refers to the ability of the information-pervaded environment to efficiently maximize resource allocation and utilization to meet end-users needs with a minimal human intervention. In the near term, self-optimization primarily addresses the complexity of managing system performance.

6.5 Self-protecting

The goal of self-protecting environments is to provide the right information to the right users at the right time through actions that grant access based on the users role and predefined privileges. A self protecting information-pervaded environment can detect hostile or intrusive behavior as it occurs and take autonomous actions to make itself less vulnerable to unauthorized access and use, viruses, denial-of-service attacks and general failures.

6.6 Self and context-aware

These features refer to perception and cognitive reaction to an event or more generically to a condition, relevant to the same intelligent node or, respectively to the environment. Context-awareness is a foundation for the rest of the operational features: self-configuring, self-healing, self-optimizing, and self-protecting.

VII. Cognitive Radio in Autonomic domain

Cognitive architectures are characterized by certain properties: the cognitive behavior should not be implemented partly but it should concern the complete system implementation of various aspects of the cognitive behavior as well as the complete system (Holistic view of the cognitive system). The system aims to learn and adapt its future reactions based on statistics about the previous executions. DARPA proposes architecture for cognitive nodes that introduce 3 processes related to cognitive behaviors: reactive, deliberative and reflective (meta management) reasoning process.

The "reactive" process provides the cognitive network with capabilities that allows reacting automatically to some events or perception of the environment. However, that sense of "reactive" excludes any possibility to take into account future possibilities, hypotheses about what might have been the case, or formulate hypotheses about what exists in some part of the system that is not currently being perceived. Deliberative process permits to leverage this limitation providing the system with the ability to represent and reason about, and to compare and evaluate, possible situations that do not exist, or exist but are not known , either because they are future possibilities, or because they are remote or hypothetical possibilities or because they might have occurred in the past [5]. Finally, the reflective process enables the cognitive elements to monitor and control their own progress to adapt their future behavior in response to previous performance. Statistics, or executions history, are an important parameter in the decision-making process.The architecture aims to reproduce the behavior of the modeled system (i.e. human nervous system) at different level of time (short term, long term reactions) to exhibit a robust behavior to errors and unknown/unexpected events.

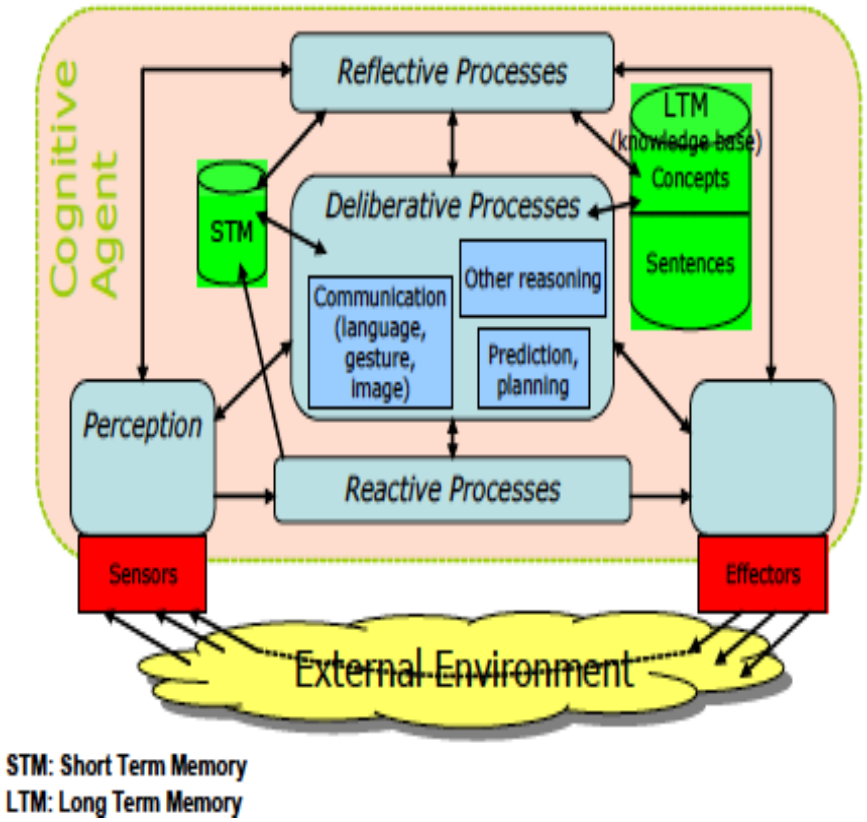

Fig 2: DARPA Cognitive Architecture

VIII. Informed Decision in Wireless Ecosystem

An extended model for autonomic computing and autonomic communications is proposed and depicted in figure 3.

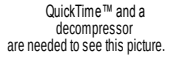

Fig 3: Model of Autonomic Computing and Autonomic communication

The model introduces some new features. The first is the functionality for sensing, assessing risks and determining behavior [3]. The second issue is the computation of fitness functions.

The following is a list of main concepts that have to be developed for the introduction of cognitive wireless networks.

1. Etiquettes: Means of interaction with environments.
2. Cross-layer abstractions: Autonomic computing and communication are highly related to cross-layer optimizations. Decisions that optimize a certain layer, directly or indirectly, can optimize the performance of other layers.
3. XDMA: Development of context Division MultiAceess Concept.

VIII. Conclusion

Continuous developments of mobile technologies and their use in everyday life increase our need to be continuously connected to others and to the Internet, anywhere and at any time. However, in mobile, pervasive environments user connectivity is mainly affected by wireless-communications constraints and user mobility. These boundary conditions do not allow us to design communication environments based on unique and fully connected networks or assume a stable path between each pair of users wishing to communicate. The present paper has overviewed state of art of autonomic communication and its impact in digital global world.

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Reduction of Delay Propagation in Parallel Architecture Based on FNT for High Speed Cyclic Convolution

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Abstract- This project deals with a cyclic convolution of high speed parallel architecture based on Fermat number transform (FNT). Cyclic convolution architectures are implemented for operands in diminished-1 representation. The code conversion (CC) method is mandatory to convert normal binary numbers into their diminished-1 representation. In this paper we discuss about the FNT (Fermat Number Transform) and IFNT (Inverse Fermat Number Transform) operations which are performed by CCWA (Code Conversion without Addition) & BOWA (Butterfly Operation without Addition). The convolution which is the point wise multiplication is implemented by modulo 2n+1 partial product multipliers (MPPM) and these output partial products are the inputs to the IFNT and modulo 2n+1 carry propagation additions are avoided in the FNT and the IFNT except in their final stages

The reduction of modulo 2n+1 carry propagation addition reduces the execution delay of the parallel architecture. The proposed one has better throughput performance and involves less hardware complexity than the existing cyclic convolution architecture.

Index Terms- Fermat Number Transform (FNT), Code Conversion Method without

-Addition (CCWA), Butterfly Operation without -Addition (BOWA), Modulo 2n+1 Partial Product Multipliers (MPPM).

I. INTRODUCTION

The cyclic convolution based on FFT is a widely used operation in signal processing, which needs to be performed in a complex domain even if both of the sequences to be performed would be real. Additionally, the dynamic range of the numbers varies widely so that one need to use floating point numbers to avoid scaling and quantization problems. Some architecture for efficient cyclic convolution has been developed to overcome the problems based on Number Theory Transform (NTT). They replace the complex domain with a finite field or a finite residue ring and can be defined by the FFT-like formula. All arithmetic operations are performed modulo *m* and the convolution results are exact without rounding errors. When the modulus in NTT is a Fermat number ($F_t=2^{2^t}+1$, t[th] Fermat; *t* is an integer), the NTT turns into the Fermat Number Transform (FNT). The multiplication in the FNT and its inverse (IFNT) can be converted into bit shifts when the transform kernel is 2 or its integer power. Though the modulus of the FNT has a strict relationship with its maximum transform, the cyclic convolution based on FNT is more attractive than the conventional method in some areas.

1.1 Novel Cyclic Convolution Structure

Cyclic convolution architectures based on FNT are implemented for the operands in the diminished-1 representation. Thus the code conversion (CC) stage which converts the normal binary numbers into their diminished-1 representation is compulsory. Other arithmetic operations described originally by Leibowitz includes modulo 2^n+1 negation, addition ion, subtraction, multiplication operations in the dimis-nished-1 number system. These operations constitute the butterfly operation (BO) which is the most important element in the FNT.

The CC and the BO are both mainly composed of modulo 2^n+1 adders of which the fastest one in the diminished-1 number system is proposed by Vergos s so far. The fast modulo 2^n+1 adder involves the carry-propagation addition computation and is used in the recent FNT implementations.

In this paper, a code conversion method without addition (CCWA) and a butterfly operation method without addition (BOWA) which take full advantage of the carry-save adder are proposed to accomplish the cyclic convolution with the unity root 2 or its integer power.The modulo 2^n+1 partial product multiplier (MPPM) is used to accomplished the point wise multiplication so that the final carry-propagation addition of two partial product in the multiplier is avoided. Thus the execution delay of the architecture is reduced evidently. Model estimations and experiment results show that the proposed architecture is faster than the existing one when the modulus of the FNT is no less than 2^8+1. For wider modulus, the proposed parallel architecture leads to considerably faster hardware implementations than those presented.

Systolic arrays and distributed arithmetic are widely used for hardware implementation of cyclic convolutions. Distributed arithmetic is slow because of its inherent bit-serial nature. Many different systolic array based designs for the implementation of convolution were presented in, which can be used for implementation of cyclic convolutions.

These designs vary in terms of moving and staying parts (inputs, outputs, and weights) and the number of required delay elements. N-length cyclic convolution, all these designs will need multiplications, additions, and clock cycles.

1.2 Area efficient fault tolerant convolution

An efficient fault tolerant number theoretic transform (NTT) implementation of a convolution using the Redundant

Residue Number System (RRNS). The system is based on a recent method by Conway where a Modified Overlap Save (MOS) technique allows use of transform lengths of differing sizes in an NTT implementation of a convolution without fault tolerance. The proposed scheme adds fault tolerance and shows significant area improvements over RRNS methods combined with the traditional OS method.. The MOS system is then described and then the proposed RRNS MOS system is detailed. The final main Section compares the area of the proposed system with that of the conventional RRNS OS system across a range of convolution sizes.

1.3 Proposed RRNS MOS system:

The proposed system uses the MOS method to realize a convolution in the RRNS. The flexibility with respect to transform length provided by the MOS method allows design of a system where a convolution is realized using a combination of Mersenne and generalized Fermat moduli implemented using WSCA and NTTs, respectively. Table 1.1 shows the moduli and transforms lengths used. The outputs are reconstituted using the Chinese Remainder Theorem. A computer program was developed in Matlab to identify the minimum area implementation for the convolution lengths and bit widths shown in Table 1.2 The number of multiplications and additions required for the WSCA is calculated using the method described by Agarwal and Cooley and using the performance figures given in. The areas of the Fermat transforms are calculated on the basis of the number of additions in the NTT and inverse NTT together with the area of the post-multiplication stage. It is assumed that the filter coefficients are pre-computed.

All NTT kernels, v, used are either of the form 2k requiring two additions per butterfly or ±2k ($2^{M/2}$ + 1) which requires three additions per butterfly. In the latter case, even powers of the kernel are powers of 2 and only one stage in an NTT requires odd powers of v, involving three additions per butterfly.

A sequence of length N has log2 N stages; each stage has N/2 butterflies. Thus, the number of adds for an NTT is calculated as the product of the number of adds per butterfly, the number of stages and the number of butterflies per stage. For example, to implement an NTT with modulus the generalized Fermat number 2^{12} +1 and transform length 16, the more complex kernel, $2^4(2^6+ 1)$ is needed. There are four stages and eight butterflies per stage but one stage has three additions per butterfly and three stages have two additions per butterfly. This gives 3 X 1 X 8 + 2 X 3 X 8= 72 adds for this NTT and a total of 144 adds for the convolution together with 16 multiplications.

The gate equivalents (GE) for addition and multiplication modulo 2^n ±1 as given in Table 3 are used and the area is calculated on the basis of GEs per unit output. Following, the area per unit output for a channel is calculated as the total area for the channel 1/(N - Q + 1), where N is the transform length associated with the particular modulus used in that channel and Q is the filter length.

Moduli	Transform lengths	Moduli	Transform lengths
2^4+1	8,16	2^{28}+1	8,16
2^8+1	8,16,32	2^{32}+1	8,16,32,64,128
2^{12}+1	8,16		
2^{16}+1	8,16,32,64	2^5-1	6,10,15,30
2^{20}+1	8,16	2^7-1	7,14,21,28,35, 42
2^{24}+1	8,16,32	2^{13}-1	15,30,35,42

2^n+1	Adder.9n/2log$_2$n +n/2+6	Multiplier 8n^2+ n-2+adder
2^n-1	Adder. 3nlog$_2$n +4n	Multiplier 8n^2-14n+adder

Table 1.1 Moduli transform lengths and kernel types for proposed system

Word length	Sequence length	Reduction (%)
8	15,19,23,27,31	64,60,64,69,80
12	15,19,23,27,31	59,59,64,72,86
16	15,19,23,27,31	46,49,53,63,86
20	15,19,23,27,31	82,100,100,100,100
24	15,19,23,27,31	84,100,100,100,100
28	15,19,23,27,31	100,100,100,100,100
32	15,19,23,27,31	100,100,100,100,100

Table 1.2 Percentage area reduction achieved in MOS RRNS system.

II. Fermat Number Theoretic Transform

The cyclic convolution via the FNT is composed of the FNTs, the point wise multiplications and the IFNT. FNTs of two sequences {*ai*} and {*bi*} will produce two sequences {*Ai*} and {*Bi*}. Modulo 2^n+1multipliers are employed to accomplish the point wise multiplication between {*Ai*} and {*Bi*} and produce the sequence {*Pi*}. The final resulting sequence {*pi*} can be obtained by taking the inverse FNT of the product sequence {*Pi*}.Each element in the {*pi*} is in the diminished-1 representation.

The FNT of a sequence of length N $\{x_i\}$ (i= 0,1, …N- 1) is defined as :

$$X_k = \sum_{i=0}^{N-1} x_i \alpha_N^{(ik)} \bmod(n) F_t (k = 0,1....N-1) \quad \ldots (2.1)$$

where $Ft=2^{2t}+1$, the t^{th} Fermat; N is a power of 2 and α is an Nth root unit (*i.e.* α^N_N mod Ft=1 and α^M_N mod Ft ≠1 ,1≤m <N). The notation < ik > means ik modulo N.

The inverse FNT is given by

$$x_t = \frac{1}{N}\sum_{k=0}^{N-1} x_k \alpha_N^{-(ik)} \bmod F_t (i = 0,1....N-1) \quad \ldots (2.2)$$

Where 1/N is an element in the finite field or ring of integer and satisfies the following condition:

$(N.1/N) \bmod F_t=1$ …..(2.3)

Parameters α, *Ft*, *N* must be chosen carefully and some conditions must be satisfied so that the FNT possesses the cyclic convolution property. In this project, α=2, $Ft=2^{2t}+1$ and $N=2.2^t$ where t is an integer.

Table 1.3 Gate equivalents for implementation of arithmetic modulo $2^n \pm 1$.

III. Important Operations In Cyclic Convolution Based On FNT

Important operations of the cyclic convolution based on FNT with the unity root 2 include the CCWA, the BOWA and the MPPM. The CCWA and the BOWA both consist of novel modulo 2^n+1 4-2 compressors mainly which are composed of the 4-2 compressor introduced by Nagamatsu. The 4-2 compressor, the novel modulo 2^n+1 4-2 compressor and the BOWA are shown in Fig.3.2 (b). In the figure, "*X**" denotes the diminished-1 representation of X, i.e. X* =X-1.

3.1 Code Conversion without Addition

The CC converts normal binary numbers (NBCs) into their diminished-1 representation. It is the first stage in the FNT. Delay and area of CC of a 2n-bit NBC are no less than the ones of two n-bit carry propagation adders.

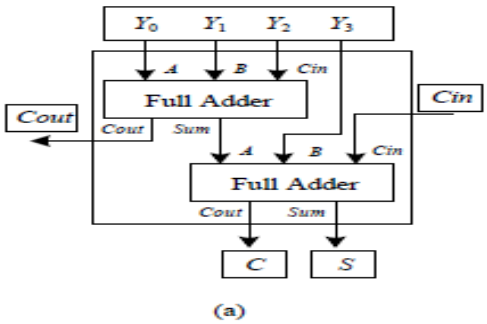


Fig.3.1 Elementary operations of FNT architecture with unity root 2, (a) 4-2 compressor

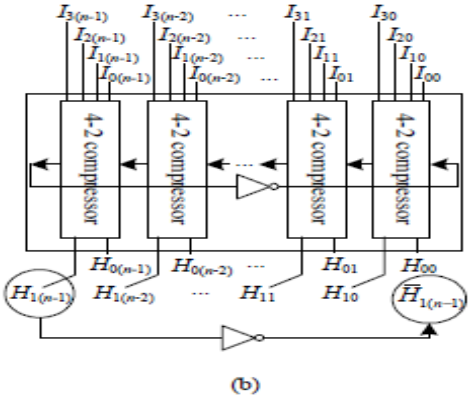


Fig.3.2 (b) modulo 2^n+1 4-2 compressor

To reduce the cost, we propose the CCWA that is performed by the modulo 2^n+1 4-2 compressor. Let A and B represent two operands whose widths are no more than 2n bits. We define two new variables:

$$A = 2^n A_H + A_L$$
$$B = 2^n B_H + B_L \quad (3.1)$$

And

$$M_0 = (2^n - 1) - A_H = \bar{A}H$$
$$M_1 = (2^n - 1) - B_H = BH$$
$$M_2 = (2^n - 1) - B_L = B_L \quad (3.2)$$

If the subsequent operation of CC is modulo 2^n+1 addition, assign A_L, M_0, B_L and M_1 to $I0$, $I1$, $I2$, $I3$ in the modulo 2^n+1 4-2 compressor respectively. $I0$, $I1$, $I2$, $I3$ are defined as follows:

$$I_0 = I_{0(n-1)}\, I_{0(n-2)} \ldots\ldots I_{01} I_{00}$$
$$I_1 = I_{1(n-1)}\, I_{1(n-2)} \ldots\ldots I_{11} I_{10}$$
$$I_2 = I_{2(n-1)}\, I_{2(n-2)} \ldots\ldots I_{21} I_{20}$$
$$I_3 = I_{3(n-1)}\, I_{3(n-2)} \ldots\ldots I_{31} I_{30} \quad (3.3)$$

We obtain the sum vector H_O^* and carry vector H_1^* in the diminished-1 number system. The most significant bit of H_1^* is complemented and connected back to its least significant bit. That is to say

$$H_O^* = H_{0(n-1)}\, H_{0(n-2)} \ldots\ldots H_{01} H_{00}$$
$$H_1^* = H_{1(n-2)} \ldots\ldots H_{11} H_{10}\, H_{1(n-1)} \quad \ldots\ldots (3.4)$$

The result of modulo 2^n+1 addition of A* and B* is equal to the result of modulo 2^n+1 addition of H_O^* and H_1^* in this way, A and B are converted into their equivalent diminished-1 representations H_O^* and H_1^*.

Let $|A^* + B^*|_{2^n+1}$, $|\bar{A}^*|_{2^n+1}$, $|A^* - B^*|_{2^n+1}$, and $|A^* + 2^i|_{2^n+1}$ denote modulo 2^n+1 addition, negation, subtraction and multiplication by the power of 2 respectively which are proposed

by Leibowitz originally. The CCWA for subsequent modulo 2^n+1 addition can be described as follows.

$$\left|A^*+B^*\right|_{2n+1} = \left|AL+M0+BL+M1\right|_{2n+1} = \left|H_O^* + H_1^*\right| \quad \ldots (3.5)$$

If the subsequent operation is modulo 2^n+1 subtraction, we assign A_L, M_0, M_2 and B_H to I_0, I_1, I_2, I_3 respectively. Then H_O^* and H_1^* in the modulo 2^n+1 4-2 compressor constitute the result of the CCWA. The conversion is described as follows:

$$\left|A^*-B^*\right|_{2^n+1} = \left|A-B\right|_{2^n+1} = \left|A+B\right|_{2^n+1} = \left|A_L+M_0+B_L+M_1\right|_{2^n+1} \quad \ldots (3.6)$$
$$= \left|H_O^* + H_1^*\right|_{2^n+1}$$

After CCWA, we obtain the result consisting of two diminished-1 numbers. The result also includes the information of modulo 2^n+1 addition or subtraction in the first stage of previous BO.

3.2 Diminished-One modulo 2^n+1 Adder Design

Modulo arithmetic has been used in digital computing systems for various purposes for many years. In particular, modulo 2^n+1 arithmetic appears to play an important role in many algorithms. A first application field is in Residue Number Systems (RNS) . In an RNS based application, every number X is represented by a sequence of residues {X1, X2 . . . XM.} where Xi=X mod pi. The pies, $1 \leq i \leq M$, comprise the base of the RNS and are pair wise relative prime integers. A two operand RNS operation, suppose}, is defined as (Z1, Z2 . . . ZM.)=(X1, X2 . . . XM.) $\Diamond$ (Y1, Y2. YM) where $Z_i = (Xi \Diamond Yi)$. Mod pi. For most RNS applications $\Diamond$ is addition, subtraction, or multiplication. Since the computation of Zi only depends upon Xi, Yi, and pi, each Z_i is computed in parallel in a separate arithmetic unit, often called channel. Moduli choices of the form$\{2^n-1, 2^n, 2^n+1\}$ have received significant attention because they offer very efficient circuits in the area x time^2 product sense Addition in such systems is performed using three channels that, in fact, are a modulo 2^n-1 (equivalently, one's complement), a modulo 2^n, and a modulo 2^n+1 adder . The addition delay in an RNS application which uses the above moduli is dictated by the modulo 2^n+1 channel. The latter means that, if we can cut down the time required for modulo 2^n+1 addition, we also cut down the addition time in an RNS application.

Modulo 2^n+1 adders are also utilized as the last stage adder of modulo 2^n+1multipliers. Modulo 2^n+1multipliers find applicability in pseudorandom number generation, cryptography, and in the Fermat number transform, which is an effective way to compute convolutions Leibowitz has proposed the diminished-one number system. In the diminished-one number system, each number X is represented by $X^*=X-1$. The representation of 0 is treated in a special way. Since the adoption of this system leads to modulo 2^n+1 adders and multipliers of n bits wide operands, it has been used for many residue number system implementations; Efficient VLSI implementations of modulo 2^n+1 adders for the diminished-one number system have recently been presented.

The adders although fast, are, according to the comparison presented, still slower than the fastest modulo 2^n adders or the fastest modulo 2^n-1 adders.. Therefore, their use in an RNS application would still limit the performance of the system.

In this paper, we derive two new design methodologies for modulo 2^n+1 adders in the diminished-one number system. The first one leads to traditional Carry Look-Ahead (CLA), while the second to parallel-prefix adder architectures. Using implementations in a static CMOS technology, we show that the proposed CLA adder design methodology leads to more area and time efficient implementations than those presented, for small operand widths. For wider operands, the proposed parallel-prefix design methodology leads to considerably faster adder implementations than those presented and as fast as the integer or the modulo 2^n+1 architecture presented.

3.3 Butterfly operation without addition

After the CCWA, we obtain the results of modulo 2^n+1 addition and subtraction in the diminished-1 representation. Each result consists of two diminished-1 values. The subsequent butterfly operation involves four operands. The proposed BOWA involves two modulo 2^n+1 4-2 compressors, a multiplier and some inverters as shown in Fig. 3.3.

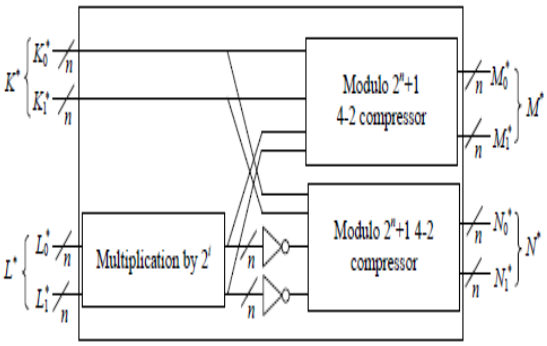


Fig.3.3 Butterfly operation without addition

The multiplication by an integer power of 2 in the diminished-1 number system in the BOWA is trivial and can be performed by left shifting the low-order *n-i* bits of the number by *i* bit positions then inversing and circulating the high order *i* bits into the *i* least significant bit positions. Thus the BOWA can be performed without the carry-propagation chain so as to reduce the delay and the area obviously. *K**, *L**, *M**, *N** are corresponding to two inputs and two outputs of previous BO in the diminished-1 number system respectively and given by

$M^*=|M_0^* + M_1^*|_{\cdot 2^n+1}=|K_0^* + K_1^* + L_0^* X\ 2^t + L_1^* X\ 2^t|_{\cdot 2^n+1}=$
$|K_*^* + L_0^* X\ 2^t|_{\cdot 2^n+1}$

$N^*=|N_0^* + N_1^*|_{\cdot 2^n+1}=|K_0^* + K_1^* - L_0^* X\ 2^t - L_1^* X\ 2^t|_{\cdot 2^n+1}=$
$|K_*^* - L_0^* X\ 2^t|_{\cdot 2^n+1}$
$=|K_*^* + L_0^* X\ 2^t|_{\cdot 2^n+1}$
… (4.1)
Where $K_*^* = |K_0^* + K_1^*|_{\cdot 2^n+1}$, $L_*^* = |L_0^* + L|_{\cdot 2^n+1}$

3.4 Modulo 2^n+1 Partial Product Multiplier

For the modulo 2^n+1 multiplier proposed by Efstathiou, there are n+3 partial products that are derived by simple AND and NAND gates. An FA based Dadda tree that reduces the n+3 partial products into two summands is followed. Then a modulo 2^n+1 adder for diminished-1 operands is employed to accept these two summands and produce the required product.

In the proposed parallel architecture for cyclic convolution based on FNT, the BOWA can accept four operands in the diminished-1 number system. Every point wise multiplication only needs to produce two partial products rather than one product. The operation can be accomplished by taking away the final modulo 2^n+1 adder of two partial products in the multiplier. Thus the final modulo 2^n+1 adder is omitted and the modulo 2^n+1 partial product multiplier is employed to save the delay and the area.

IV. The FNT Architecture

In the previous sections, we have presented the reconfiguration at a rather low level. The Butterfly constitutes a high parameterized function level. The fact to have this parameterized function allows designing a reconfigurable operator who's Butterfly forms the highest level operator.

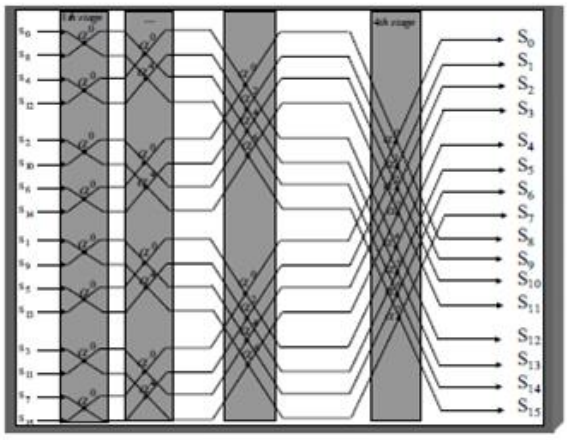


Fig 4.2 the architecture of FNT operator

V. Parallel Architecture For Cyclic Convolution

Based on the CCWA, the BOWA and the MPPM, we design the whole parallel architecture for the cyclic convolution based on FNT as shown in Fig.6.1. It includes the FNTs, the point wise multiplication and the IFNT mainly. FNTs of two input sequences {*ai*} and {*bi*} produce two sequences {*Ai*} and {*Bi*} (*i*=1, 2 …N- 1). Sequences {*Ai*} and {*Bi*} are sent to *N* MPPMs to accomplish the point wise multiplication and produce *N* pairs of partial products. Then the IFNT of the partial products are performed to produce the resulting sequence {*pi*} of the cyclic convolution.

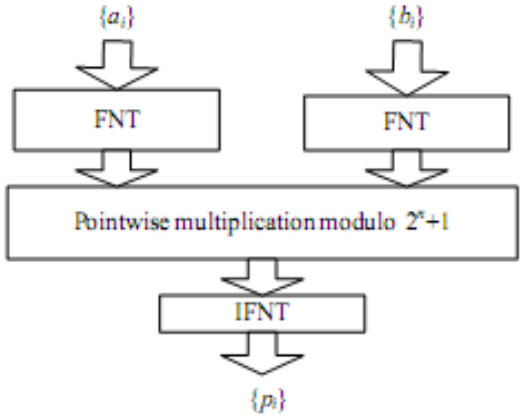
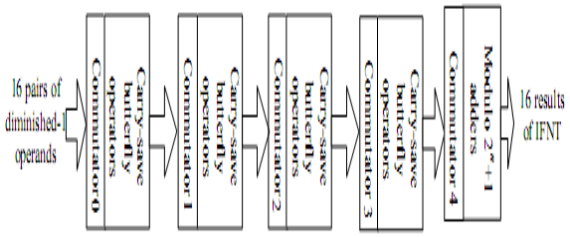


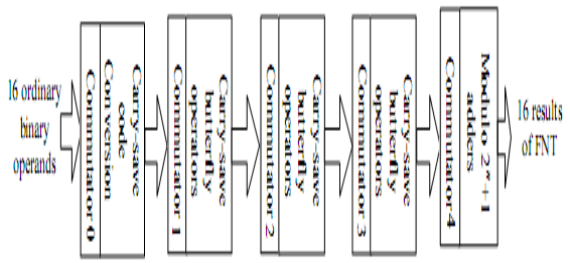
Fig.5.1 Parallel architecture for the cyclic convolution based on FNT

In the architecture, the radix-2 decimation-in-time (DIT) algorithm which is by far the most widely used algorithm is employed to perform the FNT and the IF*N*T.

Illustrative examples of the FNT and the IFNT are shown in Fig. 6.2(a) in the case the transform length is 16 and the modulus is 2^8+1. Commentators in Fig. 6.2 are used to adjust the operand order of every stage of FNT and IFNT according to the radix-2 DIT algorithm.



(a)Parallel FNT structure



(b) Parallel IFNT structure
Fig. 5.2(b) Structures for FNT and IFNT

The efficient FNT structure involves $\log_2 N$+ 1 stages of operations. The original operands are converted into the diminished-1 representation in the CCWA stage, containing the information of modulo 2^n+1 addition or subtraction in the first butterfly operation stage of the previous FNT structure. Then the

results are sent to the next stage of BOWA. After $\log_2 N$-1 stages of BOWAs, the results composed of two diminished-1 operands are obtained. The final stage of FNT consists of modulo 2^n+1 carry-propagation adders which are used to evaluate the final results in the diminished-1 representation. The CCWA stage, the BOWA stage and the modulo 2^n+1 addition stage in the FNT involves $N/2$ couples of code conversions including the information of modulo 2^n+1 addition and subtraction, $N/2$ butterfly operations and $N/2$ couple of modulo 2^n+1 additions respectively.

From the definition of FNT and IFNT in section 2, the only difference between the FNT and the IFNT is the normalization factor $1/N$ and the sign of the phase factor α_N. If ignoring the normalization factor $1/N$, the above formula

(Ft=2^8+1) is the same as that given in the FNT except that all transform coefficients $\alpha_N^{\ ik}$ used for the FNT need to be replaced by $\alpha_N^{-(ik)}$ for the IFNT computation. The proposed FNT structure can be used to complete the IFNT as well with little modification as shown in Fig. 5.2(b). After the IFNT of N-point bit reversed input data, the interim results are multiplied by $1/N$ in the finite field or ring. Then $x[j]$ and $x[j+N/2]$ (j=1,2,…,$N/2$-1) exchange their positions to produce the final results of the IFNT in natural order. Our architecture for the cyclic convolution gives a good speed performance without requiring a complicated control. Furthermore, it is very suitable for implementation of the overlap-save and overlap adds techniques which are used to reduce a long linear convolution to a series of short cyclic convolutions.

VI. Comparison and Results

K	5….6	7…9	10…13	14…19
D(k)	3	4	5	6
K	20…28	29…42	43….63	64…49
D(k)	7	8	8	10

In this section, we compare the proposed parallel architecture for the cyclic convolution against that introduced by Conway. The modulo 2^n+1 addition for the diminished-1 number system is the crucial operation which contains a standard n-bit carry propagation computation such as a parallel-prefix adder with a carry-logic block and a zero indicator of the diminished-1 operand to determine whether to perform subsequent operations. It produces the longest execution delay and requires large area in the previous solution. The proposed CCWA and BOWA overcome the disadvantage of the carry-propagation adder and don't require a zero indicator. Thus our architecture is faster and more efficient than the existing one.

This model assumes that each two-input gate excluding XOR is equivalent to one elementary gate for both area and delay. An XOR gate counts for two gates for both area and delay. Thus, a full adder has an area of seven gates and a delay of four gates. This model does not involve the cost of buffering and routing, but achieve a reasonable accuracy for the purpose of comparison.

The delay and the area estimations of modulo 2^n+1 adder and modulo 2^n+1 multiplier in the cyclic convolution are given in Table 1 as a function of the operand size n. "$D(n+3)$" in Table 6.1 is defined as shown in Table 6.2

Table 6.1 Area and delay estimations for arithmetic modulo 2^n+1 "MA" and "MM" represent modulo 2^n+1 adder and multiplier respectively.

Ft	Area(μm^2)		Delay(ns)	
	This project	[3]	This project	[3]
2^8+1	3.5 x 10^5	3.5 x 10^5	8.9	9.9
2^{16}+1	1.86 x 10^6	2.05 x 10^6	11.6	14.4
2^{32}+1	1.08 x 10^7	1.24 x 10^7	15.1	20.4

Table 6.2 D (k) as a function of k

To obtain more accurate results, we describe the proposed parallel cyclic convolution in verilog for Ft=2^8+1,2^{10}+1,2^{32}+1.The validated Verilog code is synthesized using a 0.13-μm CMOS standard cells library in the worst operating condition by the Synopsys Design Compiler. The units of area and delay are μm^2 and ns respectively. Each design was recursively optimized for speed until the EDA software can't provide a faster design. The results for the fastest derived implementation are listed in Table 6.3.

Table 6.1 and 6.3 indicate that for values of Ft ≥2^8+1 the proposed architecture comprising the CCWA and the BOWA require less delay and area than the previous one. The former results in a 12.6% reduction in area and a 26% reduction in delay respectively compared with the latter in the case Ft is 2^{32}+1 and the transform length is 64. Moreover, our algorithm will be more and more advantageous with the growth of modulus width.

Table 6.3 Area and delay results of cyclic convolution based on FNT

VII. Conclusion and Future Scope

7.1 Conclusion

A novel parallel architecture for the cyclic convolution based on FNT is proposed in the case the principle root of unity is equal to 2 or its integer power. The FNT and the IFNT are accomplished by the CCWA and the BOWA mainly. The point wise multiplication is performed by the modulo 2^n+1 partial product multiplier. Thus there are very little modulo 2^n+1 carry-propagation addition compared to the existing cyclic convolution architecture.

operator	Area		Delay	
	This project	[3]	This project	[3]
MA	14n	9/2nlogn+n/2+6	8	2logn +3
MM	$8n^2$+n-1	9/2nlogn+$8n^2$+n/2+4	4D(n+3)+1	4D(n+3) $2\log_2 n$ +3

A theoretical model was applied to access the efficiency independently of the target technology. VLSI implementations using a 0.13 um standard cell library show the proposed parallel architecture can attain lower area and delay than that of the existing solution when the modulus is no less than 2^8+1.

7.2. Future Scope

This is the High speed parallel architecture for cyclic convolution based on FNT. Here in this architecture we used modulo adders and modulo multipliers to reduce the round off errors and also to reduce the delay and also that CCWA and BOWA blocks are used for further reduction of the delay. Based on the number of operands as the butterfly stages generally increases there is no other alternative for finding the convolution.

The changes can be made either for the blocks like multiplication by 2^n or for modulo adder or for modulo multiplier in order to reduce the size. Also that this can be implemented in low power technology by carefully following the parameters and the circuit elements used for the implementation. So as this is the high speed architecture, we further work on for the size or power consumption and others. VLSI implementations using a 0.13 um standard cell library show the proposed parallel architecture can attain lower area and delay than that of the existing solution when the modulus is no less than 2^8+1.

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Do We Need Patent Protection to Biotechnology Inventions?

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Abstract- The growing research and development activities in the field of biotechnology are a new addition to intellectual property regime. Serious efforts are being made to increase the number of inventions and filling of applications for patents. The inventions, which involve more of living organisms, became controversial issues. The standard tests of patent law are obstacles for grant of patents to the biotech inventions. Non-grant of patents encourages uncontrolled piracy of new inventions and the original owners suffer economic losses because of low returns on their huge investments in the research. TRIP, agreement seeks to enforce patent laws around the world including biotechnological inventions. In India, the existing patent laws were amended to comply with TRIPs agreement. But do we really need patent protection to the genetically modified animals is the question of the day.

Index Terms- Biotechnology, TRIPs, Patent protection, IPR Laws

I. INTRODUCTION

For thousands of years human beings have exploited biological resources for medical, agricultural and other purposes. Modern biotechnology is only the latest fashion by which man is able to use his natural environment to feed, cure and house him. In the last decade the application of modern biotechnology for agricultural, ecological and medical purposes has sparked great hoes for the extent to which man can explore and exploit biological resources for his wellbeing. Biotech inventions are gaining popularity as they help identify the root cause of chronic diseases and suggest remedies. Biotechnology is a new area of science that compounds life sciences of chemical sciences. Genes in plants, providing for certain special features, would be isolated and put together into one plant with the help of biotechnology. Through the process of biotech by manipulation of genes new variety of animals can be invented. These new variety of animals can be used for different purposes such as production of medicines, for experimental for testing of medicines and treating methods.

Earlier patents were not granted to genetically modified organisms and other products derived from living systems. With the advancement of the technology, protection of biotechnological invention also started. But protection to the inventions, which involve more of living organisms, became more complex and controversial. The standard tests of patent law are obstacles for grant of patents to the biotech inventions. Non-grant of patents encourages uncontrolled piracy of new inventions and the original owners suffer economic losses because of low returns on their huge investments in the research. Now TRIPs, agreement seeks to enforce US style patent laws around the world. This agreement covers everything from pharmaceuticals to information technology software and human gene sequences. At this juncture, the author points out impact of extension of patent protection to genetically modified organisms in India.

II. WHAT IS BIOTECHNOLOGY?

The word 'biotechnology' was actually coined early in the 20th century by an agricultural engineer from Hungary, named Karl Earky, who explained it in such a way that the technology which include all such work by which the products are produced from raw materials with the aid of living organisms. Subsequently, over the period, the definition of biotechnology acquired a confusing status due to various interpretations[1].

Classical biotechnology may be defined loosely as the production of usual products by living organisms, and as such it has been with us for a long time. The first official broad definition given by the US Office of Technology Assessment states, "biotechnology includes any technique that uses living organisms(or parts of organisms) to make or modify products, to improve plant or animals or to develop microorganisms for specific use"[2].

Biotechnology can be traced back to various stages of its development. the first generation of biotechnology can be based on the traditional knowledge in various tribes like preparing fermented foods, medical distillates etc. second generation of biotechnology may be considered when the utilisation of micro-organisms started on industrial scale during the Pasteur era which involved mass production of alcohol, fermentation of antibiotics, development of classical vaccines like for cholera, typhoid, yellow fever etc.

The third generation of biotechnology, as distinct from classical fermentation technology, began in 1970s with the two basic techniques of recombinant DNA technology and Hybridoma technology. In the first of these, also referred to as gene splicing or genetic engineering, genetic material from an external source is inserted into a cell in such a way that it causes the production of a desired protein by the cell; in the second,

[1] Dr. K. K. Tripathi, *Biotechnology and IPR Regime: In the Context of India and Developing Countries*, "Biotech Patent Law",1st Ed.(the Icfai University Press,2007) p.,187

[2] Philip W. Grubb, *Patents for Chemical, Pharmaceuticals and Biotechnology*, 4th Ed,(New Delhi: Oxford University Press) 2006. P. 245-246

different types of immense all are fused together to form a hybrid cell line producing monoclonal antibodies.[3]

The fourth generation of biotechnology would see further advances where interdisciplinary techniques like information technology and nano-technology would get involved in further advancement of this discipline, especially utilising the bioinformatics which is the foundation of modern biotechnology. Bioinformatics can be broadly defined as the use of computers to handle biological information.

III. Patents and Biotechnology

The ownership and exploitation of intellectual property rights are the key factors in determining the success of any technological invention introduced in the market that provide the means for technological progress to continue or to be made thereby support the competitiveness of the industry of the country. The regulatory mechanisms in IPR have their own problems in the coming scenario of emerging technologies especially in biotechnology.

Biotechnology is the result of efforts of intellect, the application of human intelligence and knowledge to the biological processes. These human intellectual efforts deserve protection. The new plans, animal varieties, new methods of treatments, new crops producing food articles as such are the inventions of biotechnology. These inventions have to be protected for obtaining the fruits of biotechnology.

Biotechnology has based a whole new industry and patent protection for biotechnological inventions is of immense commercial importance. But patent law and practice have had serious difficulties in keeping up with the rapid scientific progress in this field and issues such as inventions steps, sufficiency of disclosure and permissible breadth of claims have proud troublesome. There has been much litigation of biotech plants and courts have found it difficult in such a rapidly moving field to determine what the general knowledge of the skilled person was at the time invention was made. A procedure to find and clone a specific gene and to express it in a suitable has may have been a breakthrough at the time when it was first done and purely routine work, not many years later. There is also the problem of opposition by special interest groups against anything to do with genetic engineering and particularly against the existence of patents in this area[4].

IV. Patenting of Micro-organisms and Cells

Patents on biotechnological processes date from the early days of the United States. Louis

Pasteur received a patent for a process of fermenting beer. Acetic acid fermentation and other food patents date from the early 1800s, while therapeutic patents in biotechnology were issued as early as 1895. The development of recombinant DNA technology (rDNA), i.e., the controlled joining of DNA from different organisms has resulted in greatly increased understanding of the genetic and molecular basis of life. Following the first successful directed insertion of recombinant DNA into a host micro-organism in 1973, scientific researchers began to recognize the potential for directing the cellular machinery to develop new and improved products and processes in a wide variety of industrial sectors. Many of these products were micro-organisms (microscopic living entities) or cells (the smallest component of life capable of carrying on all essential life processes). With the development of recombinant DNA technology, the potential of patenting the living organism resulting from the technology arose.

Prior to 1980, patents were not granted for such inventions, deeming them to be "products of nature" and not statutory subject matter. Although patent applications were rejected if directed to living organisms per se, patent protection was granted for many compositions containing living things (e.g., sterility test devices containing living microbial spores, food yeast compositions, vaccines containing attenuated bacteria, milky spore insecticides, and various dairy products). In the absence of congressional action, it took a catalytic court decision to clarify the issue of patentability of living subject matter.

The Supreme Court's single foray into biotechnology occurred in 1980 with its ruling in the patent law case *of Diamond v. Chakrabarty*[5]. Chakrabarty had developed a genetically modified bacterium capable of breaking down multiple components of crude oil. Because this property was not possessed by any naturally occurring bacteria, Chakrabarty invention was thought to have significant value for cleaning up oil spills. Chakrabarty's claims to the bacteria were rejected on two grounds:

1. Micro-organisms are "products of nature;" and
2. As living things, micro-organisms are not patentable subject matter.

Following two levels of appeals, the case was heard by the U.S. Supreme Court, which in a 5-4 ruling, held that a live, human-made microorganism is patentable subject matter. The *Chakrabarty* decision provided great economic stimulus to patenting of micro-organisms and cells, which in turn provided stimulus to the growth of the biotechnology industry in the 1980s. [6]

V. Patenting of Animals

The first animal patent was issued in April 1988 to Harvard University for mammals genetically engineered to contain a cancer-causing gene (U.S. 4,736,866). The patented mouse was genetically engineered to be unusually susceptible to cancer, thus facilitating the testing of carcinogens and of cancer therapies. Specifically, the patent covers "a transgenic non-human eukaryotic animal (preferably a rodent such as a mouse) whose germ cells and somatic cells contain an activated onco gene sequence introduced into the animal which increases the probability of the development of neoplasm (particularly malignant tumors) in the animal."

[3] *http://www.iprcommission.org*

[4] C.B.Raju, *Intellectual Property Rights*, 1st Ed., (New Delhi: Serials Publications), 2006.

[5] *447* U.S. 303 (1980).

[6] http://www.fas.org/ota/reports/8924.pdf

The claim in this patent was a transgenic non-human mammal all of whose germ cells and somatic cells contain a recombinant activalid Onco gene sequence introduced into the said mammal or an ancestor of the said animal at an embryonic stage. The rat race continued and at least 16 patents have been awarded on inventions related to transgenic mice as models exhibiting specific pathologies, such as ulcers, photo Parkinson's syndrome, inflammation, sickle cell anaemia, Alzheimer's disease, HIV infection, Cutnieous melanoma, leukaemia, thrombocytopenia etc[7].

VI. Transgenic Animals and Patenting

Most potentially patentable animals are likely to be transgenic animals produced via recombinant DNA techniques or genetic engineering. Transgenic animals are those, whose DNA or hereditary material has been augmented by adding DNA from a source other than parental germplasm, usually from different animals or from humans. Laboratories around the world are conducting research that involves inserting genes from vertebrates (including humans, mammals, or other higher organisms) into bacteria, yeast, insect viruses, or mammalian cells in culture. A variety of techniques, most developed from early bacterial research, can now be used to insert genes from one animal into another. These techniques are known by a number of exotic names: microinjection, cell fusion, electroporation, retroviral transformation, and others[8].

VII. TRIPs and Biotechnology Patenting

Genetic resources have in the past been declared "a common heritage of mankind to be preserved and to be freely available to all, for use for the benefit of present and future generations". This philosophy has done well to the country in general and to the society in particular, in the long run by enabling access to such creations and knowledge to all without discrimination. But in recent times the industrial countries are busy in the protection and privatization of inventions in the area of living objects or substances such as patenting of micro-organisms and animals. Such steps were generally not accepted by the developing countries including India.

However, TRIPs agreement encourages protection of knowledge and it seeks to enforce US style patent laws around the world. This agreement covers everything from pharmaceuticals to information technology software and human gene sequences, and is emerging as a major issue dividing North and South. TRIPs agreement forces all countries to accept a medley of new biotech patents covering genes, cell lines, organisms and living processes that turn life into commodities. Governments all over the world have been persuaded into accepting these 'patents on life' before anyone understood the scientific and ethical implications.

The patenting of life-forms and living processes is covered under Article 27.3(b) of TRIPs. The TRIPS Article 27.3(b) is designed to allow the broadest categories of patents from genetic engineering and other new biotechnologies. The patenting of life-forms and living processes is covered under Article 27.3(b) of TRIPs. This scientific briefing explains why such patents should be revoked and banned on the following grounds:

- All involve biological processes not under the direct control of the scientist. They cannot be regarded as inventions, but expropriations from life.
- The hit or miss technologies do not qualify as 'inventions', and are inherently hazardous to health and biodiversity.
- There is no scientific basis to support the patenting of genes[9], genomes[10], cells and microorganisms[11], which are discoveries at best.
- Many patents are unethical; they destroy livelihoods, contravene basic human rights, create unnecessary suffering in animals or are otherwise contrary to public order and morality.
- Many patents involve acts of plagiarism of indigenous knowledge and bio-piracy of plants (and animals) bred and used by local communities for millennia[12].

In India, as it is mandatory for all the member countries of the WTO to adopt the agreements of WTO, the Ministry of Science and Technology has issued guidelines "Instructions for Technology transfer and Intellectual Property Rights", which would help in enhancing the motivation of the scientists, research institutions and universities in various research and development projects funded by various departments of the Ministry of Science and Technology. The salient features of these guidelines are as follows:

a) Ownership of Intellectual Property: the institution shall be encouraged to seek protection of IPR rights in respect of the results of R&D. they may retain the ownership of such IPRs. Institution means any technical, scientific and academic establishment where the research

[7] http://www.slwk.com/SLWK/web/company/papers/paper14.html

[8] Arnold,Beth E. and Eve Ohgeilski-Zei, "Patenting Genes and Genetic Reasearch Tools: Good or Bad for Innovation", Annual Review of Genomics and Human Genetics(2002), P. 416.

[9] Gene is a stretch of genetic material (DNA or RNA) with a defined function in the organism or cell. It usually codes for a protein.

[10] A genome is the totality of all the genetic material (deoxyribonucleic acid or DNA) in an organism, which is organised in a precise, though by no means fixed or constant way. In the case of viruses, most of them will have ribonucleic acid or RNA as the genetic material.

[11] A micro-organism is an organism that can be seen only under a microscope, usually, an ordinary light microscope. It includes bacteria, mycoplasm, yeasts, single-celled algae and protozoa.

[12] Cooper, I. P., *Biotechnology and the Law, (New* York, NY: Clark Boardman, 1985).

is carried through funding by central or state governments.

b) Transfer of Technology: the institutions would take necessary steps to commercially exploit patents on exclusive or non-exclusive basis.
c) Royalty to inventors: owner institutions are permitted to retain the benefits and earnings generated out of the IPR. Institutions may determine the share of the inventors and other associated persons from such earnings. However such sharing is limited to one third of such earnings.
d) Norms for the private industry: IPR generated through joint research by institutions and industrial concerns through joint research efforts can be owned jointly by them on mutually agreed terms through a written agreement. The institution or industrial concern may transfer the technology to the third party for commercialisation on exclusive or non-exclusive basis.
e) Patent Facilitating Fund: the owner institution shall set apart not less than 25% of the revenues generated from IPR to create a patent facilitating fund. The fund shall be utilised by the owner for updating the inventions, filing applications for new patents and protecting the IPR against infringement and for building the competency in the area of IPR and related issues.
f) Information: the institution shall submit the information relating to the details of the patents obtained, the benefits and earnings arising out of the IPR and the turnover of the products periodically to the Department/ministry, which has provided the funds[13].

VIII. Do We Need Biotechnology Patents?

The industries that utilise biotechnology are convinced that intellectual property protection should be obtainable for the inventions that stem from research and which have commercial potential. Biotechnology research workers in academic institutions increasingly share this view because of their need for research funding which is in part conditional on patentability.

But many people are not in favour of biotechnology patents. For many such groups "patenting life" is considered unethical in principle. The opposition extends also to possible structural change in the agricultural industry which might stem from biotechnology and especially from the acquisition by the larger corporations of legal rights on the advances that are being made.

A legally permissible ground of objection is that genes are naturally occurring entities and that the methods for transferring them to plants or animals are well-known and straightforward. This is a challenge to the inventiveness content of the particular patent at issue; it is an argument that industrial competitors will sometimes use against each other's patents but so far it has not achieved a high success rate. The argument also lies at the heart of the moral objections many with religious beliefs have to patenting genes. They regard claims of invention, instead of discovery, tantamount to claiming to be God.

Some feel that patenting living things change the relationship between humanity and the rest of nature. This is particularly sensitive as regards animals, where patents are seen as conferring "ownership", thereby undermining the animal's right to independence of being and relegating it to the status of a mere object. However, plants and animals are owned by the farmers who produce them and use them as agricultural commodities. All such owners, whether of patented or unpatented organisms, are bound to respect animal welfare legislation.

The objection to animal suffering may also apply to the genetic modification of farm animals. One early experiment to insert a growth hormone gene into a pig in order to increase growth rate succeeded but caused severe unforeseen side effects including arthritis. Animal welfare groups argue that patents will encourage more research on animal genetic modification, which they oppose on grounds of possible suffering and of principle. Intended to prevent undue suffering, legislation requires the granting of animal experimentation licenses and full disclosure of the experimentation.

The patentability of genes and other nucleic acid sequences is justified on the ground that they have been subject to a microbiological or non-biological process, i.e, gene sequencing, which is itself a standard process patentable and patented under existing patent laws for invention. So, the actual patented entity is the nucleic acid sequence itself and its putative function. However, the DNA or RNA sequence is subject to change by mutation, deletion, insertion and rearrangement.

The patenting of genomes raises the question of the function of the genomes. Again, the isolated genome can do nothing by itself, while its "function" in the organism cannot be considered separately from the totality of the organism.

IX. Conclusion

It can be concluded that, genetically modified microorganisms and even higher level animals like mammals can be patentable under the Patent laws of several Countries. The WTO, under TRIPs Agreement seeks to enforce US type patents in all the member countries. The concept of patenting the biotechnological inventions has been opposed by many groups on moral grounds. The author opines that, all biotech patents should be rejected on the following grounds:

- All involve biological processes not under the direct control of the scientist. They cannot be regarded as inventions, but expropriations from life.
- The 'hit or miss technologies' associated with many of the 'inventions' are inherently hazardous to health and biodiversity.

[13] Tripathi K. K., *Biotechnology: Government of India Initiatives*, Saketlnd Digest,Feb.2002.pp.49-53.

- There is no scientific basis to support the patenting of genes, genomes, cells and microorganisms, which are discoveries at best.
- Many patents are unethical; they destroy livelihoods, contravene basic human rights, create unnecessary suffering in animals or are otherwise contrary to public order and morality.
- Many patents involve acts of plagiarism of indigenous knowledge and bio-piracy of plants (and animals) bred and used by local communities for millennia.

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[4] Arnold,Beth E. and Eve Ohgeilski-Zei, "Patenting Genes and Genetic Reasearch Tools: Good or Bad for Innovation", Annual Review of Genomics and Human Genetics(2002), P. 416.

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[6] Tripathi K. K., *Biotechnology: Government of India Initiatives,* Saketlnd Digest,Feb.2002.pp.49-53.

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Comparison of Pole Placement & Pole Zero Cancellation Method for Tuning PID Controller of A Digital Excitation Control System

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Abstract- Modern Digital Excitation control system now started utilizing the power flexibility and cost advantage of Microprocessors for control. Many rotary excitation control system are still using analog type of Automatic voltage regulator which now started to replace with the digital Automatic voltage regulator. Modern Digital automatic voltage regulator is provided with PID controller in the forward path and tuning of PID controller is a challenging task. This paper discusses two methods of tuning PID controller i.e. Pole placement method and pole zero cancellation method. GUI (Graphical user interface) prepared for both the methods on the platform of MATLAB. Using this GUI performance results and time required for tuning for both the methods are compared.

Index Terms- Digital excitation system, Automatic voltage regulator, Pole placement method, Pole Zero cancellation method

I. INTRODUCTION

The term "excitation control system" refers to the entire control system including the synchronous machine and power system as well as the excitation system.[4]

Reliability of the excitation system has become an issue, especially where many of the generation plants may be critical to the internal processes used for manufacturing. A modern static excitation system is often the favored solution for old rotary excitation system since it has problems associated with it which are primarily mechanical.

Normal Industrial practice is to use static excitation system for generators requiring excitation current of 100A to 10000A and to use rotary excitation system for generators requiring excitation current of 1A to 200A .

Modern excitation systems are beginning to utilize the power, flexibility, and the cost advantage of digital electronics. There are a large number of aging excitation systems with rotary exciters in need of upgrading there are also cases where the rotary exciters are in good condition but need upgrading of the controls. In such cases only analog voltage regulator needs to be replaced with modern digital automatic voltage regulator.[3]

In the recent practice of the consultants, offering services in the field of maintaining & improving performance of excitation control system replacing Analog AVR's of rotary excitation systems with modern digital AVR's.

For the effective tuning of the rotary excitation system involving modern digital AVR there is need for models that accurately simulate the operation of excitation control systems during system disturbances in absence of precise data. It is because precise data for the old rotary excitation is generally not available.

There are many microprocessor based AVR has been discussed in the literature, These works on two types of algorithm i.e Arithmetic based algorithm and Logic based algorithm . Like Fuzzy logic controllers, ANN controllers etc.[10,11,12,13,14]

The success of the logic based algorithm based on the availability of the systems data like exciter time constant , generator time constant , Loop gain etc.

Two popular methods available for tuning of PID controller are Pole Placement method and Pole zero cancellation method For Modern digital excitation control system ,if the algorithm for both the methods are prepared using arithmetic approach then the model resulted for the digital excitation control system will work satisfactorily even in the absence of precise data of the system.[15]

GUI (Graphical User Interface) for both the methods have been prepared on the MATLAB platform using GUIDE function

Performance of the tuned Modern digital AVR by using both the method i.e. Pole placement method and Pole Zero cancellation method have been compared

II. EXCITATION CONTROL SYSTEM

The term "excitation control system" refers to the entire control system including the synchronous machine and power system as well as the excitation system. [4] . Block diagram is shown in fig 1.

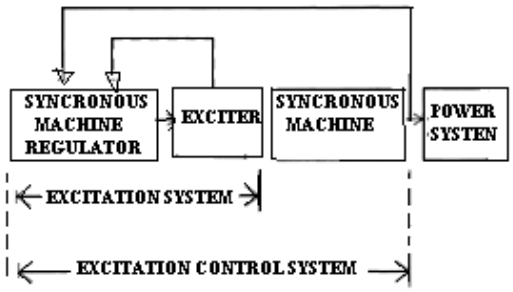


Fig .1. Block Diagram of the excitation control system

2.1. Types of excitation system.

Broadly two types of excitation systems are only recognized in the industry.

1. Static Excitation system.
2. Rotary excitation system.

These are further classified depending on following.

- **Depending on supply given to AVR (Automatic voltage Regulator) .**

a) Self excited excitation system

For this type of system automatic voltage regulator does not require supply from the external sources like PMG (permanent magnet generator) or battery, it takes supply directly from the output of the main generator.

b) Separately excited excitation system

For this type of system automatic voltage regulator requires supply from the external sources like PMG (permanent magnet generator) or battery,

- **Depending on the technology used for making AVR.**

a) Analog Excitation system.

This type of excitation system uses Analog IC'S like OP AMP

b) Digital Excitation system.

This type of excitation system uses chips like DSP or Microprocessor

III. TUNING OF PID CONTROLLERS

There are two popular methods of tuning PID controllers Pole Placement Method and Pole Zero Cancellation Method

3.1. Pole placement method.

This needs to determine Transfer function model of the exciter with generator G(s).Then determine transfer function H(s) of AVR with PID whose gains are unknown. Using characteristic equation G(s) H(s) = -1. Assume S = (a + jb) and equate real and imaginary terms we get two equations. Third equation is obtained by putting S= c .Now there will be three equations and three unknowns (K_P K_I and K_D) we can find K_P K_I and K_D. Using this known values of K_P K_I and K_D. We will find TF of system. Using this T.F. unit step voltage response, open loop frequency response, closed loop frequency response and root locus of the system will be obtained using GUI based on MATLAB.

This process will be repeated till we get desired transient response specifications.

Basically above process involve the placement of two dominant poles which are near the imaginary axis and belongs to generator and controller are made complex conjugate that makes response of the system underdamped.

The third pole of the exciter is kept a real pole and is placed far away from the imaginary axis so that it should not affect the natural mode of the voltage response.

The closed loop transfer function also involves The placement needs some trial and error in order to reduce the effect zero on the dynamic response of the performance..

3.2. Pole Zero cancellation method.

The PID equation can be written in this form:

$$G(s) = K_d \frac{s^2 + \frac{K_P}{K_D}s + \frac{K_I}{K_D}}{s}$$

When this form is used it is easy to determine the closed loop transfer function.

$$H(s) = \frac{1}{s^2 + 2\zeta\omega_n s + \omega_n^2}$$

If

$$\frac{K_P}{K_D} = 2\zeta\omega_n$$

and

$$\frac{K_I}{K_D} = \omega_n^2$$

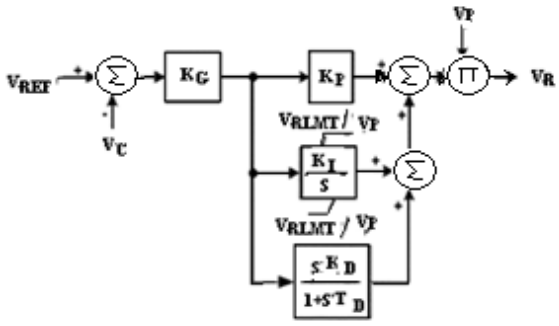
Then

$$G(s)H(s) = \frac{K_D}{S}$$

This can be very useful to remove unstable poles.

IV. MODELS DESIGNED FOR AVR USING BOTH THE METHODS.

IEEE Std. 421.5 Type AC5A - Simplified Rotating Rectifier Excitation System Representation has been considered which then simplified to a Modern Digital AVR. The simplified model of the digital AVR is shown in fig 2.



V_{REF} is the generator reference voltage
Vc is sensed voltage
V_{RLMT} is max field forcing
Vp power input voltage
V_R is Regulator output voltage
Fig 2: Simplified block diagram of Digital AVR

4.1: System specification

The system considered for the analysis has following specification

- 500 KVA Diesel generator set (CUMMINS POWERICA)
- Type of ac generator : brushless ac generator (STAMFORD make , Model : HCI544D1)
- Type Of Excitation System : Self excited rotating rectifier type excitation system.
- Type of AVR : Analog type (AS 440)
- Only generator time constant tg = 2.2 sec is available
- Exciter time constant is not available therefore it is considered as te = tg/6 sec

The generator under control is fed by a rotary exciter. The plant transfer function G(s) is given as

$$G(s) = \frac{1}{(1+st_g)(1+st_e)}$$

where
tg = the generator open-circuit time-constant
t_e = the exciter open-circuit time-constant.

The continuous-time PID controller model can be written as

$$G_C(s) = K_P + \frac{K_I}{S} + K_D S = \frac{K_D S^2 + K_P S + K_I}{S}$$

4.1 Model of the AVR using pole placement method.

The characteristic equation of the resulting closed-loop system is expressed in different forms as follows.

It is desired that

$$G(s)G_C(s) = -1$$

$$\therefore \quad \frac{K_D S^2 + K_P S + K_I}{S(1+St_g)(1+St_e)} = -1$$

$$\therefore K_D S^2 + K_P S + K_I = -S(1+St_g)(1+St_e)$$

The closed-loop system be dominantly second order. The dominant closed-loop poles (s = a & jb) are therefore selected from the peak overshoot and settling time specifications. The third pole is selected to be real (s =c) and in the far left-half plane, so that the closed-loop system is dominantly second-order. Substituting s=a+jb in the characteristic equation, and equating the real and imaginary parts, results in two equations. The third equation is obtained by substituting S = C in equation. The three unknown values (Kp , K_I , K_D) are then obtained from the three equations. A word of caution regarding the system zeroes is in order. The PID controller designed via pole placement method will force the closed loop poles to lie at the selected locations. The placement of the poles is achieved via appropriate choice of the controller settings (K_P, K_I, K_D). These controller settings give rise to two zeroes. The locations of the zeroes of the controller may be real or may even turn out to be a complex conjugate pair. The controller zeroes are not only the open-loop zeroes but also the closed-loop zeroes. Since the zeroes do affect the transient response some experienced engineering adjustment required in the design.

4.2 Model of the AVR using pole Zero cancellation method.

The characteristic equation of the resulting closed-loop system can be written as

$$G_C(s).G(s) = \frac{K_D(s^2 + \frac{K_P}{K_D}s + \frac{K_I}{K_D})}{t_g t_e s\,(s+\frac{1}{t_g})(s+\frac{1}{t_e})}$$

For the pole zero cancellation , we set

$$K_I = \frac{K_D}{t_g t_e}$$

$$K_D = K_D\left(\frac{t_g + t_e}{t_g t_e}\right)$$

Thus the transfer function gets reduced to

$$G_C(s).G(s) = \frac{K_D}{t_g t_e s}$$

The closed loop transfer function then becomes

$$\frac{C(s)}{R(s)} = \frac{G_C(s).G(s)}{1+G_C(s).G(s)} = \frac{K_D/t_g t_e s}{s+K_D/t_g t_e s}$$

The time response to a unit step input is as follows

$$C(t) = 1 - e^{-\frac{K_D}{t_g t_e}t}$$

If t_r is the specified rise time which is defined as the time required for the response to rise from 10% - 90% of its final value, the value of K_D is obtained by

$$K_D = \frac{t_g t_e \ln 9}{t_r K_C}$$

It can be seen that K_D depends on the plant parameters and the desired rise time. Once we establish K_D we can calculate K_I and K_P from the equation discussed above

At first, the idea of pole-zero cancellation might seem academic since the exact pole-zero cancellation is virtually impossible.

The root locus plots for cases where the actual and estimated time constants are off by ±*20% appear significantly* different. Experiment shows that in spite of these differences, the designed controller parameters result in performance that is acceptable for most generator sets with the exciter time constant about one-tenth of the generator time constant.

V. GRAPHICAL USER INTERFACE (GUI)

Graphical user interface prepared for Pole placement method which is also called direct design approach and pole Zero cancellation method is prepared in the MATLAB platform using GUIDE function

5.1 Acceptable values of Perfomance Index As per IEEE std

- Rise time = 0.15 to 2.55 sec
- Overshoot = 0 to 15%
- Settling time = 0.2 to 10 sec
- Bandwidth = 0.3 to 12 Hz

5.2: GUI For Tuning PID Controller By Pole Placement Method

GUI (Graphical user interface) to be used for pole placement method is shown in fig 3

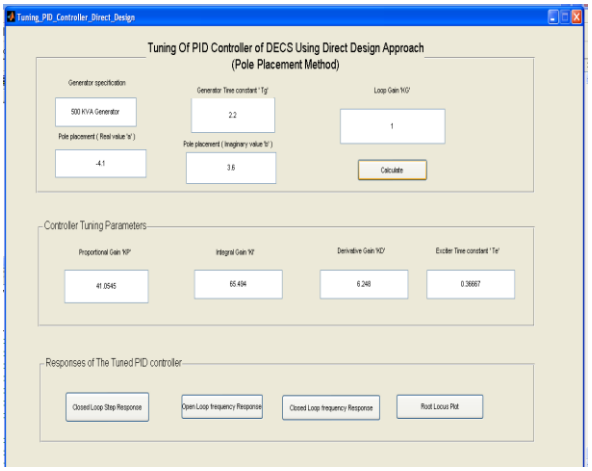


Fig3 . GUI for DECS using Pole Placement method

Pole placement method applied for rotary excitation system of 500KVA Diesel Generator set whose generator time constant is only known

The desired performance index would be

- Overshoot should be less than 15%
- Settling time should be below 1.5sec
- Bandwidth should be upto 12 Hz

Table 1. shows that the desired location of poles which meets desired performances indexes are at -3.9+ j2.6. GUI gives values of K_P =33, K_I=47 and K_D=6 for this location.

Fig 4, shows the closed loop unit step response with settling time = 1.5sec and overshoot is 14.2%, rise time = 0.188sec

Table 1. Search for better pole placement

Position of Pole	Rise time 'tr' (sec)	% overshoot '% Mp'	Settling time 'ts' (sec)
-3.1+j3	0.213	17.4	1.15
-3.6+j3	0.19	16.1	1.07
-3.6+j3.4	0.813	17.5	1
-3.6+j3.8	0.176	19	0.936
-3.8+j3	0.182	15.8	1.04
-4.1+j2.9	0.172	15	1.01
-4.2+j3	0.168	15.2	0.981
-3.8+j2	0.198	13	1.19
-3.8+j2.6	0.188	14.6	1.1
-3.9+j2.6	0.184	14.4	1.08

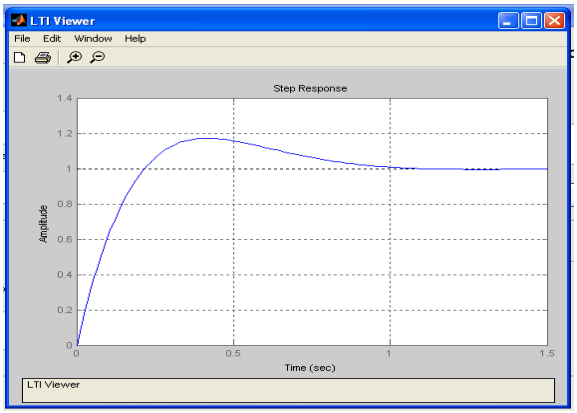


Fig.4. Closed loop step response on the 500 KVA generator for K_P =33, K_I=47 and K_D=6

Open loop frequency response Fig 5.Shows that at 0 db, crossover frequency is 40.4 Hz the phase lag is 87.5 degree

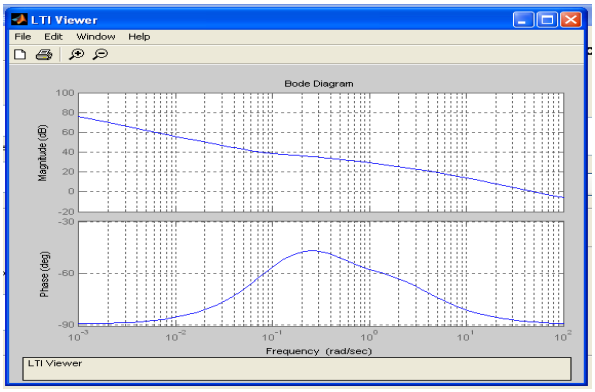


Fig 5. Open Loop frequency response on the 500 KVA generator for K_P =33, K_I=47 and K_D=6

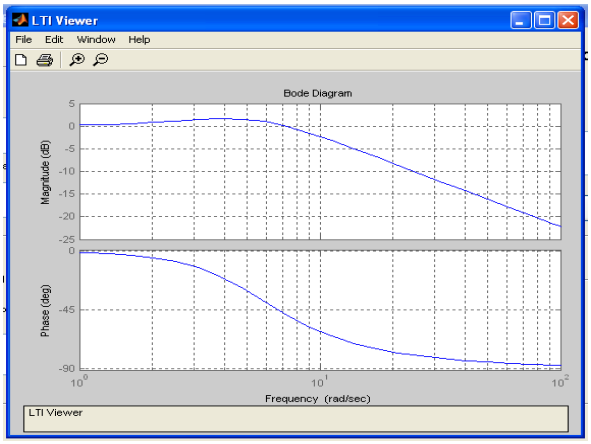


Fig.6. Closed Loop frequency response on the 500 KVA generator for K_P =33, K_I=47 and K_D=6

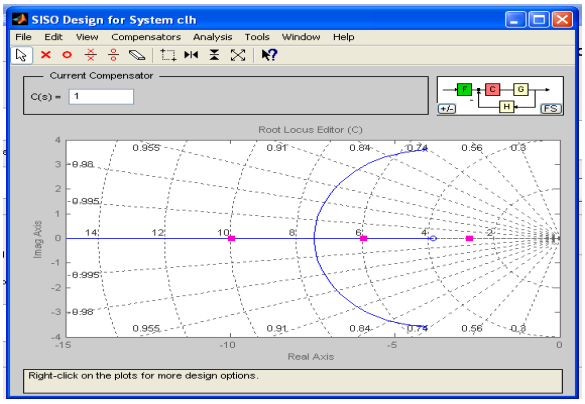


Fig.7. Root locus of the controller designed by pole placement method.

Closed loop frequency response Fig6, shows the phase lag of -61.1 degree at -3db. The decibel rise prior to roll off indicates that during step voltage test voltage overshoot is noted . The bandwidth is 9.55Hz

Root locus of the controller designed by pole placement method is shown in fig 7 confirms that in spite of oscillations system will be stable.

There is again a scope for a commissioning engineer to perform trial & error and also implement his field experience for making a system more stable. This is for making placement of zeroes in such a way that transient response is reduced by certain amount.

5.3: GUI for Pole Zero cancellation Method

GUI (Graphical user interface) to be used for pole placement method is shown in fig 8

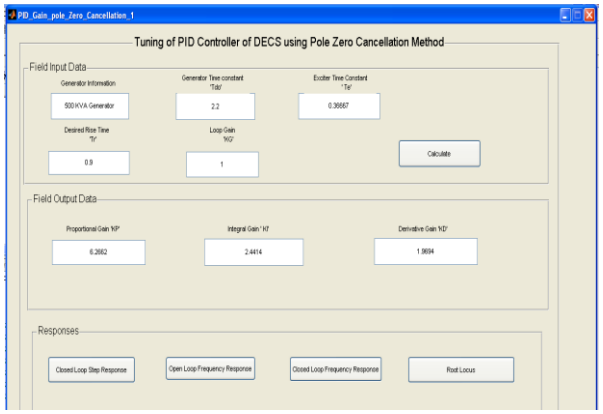


Fig.8. GUI for DECS using Pole Zero cancellation method

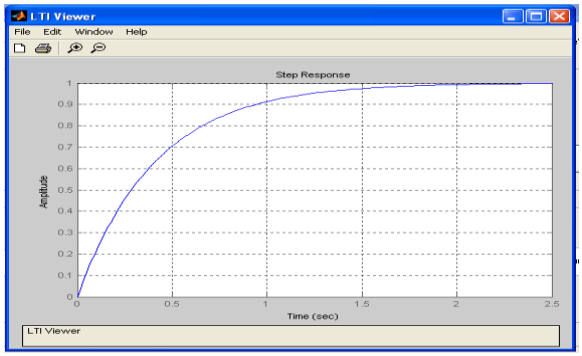


Fig.9. Closed loop step response on the 500 KVA generator for K_P =30, K_I= 11 and K_D= 9

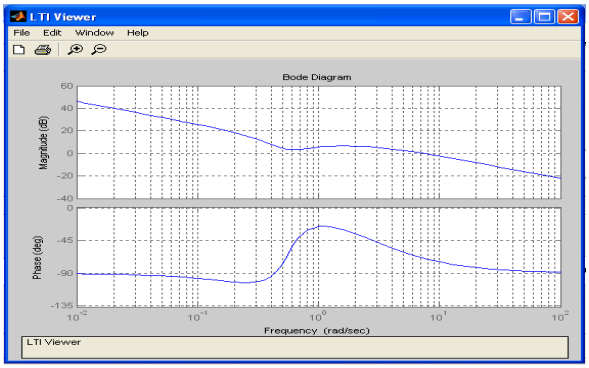


Fig.10. Open loop frequency response on the 500 KVA generator for K_P =30, K_I= 11 and K_D= 9

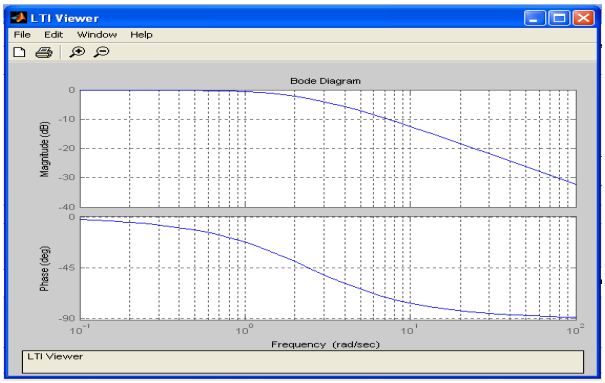


Fig.11. Closed loop frequency response on the 500 KVA generator for K_P =30, K_I= 11 and K_D= 9

Closed loop step response of the controller as shown in Fig9. is as per the desired rise time of 0.188 sec, settling time is 0.335 sec and almost no overshoot

Fig 10 .Open loop frequency response Shows that at 0 db, crossover frequency is 37 Hz the phase lag is -85.7 degree (less than what We got using pole placement method-87.5 degree)
Fig11.Closed loop frequency response shows the phase lag of -44.9 degree at -3db. There is no decibel rise prior to roll off Confirms that during step voltage test no voltage overshoot noted The bandwidth is 11.6Hz (more than that we got using pole placement method 9.55Hz)

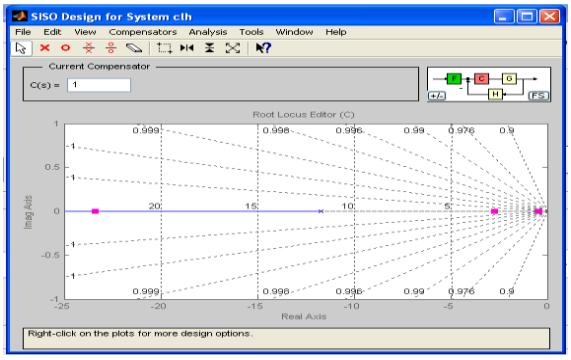


Fig 12. Root locus of the controller designed by pole Zero cancellation method

Fig 12.shows that the locus is on the real axis confirms that there will be no oscillation and hence system is more stable.

VI. Comparison of Performance

The performance of the controller designed using pole placement method and pole Zero cancellation method has been compared and presented in Table.2.

Table,2. Comparison of Performance

Methods/ Performance Index	Pole Placement Method	Pole Zero Cancellation Method
Rise Time(sec)	0.188	0,188 (desired)
Settling Time(sec)	1.1	0.335
% Overshoot	14.2	0
Bandwidth (rad/sec)	9.55	11.6
Degree of phase lag	61.1	44.9

VII. Conclusion

The comparison for pole-placement and the pole–zero cancellation methods for tuning PID controllers for digital excitation systems have been presented with the help of GUI using MATLAB. The Pole zer cancellation method requires less time for tuning PID gains therefore quick commissioning can be accomplished with excellent performance results.

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Preparation of Stable and Highly Active Hemoglobin by Using Antiglycosylspecific Hemoglobin Polyclonal Antibodies

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Abstract- The highly purified oligosaccharide chains obtained from glycoprotein invertase were coupled with bovine hemoglobin with the help of cross linking agent .The neoglycoprotein hemoglobin was highly purified to homogeneity by column chromatography and characterized on SDS- PAGE . Rabbits were immunized to produce glycosylspecific antihemoglobin polyclonal antibodies. Antibodies were purified by $(NH_4)SO_2$ precipitation followed by DEAE cellulose chromatography. The IgG-Sepharose was prepared by covalently coupling the purified polyclonal antibodies to CNBr-activated Sepharose 4B. The large amount of hemoglobin was immobilized on IgG-Sepharose by alternate incubation of hemoglobin and glycosylspecific antihemoglobin polyclonal IgG. The immunoaffinity based layered assembled preparation were highly active .The amount of hemoglobin immobilized could be raised 55 fold after four binding cycles. A layer by layer immobilization of hemoglobin significantly increases its life span and improved in stability against high temperature and other several denaturants (pH and salicylate). The results show the excellent platform for protein immobilization with high affinity. The polyclonal antibodies as supporting material has been found challenging for immobilizing the proteins. Further antiglycosylated hemoglobin can be used as a nanoparticle for immobilizing high amount of hemoglobin with no cytotoxicity to cell and this technique enhances the incorporation of hemoglobin properties (recognition, transport and catalytic properties). The applications of the immobilized glycosylated hemoglobin are found in the field of medicine, biotechnology, nanotechnology and biosensors.

Index Terms- Immobilization, glycosyl specific antibodies, Hemoglobin, stability, Application
Abbreviation : ConA, concanavalinA

I. Introduction

Hemoglobin is an important respiratory protein in red blood cells has a molecular weight of approximately 64,500 and consists of four polypeptide chains each with electroactive iron heme group [1] . It is an iron containing oxygen transport matalloprotein used extensively as an ideal electron transfer molecule in biosensor [2], [3]. The measurement of hemoglobin is of great importance due to its significance as a mediator in clinical, pharmaceutical, and as electrochemical biosensors [4],[5]. The polyclonal antibodies as supporting material has been found challenging for immobilizing the enzymes and proteins. In the authors laboratory, the high yield and very stable immobilized glucose oxidase preparations were obtained by immunoaffinity– layering of enzyme with polyclonal antibodies and $(Fab)_{2\epsilon}$ fragments [6],[7].

The carbohydrate specific antibodies have also been successfully employed for the high yield immobilization and stabilization of glycoenzymes [8]. In the present study an efforts have been made to obtain high yield immobilized neoglycohemoglobin preparation by assembling the proteins in layers by using glycosyl specific antihemoglobin polyclonal antibodies. These various immunoaffinity layered preparations of hemoglobin have been investigated for its stability against pH, temperatures and few more denaturants.

The procedure of bioaffinity based layering using an antibody support find its application in nanotechnology and biotechnology which give rise to newly emerging field: nanobiotechnology [9].

II. Materials and Methods

Hemoglobin (bovine blood was purchased from sigma),Invertase (250 units /mg), Sepharose 4B were obtained from Sigma (USA), PD-10 column, Sephadex G-10 and DEAE-52 (Amergham Biosciences), Pronase (Calibiochem Novabiochem Corporation ,USA).

Preparation and purification of neoglycoprotein

Commercially available invertase (150 mg) was digested with 8.0mg of pronase at 37 ^{0}C in 5ml of 0.1M Sodium phosphate buffer, pH 7.2 for 72 hours. The pronase digest was boiled at 100^{0}C for 10 min and then centrifuged at 1600g for 30min and was fractionated to isolate glycopedtide by loading on PD10 gel filteration column(7.5cm x 1.7cm) [6]. The preparation obtained was further purified by affinity column chromatography on the ConA-Sepharose support. The bound glycopeptides were eluted with 0.2 glycine/HCL containing 0.05 M NaCl buffer, pH 2.0, after 30min of continuous agitation. Excess glycine /HCL were removed by passing the sample through a Sephadex G-10 gel filteration column (1.8cmx20cm) [8]. Purified glycopeptides were treated with 0.5%glutaraldehyde for 5 hours a at4^{0}C These treated glycopeptides were then treated with 50 mg of hemoglobin dissolved in total volume of 5ml of sodium

phosphate buffer, pH 7.0. The obtained neoglycoprotein hemoglobin was separated from uncross reacted hemoglobin by loading the sample on Con A- Sepharose .The neoglycoprotein was then eluted with 0.2 M glycine / HCL buffer, and the preparation was immediately neutralized with 0.1 M NaoH [6]. Glycine /HCL were removed by passing purified neoglycoprotein through a Sephadex G-10 gel.

Production of polyclonal antibodies against neoglycoprotein.

Albino rabbits were immunized with 300µg of neoglycoprotein dissolved in 0.5ml of mM sodium phosphate buffer, pH7.2 mixed with equal volume of Freunds complete adjuvant as a first injection [10]. Booster doses of 150µg neoglycoprotein with Freunds complete adjuvant were administered weekly after resting the animal for 15 days. After each booster blood was collected from the ear vein of the animal and allowed to clot at room temperature for 3 hrs. Serum was collected by centrifugation after each booster dose and purified on DEAE cellulose column (1.2cmx10cm) [11].

Immunodiffusion

Ouchterlony double diffusion was used to detect the presence of antibodies against the glycosyl moiety of the hemoglobin. The cross reactivity of antineoglycoprotein with other proteins were performed in 1% agarose gel [12].The purified carbohydrate specific antibodies were employed for immunoaffinity layering of hemoglobin on sepharose 4B.

Preparation of layered assembly of hemoglobin by using glycosyl-specific anti hemoglobin polyclonal antibodies .

Purified polyclonal antibodies were coupled to CNBr-activated Sepharose 4B[13].The IgG bound matrix was then incubated with excess of hemoglobin in 20 ml of sodium phosphate buffer ,pH7.2 in a total volume of 3.0 ml overnight at 4^0C .The matrix bound enzyme was then washed thoroughly with buffer to remove the unbound protein . The preparation thus obtained was considered as first affinity layer. This preparation was then incubated with excess and appropriate amounts of glycosyl specific atihemoglobin polyclonal IgG and hemoglobin, alternately up to four layers.

Protein Assay

Protein concentration was determined by using method of Lowery etal. [14].
Hemoglobin concentration by Drabkin etal Method [15].

III. Results and Discussion

Pronase – digested preparation of commercially available invertase was used to isolate the oligosaccharide chains of the enzyme. The glycosyl moieties were purified by ConA-Sepharose chromatography and coupled to hemoglobin with the help of glutaraldehyde to synthesize the neoglycoprotein [8].The newly synthesized glycoconjugate was further purified by conA-Sepharose chromatography and characterized on SDS/PAGE by periodic acid –Schiffs staining [16].The neoglycoprotein moved as a single band on the gel .The neoglycoprotein thus obtained was highly immunogenic and precipitating antibody produced against this antigen. The glycosyl-specific antihemoglobin antiserum showed cross reactivity with neoglycoprotein and hemoglobin, but not with the digested peptides of invertase (fig1). The results confirm that the specific antibodies were raised against peptide epitopes of neoglycoprotein and native hemoglobin. The results suggest that presence of specific antibodies for neoglycoproteins and hemoglobin. In earlier studies it has already been demonstrated that neoglycoproteins synthesized by coupling the glycosyl specific moieties of yeast glucose oxidase to BSA were highly specific. The immunoadsorbant for immobilization of hemoglobin was prepared by coupling IgG purified from the sera of rabbits immunized with neoglycoproreins on activated Sepharose-4B [8]. The activated support bound 7.6 mg of protein /g of the gel. The IgG-Sepharose preparation was incubated overnight with excess of hemoglobin at 4^0Cin 20 mM buffer, pH 7.2.The preparation thus obtained was considered as first layer. The IgG immobilized hemoglobin preparation was then incubated alternately with glycosyl specific polyclonal antibodies and hemoglobin .The results obtained after four affinity layers are summarized in table 1.

Enzyme immobilized on Sepharose-4B support increased gradually with successive incubation of protein antibodies and after four desired cycles, the amount associated with matrix was 55 fold as compared with protein directly bound on IgG-Sepharose4B. The effectiveness factor η also increased linearly with the formation of successive affinity layers of IgG and protein. Protein assembled in the form of layers on Sepharose -4B exhibited enhancement in the incorporation of hemoglobin properties (recognition, transport and catalytic properties).

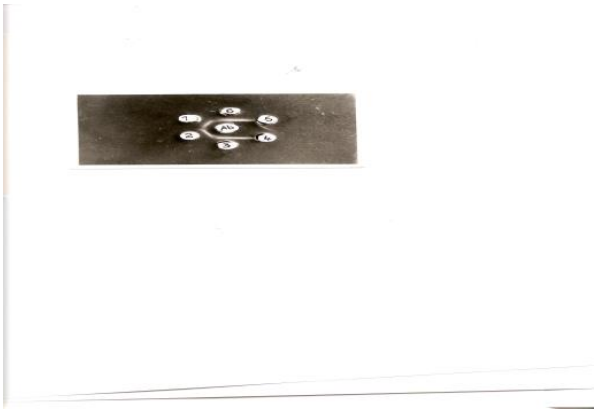

Figure 1 shows antiglycosylspecific polyclonal antibodies (Abs) at the center and peripheral well 1, 2 Hemoglobin 3,5and 6well neoglycohemoglobin and well 4 with digested invertase peptides. .

Study the properties of immunoaffinity layered glycosylated hemoglobin.

Figure 2 demonstrates the effect of temperature on stability of various immunoaffinity layered hemoglobin. The preparation bearing 4 layers was significantly superior in stability and retained nearly 85% activity while the first affinity layer preparation exhibited only 55% of the original protein activity after 2 h of pre incubation at 60^oC. The native protein lost 99% of the initial activity. The increase in stability was also layer –

dependent and was of relatively highly magnitude in all the immobilized preparation as compared to the soluble protein. The immunoaffinity support holds the protein at multiple points of attachment in three dimensional antigen-antibody complexes, thus increased the thermal stability of protein. The enhancement in thermal stability was further proved to a greater extent with bioaffinity layering of proteins on the matrix [8].

Table 1
Immunoaffinity layering of hemoglobin on Sepharose -4B using purified glycosyslpecificIgG. Each value represents the mean from three independent experiments performed in duplicate not exceeding 100% of the mean value.

Amount of Hemoglobin bound mg /g of the gel					
Layer	Theoritical Value (A)	Actual Value (B)	Effectiveness factor (B/A)η	Increase in binding over 1(fold)	**Protein bound in (mg/g of gel)**
I	**10**	**7.6**	**0.76**	**1**	**7.6**
II	**100**	**60**	**6.0**	**8**	**60**
III	**250**	**235**	**23.5**	**31**	**235**
IV	**500**	**418**	**41.8**	**55**	**418**

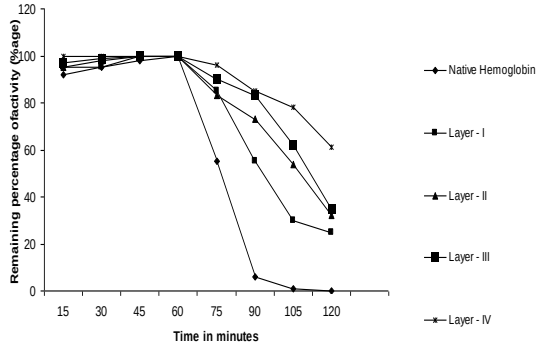

Figure2. Thermal stabilty plot for native and immunoaffinity bound hemoglobin. Native and immunoaffinity layered hemoglobin was incubated at 60 ^{o}C for 2 h for various times and its activity was measured as described in text by using Drabkins method [15]. The activity of unincubated hemoglobin was considered as control (100%).

Effect of pH on the immunoaffinity layered preparation of protein and native protein was was investigated in pH range from pH 4 to pH 13 [Fig.3]. There was a remarkable increase in stability of immunoaffinity layered preparations of hemoglobin as compared to the native protein at highly acidic and alkaline side of the pH. Several earlier investigators have also documented for the stability of normal immobilized hemoglobin [17]. Khan et al. [18] have also shown that bioaffinity bound enzymes retained high activity at wide range of acidic and alkaline pH. It exhibited that the stability of hemoglobin enhanced due to binding of protein with antibodies at multiple sites [19],[20]. The structural basis for stabilization of hemoglobin is associated with decrease in the ionic character and increase in the covalent nature of the bonds in the bioaffinity layered preparation of hemoglobin.

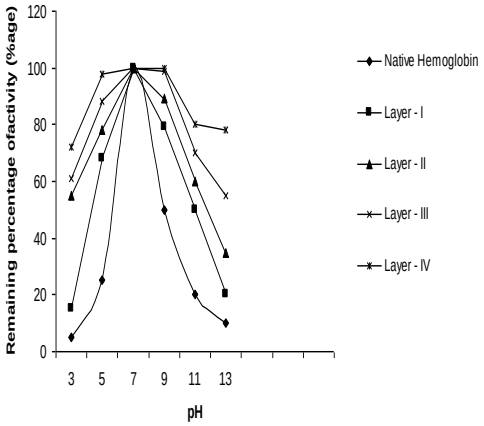

Figure 3.pH-activity profiles for native and immunoaffinity bound hemoglobin. Native and immunoaffinity bound hemoglobin were incubated in the buffers of various pH ranges for 2 h at 37^{0}C. The activity for native and immunoaffinity layered neoglycoprotein at pH 7.0 was considered as control (100%).

Four immunoaffinity layered preparation of hemoglobin showed no sign of its precipitation when it was incubated with 0.5 M salicylate in a neutral phosphate buffer for 2 h at 37^{o}C while the native and single layered protein preparations were precipitated when these preparations were incubated under similar experimental conditions. The 10% of precipitation was observed for one layered preparation whereas no precipitation was seen in the fourth layered preparation. The results show the dense and visible precipitation in the native protein hemoglobin

under the identical conditions. Earlier studies were documented with differences between native and denatured hemoglobin [21].

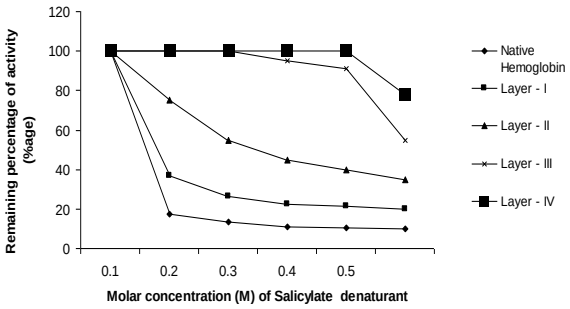

Figure 4. Stability of native and immunoaffinity layered hemoglobin against denaturant salicylate. Native and immunoaffinity layered hemoglobin preparations were incubated with various concentration of salicylate for 2 h 37^{o}C. The activity of unincubated native and immunoaffinity bound hemoglobin was considered as control (100%).

Storage stability

Storage stability the native and the bioaffinity layered preparation of neoglycohemoglobin were stored at 37^{0}C and 4^{0}C for 30, 60, 90 and 120 days. In the native protein hemolytic and rapid growth of bacteria were observed after 30 days in both 4^{0}C and 37^{0}C. However, The bioaffinity layered preparation of hemoglobin did not show any hemolysis and growth of bacteria after a long term storage for 120 days .

Applications

The large amount of immobilized hemoglobin molecule on a small surface of matrix is one of the important factors for development of immunosensors and biosensors which is clinically relevant for diabetic patients. The denaturation of hemoglobin with denaturing adducts present in the blood can be reduced by using the immunoaffinity layered preparation of protein .The study is also applicable for the patients of family with glycated hemoglobin formed by non enzymatic reaction of glucose with reactive amine groups on hemoglobin

The work done can be carried out to study the immunoaffinity matrix for development of HbA1c immunosensors.

IV. Conclusion

Immunoaffinity layered hemoglobin showed very high stability against extreme conditions of pH and temperature and this preparation did not exhibit any sign of precipitation when exposed to high concentration of salicylate for very long time. In view of the stability of the immobilized hemoglobin it can be suggested that these types of preparations can be used in biosensors. The proposed work showed excellent reproducibility, stability and long term storage and it indicated that such preparations can successfully be used for continuous analysis.

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A Fuzzy Switching Filter for Enhancing Digital Images Corrupted by Impulse Noise

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Abstract- In this paper, fuzzy based median filtering technique is proposed for enhancing highly corrupted digital images. This filter is obtained in two steps; in first step fuzzy decision rule is applied to detect the impulse noise on input image (noisy image). In second step, noisy pixels are removed using decision based filters. The performance of proposed filter is compared with other existing filters and shown to be more effective in terms of eliminating impulse noise and preserving edges and fine details of digital images.

Index Terms- Fuzzy logic, Impulse Noise, Membership function and Nonlinear Filter

I. INTRODUCTION

Removal of noise is one of the challengeable tasks in image processing. commonly, linear techniques are used because linear filters are easy to implement and design. Further, they are optimal among the class of all filtering operations when the noise is additive and Gaussian [1-3]. However, if the images are contaminated by impulse noise, these assumptions are not satisfied. Thus the linear filters deteriorate severely. To address these issues, a large number of nonlinear methods had been investigated [4, 5]. The most popular nonlinear filter is the median filter; it is computationally efficient and has proved extremely successful for removing noise of impulsive nature. Unfortunately, it suffers from the fact that the signal details become blurred.

To overcome these problems, Existing switching-based median filters had been investigated [6-16]. However these are commonly found to be non-adaptive to noise density variations and prone to misclassifying pixel characteristics at high noise density interference. To overcome these, it is investigated to study and examine new decision based filtering schemes [17-25]. The decision based filtering procedure consists of the following two steps: (i) an impulse detector that classifies the input pixels as either corrupted pixels or uncorrupted pixels, (ii) a noise removal filter that aims to restore those pixels that are classified as corrupted pixels. These techniques aim to achieve optimal performance over the entire image. A good noise filter is required to satisfy two criteria, namely, suppressing the noise and preserving the useful information in the signal.

Fuzzy switching median filters have been investigated to satisfy the above mentioned criteria [26-30]. In many image-processing applications, expert knowledge is used to overcome the difficulties (e.g. object recognition, scene analysis). Fuzzy set theory and fuzzy logic offer powerful tools to represent and process human knowledge in the form of fuzzy if-then rules. On the other side, many difficulties in image processing arise because the data/tasks/results are uncertain. This uncertainty, however, is not always due to the randomness but to the ambiguity and vagueness. Beside randomness which can be managed by probability theory can distinguish between three other kinds of imperfection in the image.

Fuzzy filters eliminate impulse noise satisfactorily. Even though, these are commonly found to be non-adaptive to noise density variations and prone to misclassifying pixel characteristics. In order to address the above mentioned issues, a new fuzzy switching filter is proposed. In this paper, gradient vector is determined between noisy image and two decisions based nonlinear filters separately for impulse detection. In addition to this, fuzzy gradient is calculated. Impulse detection is obtained using fuzzy IF-THEN rule and detected impulse noise is eliminated using two decision based filters. The proposed filter exhibit superior performance in terms of impulse noise detection and reduction than the other existing filtering techniques. The newly proposed fuzzy switching filter is a powerful tool for knowledge representation and processing and it can manage the vagueness and ambiguity efficiently.

In this paper, noise model and two special of decision based filters are explained for clarity of work and then the proposed filter is explained. In section2, noise model is described. Section 3, described the image denoising. In section 4, proposed filtering algorithm is discussed. Section 5 discusses the experimental results. Conclusion is drawn in section 6.

II. NOISE MODEL

Fundamentally, there are three standard noise models, which model the types of noise encountered in most images; they are additive noise, multiplicative noise and impulse noise. Digital image are often corrupted by salt and pepper noise (or impulse noise). Impulse noise is considered for proposed work. For images corrupted by salt-and-pepper noise (respectively fixed-valued noise), the noisy pixels can take only the maximum and the minimum values (respectively any random value) in the dynamic range. In other words, an image containing salt-and-pepper noise will have dark pixels in bright region and bright pixels in dark regions. A digital image function is given by $f(i,j)$ where (i,j) is spatial coordinate and f is intensity at point(i,j). let $f(i,j)$ be the original image, $a(i,j)$ be the noise image version and η be the noise function, which returns random values coming

from an arbitrary distribution. Then the additive noise is given by the equation (1)

$$a(i, j) = f(i, j) + \eta(i, j) \quad (1)$$

Impulse noise is caused by malfunctioning pixels in camera sensors, dead pixels, faulty memory locations in hardware, erroneous transmission in a channel, analog to digital converter, malfunctioning CCD elements (i.e. hot and dead pixels) and flecks of dust inside the camera most commonly cause the considered kind of noise etc. It also creeps into the images because of bit errors in transmission, faulty memory locations and erroneous switching during quick transients. Two common types of impulse noise are the salt and pepper noise and the random valued noise. The proposed filter first detects the Salt and pepper noise present in digital images in very efficient manner and then removes it. As the impulse noise is additive in nature, noise present in a region does not depend upon the intensities of pixels in that region. Image corrupted with impulse noise contain pixels affected by some probability. The intensity of grayscale pixel is stored as an 8-bit integer giving 256 possible different shades of gray going from black to white, which can be represented as a [0,*L*-1] (*L* is 255) integer interval. In this paper the impulse noise is considered. In case of images corrupted by this kind of salt and pepper noise, intensity of the pixel A_{ij} at location (i,j) is described by the probability density function given by the following equation (2)

$$f(A_{ij}) = \begin{cases} p_a & for \;\; A_{ij} = a \\ 1 - p & for \;\; A_{ij} = Y_{ij} \\ p_b & for \;\; A_{ij} = b \end{cases} \quad (2)$$

where *a* is the minimum intensity (dark dot); *b* is the maximum intensity (light dot); p_a is the probability of intensity (a) generation; p_b is the probability of intensity (*b)* generation; *p* is the noise density, and Y_{ij} is the intensity of the pixel at location (i,j) in the uncorrupted image. If either p_a or p_b is zero the impulse noise is called unipolar noise. If neither probability is zero and especially if they are equal, impulse noise is called bipolar noise or salt-and-pepper noise.

III. IMAGE DENOISING

In this section, two special classes of decision based filtering techniques are described. Decision based filter 1 (DBF1) is used to preserve the edge region of the digital image and homogeneous region is preserved using Decision based filter 2 (DBF2). These are explained in subsection 3.1 and 3.2.

A. Decision based filter 1

This filtering technique has been investigated in the literature [20] for edge preserving properties of digital images. The filtering operation is experimented by way of edge preserving method to improve visual perception and then filtering is performed at the current pixel within the sliding window on digital image. Edge detection is a fundamental tool in image processing and computer vision, particularly in the areas of feature detection and feature extraction, which aim at identifying points in a digital image at which the image brightness changes sharply or more formally has discontinuities. Edges play an important role with these shapes because transition of one shape to another is controlled by the edge and its quality. The concept of hard, soft or lost edges control shape is depend upon a conceptual basis. There is no analytical reality for human visual perception. In this paper, edges on the noisy image are identified using one of the properties of edge detection and preserved.

In this filter, the central pixel is identified as corrupted one; it is replaced by the proposed edge preserving method. Therefore, edges on the image is detected by computing gradient value in the direction of horizontal, vertical, left diagonal and right diagonal within the filtering window respectively. Based on neighbourhoods within the filtering window, the gradient value is obtained by determining the difference of two pixel intensities in direction of vertical, horizontal, left diagonal and right diagonal respectively. These four gradient values (according to the four different directions or neighbours) are considered for making the decision to eliminate impulse noise as well as preserve the edges of the image. If the gradient value is quite large, any one of the pixel is affected in the corresponding direction with minimum/maximum value of impulse noise. The minimum gradient value is a good indication that those pixels are noise free edge pixels in the direction of orientation. The minimum gradient value with respect to (i,j) can be used to determine direction of orientation of edge pixel. In order to preserve the edges, the corrupted central pixel is replaced by the average of two intensities which are obtained with respect to the direction of minimum gradient value.

B. Decision based filter 2

In this section, homogeneous region of image is preserved by applying decision based switching median filtering technique [21]. The pixels inside the sliding window are classified as corrupted and uncorrupted pixels by comparing their values with the maximum (255) and minimum (0) values. Consider an image of size M×N having 8-bit gray scale pixel resolution. A two-dimensional square filtering window of size 3 x 3 is slid over the noisy image. As the window move over the noisy image, at each point the central pixel inside the window is checked whether it is a corrupted pixel or not. If the pixel is an uncorrupted one, it is left undisturbed and the window is moved to the next position. On the other hand, if the pixel is detected as a corrupted one, the filtering procedure is performed by following the further steps described below.

Step1) Separate the corrupted and uncorrupted pixels inside the filtering window at its current position.

Step2)Check if the uncorrupted samples are an odd number inside the filtering window. If so, the median of the uncorrupted samples is set out as the filter output.

Step3)If the uncorrupted samples are an even number, then the Range estimator (RE) is determined for the uncorrupted samples. Range estimator for uncorrupted pixels is the difference between the last and the first of the sorted out uncorrupted pixels. A suitable threshold value T is chosen for determining the presence of an edge at the central pixel. In this work, 125 is selected as suitable threshold.

Step4) If RE is greater than the threshold value T, the central pixel is declared an edge and therefore, it left unaltered; otherwise, the central pixel is replaced by the arithmetic average of the uncorrupted pixels inside the filtering window.

Then the window is moved to form a new set of values, with the next pixel to be processed at the centre of the window. This

process is repeated until the last image pixel is processed. It may be noted that the filtering is performed by either taking the median or the arithmetic mean of uncorrupted pixels of the filtering window.

IV. PROPOSED ALGORITHM

The filtering technique proposed in this paper detects the impulse noise on the image using a fuzzy decision rule. An image gradient vector is determined between central pixel from an input image and filtered output image for two decision based filters. These gradient vectors are given by the following equation:

$$\nabla(k,l)D1 = YDBF1(i,j) - A(i,j) \quad with k,l \in \{-1,0,1\} \tag{3}$$

$$\nabla(k,l)D2 = YDBF2(i,j) - A(i,j) \quad with k,l \in \{-1,0,1\} \tag{4}$$

where, is gradient vector between $\nabla_{(k,l)}D_1$ original pixel value and decision based median filter1 (DBF1), $\nabla_{(k,l)}D_2$ is gradient vector between input pixel value and decision based median filter2 (DBF2)output, *A(i,j)* is the Noisy image, Y_{DBF1}(i,j) is the decision based median filtered (DBF1) output, Y_{DBF2}(i,j) is the decision based median filtered (DBF2) output. The proposed filtering algorithm is applied on noisy image and is described in steps as follows:

Step 1) A two-dimensional square filtering window of size 3 x 3 is slid over the noisy image.

Step 2) As the window move over the noisy image; at each point, the central pixel inside the window is checked whether it is a corrupted pixel or not. If the pixel is an uncorrupted one, it is left undisturbed and the window is moved to the next position. On the other hand, if the pixel is detected as a corrupted one, the filtering procedure is performed by following the further steps described below.

Step3) Determine gradient vector between central pixels from filtered image and processing pixels from noisy image (i.e. input image)

Step4) Fuzzy IF THEN rule is applied for impulse noise detection and noisy pixels are replaced by two decision based filters. Then the window is moved to form a new set of values, with the next pixel to be processed at the centre of the window. This process is repeated until the last image pixel is processed.

In this paper, fuzzy gradient () is obtained by fuzzy set and is used for impulse detection. This gradient is used to determine if a central pixel switching the filtering window is corrupted with impulse noise or not, because if this gradient is small then It is good indication of noise free pixel. Suppose if gradient is medium then the processing pixel is either an edge pixel or noisy pixel. If gradient is large then it is an indication of noisy pixel.

Finally, fuzzy gradient value is used to detect the impulse noise. The proposed fuzzy switching median filter is constructed using Sugeno fuzzy system, Gauss membership function with 3 input, two input variables and 9 fuzzy rules. Noisy pixels are detected by the following fuzzy rules:

IF D_1 is low AND D_2 is medium THEN $\nabla_{(k,l)}D_F$ *is high*

IF D_1 is medium AND D_2 *is high* THEN $\nabla_{(k,l)}D_F$ *is low*

IF D_1 *is high* AND D_2 *is low* THEN $\nabla_{(k,l)}D_F$ is medium

IF D_1 *is low* AND D_2 is medium THEN $\nabla_{(k,l)}D_F$ *is high*

IF D_1 is medium AND D_2 *is high* THEN $\nabla_{(k,l)}D_F$ *is low*

IF D_1 *is high* AND D_2 *is low* THEN $\nabla_{(k,l)}D_F$ is medium

Here, Low, medium and high are the nondeterministic features, these terms can be represented as fuzzy sets. Fuzzy sets can be represented by membership function. These membership functions are described by the following fig.1.

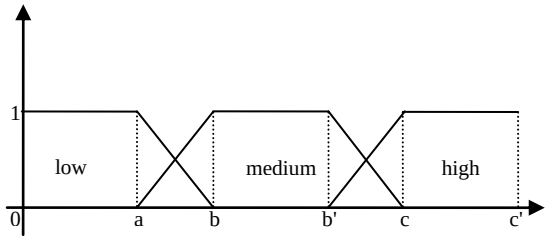

Fig.1 Membership function low, medium and high respectively

These functions represents all possible gradient values in the range of 0 to 255 and membership degree is represented in the range of 0 to 1 (i.e. $\in$[0,1]). A membership degree indicates the degree in which a certain gradient value matches the predicate. If the gradient value has membership degree one for the fuzzy set high, it means that it is large for noisy pixels, low for noise free pixels and medium for noisy/edge pixelsFuzzy rule contains some conjunctions and disjunctions. In fuzzy logic, AND operator is used for conjunctions and OR used for disjunctions. These three membership function depend on the parameters 0-a, b-b' and c-c'. According to the following observations these parameters values are determined.

1) Gradient value $\nabla_{(k,l)}D_1$ or $\nabla_{(k,l)}D_2$ *lie in the interval [0-45] indicates* the noise free pixels and nonedge pixels. So these pixels are categorized as noise free pixels and there is no uncertainty is happen, resulting in a zero membership degree in the fuzzy set.

2) Gradient value $\nabla_{(k,l)}D_1$ or $\nabla_{(k,l)}D_2$ *lie in the interval [45-125] indicates* the noisy pixels and edge pixels. So these pixels are categorized as either noisy pixels or edge pixels. Here, some kind of uncertainty is expressed by membership degrees in fuzzy set(lies in the interval [0-0.5].

3) Gradient value $\nabla_{(k,l)}D_1$ or $\nabla_{(k,l)}D_2$ *lie in the interval [125-255] indicates* the noisy pixels. So these pixels are categorized as noisy pixels. In this case, membership degree in the fuzzy set is one.

In fig.1 the membership degree is represented for noise free pixel in the range of [0-a], noisy or edge pixels are represented in the range of [b-b'] and noisy pixels are represented in the range of [c-c']. Based on image quality only, the membership degree and fuzzy gradient $\nabla_{(k,l)}D_F$ value is chosen. After choosing the membership degree and value, activation degree is calculated for fuzzy IF-THEN rule. This activation degree indicates the low, medium and large $\nabla_{(k,l)}D_F$ value for impulse noise detection on digital images. Finally, the detected impulse noise is replaced by either DBF1 or DBF2. If the indentified pixel is either noisy or edge pixels then it is replaced

by DBF1 and otherwise it is a corrupted one and is replaced by DBF2.

V. SIMULATION RESULTS

The filtering technique is tested using 3 x 3 windows with Lena image of size 256 x 256. In this paper, Lena image is used as a test images. In order to analyze the performance of the proposed filter approach, the performance evaluation factors like Peak Signal to Noise Ratio (PSNR) is used. This performance evaluation is based on threshold values and noise levels. Filter has higher PSNR values are considered to be superior filter in terms of noise elimination and restoration of image features. A New Tristate Switching Median Filtering Scheme (NTSMF) is quantitatively evaluated using objective measures are defined as:

$$PSNR = 10log_{10}\left[\frac{255*255}{MSE}\right] \quad (5)$$

where,

$$MSE = \frac{\Sigma\left|X(i,j) - F(i,j)\right|^2}{row*col} \quad (6)$$

(i,j) denotes the number of rows and columns in the image data, *X(i,j)* represents the pixel intensities of the original image at the position of *X(i,j)*, *F(i,j)* represents the output intensities in the filtered image at the position of (i,j). The proposed filter has very good subjective improvements for lower level of mixed impulse noise (i.e. fine details preservation of the image).

TABLE.1

PSNR values obtained using proposed filter and compared with different filtering techniques on Lena image corrupted with various densities of impulse noise

Filtering Techniques	10	30	50	70	90
MF	31.74	23.20	15.28	9.98	6.58
WMF	23.97	22.58	20.11	15.73	8.83
CWMF	28.72	20.28	14.45	10.04	6.75
TSMF	32.89	24.96	16.82	11.33	7.58
NID	37.90	28.75	23.42	14.65	7.77
MDBSMFS	34.83	24.79	16.99	11.28	6.97
IDBA	36.5	29.72	26	23.5	19.3
NTSMF	42.57	35.38	29.34	19.52	10.13
DBF1	40.8	31.0	22.6	13.42	7.06
DBF2	38.42	30.47	24.92	18.84	10.03
Proposed fuzzy filter	46.07	39.64	30.64	19.93	10.76

The enhancement result for the corrupted 'Lena' image by different level of impulse at suitable threshold has been estimated. The estimated values are tabulated and are given in the Table1. Figure 2 graphically illustrates the objective improvement of the proposed filter with respect to other switching schemes and this fuzzy filter exhibit good performance when images are corrupted up to 60% of noise. In order to prove the effectiveness of this filter, existing filtering techniques are experimented and compared with proposed filter for visual perception and subjective evaluation on Lena image including the standard medians filter (MF), the Weighted median filter (WMF), Centre weighted median filter (CWMF), the Tristate median filter (TSMF), New impulse detector (NID), Multiple decision based switching median filter (MDBSMF), Improved decision based algorithm (IDBA), New tristate switching median filter (NTSMF), Decision based filter1 (DBF1), Decision based filter2 (DBF2) and the proposed filter in Fig.2.

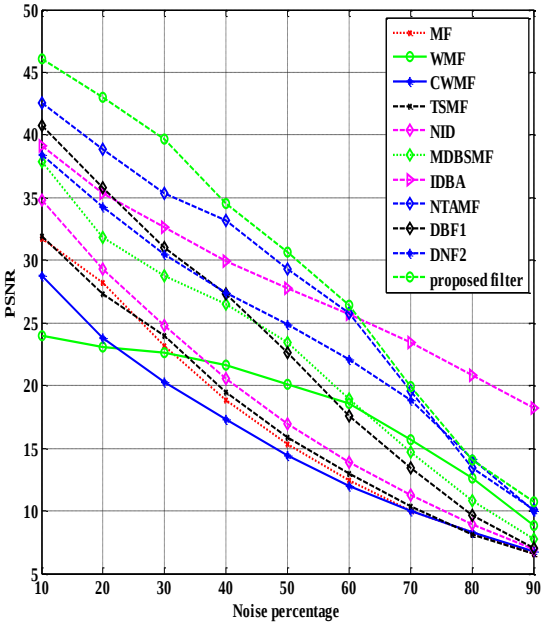

Fig.2 PSNR obtained using proposed filter on Lena image corrupted with different densities of impulse noise and compared with other existing filtering techniques

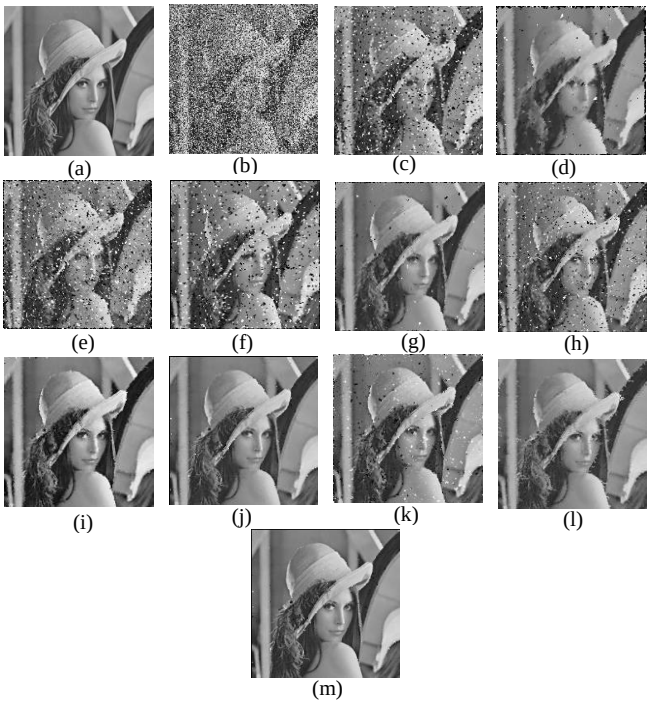

Fig.3 Subjective Performance comparison of proposed filter with other existing filters on test image Lena (a) Noise free images, (b) image corrupted by 50% impulse noise, (c) images restored by MF, (d) images restored by WMF, (e) images restored by CWMF, (f) images restored by TSMF, (g) images restored by MDBSMF, (h) images restored by NID, (i) images restored by IDBA,

(j) image restored by NTSMF , (k)images restored by DBF 1, (l) image restored by DBF 2 and (m) image restored by the proposed fuzzy filter

Figure 3 illustrates the subjective performance for human visual perception. The performance of this filter is evaluated using various impulse corruption ratios from 10% to 90% with suitable threshold. This filter has better performance than the other filtering schemes for the noise densities up to 60%. It shows that the better performance in removing impulse noise from digital images without distorting the useful information in the image.

Table 2
PSNR values obtained by applying proposed filtering technique on different test images corrupted with various densities of impulse noise

Noise %	Baboon	Lena	Pepper	Rice
10	40.70	46.07	49.52	47.65
20	35.48	43.04	45.82	45.93
30	31.64	39.64	42.25	39.74
40	28.43	34.54	37.74	36.03
50	25.34	30.64	34.26	33.73
60	22.23	26.44	29.78	28.01
70	18.45	19.93	23.56	22.65
80	13.74	14.04	18.78	17.57
90	9.89	10.76	14.47	14.95

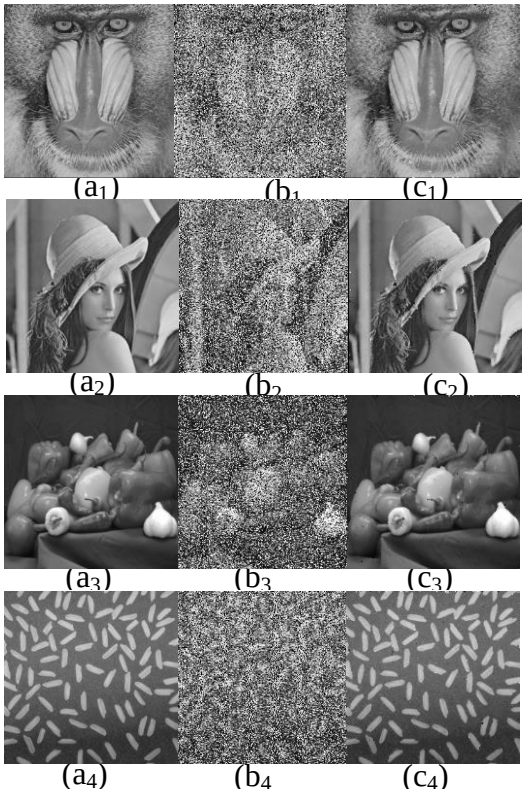

Fig.4 Performance of test images: ($a_{1,2 \text{ and } 3}$) original images,($b_{1,2 \text{ and } 3}$) images corrupted with 50% of noise and ($d_{1,2 \text{ and } 3}$) images enhanced by proposed filter

Digital images are nonstationary process; therefore depends on properties of edges and homogenous region of the test images, each digital images having different quantitative measures. Fig.4 illustrate the subjective performance of the fuzzy filtering technique for Baboon, Lena, Pepper and Rice images: noise free image in first column, images corrupted with 50% impulse noise in second column, Images restored by proposed Filtering technique in third column. This depicts properties of digital images. Performance of quantitative analysis is evaluated and is summarized in Table.2. This is graphically illustrated in Fig.5. This qualitative and quantitative measurement shows that the proposed filtering technique outperforms the other filtering schemes for noise densities up to 60%. Since there is an improvement in PSNR values of all images up to 60% while compare to PSNR values of conventional filters output.

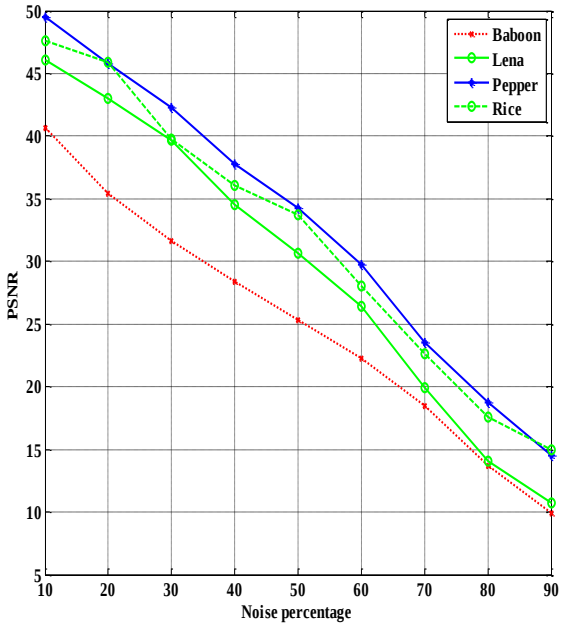

Fig. 5 PSNR obtained by applying proposed filtering technique for different images corrupted with various densities of impulse noise

The qualitative performance of Pepper and Rice images are better than the other images for the noise levels ranging from 10% to 50%. But for higher noise levels, the Pepper image is better. The Baboon image seems to perform poorly for higher noise levels. Based on the intensity level or brightness level of the image, it is concluded that the performance of the images like pepper, Lena, Baboon and Rice will change. Since digital images are nonstationary process. The quantitative performance of Pepper and Rice images are better than the other images for the noise levels ranging from 10% to 60%. The proposed filtering technique is found to have eliminated the impulse noise completely while preserving the image features quite satisfactorily. This novel filter can be used as a powerful tool for efficient removal of impulse noise from digital images without distorting the useful information in the image and gives more pleasant for visual perception.

In image processing, some objective quality criteria are usually used to ascertain the goodness of the results (e.g. the image is good if it possesses a low amount of fuzziness indicating high contrast). The human observer, however, does not perceive these results as good because their judgment is subjective. This distinction between objectivity and subjectivity is the first major problem in the human machine-interaction. Another difficulty is the fact that different people judge the image quality differently. This inter-individual difference is also primarily due to the aforesaid human subjectivity.

VI. CONCLUSION

In this paper, the efficiency of this fuzzy filtering technique is examined and is well suited for digital images when the images are contaminated by impulse noise up to 60%. Since the fuzzy based impulse noise detection mechanism can accurately detect the corrupted pixels on digital image and are replaced with the estimated central noise-free ordered median value from decision based median filter. Based on the expert knowledge and fuzzy if then rules, misclassification is avoided. Simulation results show that the filtering technique has better performance in terms of both quantitative and qualitative measurements.

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Social Networking Sites and Their Security Issues

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Abstract- Social networking sites offer a straightforward way for people to have a simple social presence through web. They provide a virtual environment for people to share each and every activity, their interests, and their circle of acquaintance with their family, friends, or even the unknown. With so much sharing, hackers and thieves have found very easy ways to steal personal information through these networking sites. This calls for advances in security protocols to safeguard against hackers which form the basis of this research. In this paper, we will discuss some of the privacy and security concerns, attacks and their respective prevention techniques. In this paper we propose an architecture for secure request response exchange of data between users. This architecture improves the customization of profiles. Our research suggests that only a proper knowledge of the hacking strategies will prove the best defence in the war against cyber-attacks.

Index Terms- Social Networking sites, security, privacy, secure request-response data exchange , prevention strategies, survey.

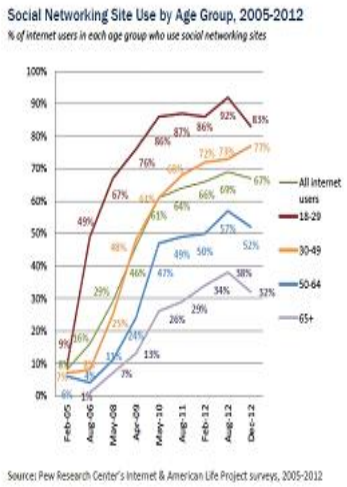
I. INTRODUCTION

Social networks are one of the easiest forms of communication these days. They reflect the social image of a person. They can keep you glued to your *avatar* for hours together and make you forget about the whole physical world around you. The network of social relations that build up during your everyday life can be simply translated onto your "*profile*" and made available for the whole of your friends to see. Then there is a concept of "*following"* that can turn a nomad into a rockstar. The world of pictures you share *live* has only made your presence felt more. It all seems so entertaining that one would seldom think of leaving this "world" and becoming an offline monk. But the more comfortable and attached we become with these sites, the more casual and careless we are to share personal details about ourselves.

People, hundreds of millions of them, use a wide variety of social networking sites (SNSs) that seem no less than a menu card in a restaurant. Facebook, the world's leading social networking site, for example, has more users than the population of many of the countries combined. There is absolutely no doubt that social networks have become a part of every internet user these days and the trend is only set to increase. Figures suggest that there were about 1 billion social network users in 2012, representing a 19.2% increase over 2011 figures.

Table 1
Five biggest social networking sites as of May 2012

Rank	Network	Number of Users (in millions)	Monthly Visits (in millions)
1	facebook	901	7012.9
2	twitter	555	182.1
3	Google+	170	61
4	LinkedIn	150	85.7
5	Pinterest	11.7	104.4



Even though the use of social network web sites and applications is increasingly day by day but users are not aware of the risks associated with uploading sensitive information. The reason why cyber-conspirators prey on these networks is because users upload their personal information that commonly include their interests, social relationships, pictures, confidential information and other media content, and share this information to the whole world via SNSs which are very easily accessible. Employees, too, unknowingly share plethora of personal information on SNS thus putting their corporate infrastructure and data at a risk. The volume and ease of accessibility of personal information available on these sites have attracted malicious people who seek to exploit this information. Due to the sensitivity of information stored within social networking sites,

intensive research in the area of information security has become an area of paramount importance.

Facts reveal that the majority of social media users post risky information online, unaware of the privacy and security concerns. Social networking sites are meant to get as many users in one place as possible on one platform and for attackers there's a lot of return-on-investment in going after them. The values at the core of networking sites – openness, connecting, and sharing with others - unfortunately are the very aspects which allow cyber criminals to use these sites as a weapon for various crimes. Without a careful security policy in place, the entertaining face of social networking could easily compromise on the social stature of an individual.

The dramatic rise in attacks in the last year tell us that social networks and their millions of users have to do a lot more to protect themselves from organized cybercrime, or risk failing to identity theft schemes, scams, and malware attacks. Understanding these risks and challenges should be addressed to avoid potential loss of private and personal information. Social networking definitely needs to be integrated into the information security policy and user education.

II. PRIVACY ISSUES

Security risks

With increasing use of SNSs, the associated security risks are also increasing tremendously. Some of the security risks are identity theft, phishing, scam, cyber bullying etc. People use to provide their personal data on SNSs like facebook, twitter etc. This data is stored in SNS and in lack of proper security techniques implemented in SNSs, It is not secure.

Identity Theft

Some of the attackers attack through the application in which they ask permission for accessing the information provided in the profile of SNS. When a user allows to do so, they get all the information and can misuse that easily without the user knowledge or permissions.

Phishing

Phishing in SNS began in 2007[3]. The purpose of phishing is to harm economically that is the phishers try to retrieve the profile information to know about the banking or the financial information of the users.

Profiling Risk

Profiling risk is the risk associated with profile cloning. The attackers retrieve the personal information of the users and make a clone of the profile [2]. They do so to make their social image bad or for other purposes like knowing about friends of victims. This is the most popular security risk associated with the SNSs because it is very easy to do without the permission of the user. There is nearly no security for profile cloning in SNSs.

There is another way of profile cloning that is "cross-site profile cloning". In this the attacker steals information from one social networking site and uses this information to make a profile on another social networking site.

Fake Product Sale

The attacker advertise on the SNSs for selling the products offering huge discount and when the user clicks on the products advertisement their profile information goes to the attackers. Sometimes when user tries to purchase and give their account information for payment, all the account information is retrieved by the attackers and they misuse this information.

III. ATTACKING SCENARIOS

Conventional Attacking Scenarios

1. CBIR(Content based Image Retrieval)

In this scenario, the attacker can know the location of a user by matching the patterns of the images associated with the profile of the user [1]. Thesetype of attacks are done to know the current location of the user.

2. Click jacking

This is another type of attack scenario in which attacker posts some videos or post to the victim and when victim clicks on the page some malicious actions are performed. This is common in Facebook with the name like jacking that is when a user likes a page, a picture or a video the user is trapped by the attackers[4]. This type of attacks are done to do malicious attack or to make some page popular.

3. Neighbourhood Attack

The neighbourhood attacks are done by the attackers by knowing the victim's neighbourhood[4]. It means the attacker knows the friends of the victim. Attacker uses the relationship among these friends and based on this relationship tries to identify the victim.

B. New attack Strategy

Watering Hole

In January 2013, the attackers used to a new approach to make SNSs user insecure. The attack was done on Facebook. The attackers hacked a mobile developer forum and when developers visited the forum their system got infected with a MAC trajon [5]. This attack was not done to steal profile information or funds, but it was done to infect the system of developers. After attacks on facebook, the same attack was done on many other company, not only on SNS, but on their insecure sites as well.

IV. PREVENTION STRATEGIES

Limit the "amount" - Limit the amount of personal information you post. Do not disclose information such as your residential address or information about your upcoming schedule or your daily routine. Also be considerate when posting information, including photos, videos and other media content.

Internet is always "public" – Always remember that anything that you post on the internet is always available to the public. Thus, it is your responsibility to post information that you are comfortable with anyone seeing. This includes your personal information and photos you post and those in which you are tagged in. Also, once you post information online, you can't delete it. Even if you remove the information from a site, cached

versions remain on the world wide web and also on other people's computers that may be later retrieved as well.

Beware of strangers - The internet makes it really easy for people to misrepresent their personal identities and motives. It is always recommended to limit the people who are allowed to contact you on these sites. If you interact with unknown persons, be cautious about the amount of information you reveal or even agreeing to meet them in person. Common sense should prevail and dominate in such situations no matter how alluring it may appear.

Be sceptical - Don't believe in all that you read online. People make many mistakes and do post false or misleading information about different topics, including their own identity information. This is not necessarily done with a malicious intent since it could be unintentional, an exaggeration of any topic, or simply a joke that one may misinterpret. Take appropriate precautions, though, and make sure you verify the authenticity of any information before taking any action. As said before, common sense should matter more.

Evaluate your settings – Make sure you stay updated with the site's privacy settings. The default settings may allow anyone to see your "profile", but you may have an option to customize your settings to restrict access to only certain people. Sites may change their features periodically, so make sure you review your privacy/security settings regularly to make sure that your choices are still appropriate.

Beware of third-party applications - Third-party applications may provide entertainment or functionality, but use caution and common-sense when deciding which applications can access your personal information. Avoid applications that seem suspicious, and make sure to modify your settings to limit the amount of information which the applications can access.

Use strong passwords - Protect your account with passwords that are hard to be guessed. If your password is compromised, someone else may access your account and pretend to be you or can do virtually anything on your behalf, without your knowledge. Combining capital and lowercase letters with numbers and symbols creates a more secure password. Different password for different accounts always confuses the cyber-criminals.

Keep software, particularly your web browser, up to date - Install the latest software updates so that attackers cannot take advantage of known problems or vulnerabilities. Almost all operating systems and software offer automatic updates. If this option is available, it is always recommendable to enable it.

Use an Anti-virus - Anti-virus software helps protect your computer against known viruses. Since the attackers are continually creating new viruses, it is important to keep your virus definitions up to date. Making sure you have the latest security software, web browser is the best practice against online threats.

Keep an eye on your children - Children are quite susceptible to the threats in social networking sites. Although many of these sites have age restrictions, children are smart enough to misrepresent their ages so that they can join. By teaching children about internet usage, being aware of their online habits, and guiding them to proper and safe sites, parents can make sure that the children become responsible and safe internet users.

Once posted, it cannot be removed: Protect your social reputation on these networks. What you post anything online, it stays online even if you are not able to see it. It is always advisable to think twice before posting pictures you wouldn't want your parents or future employers to see. Recent research found that 70% of job recruiters rejected candidates based on information they found online.

Create an online reputation : A recent research conducted by Microsoft also found that recruiters respond positively to a strong, attractive personal brand online. So show your smartness, thoughtfulness and creativeness to create an impression on your recruiter.

Know and manage your friends : Online friends should not be considered as real friends unless you have met them personally or have spent some time together. Beware of what you share with these "pseudo-friends". If you're trying to create a public image like blogger or expert, create an open profile or a "fan" page that encourages broad participation and also limits personal information. Use a personal profile to keep your real friends more synched up with your daily life.

Be open if you're uncomfortable: If a friend links you to a post and it makes you uncomfortable or you think it is inappropriate, ask them to remove it immediately. Likewise, stay broad-minded and co-operative if a friend asks you to remove something you posted that makes him or her uncomfortable. People have different tolerances and sentimental levels. Respect those differences.

Know what to do : If someone is harassing or threatening you, make sure you use proper measures to remove them from your friends list, block them, or report them to the site administrator using proper channels.

When in doubt, take the safer path : Cyber-criminals compromise your computer by sending links in emails, tweets, posts, and online advertising. If it looks suspicious, it's best to delete or if appropriate, mark as spam and reporting to others as well through proper channels and be a responsible internet citizen.

Other Ways to Secure an Account Typing a username and password into a website isn't the only way to identify yourself on the web services you use.

a) **Multi-factor authentication** uses more than one form of authentication to verify an identity. Some examples are facial recognition, iris recognition, voice ID, and finger scanning.
b) **Two-factor authentication** uses a username and password and another form of identification, often a security code in the form of a "*Captcha",* or likewise.

One of the main reasons why social media has so many loopholes is the trust factor. We think that the people we are dealing with are actually our friends, our colleagues, our favourite sports teams, magazines, or food brands and thus they cannot be "fake" or "criminals". This is the point where the actual criminals take advantage of your trust to retrieve your information.

V. PROPOSED ARCHITECTURE

Secure Request-Response Application Architecture. It is an architecture developed for the secure exchange of data between SNSs users. This architecture allows a user to accept or reject the request of accessing information from his profile. The user can reject the request of friend as well as the visitors. The second functionality of this architecture is that user can have two different databases with different information provided. The user may select data from any one of the two databases to response a particular request. This architecture improves the degree of customization of the profile of a user.

According to this architecture the visitors or friends request for any information to the application between the visitor and the user. The application requests to the user for the response then the user can response from any one of the databases according to his trust on the person who has requested for the information.

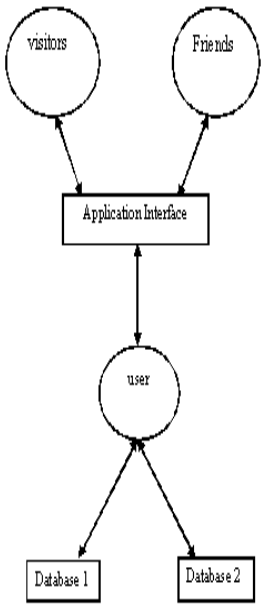


Figure 2 : Secure request response data exchange

Benefits of architecture

The proposed architecture improves the customization of user profile and gives the ability to user to show his profile and information to the others in more customized manner. This will benefit the user to hide his profile information from unwanted visitors and friends.

Limitations

The proposed architecture only adds value to the customization but it is unable to protect from profile cloning. If any one gets information after approval of the user then he may use the information to make a clone of the profile. In this case it is the responsibility of user to provide the information only to trustworthy persons.

VI. CONCLUSION

In the end, the only solution to social network privacy and security issues is to have some knowledge of the ways in which one can get fooled. Don't post anything you would want to hide from a stranger. Be careful who you add as a "friend" since there's simply no way of verifying a user's actual identity online. We have proposed a architecture for secure communication between the users and a secure request-response architecture for exchange of information between the users. Keep your system clean and updated. Keep your senses open while using the internet and never jump to conclusions. Analyse the content thoroughly before doing anything. And remember, there are no free lunches in this world. And, internet is no different.

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India-Afghanistan Strategic Partnership: An Analysis of India, Afghanistan and Pakistan Perspectives

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Abstract- The India-Afghanistan Strategic Partnership signed on October 04, 2011 was the first agreement that Afghanistan has formally entered into with any country to reconstruct Afghanistan, help guarantee its security and is linked with the drawdown of US forces from Afghanistan. Afghanistan views the strong ties with India as a means to assuage the sense of insecurity and capable of assisting the war-torn nation to stabilize the helpless nation on account of the debilitating power struggle anticipated in the power vacuum following the withdrawal in 2014 of the US and NATO forces. Pakistan is concern about the agreement and is quite suspicious of encirclement by India. Perspectives of three countries i.e., India, Afghanistan and Pakistan on India-Afghanistan Strategic Partnership 2011 are attempted to analyze in the present article.

Index Terms- India-Afghanistan Strategic Partnership, Perspective, Pakistani Suspicion, Indo-Afghan Pact

I. India Afghanistan Strategic Partnership Agreement

On October 04, 2011, during the visit of Afghanistan Hamid Karzai to India both countries signed Strategic Partnership agreement "based on mutual understanding and long-term trust" outlines areas of common concern including trade, economic expansion, education, security and politics. This is the first such agreement that post-Taliban Afghanistan has ever formally entered into with any country to help guarantee its security as international troops begin withdrawing from the war-torn country after more than a decade of fighting and is designed to address the challenges of transition as much as prepare ground for preventing the reversal of gains beyond 2014.[14] This agreement is linked with the drawdown of US forces from Afghanistan. It covers security cooperation, trade and economic ties, as well as social and cultural exchanges. The two leaders have also signed two separate deals increasing links in mining and energy which represented "a new dimension in economic relations" to enable Kabul to integrate more effectively with the Indian economy and other economies in South Asia.

The agreements come at a time of severely strained ties between Pakistan and Afghanistan. The people of India sympathized with Afghanistan as it sought to cope with "acts of terrorism, particularly the assassination of [peace envoy] Burhanuddin Rabbani". Karzai has accused Pakistan of supporting militant networks in his country and of having links to the assassination of peace envoy and former president Burhanuddin Rabbani.

This historic document is symbolic of the mutual trust and confidence between the two nations; it has also served to introduce a new twist to the already complex geo-political situation in the region. More specifically, it has added a new dimension to the ongoing turmoil in the somewhat unpredictable relationship between India and Pakistan as also has aggravated tensions between Afghanistan and Pakistan.

II. Afghanistan Perspective

Afghanistan views strategic partnership agreement is to support, to reconstruct Afghanistan and future commitment at a time when other countries are talking of downsizing or even complete withdrawal. During Karzai's visit to India, he told the media that "This strategic partnership is not directed against any country. He also stated that "Pakistan is our twin brother, India is a great friend. The agreement we signed with our friend will not affect our brother." The Government and the common people in Afghanistan view the agreement on strategic partnership with India as a reiteration of the strong traditional bonds between the two nations that have existed for centuries. Afghanistan sees India, a better position to play a more constructive role in rebuilding the shattered economy of Afghanistan, a nation devastated by the prolonged and seemingly interminable conflict than Pakistan. In which Afghanistan views Pakistan not in a position to compete with India in respect of economic cooperation as India for years of economic aid without conditions or in the realm of foreign policy where India is being seen as a mature, seasoned and proactive player on the global scene. Afghanistan has clearly opted to go along with a partner that can effectively contribute to stability and prosperity. In the past decade, India has already donated nearly $2 billion by way of economic assistance. India has accumulated a fund of goodwill amongst the common people through her strong focus on programmes related to the development of infrastructure such as roads, telecommunication facilities and power generation and also invested in education, human resource development as investment in the future leadership and healthcare.

But beyond the lure of funds for reconstruction, the Government of Afghanistan also views the strong ties with India as a means to assuage the sense of insecurity and capable of assisting the war-torn nation to stabilize the helpless nation on

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account of the debilitating power struggle anticipated in the power vacuum following the withdrawal in 2014 of the US and NATO forces.

III. INDIAN PERSPECTIVE

For India the strategic partnership agreement is more than just that as it aims to propel the relationship beyond a mere aid-donor equation to a much higher plane with training of the Afghan National Security Forces and the Afghan National Police included as an important and integral part of the agreement. India sees a strong, independent, stable, prosperous and democratic Afghanistan as being critical to her security interests and for overall stability of the region in the evolving geo-political and geo-strategic scenarios. Building up and sustaining the capability of the Government of Afghanistan through external assistance to provide for her own security is therefore the first and indispensable step in the pursuit of this objective. For India, Afghanistan is undoubtedly a very important partner in the region to project its power status and to secure its entry into energy rich and untapped markets of Central Asia.

At a press conference in Delhi, Mr Singh said that the strategic partnership between the two countries will create an "institutional framework" so that India can help in Afghan "capacity building" in the areas of education, development and people-to-people contacts. But the real motive was to secure its entry into the market of Central Asia and to block Pakistan.

India helps Afghanistan's reconstruction in light of the enduring security competition between India and Pakistan. On the one hand are those who want to expand India's presence in Afghanistan through increased Indian training of Afghan civilian and military personnel, development projects, and expanded economic ties. India's interest in Afghanistan is Pakistan-specific on one side, but more importantly, tied to India's desire to be and to be seen as an extra-regional power moving toward great power status.

IV. PAKISTANI SUSPICIONS

Pakistan is deeply suspicious of Indian influence across its western border. For much of its history, Pakistan has pursued a policy of "strategic depth" in Afghanistan by training, funding, and arming groups that can act as proxies for Pakistani interests. Pakistan is concerned that India's economic and political linkages are building up Indian capacity to destabilize Pakistan. It is widely believed in Pakistan that the Indian consulates in Jalalabad and Kandahar are covertly supporting Balochi insurgents against Pakistan.

Pakistan's concerns that India is trying to encircle it by gaining influence in Afghanistan has in part led to "continued Pakistani ambivalence toward the Taliban," argues a new report by the independent, U.S.-based Pakistan Policy Working Group. The report says Pakistani security officials calculate that the Taliban offers the best chance for countering India's regional influence. Pakistan's support for the Taliban has led to increased instability in Afghanistan, from the growth of terrorism to upped opium cultivation. But Islamabad denies any support for the Taliban and says it is committed to fighting terrorism. U.S. military and intelligence officials have repeatedly warned that Pakistan's tribal areas along the Afghan border continue to serve as safe havens for the Taliban and al-Qaeda to stage attacks against Afghanistan. Experts say Pakistan's cooperation in counterterrorism is vital to winning the war in Afghanistan. Pakistan also competes with India for access to consumer markets in Afghanistan. Pakistan sees Iran's Chabahar port, which India hopes to use as its route for trade with Afghanistan, as a rival that would compete with its new port at Gwadar, which was been built with Chinese assistance.

Despite assurances from President Hamid Karzai that, "the new partnership with India was not meant as a form of aggression towards Pakistan" and readiness on the part of both the signatories to the agreement to accommodate Pakistani interests and address her apprehensions, the overt expression of strategic partnership is bound to cause serious discomfort to Pakistan especially on account of the "strategic" connotation of the agreement. Indian involvement in the training of Afghan National Security and Police Forces, in all likelihood, will be unpalatable to Pakistan. In her perception, the agreement will facilitate direct access to Afghanistan for Indian forces with the possibility of the country being "sandwiched" between two not-so-friendly neighbours. Pakistan sees the move by India as a new "great game" directed against herself and her mentor, China. She also views the agreement as a major impediment to her vision of the establishment of a bloc consisting of Afghanistan, Iran, Pakistan and Turkey duly patronized by China to counterbalance India's rise as a regional power and contain US hegemony. It would not be surprising therefore that in the new situation, Pakistan brands Afghanistan as an enemy equated with India and undertakes a complete review of her foreign policy.

V. INDO-AFGHAN PACT AND PAKISTAN CONCERN

The India and Afghanistan Pact came on the heels of the killing of former President Burhanuddin Rabbani and the subsequent suspension of reconciliation talks with the Taliban, leading many to conclude that it was signed in order to isolate Pakistan. But the agreement was more than five months in the making, designed to address the long-standing demands of the Afghan people. A series of official visits and private deliberations since January of that year culminated in Prime Minister Manmohan Singh's announcement in May of the two countries' plans for a strategic partnership. During an interview in Kabul in the days following the establishment of the pact, former Interior Minister Ali Jalali said he "recognizes the agreement as a document making official the close ties that already exist between the two countries." Shah Mahmood Miakhel, former Deputy Minister of Interior, strongly supported the agreement as "useful for reconstruction and stability of Afghanistan to prevent civil war or proxy war."

Indeed, Sajjad Ashraf, a former Pakistan ambassador to Singapore and now a professor at the Lee Kuan Yew School of Public Policy, cautioned against a knee-jerk Pakistani reaction to the Indo-Afghan treaty. In a paper for the Institute of South Asian Studies, he said that a careful reading of the pact suggests that the countries involved want to develop Afghanistan as a hub linking South and Central Asia since it sits in both regions; which isn't such a bad thing for the countries of south Asia but

especially Pakistan which by its geography as landlocked Afghanistan's neighbor with the longest border has a key role to play. Ashraf said: "If the three countries can reach an understanding and let India develop Afghan capacity leading to regional economic integration, Pakistan too becomes a winner. In the age of globalization, following any other course will result in Pakistan lagging behind.

VI. Future implication of India Afghanistan strategic partnership agreement

The signing of a strategic partnership between India and Afghanistan on October 4, 2011 during Afghan President Hamid Karzai's visit to India was a landmark event. The document is significant for its implications for Indo-Afghan relations as well as for India's wider neighbourhood policy. The agreement positions India and Afghanistan for the post-2014 situation when the international forces are scheduled to withdraw and hand over security responsibilities to Afghan forces. Prime Minister Manmohan Singh was categorical in his support for the Afghan people when he stated at a news conference that "India will stand by the people of Afghanistan as they prepare to assume the responsibility for their governance and security after the withdrawal of international forces in 2014."

Pakistan takes India-Afghan relations as detrimental to its own interests. Its zero-sum attitude to regional cooperation creates many security dilemmas in the region. President Karzai is in an unenviable position. On the one hand, he sees Pakistan as playing a destabilizing "double game" in Afghanistan; and, on the other, he regards Pakistan as a "brother", while describing India as a "great friend". The nuance to be underlined here is that friends always help while brothers can sometimes do great harm. Pakistan is singularly placed to hurt Afghanistan's interest. This is well recognized in Afghanistan where India enjoys warm welcome while Pakistan often comes for stinging criticism. Pakistan, concerned over the India-Afghanistan strategic partnership, is likely to step up pressure on the Afghan government. Pakistan's negative reaction to the Indo-Afghan Strategic partnership was expected.

There is a politico-security component to the strategic partnership but the agreement does not tantamount to a security alliance. The agreement states clearly that the strategic partnership is not directed against "any other state or group of states". India has merely agreed to assist in the "training, equipping and capacity building programmes for Afghan national security forces." Going beyond the security dimension, the partnership arrangement also dwells on trade and economic cooperation, capacity development and education, social cultural & civil society and people to people relations. Significantly, the agreement provides for a high powered implementation mechanism. A "Partnership Council" at the Foreign Ministers' level with four separate joint working groups, on political & security consultations, trade and economic cooperation, capacity development and education, and social cultural & civil society interactions, will be set up. The numerous existing dialogue mechanisms between the two countries will be consolidated and brought under the Partnership Council. The two sides will also have a regular strategic dialogue. The setting up of a Partnership Council will ensure that bilateral relations get sustained attention.

VII. Conclusion

The Strategic Partnership agreement which was signed during the visit of Afghanistan President Hamid Karzai to India on October 04, 2011 was the first agreement that Afghanistan has formally entered into with any country to help guarantee its security and is linked with the drawdown of US forces from Afghanistan. It was meant to reconstruct Afghanistan. But beyond the reconstruction, the Government of Afghanistan also views the strong ties with India as a means to assuage the sense of insecurity and capable of assisting the war-torn nation to stabilize the helpless nation on account of the debilitating power struggle anticipated in the power vacuum following the withdrawal in 2014 of the US and NATO forces. For India the strategic partnership agreement is more than just that as it aims to propel the relationship beyond a mere aid-donor equation to a much higher plane with training of the Afghan National Security Forces. India sees a strong, independent, stable, prosperous and democratic Afghanistan as being critical to her security interests and for overall stability of the region in the evolving geo-political and geo-strategic scenarios. It also includes coping with "acts of terrorism. Even though Karzai made his remark that the agreement with India will not affect Pakistan. Pakistan views India as trying to isolate Pakistan through this agreement and is bound to cause serious discomfort to Pakistan especially on account of the "strategic" connotation of the agreement. Indian involvement in the training of Afghan National Security and Police Forces, in all likelihood, will be unpalatable to Pakistan. It is a new dimension in economic relations to enable Kabul to integrate more effectively with the Indian economy and other economies in South Asia.

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Nested Digital Image Watermarking Technique Using Blowfish Encryption Algorithm

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Abstract- Digital watermarking is referred to a method used for copyright protection and authentication. In this paper, we present a method of nested digital watermark embedding and extraction in which a nested watermark (a watermark inside another watermark) is embedded into the main image. This concept of nested watermarking is used to increase the watermark embedding capacity. In this method, a watermark is embedded into another watermark which is considered as the main watermark and then this main watermark is embedded into the cover image (main image). These watermarks are encrypted before embedding in order to have increased safety, thus to perform encryption and decryption process we used the Blowfish algorithm. Therefore, our research work is focused on increasing the embedding capacity and improving security of the watermarks.

Index Terms- Digital Watermarking, Blowfish, Image Watermarking, copyright protection, Least Significant Bit, Nested Watermark

I. INTRODUCTION

With the widespread use of Internet and increasing rate of development in Information Technology industry, the digital media files (such as images, audio files, documents and videos) are easily accessible and acquired in our daily life from the internet. Due to this, original digital multimedia contents suffer from infringement of their copyrights, easy modification and fast content transfer over the Internet. As a result, data piracy and copyright protection has become a serious issue in order to protect one's ownership rights. Hence, some protection measures are needed to be employed to conquer this major problem that would only be increasing as time progresses. Sensing the need, many new techniques have been developed in recent times and the research is still ongoing for new techniques of information hiding and security. A very important technique that had come into existence for the first time in Italy, in the 13th century was *Digital watermarking*. Digital watermarking is the process of embedding data referred to as a watermark, tag or a label into a multimedia file in such a way that it can be detected or extracted by the owner to make necessary assertions about the illegal modifications of the media that were done without owner's approval. The media file can be any digital file such as image, video, audio or text. In other words, a watermark can be explained as a carrier of information (data). This Information can constitute the copyright details of the file, license, tracking and authorship details of the media etc. The watermark may or may not be visible to the user/viewer of the file.

HISTORY OF WATERMARKS

The term "watermark" was originated from the German term "wassermarke". The name 'watermark' is given because these marks have a resemblance towards the effects of water on paper [21]. The ancestors of Digital watermarks were the paper watermarks, as papers were invented in China over a thousand years ago. However, the first paper watermark was created in 13th century, in Italy [21]. In the 18th century, the watermarks had been used as trademarks to record the manufactured date, or to indicate the size of original sheets. These days, watermarks are commonly used on bills to avoid counterfeiting, documents for copyright protection. Most of the developed countries today use watermarks in their paper currencies and postage stamps to make forgery/infringement more difficult.

The digitisation of our world has expanded our concept of watermarking to include immaterial digital impressions for use in authenticating ownership claims. However, Digital Watermarks have replaced paper watermarks almost completely.

II. EXISTING TECHNIQUES

- A method in which the multimedia content is encrypted with one key and can be decrypted with several other keys, the relative entropy between encrypt and one specific decrypt key was developed by Miroslav Dobsicek [1].
- In 2001, a web based authentication system was developed by Yusuk Lim, Changsheng Xu and David Dagan Feng. In case of watermark embedding, it is installed in the server as an application software that any authorized user (who has the access to the server) can generate watermarked image [2].
- In 2003, a new method which manipulates "flappable" pixels to enforce specific block based relationship in order to embed a significant amount of data without causing noticeable artifacts was developed by Min Wu and Bede Liu [3].
- In 2005, Rehab H. Alwan, Fadhil J. Kadhim, and Ahmad T. Al- Taani explained a method consisting of three steps. First, the edge of the image is detected using Sobel mask filters. Second, the least significant bit LSB of each pixel is used. Finally, a gray level connectivity is applied using a fuzzy approach and the ASCII code is used for information hiding. The given method embeds three images in one image and includes, as a special case of data embedding, information hiding, identifying and authenticating text embedded within the digital images [6].

- In 2006, Harpuneet Kaur, R. S. Salaria proposed a method of nested watermark embedding and extraction for increasing the embedding capacity of watermarks and increasing security of the watermarks by their encryption [4].
- In 2007, a method was proposed by Nameer N. EL-Emam in which data security using LSB insertion steganographic method was introduced. In this approach, high security layers have been proposed through three layers to make it difficult to break through the encryption of the data [5].
- In 2008, an approach to hide huge amount of data using LSB Steganography technique was proposed by Prof S. K. Bandyopadhyay, Debnath Bhattacharyya, Swarnendu Mukherjee, Debashis Ganguly, Poulami Das. They have given much emphasis on space complexity of the data hiding technique [8].
- In 2008 , G. Sahoo and R. K. Tiwari proposed a method that works on more than one image using the concept of file hybridization. This particular method implements the cryptographic technique to embed two information files using Steganography [10].
- In 2012, Preeti Gupta proposed a cryptography based digital watermarking method in which the embedding and extraction of nested watermarks was done. Encryption and decryption of watermarks was done using XOR operation [7].

A) WATERMARK EMBEDDING TECHNIQUES

There are various watermark embedding techniques that have evolved in recent times but we specifically used LSB technique [20] for embedding watermarks into main image.

Least Significant Bit (LSB) Technique

The most straightforward method that can be used for watermark embedding would be to embed the watermark into the least significant bits of the main object (cover object) [20]. Given the extraordinarily high channel capacity of using the entire cover for transmission in this method, a smaller object may be embedded multiple times. Even if most of these are lost due to attacks, a single surviving watermark would be considered a success. An example is given below:

Let us consider:
A= Original Image
B= Watermark
C= Watermarked Image

Embedding Procedure

A: 10000100 00100101 10001000 01010001…
B: 1 0 0 1…
C: 10000101 00100101 10001000 01010001…

It can be inferred from the above example that the watermark's (in this case, B) binary bits are added to the least significant bits of the original image (in this case, A). Least Significant Bit substitution will survive transformations such as cropping effectively, but with any addition of noise or lossy compression it will adversely affect the watermark.

B) ENCRYPTION AND DECRYPTION TECHNIQUES

Blowfish Algorithm

Blowfish is a symmetric encryption algorithm, which makes the use of the same secret key for both encryption and decryption of the message. It is a block cipher which means that it divides the message into blocks of fixed length during encryption or decryption This algorithm is used as an alternative for DES (Data Encryption Standard) or IDEA International Data Encryption Algorithm) [17]. It takes a variable-length key, ranging from 32 bits to 448 bits. Blowfish was designed in 1993 by Bruce Schneier as a fast, free alternative to existing encryption algorithms. Since then, it has been slowly gaining acceptance as a strong encryption algorithm. Blowfish is not patented, is license-free, and is available free for all uses [12].

Blowfish Algorithm is a Feistel Network, in which a simple encryption function is repeated 16 times. The block size is 64 bits, and the key can be of any length up to 448 bits [13]. This algorithm is suitable for applications in which the key does not change very frequently. It is comparatively faster as compared to other encryption algorithms over handling of larger data. A Feistel network is a general method of transforming any function into a permutation. It was invented by Horst Feistel and has been used in many block cipher designs [13].

This algorithm consists of two parts :

- *Key-expansion part:* In Key expansion, a key of at most 448 bits is being converted into several subkey arrays totalling 4168 bytes.
- *Data- encryption part:* In Data encryption, 16-round Feistel network is used. Each round consists of a key dependent permutation, data-dependent substitution and a key. All operations are XORs and additions on 32-bit words. The only additional operations are four indexed array data lookups per round [13].

Blowfish uses a large number of subkeys during its execution. These keys are computed before data encryption or decryption [12].

- The P-array consists of 18 subkeys (32 bit): P1, P2,..., P18.
- There are four S-boxes (32 bit) with 256 entries each:

S1,0, S1,1,..., S1,255;
S2,0, S2,1,..,, S2,255;
S3,0, S3,1,..., S3,255;
S4,0, S4,1,..,, S4,255.

Encryption

Blowfish has 16 rounds. The input is a 64-bit data element, x.

Divide x into two 32-bit halves: xL, xR.
Then, for i = 1 to 16:
xL = xL XOR Pi
xR = F(xL) XOR xR
Swap xL and xR

After the sixteenth round, swap xL and xR again to undo the last swap.

Then, xR = xR XOR P17 and xL = xL XOR P18.

Finally, recombine xL & xR to get the cipher text [12]. Process of encryption is shown in figure 1.

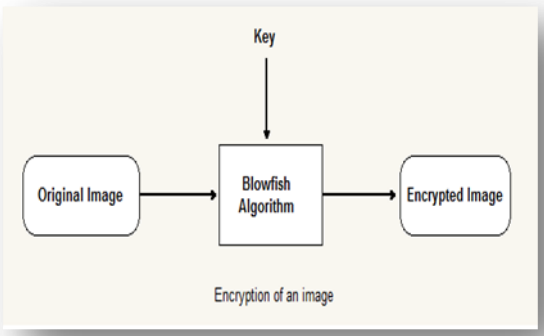

Figure 1: Encryption of image using Blowfish Algorithm

Decryption

The process of decryption using blowfish algorithm is shown in figure 2. In this process, an already encrypted image is decrypted using the same key that was used at the time of encryption. Decryption process is similar to encryption except that in decryption, P1,P2, … P18 are used in reverse order [13].

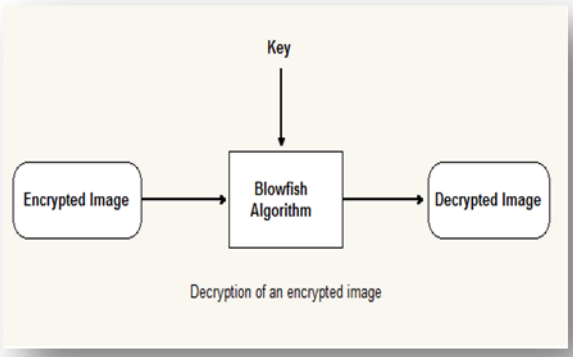

Figure 2: Decryption of image using Blowfish Algorithm

III. PROPOSED METHOD

In this paper, we are proposing a nested digital watermarking technique of watermark embedding and extraction. 'Nested' here basically means a watermark embedded into another watermark. In this method, one watermark is encrypted (using *Blowfish Algorithm*) and embedded into another watermark and this nested watermark (watermark embedded into another watermark) is again encrypted and finally embedded into the main cover image. This concept is shown in figure 3. This method provides an additional level of security for watermarks due to encryption of watermarks. It also increases the embedding capacity of a watermark (due to use of nested watermarks) as compared to the method in which a single watermark is embedded into an image for protecting copyright infringement.

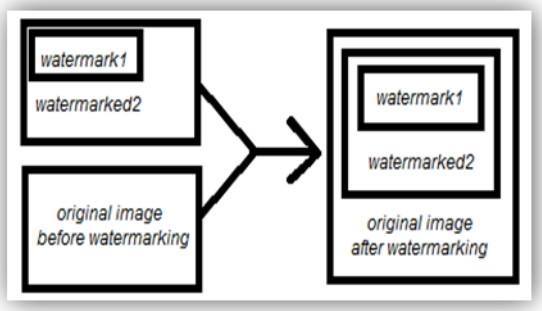

Figure 3: Watermark Embedding Process

For embedding first watermark in second, we used the spatial domain technique (LSB) as it is concise and less time consuming as compared to the frequency or wavelet domain techniques. *Least Significant Bit (LSB)* technique basically inserts the watermark bits to the least significant bits of the main image. By using LSB technique, first watermark is embedded into the second watermark. And then this second watermark is embedded into the main image. This technique is explained above in section II(A).

For encryption and decryption of watermarks, the *Blowfish Algorithm* is used as it is an efficient technique. The *Blowfish Algorithm* is explained above in section II(B). The first watermark is encrypted before embedding it to the second watermark. After first watermark is embedded into the second watermark, the nested watermark (one watermark inside another watermark) is encrypted again and finally embedded into the main image.

Watermark Embedding

Abbreviations that are used in this algorithm:
Watermark1: Primary watermark image.
Watermark2: Secondary watermark image (Main watermark).
Main Image: Original image that is to be watermarked.
Y1: Key for encrypting *Watermark1*.
Y2: Key for encrypting *Watermark2*.
W1: Key for embedding *watermark1*.
W2: Key for embedding *watermark2*.

Embedding Algorithm

1. *Watermark1* is encrypted by using *Blowfish Algorithm* with key *Y1*. Output is called *Encrypted1*.
2. *Encrypted1* is embedded into Watermark2 using key *W1*. Output image is *Watermarked1*.
3. *Watermarked1* is encrypted by using *Blowfish Algorithm* using key *Y2*. Output image is *Encrypted2*.
4. *Encrypted2* is embedded into *Main image* using key *W2*. Output is *Watermarked2*.

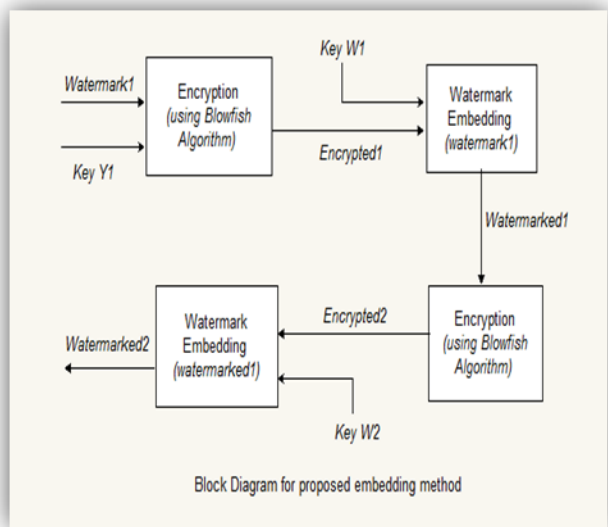

Figure 4: Block Diagram for embedding method

Output

Watermarked Image: *Watermarked2*.

Watermark Extraction

Inputs

Watermarked2': Already watermarked image.
S1: Size of watermark1.
S2: Size of watermark2.
Y2: Key for decrypting recovered watermark from cover Image.
Y1: Key for decrypting recovered watermark from main watermark (secondary watermark).
W2: Key used for recovering encrypted watermarked watermark from Main Image (cover image).
W1: Key for recovering encrypted watermark from the main watermark.

Extraction Algorithm

1) Extract encrypted *Watermark2* from *Watermarked2* using key *W2*. Output is *Encrypted2'* (recovered image).
2) Decrypt *Encrypted2'* using *Blowfish Algorithm* with key *Y2*. Output is *Recovered2*.
3) Extract encrypted *Watermark1* from *Recovered2* using key *W1*. Recovered image is called *Encrypted1'*.
4) Decrypt *Encrypted1'* using *Blowfish Algorithm* with key *Y1*. Output is *Recovered1*.

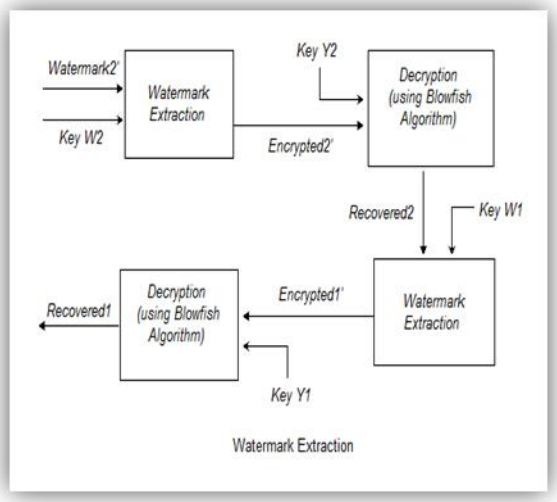

Figure 5: Block Diagram of Watermark extraction process

Output

Recovered2: Main watermark recovered from *Watermarked2* (Already watermarked image)

Recovered1: Watermark recovered from the main watermark.

IV. EXPERIMENTAL RESULTS

In our experiment, we used 450x300 sized images for executing our method of watermarking.

Simulation Tool Used: Matlab R2012b

The above proposed method was executed using Matlab R2012b software. Matlab is a registered product of Mathworks Inc, Copyright 1984-2012.

Figure 6: Watermark1

Figure 7: Watermarked watermark (watermark2)

Figure 8: Image before Watermarking (450 x 300)

Figure 9: Image after watermarking (*Watermarked2*)

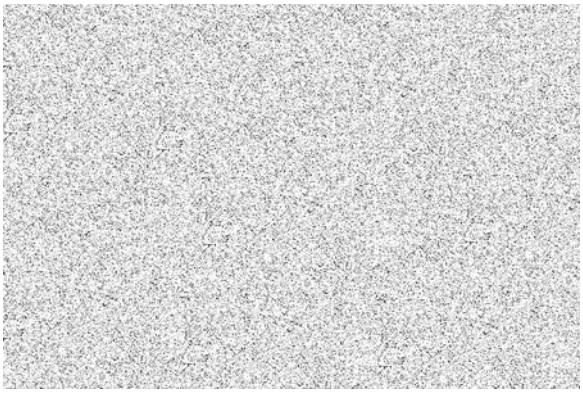

Figure 10: Difference of images (original and watermarked)

We used a 450x300 gray-scale image and performed this method. First, we took watermark1 as shown in figure 6. This watermark1 is encrypted and then this encrypted watermark is embedded into another watermark (here referred as watermarked watermark) as shown in figure 7. Now, this watermarked watermark is embedded into the main image. Figure 8 shows the original image before watermarking and figure 9 shows the image after watermarking. The difference between original and the watermarked image is shown above in figure 10.

V. CONCLUSION

In this paper, a new technique of digital watermarking is proposed in which a watermark is encrypted and embedded into another watermark and this combined watermark is embedded into the main image. This phenomenon of embedding one watermark into another is known as Nested watermarking. By doing so, the level of security of the watermark increases (due to use of encryption and decryption techniques) and the embedding capacity of the watermark is also enhanced (as concept of nested watermarks is used). The Blowfish encryption and decryption algorithm is used in this method as it is suitable and efficient for hardware implementation. Besides, it is unpatented and no license is required (Open source algorithm).

Advantages of this proposed method:

- Concept of Nesting increases embedding capacity of watermark into the main image.
- Encryption of watermarks before embedding them into main image helps to increase the security of the watermark.
- Use of Blowfish algorithm helps to make the method more robust.

FUTURE SCOPE

The concept of digital watermarking is not very matured. A lot of research is going on for improving the existing watermarking techniques. In this field, the highest priority is given to the work that is initiated towards getting information from attacked watermarks. The watermarking technique proposed in this paper can be improvised in future in terms of enhancing security and embedding capacity of watermarks. In this paper, we have used Blowfish Algorithm for encryption and decryption and LSB method for embedding of watermarks. So in future, some other algorithms can be used or proposed for encryption or decryption and embedding of watermarks.

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Delineation of Target Zones for Detailed Uranium Exploration Using Secondary Dispersion Haloes (Pedogeochemical) as a Tool in the NNW Parts of Srisailam Sub Basin, Andhra Pradesh (India)

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Abstract- Pedogeochemical sampling over an area of 45 Sq.Km was initially taken up in the Chitrial outlier, which is actually detached from the main Srisailam sub basin as a result of topographic low formed through weathering and erosional processes. Pedogeochemical sampling is carried out on a grid of 1Km x 1Km and various geochemically associated trace and ore elements viz., U,Th,V,Cr,Co,Ni,Cu,Pb,Zn and As are analyzed and studied for their dispersion patterns. The pedogeochemical dispersion pattern of trace elements studied in conjunction with geology and structure has yielded an area of 11 Sq.Km which is identified with specific trends (N-S and NNE-SSW), as a potential zone for uranium mineralization and is recommended for detailed and subsurface investigations.

Index Terms- Dispersion haloe, pedogeochemical, Srisailam sub basin, uranium.

I. INTRODUCTION

The mid-Proterozoic – Archean basement unconformity is proved all over the world as the most potential zone hosting some of the world's largest uranium deposits. The Srisailam sub basin is geologically mid to late Proterozoic in age and it is unconformably overlying the Archean granite basement. The sediments are essentially arenaceous and are represented by a basal pebbly quartzite, followed by a grey medium grained quartzite with grey to buff coloured shale as intercalations. The basement is represented by a coarse to medium grained grey fractured granite. Srisailam sub basin and cuddapah basin geology with known uranium mineralization.

Unconformity-related Uranium mineralization

With the discoveries of high grade mineralisation (~ 0.5% U3O8) of the unconformity-type in the Athabasca basin of Canada and the Pine creek geosynclines of Australia, (Hoeve , J. et al., 1980) emphasis on exploration was shifted in early 1990's to locate fracture controlled unconformity related uranium mineralisation in Cuddapah basin. Workable deposits of unconformity-related type have been established at Lambapur-Peddagattu, Nalgonda district and Koppunuru, Guntur district, Andhra Pradesh, along NW margin of the Cuddapah basin.

Fracture/Shear controlled uranium mineralisation

Fracture controlled uranium mineralisation is both basement granite hosted as well as sediment hosted (Gulcheru Formation, the oldest member of the Cuddapah Super Group) and occurs along the southern margin of the Cuddapah basin.

Strata bound Uranium mineralisation

The strata bound uranium-mineralisation in southwestern part of the Cuddapah basin is unique in the sense that no such strata bound uranium deposit hosted by carbonate rocks is reported in the world. Uranium mineralisation is hosted by impure phosphatic dolostone of the Vempalle Formation of the Papaghni Group. It extends from Chelumpalli in the northwest to Maddimadugu in the southeast over a belt of 160 km with promising mineralisation at Tummalapalle, Rachak-untapalli and Gadankipalli in the central part.

Significance of unconformity uranium deposits

An unconformity is a buried erosion surface separating two rock masses or strata of different ages, indicating that sediment deposition was not continuous. In general, the older layer was exposed to erosion for an interval of time before deposition of the younger, but the term is used to describe any break in the sedimentary geologic record.

Unconformity-type uranium (Jefferson.C., et al., 2005) deposits host high grades relative to other uranium deposits and include some of the largest and richest deposits known. They occur in close proximity to major unconformities between relatively quartz-rich sandstones comprising the basal portion of relatively undeformed sedimentary basins and deformed metamorphic basement rocks. These sedimentary basins are typically of Proterozoic age, however some Phanerozoic examples exist. The most significant areas for this style of deposit are currently the Athabasca Basin in Saskatchewan, Canada, and the McArthur Basin in the Northern Territory, Australia.

The Proterozoic basins all over the world have gained importance in the light of high-grade unconformity related uranium deposits associated with them. The depositional conditions for that geologic time favored the chemical mobility of the uranium from their primary source to the sediments/host rocks and got deposited as different type of uranium deposits specific to their type and nature. In particular, the unconformity surface between the Proterozoic sediments and the Archean

basement crystalline acted as the most favourable geochemical front/ surface with host of other favourable criteria like basement faults and folds, along with the clay horizons, reductants and alteration zones associated with the rocks proximal to the unconformity surface. The hydrothermal solution activity along with structural disturbances over the geologic times would have reactivated, remobilized and concentrated uranium along these surfaces.

Proterozoic unconformity and their environs are established potential target areas of medium to high grade, large tonnage, low cost uranium deposits in parts of Canada (Cigar lake) and Australia (Athabasca basin) and more than 50% of present day world production is recovered from such geological settings.

The classical geological setting of unconformity type uranium deposits is "the contact between Meso to Neo-Proterozoic sediments and Palaeo-Proterozoic metasediments". Such deposits are mainly located at or near the unconformity surface, (Senthil kumar, P., et al 2002) as it provides an effective plumbing system for circulation of hydrothermal solutions. Precipitation due to intermixing of highly charged metal rich oxidizing solutions from above and reducing solutions from below the unconformity and repeated recycling/ remobilization are the conspicuous features of such deposits imparting them unusually high grade.

Indian Scenario of Unconformity related uranium deposits: In India, first unconformity deposit was established in the intra-cratonic Proterozoic Cuddapah basin at Lambapur- Peddagattu, Nalgonda district (Sinha, 1995), and later at Koppunuru, Guntur district (Jeyagopal, 1996) at the proximity of unconformity contact of Meso-Proterozoic Srisailam quartzite/ Neo-Proterozoic Banganapalle quartzite and the basement granite respectively.

II. Geology and Structure

The Srisailam sub basin is located in the north eastern part of the Cuddapah basin. The Chitrial, area is located in the north eastern fringes of Srisailam sub basin (fig no.1). Chitrial outlier is horseshoe shaped and is located to the north of Dindi River which is flowing from northwest to southeast along lineament and merges with Nagarjunasagar reservoir.

The study area (45 Sq.km) falls around the known / established uranium mineralization at Chitrial. The area is located in the northern periphery of the Srisailam sub-basin. The area represents small Outlier of Quartzite (Srisailam formation). The area covers both the quartzite outliers and the basement granite, dolerite dykes, shale, clays are presented along the unconformity of the Chitrial outlier (fig no. 2), considering the importance of both the source and host for uranium mineralization. The fertile nature of the granite basement is a pre-requisite for unconformity type of uranium mineralization in Proterozoic basins.

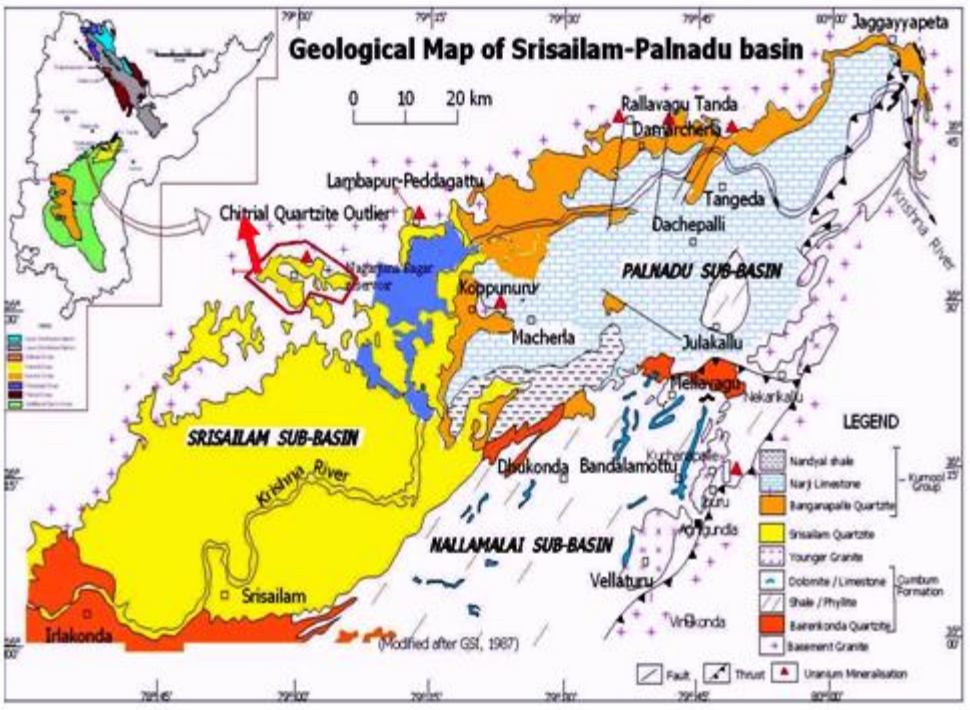

Figure 1. Location map of Chitrial outlier in Srisailam Sub basin (AMD)

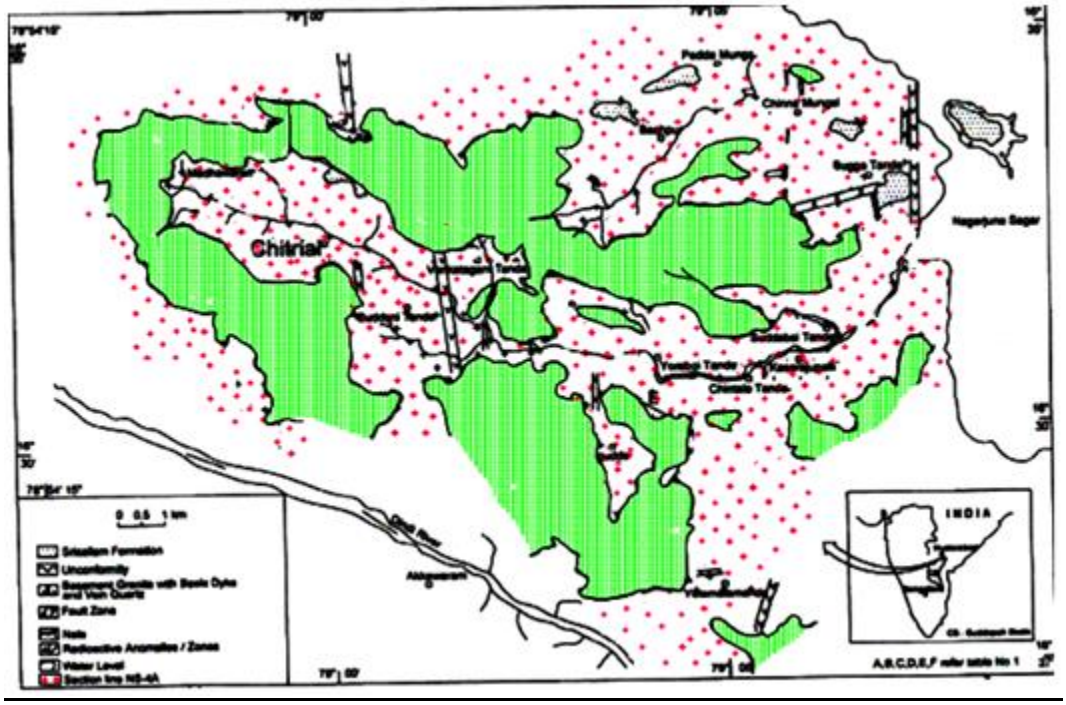

Figure 2. Geological map of Chitrial outlier (AMD, 2009)

Structure: A number of basement-dykes (dolerites) trending N-S and NNW-SSE and a minor trend represented by E -W are observed. The mineralization trend as established in the Chitrial area is also more or less in the N-S and NNW-SSE direction (fig 2). The Chitrial area has been dissected by NW-SE; NE-SW; and NS lineaments. The main drainage in the area Viz. Dindi and Krishna Rivers flow along WNW-ESE and N-S lineaments respectively. The joints in the basement show two distinct directions along NW-SE and NE-SW, with less developed directions of N-S and E-W.

Table-1 - Geological succession in the Cuddapah Basin, after King {1872}

	Kundair Group
	Paniam Group
KURNOOL SYSTEM	**Jamalamadugu Group**
	Banganapalli Group
	----------*Unconformity*----------
Kistna Group	**Srisailam Quartzite**
	Kolamnala Slate
	Iriakonda Quartzite
	---------- *Unconformity*----------
Nallamalai Group	**Cumbum Slate**
	Bairenkonda Quartzite
CUDDAPAH SYSTEM	----------*Unconformity*----------
Cheyair Group	**Tadpatri Shale**
	Pulivendla Quartzite
	Nagari Quartzite
Papaghni Group	**Vempalle Slate**
	Gulcheru Quartzite
	---------- *Unconformity*----------
	Older Precambrian

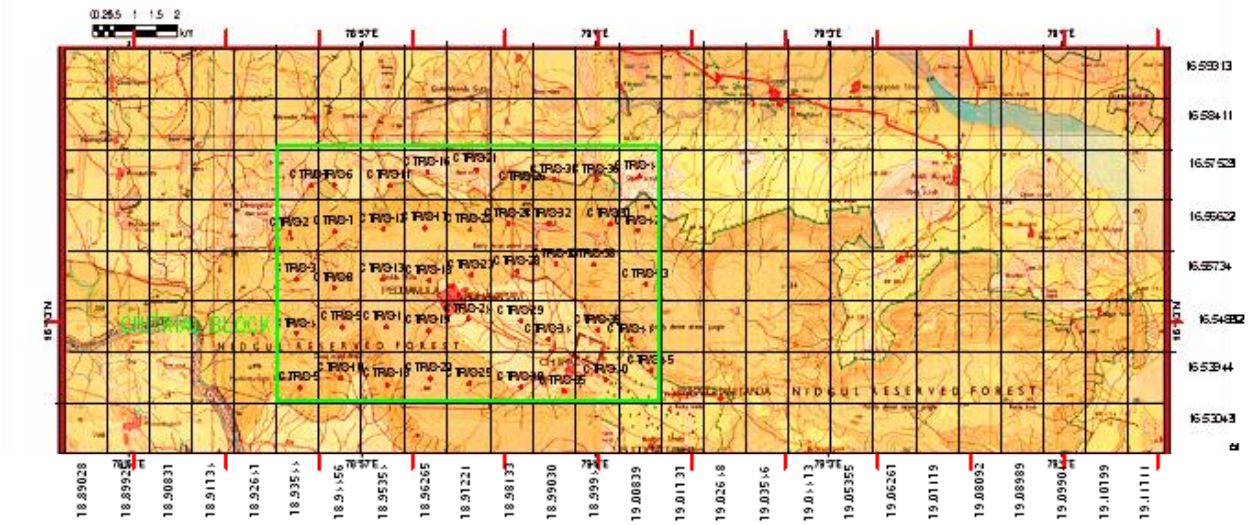

Figure3. Pedogeochemical sampling grid in Chitrial

III. PEDOGEOCHEMICAL SAMPLING

Pedogeochemical sampling is carried out on a grid of 1Km x 1Km (fig no.3) and various geochemically associated trace and ore elements viz., U,Th,V,Cr,Co,Ni,Cu,Pb,Zn and As are analyzed and studied for their dispersion patterns (Chaudary M.A. et al., 2002). The samples were sun dried and sieved to -80 mesh size for the chemical analysis in the field itself to save the time and transport cost. Before the analysis the samples were homogenized with quartering & conning, then take one gram of each of the soil samples were subjected to the cold acid partial digestion technique with 1:1 HCl to know the concentration of the transported mobile trace elements which help in knowing the adsorbed material around the soil particles into soil solution for analysis. The dispersion pattern of the trace elements w.r.t uranium mineralization. The solutions thus obtained were subjected to analysis on ICPMS at NGRI. (Balaram, V. et al., 2003).

Table no.2 Pedogeochemical Analysis results of Chitrial

longitude	latitude	sample no	V	Cr	Co	Ni	Cu	Zn	As	Pb	Th	U
78.93365	16.57418	CTR/S-1	92.15	24.88	22.02	23.48	20.51	57.29	0.00	667.43	0.00	18.84
78.935056	16.56575	CTR/S-2	118.07	24.88	33.64	21.70	21.14	51.69	0.00	35.22	6.96	22.85
78.934940	16.557656	CTR/S-3	51.78	17.55	16.00	14.72	10.20	27.34	0.00	15.22	4.68	11.85
78.935848	16.548076	CTR/S-4	107.85	31.79	27.67	37.27	21.41	84.28	0.00	40.61	54.50	25.97
78.943162	16.538358	CTR/S-5	140.93	43.02	23.42	26.18	25.67	51.15	0.00	31.80	10.31	5.92
78.943302	16.574324	CTR/S-6	54.70	19.90	18.59	17.82	11.77	60.75	0.00	30.33	4.18	3.79
78.943162	16.566116	CTR/S-7	95.45	29.57	23.66	25.21	20.01	77.06	0.00	631.42	12.38	10.67
78.944956	16.556147	CTR/S-8	41.51	11.98	16.47	10.63	7.64	28.30	0.00	21.86	7.46	13.56
78.944583	16.549082	CTR/S-9	85.08	30.62	25.67	20.09	61.73	94.36	0.00	569.11	11.76	4.43
78.955112	16.540142	CTR/S-10	77.88	20.08	28.32	27.83	18.06	65.62	0.00	33.94	5.41	14.32
78.953691	16.574187	CTR/S-11	523.98	33.89	69.61	48.88	111.25	31.03	0.00	3.67	1.19	1.41
78.953574	16.566505	CTR/S-12	84.07	29.25	26.06	35.00	13.36	41.59	0.00	15.78	1.62	5.34
78.954343	16.557656	CTR/S-13	108.41	30.60	30.88	20.46	15.65	9.06	0.00	16.64	3.93	1.43
78.954599	16.549219	CTR/S-14	60.20	22.59	13.73	16.93	7.00	9.08	0.00	14.17	0.71	0.72
78.963218	16.538999	CTR/S-15	121.96	31.07	31.82	29.41	23.72	114.99	0.00	40.22	26.63	8.05
78.962705	16.576474	CTR/S-16	65.06	21.19	6.77	12.68	9.22	16.32	0.00	8.25	0.83	0.08
78.963590	16.566619	CTR/S-17	59.63	15.56	20.97	23.02	12.30	28.89	0.00	14.41	3.54	1.14
78.963474	16.557405	CTR/S-18	71.60	20.12	17.73	26.45	12.55	27.70	0.00	12.48	2.93	0.90
78.963730	16.548465	CTR/S-19	70.99	28.12	16.04	21.00	8.69	16.14	0.00	13.00	0.89	1.13
78.973746	16.540005	CTR/S-20	72.90	29.03	26.49	24.55	10.98	64.85	0.00	671.77	14.48	9.37
78.972325	16.576977	CTR/S-21	65.73	22.67	20.69	20.47	13.18	27.78	0.00	32.25	0.96	0.84
78.972605	16.566368	CTR/S-22	135.83	37.66	32.93	52.62	25.11	21.05	0.00	16.63	1.55	0.56
78.971952	16.558296	CTR/S-23	95.05	21.31	28.05	35.66	24.66	42.58	0.00	31.79	3.34	6.73
78.972209	16.550591	CTR/S-24	118.52	25.43	35.56	34.41	26.82	143.31	0.00	705.51	39.04	3.75
78.983785	16.539113	CTR/S-25	46.70	17.49	21.51	16.47	13.37	22.68	0.00	10.55	2.02	0.15
78.980571	16.573936	CTR/S-26	112.31	30.11	41.46	24.42	22.31	21.21	0.00	18.07	3.10	0.20
78.982481	16.567374	CTR/S-27	86.93	20.93	29.92	26.83	12.29	46.65	0.00	34.88	18.07	3.08
78.983250	16.558937	CTR/S-28	46.63	13.50	19.07	32.56	7.00	37.20	0.00	18.37	3.73	2.04
78.983250	16.550339	CTR/S-29	51.19	18.25	27.47	17.66	12.39	28.30	1.98	12.64	0.79	0.88
78.990447	16.538495	CTR/S-30	77.78	24.78	37.94	31.66	15.75	25.21	2.26	18.38	1.26	3.69
78.989166	16.575193	CTR/S-31	67.44	15.21	18.75	184.27	12.74	72.17	2.10	42.95	10.36	29.85
78.990843	16.567511	CTR/S-32	97.54	35.80	16.89	21.75	15.18	15.12	2.16	13.02	0.41	1.49
78.989166	16.560194	CTR/S-33	78.43	23.23	30.75	32.80	18.85	16.41	2.23	12.93	1.35	1.55
78.992520	16.546933	CTR/S-34	75.10	20.48	23.57	33.26	19.84	31.16	2.03	18.83	0.77	3.47
78.999578	16.537855	CTR/S-35	50.11	18.61	21.27	11.77	9.72	8.11	1.51	8.95	0.88	0.88
79.002420	16.575193	CTR/S-36	132.86	48.67	52.62	41.23	30.82	65.27	1.89	43.89	11.43	8.05
78.998809	16.567374	CTR/S-37	54.96	19.47	28.31	18.61	15.70	43.60	1.45	68.00	25.71	19.46
78.999834	16.560194	CTR/S-38	65.96	36.14	20.50	39.58	15.67	23.14	1.15	17.64	3.31	2.17
79.001628	16.548579	CTR/S-39	99.34	39.97	32.30	32.02	26.39	15.65	1.24	24.06	5.91	3.16
79.008709	16.539890	CTR/S-40	35.03	17.53	25.88	39.06	14.35	28.53	1.40	19.64	3.13	18.19
79.008453	16.575834	CTR/S-41	52.69	22.80	26.38	18.86	18.81	49.84	0.99	45.41	12.04	7.70
79.010130	16.566253	CTR/S-42	64.74	26.77	35.23	29.73	26.64	92.79	1.13	61.04	9.22	36.70
79.006915	16.556787	CTR/S-43	99.27	40.14	24.22	22.94	18.26	12.26	1.06	26.26	7.08	3.58
79.011271	16.546933	CTR/S-44	113.25	36.62	60.29	50.08	36.07	30.08	0.95	34.57	5.75	2.73
79.011271	16.541514	CTR/S-45	93.81	36.95	37.46	50.58	29.84	80.23	1.09	49.00	7.77	8.11
		Avg	91.59	26.36	27.66	31.17	20.55	43.51	0.59	94.95	7.85	7.35
		St.D	0.98	0.67	0.59	0.46	0.69	0.71	0.00	13.62	0.00	2.32
		Threshold	93.55	27.71	28.83	127.17	116.55	44.93	0.59	122.19	7.85	12.00

IV. CORRELATION OF PEDOGEOCHEMICAL PARAMETERS

Correlation coefficient is commonly used measure to establish the relation between independent and dependent variable. The correlation matrix for 10 elements for Chitrial area in Srisailam sub basin.

Table no .3. Correlation matrix of 10 elements of study area.

	V	Cr	Co	Ni	Cu	Zn	As	Pb	U	Th
V	1.000									
Cr	0.434	1.000								
Co	**0.769**	0.477	1.000							
Ni	0.134	-0.010	0.116	1.000						
Cu	**0.895**	0.409	**0.739**	0.107	1.000					
Zn	0.085	0.129	0.192	0.214	0.212	1.000				
As	-0.146	0.042	0.103	0.321	-0.097	-0.224	1.000			
Pb	-0.004	0.105	0.018	0.055	0.150	**0.611**	-0.234	1.000		
U	-0.094	-0.151	-0.022	0.451	-0.068	0.413	0.067	0.168	1.000	
Th	0.035	0.150	0.154	0.090	0.061	**0.728**	-0.158	0.312	0.453	1.000

According to Karl Pearson from above table explained the Co &V, Cu & V, Cu & Co, Th & Zn shows best correlation between them. Pb & Zn shows good correlation. The remaining elements shows positive and some are related to negative correlationship.

V. RESULTS AND DISCUSSIONS

The secondary dispersion of geochemically associated elements are i.e. U, Th, V, Cr and Ni shows NE – SW trend, which is also reflected extent in the secondary environment. The secondary dispersion haloe, these elements are Cu, Pb, Co, Zn and As which are can be related to the possible hydrothermal sulphide mineralization is observed in primary environment and more dispersed with a distinct of trend to NNE – SSW in secondary environment. The mineralization trends are explained with concentration maps below. Elements like U, Th, V, Cr, Co, Ni, Cu, Pb, Zn, & As, are expressed in ppm levels. The main theme of this study area to find out the mineralization in the unconformity related uranium deposits. The maximum range of uranium is 1 ppm in soils, the study area consisting of U concentration (fig 4) is maximum 29.85 ppm and minimum is 0.081 ppm. The average value of all samples 7.18 ppm; the threshold value is 12 ppm. Thorium concentration (fig 5) containing the highest value is 54.50 and lowest value is 0 ppm average value is 8.71 ppm, the threshold value is 7.85 ppm, naturally the average abundance of thorium is 13 ppm in soils, and some samples have high values compare with desirable limit. Pb concentration (fig 11) consisting of maximum value is 705.51 ppm, minimum value is 3.6 ppm the average value is 113.46 ppm, the threshold value is 122.19, the common abundance value is 2-200 ppm in soils, 60% of soil samples showing above permissible limit. Arsenic concentration (fig 13) containing maximum value is 2.26 ppm, minimum value is nil, the average value is 0.59 ppm, the threshold value is 0.59 ppm, the permissible limit is 1-50 ppm in soils, no sample exceeds the limit. Zn concentration (fig 12) consisting of 143.30 ppm at maximum, 8.10 ppm is minimum and the average is 43.90 ppm, the threshold value is 44.93 ppm, natural abundance of Zn is present in 10-300 ppm in soils. Cu concentration (fig 10) having highest value is 111.24 ppm, the minimum value is 6.99 ppm and the average is 21.74 ppm, the threshold value is 116.55 ppm, the permissible limit is 2-100 ppm in soils, one sample is exceeds the limit. Ni concentration (fig 9) maximum value is 184.26 ppm, minimum is 10.62 ppm the average value is 33.93, the threshold value is 127.17 ppm, Co concentration (fig 8) maximum value is 69.60 ppm, the minimum value is 6.77 ppm, the average value is 27.08, the threshold is 28.83 ppm, the natural abundance of Co is 1-40 ppm in soils, and some samples exceed the permissible limits. Vanadium concentration (fig no. 6) maximum value is 523.97 ppm, minimum is 35.02 ppm, the threshold is 93.55 ppm,

the permissible limit is 20-500 ppm in soils, and some samples exceed the limit. Cr concentration (fig 7) maximum value is 48.66 ppm; minimum value is 11.98 ppm the average value is 28.79 ppm, the threshold value is 27.71 ppm, the permissible limit is 5-1000 ppm in soils, no one sample exceed the permissible limit.

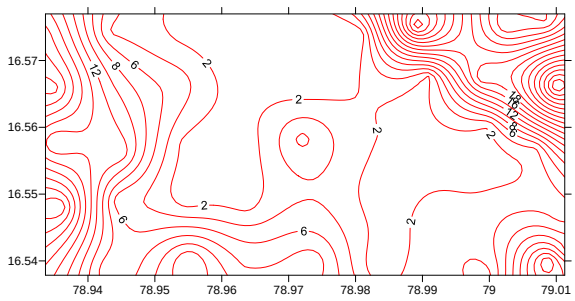

Fig no 4. U – Concentration

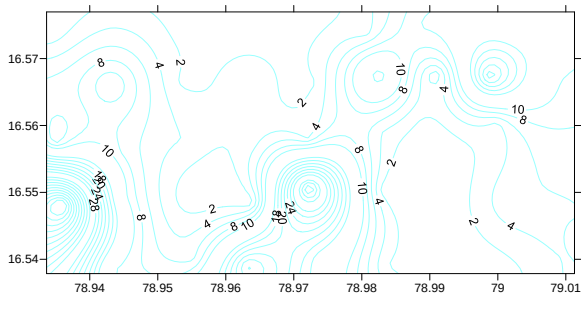

Fig no 5. Th – Concentration

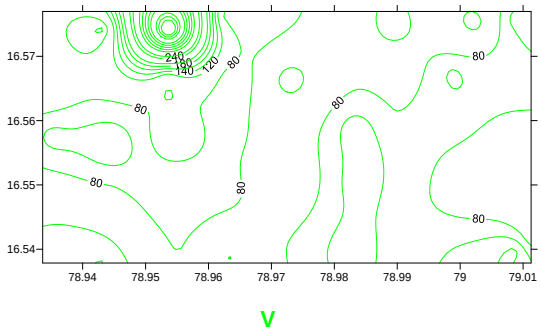

Fig no 6. V – Concentration

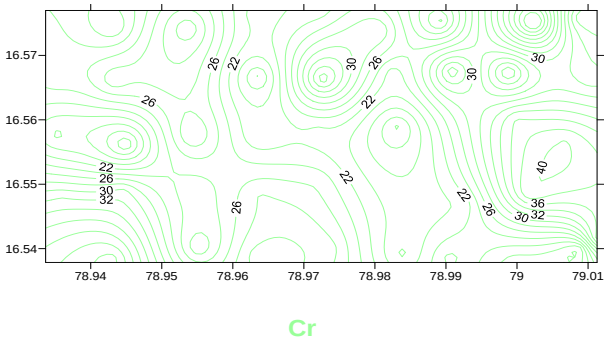

Fig no 7. Cr – Concentration

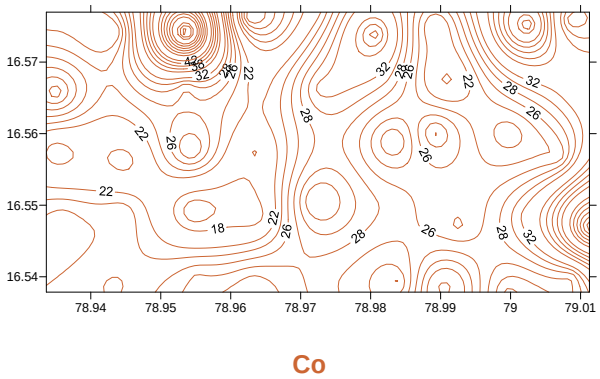

Fig no 8. Co – Concentration

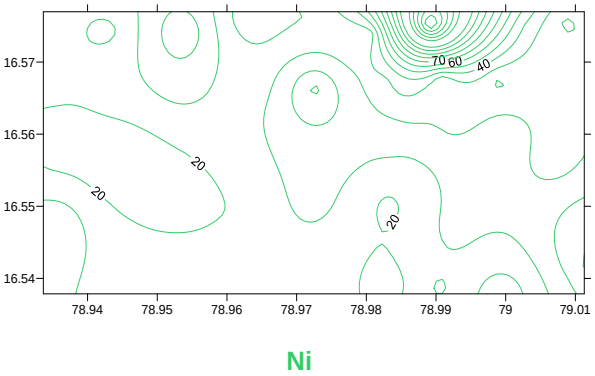

Fig no 9. Ni– Concentration

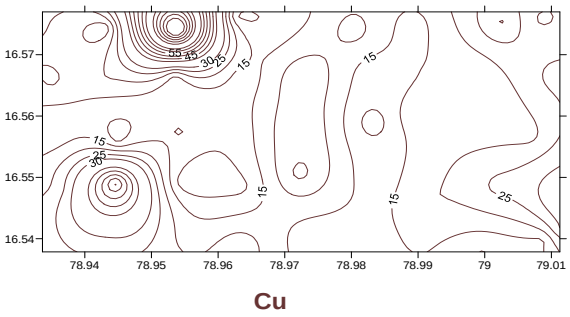

Fig no 10. Cu – Concentration

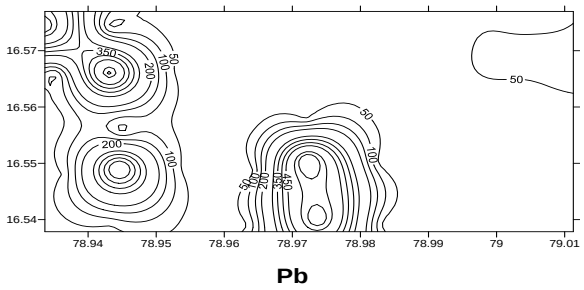

Fig no 11.Pb – Concentration

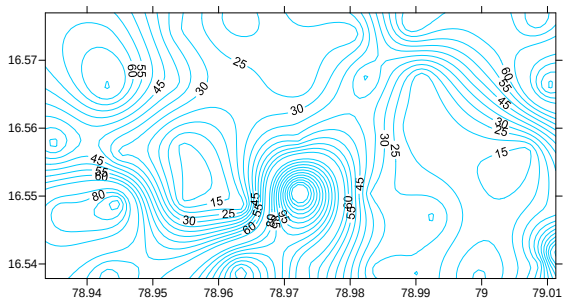

Fig no 12. Zn – Concentration

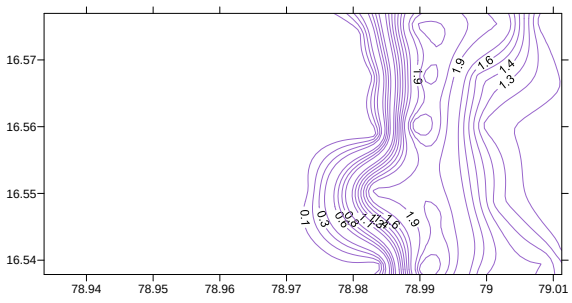

Fig no 13. As – Concentration

VI. Conclusion

Significance of sub surface mineralization located at the unconformity contact of basement granitiod with Srisailam sub basin, around the Chitrial, Nalgonda District, and Andhra Pradesh led to establishment of sizeable uranium mineralization. Thus intensive sub surface investigation resulted in delineating the persistency of mineralization over the 11 Km. An attempt has been made to synthesize the exploratory history commencing from locating uranium occurrences up to developing them into workable economic deposit in Chitrial. Based on the geology, structure and pedogeochemical (secondary dispersion haloes) results, the Chitrial area has been recommended for detailed and subsurface investigations and delineate of target zone for Uranium mineralization (11 Sq.Km) which is identified with specific trends (N-S and NNE-SSW), as a potential zone.

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SENSTIVITY AND SPECIFICITY OF MRI VERSUS ARTHROSCOPY IN INTERNAL DERANGEMENT OF KNEE

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***Abstract*- Introduction:** Traumatic knee injuries were conventionally assessed clinically, and subjected to radiographs. In modern Orthopaedics, MRI and a subsequent arthroscopic procedure has come of age. Arthroscopy of the knee is a key hole surgery performed as a daycare procedure. There are many indications however, this study is to correlate the findings subsequent to an MRI diagnosis.

Material and methods: 40 patients of different age group and sex with internal derangement of knee (IDK) were subjected to MRI for determining ACL,PCL, medial and lateral meniscus injury.and they were subjected to arthroscopy and the findings were recorded.

Results: Magnetic resonance imaging performed on the injured knees revealed 30 tears of the menisci, of which 20 were of medial meniscus and 10 tears of lateral meniscus. There were 21 tears of the cruciate ligaments, out of which 15 were tears of ACL and 6 of PCL.

50 menisci on MRI were normal, of which 20 were medial meniscus and 30 lateral meniscus. Of the 59 cruciates showing normal signal on MRI, 25 were of ACL and 34 of PCL.

Arthroscopy revealed 27 tears of menisci, 19 of these were of medial meniscus and 8 of lateral meniscus. 53 menisci were normal on arthroscopy, 21 of medial and 32 of lateral meniscus. There were 19 tears of cruciates, of which 14 were of ACL and 5 of PCL. There were 61 normal cruciates, of which 26 were ACL and 35 were PCL

Discussion: The area under the curve of magnetic resonance imaging was 87.6% percent for the medial meniscus, 89.1 percent for the lateral meniscus, 76.1 percent for ACL, and 87.1 percent for PCL in this series of 40 patients. Various intra articular pathologies of the knee such as loose bodies, chondral fractures, degenerative changes, plicae can mimic a meniscal tear. Mucoid and eosinophilic degeneration can cause false positive results on MRI [15]. Magnetic resonance imaging is useful as a diagnostic tool in internal derangements of the knee. It is useful in circumstances where there is a need for detailed differential diagnosis. Magnetic resonance imaging also helps the surgeon to plan the definitive management of a tear during the same session.

Conclusion: Magnetic resonance imaging is useful as a pre operative diagnostic tool in selected cases where a clinical examination cannot be performed as in acute injuries or in cases where clinical examination is inconclusive.The efficacy of MRI in diagnosing a tear varies among different intra articular structures.

***Index Terms* -** MRI, IDK, ACL, PCL.

I. INTRODUCTION

The knee joint is one of the most commonly injured joints, as an isolated injury or a frequent component in a multiple trauma patient. The knee is a complex joint, consisting of two condylar joints between the corresponding condyles of femur and tibia and a sellar joint between the patella and femur[1].

The principal intraarticular structures in knee are the two menisci, the two cruciate ligaments, and the two collateral ligaments. The menisci serve to distribute joint fluid, cartilage nutrition, mechanical shock absorption, increasing the surface area of the joint and therefore the stresses, serve to stabilize the joint, and a weight bearing function. The cruciate ligaments function as stabilizers of the knee in both forward and backward motions of the tibia on the femur and provide an axis around which both medial and lateral rotary movements are assisted [2]. The injury to these intraarticular structures is generally termed as "Internal derangement of knee" which was first coined by William Hey in 1784 [3].

Traumatic knee injuries were conventionally assessed clinically, and subjected to radiographs. In modern Orthopaedics, MRI and a subsequent arthroscopic procedure has come of age.

A detailed clinical examination, with the numerous stability tests provide an almoast 70% accuracy in diagnosing the pathology [4,5,6]. However, in the acute stage following injury, clinical tests may not be appropriate due to pain, thus an MRI is the preferred modality of investigation. It is non invasive, and considered to be highly sensitive to meniscal injuries [7,8,9] but is less so for the Anterior cruciate ligament injuries[10].

Arthroscopy of the knee is a key hole surgery performed as a daycare procedure. There are many indications however, this study is to correlate the findings subsequent to an MRI diagnosis.

II. AIM AND OBJECTIVES

1. To observe the sensitivity and specificity of MRI in detecting cruciate ligament tears, confirmed subsequently with Arthroscopy.
2. To observe the sensitivity and specificity of MRI in detecting meniscal tears.
3. To categorize discrepancies in findings of the menisci and cruciate ligaments between arthroscopy and MRI.

III. Material and Methods

We studied 13 females and 27 males who were consecutively referred from general practitioner presented to the department of ORTHOPAEDICS between February 2010 and January 2011, clinically diagnosed to have an internal derangement of knee, and who underwent an MRI, followed by Arthroscopy of the knee, are included in the study.

A thorough history was taken and physical examination of the injured knee was performed.

MRI examination were performed on all patients in the study, at our institute using a Siemens Sonata- Maestro class 1.5T MRI system with a gradient strength of 40mT/m. with dedicated knee coils. The scans were made in sagittal, coronal, and axial planes of the knee, with T1- , T2- ,STIR, PD and PD with fat suppression images. All the sections were 3mm thick. Patients were placed supine in MRI scanner and the involved extremity was flexed 15 degrees during the sagittal images to obtain better sections of the ACL [11].

MRI images were reported on an objective proforma by a single senior consultant radiologist, who was blinded to the clinical findings.

After reaching to a MRI diagnosis, the patients were subjected to an arthroscopy of the affected knee. All the arthroscopies were performed by a single knee surgeon. The arthroscopic findings were recorded. Then the findings of MRI and arthroscopy were compared and analyzed.

MRI diagnosis were placed into one of the four categories after arthroscopic evaluation [12] :

1. True positive: When MRI diagnosis of tear was confirmed on arthroscopic evaluation.
2. True negative: If the diagnosis of no tear was confirmed on arthroscopy.
3. False positive: If MRI showed a tear but arthroscopy was negative
4. False negative: If MRI images were negative but arthroscopy showed a tear.

Based on the above categories, five parameters were calculated to assess the reliability of the MRI results[13]:

1. Sensitivity: sensitivity of MRI is the ability of the MRI to detect an abnormality. It is determined by the equation :

True-positive/ (true-positive+true negative) X 100 per cent.

2. Specificity: specificity of MRI is the ability of MRI to give how many detected tears are usually accurate. It is determined by the equation:

True-negative/ (true negative+false positive) X 100 per cent.

3. Positive predictive value: It correlates a positive result of MRI with findings of arthroscopy. It is calculated by the equation:

True-positive/ (true-positive+false positive) X 100 per cent.

4. Negative predictive value: It correlates a negative result on MRI with the findings of arthroscopy. It is calculated by the equation:

True-negative/ (true-negative+false-negative) X 100 per cent.

Statistical analysis

All the continuous data were represented by mean with standard deviation. Categorical data were presented by frequency with percentage and it was analyzed by using Chi-square and Fischer exact test. Sensitivity, specificity, Positive predict value and Negative predictive value was used for comparison between MRI and Arthroscopy.

All the analysis was done by using SPSS 19.0 version, A p value less than 0.05 was considered as significant.

IV. Results

The results obtained from MRI were compared to those of arthroscopic findings.

Magnetic resonance imaging performed on the injured knees revealed 30 tears of the menisci, of which 20 were of medial meniscus and 10 tears of lateral meniscus. There were 21 tears of the cruciate ligaments, out of which 15 were tears of ACL and 6 of PCL.

50 menisci on MRI were normal, of which 20 were medial meniscus and 30 lateral meniscus. Of the 60 cruciates showing normal signal on MRI, 25 were of ACL and 35 of PCL.

Arthroscopy revealed 27 tears of menisci, 19 of these were of medial meniscus and 8 of lateral meniscus. 53 menisci were normal on arthroscopy, 21 of medial and 32 of lateral meniscus. There were 19 tears of cruciates, of which 14 were of ACL and 5 of PCL. There were 61 normal cruciates, of which 26 were ACL and 35 were PCL (table 1).

Table.1

	Normal		Tears	
	MRI	SCOPY	MRI	SCOPY
ACL	25(62.5%)	26(65%)	15(37.5%)	14(35%)
PCL	34(85%)	35(87.5%)	6(15%)	5(12.5%)
Medial meniscus	20(50%)	21(52.5%)	20(50%)	19(47.5%)
Lateral meniscus	30(75%)	32(80%)	10(25%)	8(20%)

With regards to tears of medial meniscus, comparison of the findings of magnetic resonance imaging with those of arthroscopy revealed 17 true positive, 18 true negative, 3 false positive and 2 false negative images. Area under the curve 87.6%, sensitivity was 89.5 per cent, specificity was 85.7 per cent, positive and negative predictive values were 85.0 and 90.0 per cent respectively.

For the lateral meniscus, there were 7 true positive, 30 true negative, 2 false positive and 1 false negative results. Area under the curve 89.1%, sensitivity was 87.5, specificity was 93.8 per cent. Positive and negative predictive values were 77.8 and 96.8 per cent respectively.

For the anterior cruciate ligament, there were 23 true positive, 10 true negative, 4 false positive and 3 false negative results. Area under the curve 76.1%, sensitivity was 88.5 percent, specificity was 71.4 percent and positive and negative predictive values were 85.2 and 76.9 percent respectively.

For the posterior cruciate ligament, there were 4 true positive and 33 true negative results. There were 2 false positive, and 1 false negative results and area under the curve is 87.1%, sensitivity is 80%, specificity is 94.3, positive predictive value 66.7% and negative predictive value were 97.1%.

V. Discussion

The area under the curve of magnetic resonance imaging was 87.6% percent for the medial meniscus, 89.1 percent for the lateral meniscus, 76.1 percent for ACL, and 87.1 percent for PCL in this series of 40 patients.

Imaging of the menisci showed 5 false positive and 3 false negative results of which 3 false positive results were of medial meniscus, and 2 of lateral meniscus. 2 false negative images were of medial meniscus and 1 of the lateral meniscus.

Among the 3 false positive results of medial meniscus, 2 menisci showed significant fraying due to degeneration, which was reported as a tear on MRI, and 1 knee showed a loose body whose signal was falsely reported as a tear in medial meniscus.

In both the false positive results of lateral menisci, the normal signal of the transverse meniscal ligament was reported as a tear of anterior horn.

The false negative results which were obtained were 2 for the medial and 1 for the lateral meniscus. Out of this 1 medial meniscus and the lateral meniscus had a tear on arthroscopy but it was reported as intra substance degeneration on MRI.

Imaging of the cruciate ligaments in this study showed 4 false positive and 3 false negative results, all of which were for the anterior cruciate ligament. There were no false positive or false negative results for the PCL in this study.

The false positive results for ACL were attributed to the presence of large ligamentum mucosum, which was reported as a tear in the substance of ACL.

In 2 of the 3 false negative results, the ACL was found to be lax on probing, which was probably due to a partial tear, and 1 tear was proximal and ACL was found attached to the PCL, which was hence reported as normal.

The receiver operating characteristic (ROC) analysis was used to validate the discriminative ability of MRI in distinguishing between a tear and normal pathology, which was confirmed or proven wrong by arthroscopy. The area under the curve (AUC) was 87.6%,89.1%. 76.1%, and 87.1% respectively for medial meniscus, lateral meniscus, ACL and PCL respectively, which is statistically significant [14].

The results of this study is in accordance to the literature which suggests an accuracy of 68 to 88 percent for the meniscal tears [15] and 80 to 94 percent for the cruciate ligament tears [16].

Various intra articular pathologies of the knee such as loose bodies, chondral fractures, degenerative changes, plicae can mimic a meniscal tear. Mucoid and eosinophilic degeneration can cause false positive results on MRI [15].

Post operative cases with metallic implants around the knee produce artifacts on MR images and an erroneous interpretation of the images is likely in such cases.

Magnetic resonance imaging is useful as a diagnostic tool in internal derangements of the knee. It is useful in circumstances where there is a need for detailed differential diagnosis. It is also an important diagnostic tool in cases of acute and painful knees, where clinical examination is difficult to perform.

Currently MRI is gaining popularity as a diagnostic tool in knee injuries due to increasing sports injuries, and road traffic accidents.

A further improvement in the techniques and increasing experience in interpretation of the images is likely to reduce the false positive and false negative results in future.

Magnetic resonance imaging also helps the surgeon to plan the definitive management of a tear during the same session.

VI. Conclusions

1. Magnetic resonance imaging is useful as a pre operative diagnostic tool in selected cases where a clinical examination cannot be performed as in acute injuries or in cases where clinical examination is inconclusive.
2. The efficacy of MRI in diagnosing a tear varies among different intra articular structures.
3. MRI has a high accuracy in diagnosing a tear of PCL.
4. Sensitivity for medial meniscal tear is higher as compared to lateral meniscus and high for PCL as compared to ACL.
5. MRI has a high positive predictive value for ACL, but has a low negative predictive value.
6. For PCL tears, MRI has a high negative predictive value which indicates that with a negative result for PCL on MRI, a diagnostic arthroscopy can be avoided.

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Utility of Non-conventional Energy Sources to Meet Increasing Power Demands by Solar Power Plant

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Abstract- Solar is a green energy that consumes small streams to generate electricity without depends on any sources of non-renewable energy. Even though the power generated is less but the benefits gained from this energy is the ability to raise the standard living of residents in remote areas and it does not emit any pollution gas which is able to give an unfavorable effect in the local environment, get free from pollution and helps to maintain sound health. It is able to support the mission of protecting the environment particularly to the ecosystem. It is helpful to reduce the green house effect also. This paper presents a use of non-conventional sources such as solar power system for the fulfillment of power demand to some extent and gives remedial measures during extreme emergencies of electrical power.

Index Terms- Light sensor, Non-conventional, Solar collector, Power

I. INTRODUCTION

Utility of solar energy i.e. non-conventional energy is the only alternative solution to reduce the pollution created by fossil fuels. By thermonuclear processes the sun produces huge amount of energy. This process produced heat and electromagnetic radiations.

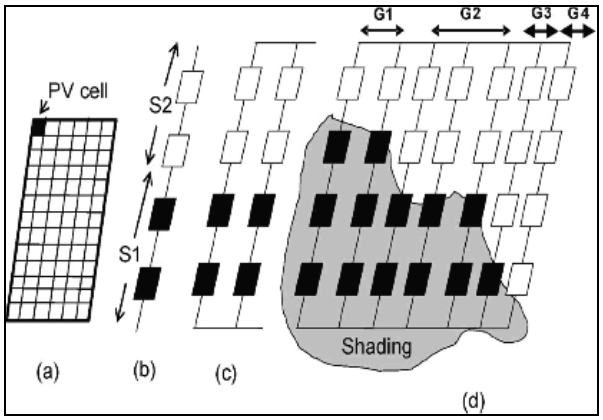
Since today, all countries around the world are concern about global climate changes and fuel sources that becoming less day after day. Green technology is intended to increase the employment of local resources and to contribute towards the national electricity supply security and sustainable socio-economic development.

Collector and storage units are two main components of solar power system. Collector collects the radiations which falls on it and convert it in other form of energy. Storage unit is required to store the heat because un-uniform radiation will be received throughout the year.

II. BACKGROUND THORY

PV Array Terminologies

Various components of a PV array are explained with the help of following Fig.



IMPLEMENTATION OF THE SUN TRACKING AND SELF-CLEANING OF SOLAR PV MODULES

Sun tracking systems are designed in a way to track the solar azimuth angle on a single axis. In single axis tracking system the collector is rotated around only one axis, the sun moves tracing an angle from the sunrise to the sunset. This angle traced by the sun is called the azimuth angle (γ) is defined as the angle between the lines due south and projection of normal to the collector as shown in Figure 1. Here we have used vertical axis with movement in the east-west (E-W) direction. The automated cleaning and tracking systems are implemented using a stepper motor, gear box (40:1), shaft, and sliding rod solar PV modules and circular metal rings for contacts as shown in Figure 3. The control of the stepper motor and the cleaning arrangement is done using a microcontroller.

A novel mechanism of sun tracking with automatic cleaning of PV modules is presented and cleaning mechanism of the PV modules consists of sliding brushes, which slides over module and cleans it twice a day, wherein PV panel makes a rotation of 360o in a day. It is observed that the daily energy generation of a flat PV module (kept stationary on ground) increases by about 30% and 15% for case of tracking-cum-cleaning and just single axis tracking, respectively. This demonstrated the effectiveness of tracking-cum-cleaning mechanism.

The mentioned tracking-cum-cleaning system is most suitable for today's industrial need. Above system can be kept inclined in the north or the south direction accordingly to achieve better energy generation from the PV modules of given wattage ratings. This system can extend to two axis tracking by rotating one axis manually and other axis automatically as rotated in this system. The other axis (north-south) can be rotated on daily or monthly

basis. This axis can also be implemented manually or automatically using motor, microcontroller etc.

`Figure 2. Schematic representation of Azimuth angle (γ) and Inclination angle (β)

A. Tracking mechanism A single axis tracking of the solar PV module is implemented along with the automated cleaning mechanism. For tracking the sun, the module is made to rotate 360o angle in a day, i.e. one rotation in 24 hours. The module starts its rotation from vertical position at the time of sunrise facing towards east (perpendicular to ground) and rotates at the rate of 15° per hour as shown in Figure 4.

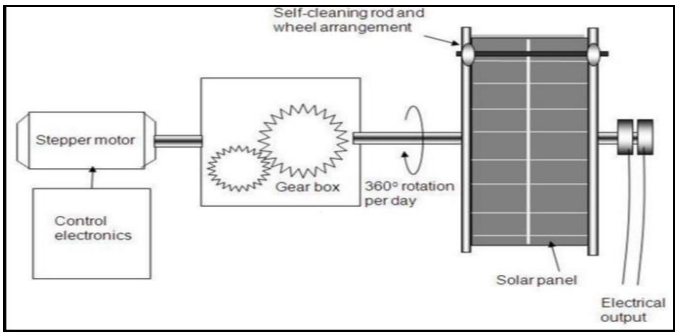


Figure 3 Schematic diagram of Sun tracking and automatic cleaning of solar PV modules

This tracking mechanism is based on the angle of rotation of earth around its own axis. The time for rotation of earth around its own axis is 24 hours which is equal to the tracking time of this system. This system is always in synchronization with the rotation of earth without any extra component because, this system starts at the time of

Figure 4. Rotation of panel throughout the day

sunrise and goes on and on as earth rotates on its own axis. That is the reason this tracking system does not require any sensor or extra component for synchronization like any other tracking system which usually comprised of.

B. Cleaning mechanism The automated cleaning mechanism is implemented using brush, rod & sliding wheels as shown in Figure 5. The brush is fitted in the rod. The rod is fitted with the wheels at both the ends, which are fitted in the channel in which they rotate. When panel comes in a vertical position at 6 am and 6 pm the brush fitted on the rod rotates on the panel from upwards direction due to gravity and cleans the panel two times in a day. In this way the cleaning mechanism works.

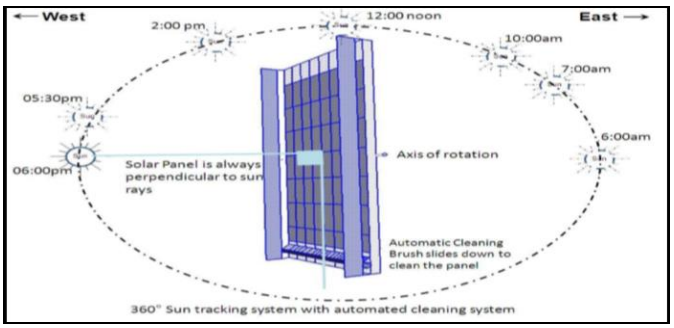


Figure 4: Rotation of panel throughout the day

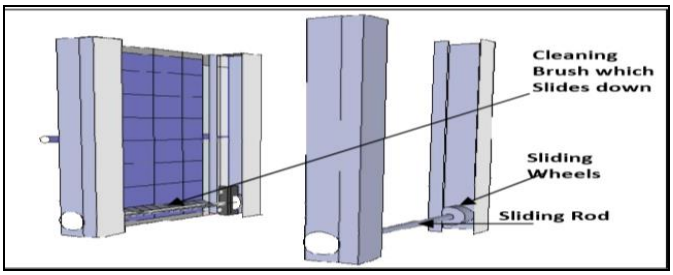


Figure 5 Sliding rod and wheels arrangement for cleaning mechanism

The proposed Sun tracking and self-cleaning of solar PV modules are a complete product and can be implemented with any existing solar PV system. This arrangement has capacity to enhance the energy output of the system and reduces the maintenance required for regular cleaning of

Sliding Wheels

Cleaning Brush which Slides down

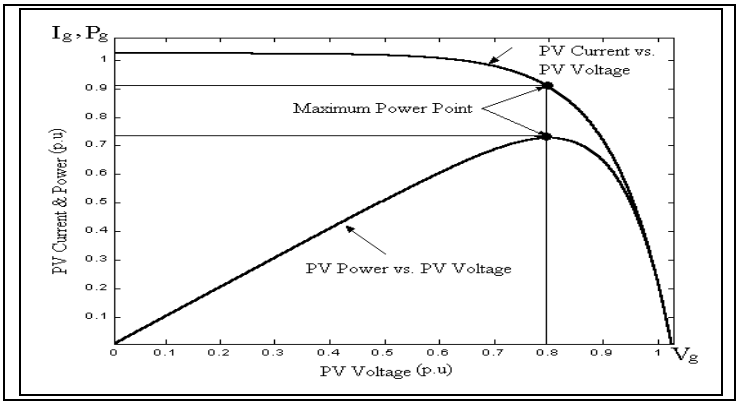
Sliding Rod the PV modules. This system cleans the modules twice in a day automatically.

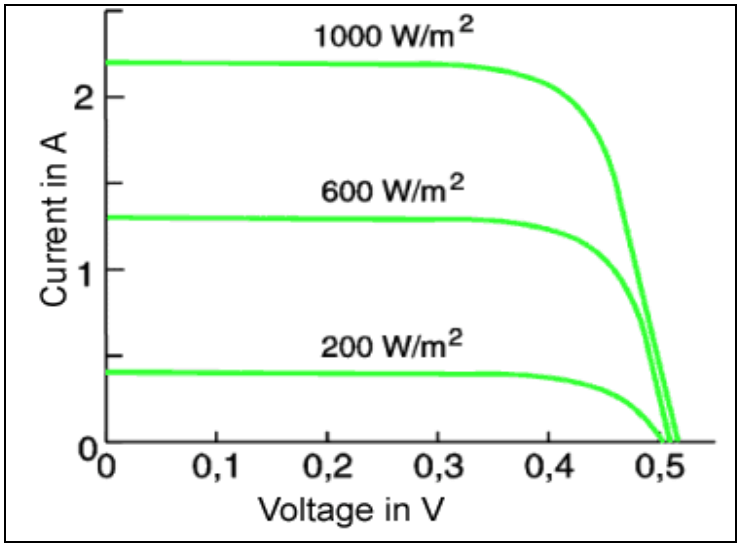
Maximum Power Point Tracking (MPPT)

The maximum power point tracking means to move the solar array voltage close to the maximum power point in order to draw the maximum power from the array under certain atmospheric conditions.

To find the percentage change in power maximum power point tracking offers by two methods:

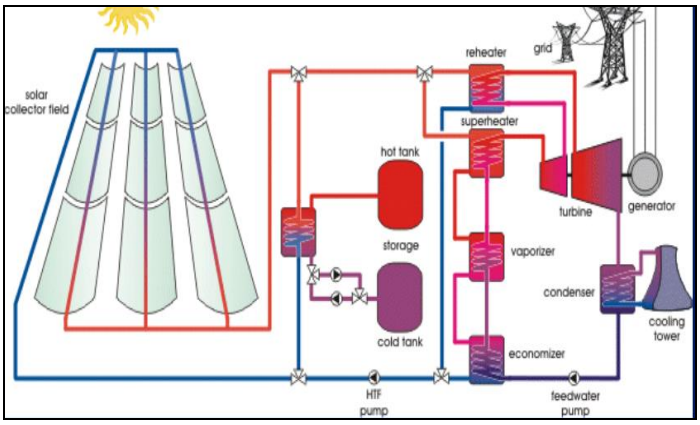
- Operating the stationary solar panel at constant load voltage and with MPPT.
- Operating the east west tacking solar panel at constant load voltage and with MPPT.
- Maximum power point tracking for improving the utilization ratio of photo voltaic cells, full of the conversion of energy is most important under the condition of changing temperature, under different sun light having the maximum power output increasingly and it is the basic principle of tracking





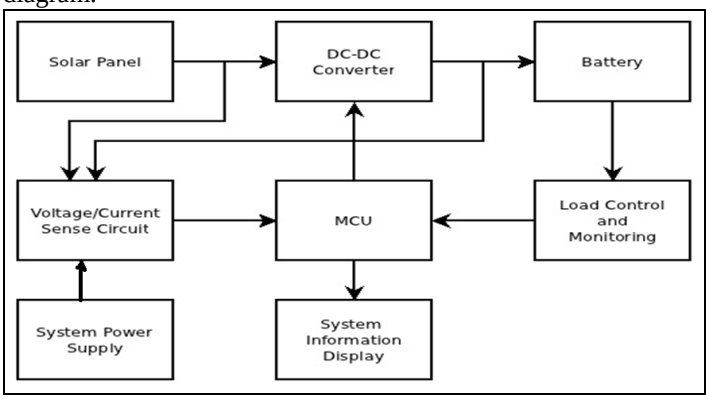
III. PROPOSED SYSTEM DESIGN

The overall design module of the pilot prototype is designed as below:



IV. CIRCUIT DESIGNING

The overall panel of solar will mainly consist of below circuit diagram:



- Solar panel - Output voltage - 9V, current - 3.5A , DIY 0.25m² solar panel built using 18 3"x6" solar cells.
- DC-DC converter, simple boost converter, solar panel voltage is lower than battery voltage
- Battery - 12V 10Ah SLA battery.
- For MPPT tracking accurate measurement is required
- Voltage and current sense circuit measures panel and battery voltage and current and disconnect the load if the battery is exhausted.
- 3.3V Power supply
- MPPT algorithm DC-DC converter MOSFET and system display is required
- System information display current, battery charge, Solar panel power, load power,etc

The main processor is microcontroller unit which provided linkages to DC-DC converter and system information display. The input to main processor is provided by Voltage/Current Sense circuit and load con and monitoring unit.

The solar panel is connected to DC-DC converter which is also connected to battery providing connection to Load control and monitoring unit. The system power supply is connected directly to Voltage/Current Sense circuit.

V CONCLUSION

As a conclusion, it should be noted that there are two input parameters in solar system which are very important in ensuing the efforts to harnessing solar power is successfully implemented, that are collectors and a storage unit. Based on the entire discussion in this paper, it is clear that the existence of solar system technology provide the peoples in remote areas a solution to the alternatives in generating electricity. Solar power system offers a technology which is small in size and less efficiency, low in cost and environmental friendly, pollution free, clean & green energy. It is recommended to all the researchers out there to explore and enhance the system of solar power for the purpose energy utilization for all the peoples in the world.

As such, use of non-conventional sources such as solar power system for the fulfillment of power demand to some extent and gives remedial measures during extreme emergencies of electrical power and to contribute towards the national electricity supply security and sustainable socio-economic development.

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Parametric Evaluation for Machining Die-Steel with WEDM

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Abstract- Wire electro discharge machining (WEDM) has become one of the most popular processes for producing precise geometries in hard materials, such as those used in the tooling industry. With the demands on the minimization of surface roughness, decrease of the producible geometric dimensions and improvement of the machining accuracy, Micro-Wire-EDM with a wire diameter of 80µm-250 µm has been a key technology for micro-machining. This work deals with the effect of the input parameter i.e thickness of the job on output parameters such as discharge current, cutting speed, spark gap/over cut, metal removal rate and surface roughness value of high carbon high chromium steel (HC-HCr), a die steel cut by wire-electrical discharge machining (WEDM). To obtain a precise workpiece with good quality, the parameters to be set on the machine are optimized experimentally. The output criteria can be estimated for a given thickness of the workpiece using the mathematical correlations developed.

Index Terms- wire electrical discharge machining, discharge current, cutting speed, spark gap, MRR

I. INTRODUCTION

During the last decade, WEDM has become an important nontraditional machining widely used in the aerospace, automotive and tool & die industries. WEDM has nearly obtained a monopoly position in some important areas, due to its capability of machining any material with electrical conductivity more than 0.01S/cm with high cutting speed, high precision and satisfying surface finish. 5-axis CNC WEDM machine has been routinely employed in the machining of complex 3-dimensional shape and the surface roughness has improved to better than 0.2mRa. The range of materials that are machined by WEDM has increased considerably, including now sintered carbides, PCD, PCBN and specific ceramics. In WEDM, the erosion mechanism has been described as melting and/or evaporation of the surface material by the heat generated in the plasma channel. A spark is produced between the wire electrode (usually smaller than 0.3mm) and workpiece through deionized water,(used as dielectric medium surrounding the workpiece) and erodes the workpiece to produce complex two and three-dimensional shapes. Usually some extra repetitive finish cuttings along the contour of a previous cut are necessary, by offsetting the wire by a value, so that the specified accuracy and good surface quality can be obtained.

II. MAJOR AREAS OF WEDM RESEARCH

The available information in the existing literature is divided into 6 categories and reviewed in the same order.

- Influence of machining parameters
- Pulse classification
- Effect of wire electrode parameters on machining criteria
- Thermal load on wire electrode
- Adaptive systems
- Parametric optimization

Parametric Optimization

In this section the authors work on machining of few ferrous materials DC53 tool steel of 27mm thick, SKD 11 tool steel, X210 Cr 12 steel of 17.3, 25, 34mm thick, stainless steel of 10, 15mm thick, En8, En31, AISI 420 steel of 31.5mm thick, non-conductive materials, ceramics like sintered carbide, polycrystalline diamond, sailon, boron nitride, silicon nitride and 5mm thick nonferrous materials, copper, brass, aluminum, graphite, tungsten carbide were discussed. The authors evaluated the optimal parameters of machining to maximize the MRR and surface finish. Hadda and Tehrani [1] designed their experiments using Taguchi L18 array and conducted wire electrical discharge turning operations on AISI D3 steel. They performed regression analysis and determined the optimum values of spark on time and electrode rotational speed for achieving higher material removal rate and better surface roughness values.

Kanlayasiri and Boonmung [2&3] optimized the parameters effecting surface finish, for machining DC53 tool steel of 27mm thick with using 0.25mm diameter wire by designing the experiments with Taguchi method. The authors developed mathematical model for optimization to predict surface roughness values and errors. The developed model has shown a maximum error of 30%. Mohammadi et al. [4] performed 54 precision turning experiments. ANOVA is used for analyzing the effect of input parameters on response. The authors considered the power, spark-on time, spark-off time, wire velocity, wire tension, wire speed and rotational speed as parameters and material removal rate as response. The authors developed mathematical relations for determining the material removal rate. Haddad and Tehrani [5] performed turning operations using L18 orthogonal array on DIN X210 Cr 12 steel. The authors derived mathematical model for material removal rate determination and

its effect on surface roughness and roundness of machined surface. The die rotational speed, power and pulse off time exhibit significant effect on material removal rate. Tarng et al. [6] optimized the cutting parameters for better cutting performance using feed forward neural net work through simulated annealing algorithm. The process parameters considered for optimization are thickness of work piece, material, spark-on time, spark-off time, machining current, voltage and capacitance on cutting speed and surface roughness as response. The authors machined SUS 304 stainless steel of 10 and 15mm thicknesses. The predicted optimized values are: For 10mm thick job: Ra value 16.1µm, cutting speed 1.63mm/min and for 15mm thick job: Ra value 1.65µm, cutting speed 1.65mm/min

Jesudas et al. [7] developed a mathematical model using Taguchi analysis for optimizing the parameters of machining bronze-alumina alloy metal matrix composite. L9 orthogonal array is followed for design. ANOVA is applied to find the velocity of the optimal parameters derived. Rajurkar and Wang [8] performed experiments on work pieces of different thickness and developed an adaptive control system which monitors and controls the spark frequency according to on line identification of work piece thickness. Lok and Lee [9] machined 10 samples of Sailon material, 40mm thick under preset conditions, evaluated MRR as 4.5-6.0 mm^3/min. The authors compared the machining rate of Sailon with that of SKD11 steel and found that Sailon material machinability is poor. It was also revealed that the material removal rate increased with increase in machining current to some extent and then decreased. Kuriakose and Shunmugam [10] designed the experiments using Taguchi L18 array and conducted on Ti6A14V material with 0.25mm diameter brass coated wire under preset conditions, 80V, 8-12A machining current, 4-8µs pulse time. Formation of oxides was observed due to high temperature generation, macro and micro level stresses induced during the process. The authors observed that when the time between two pulses is larger, non-uniform cooling and heating occurs and suggested coated wire as electrode from metallurgical point of view.

Calýk and Çayda [11] presented an experimental investigation of the machining characteristics of AISI D5 tool steel in wire electrical discharge machining process. During experiments, parameters such as open circuit voltage, pulse duration, wire speed and dielectric fluid pressure were changed to explore their effect on the surface roughness and metallurgical structure. Optical and scanning electron microscopy, surface roughness and micro hardness tests were used to study the characteristics of the machined specimens and it was found that the intensity of the process energy does affect the amount of recast and surface roughness as well as micro cracking, the wire speed and dielectric fluid pressure do not seem to have much of an influence. Kadam and Basu [12] performed experiments on 17.3mm thick HC-HCr steel with 0.25mm diameter wire by varying duty factor, machining current and wire speed. The authors optimized the cutting speed and surface roughness and developed mathematical equations using regression analysis. Rao et al. [13-16] studied the effect of process parameters on the yield criteria for machining different nonferrous materials and developed mathematical correlations to evaluate the parameters. Kuriakose et al. [17] machined titanium alloy of 40mm thickness with zinc coated brass wire, 0.25mm diameter at preset machining conditions 80V voltage, 8-16A machining current, 8-10m/min wire speed,1-1.2kN wire tension while optimizing the machining parameters, cutting speed and surface roughness using data mining technique.

Kuriakose and Shunmugam [18] explained that the influence of cutting parameters on cutting velocity and surface finish are quite opposite. The authors developed a relationship between input and output variables using multiple regression model and non-dominated sorting genetic algorithm is used to optimize the multiple objectives. The authors considered voltage, machining current, spark ON, OFF times, wire speed, wire tension, injection pressure and work piece height as parameters and determined their effect on cutting speed and surface finish. The experiments on 60mm thick Ti6A14V Titanium alloy yielded a cutting speed of 0.9735mm/min and surface roughness of 3.2µm as optimum. Hewidy et al. [19] observed that right selection of the machining conditions is the most important aspect in processes related to the WEDM of Inconel 601 material. The work highlights the development of mathematical models for correlating the inter-relationships of various WEDM machining parameters of Inconel 601 material such as: peak current, duty factor, wire tension and water pressure on the metal removal rate, wear ratio and surface roughness.

Prakash and Ranganath [20] have machined En8, En31 and HC-HCr materials and analyzed the data for optimizing the parameters to evaluate MRR and surface roughness values. Kannan et al. [21] have carried out experiments on OHNS die steel of 30mm thick. The authors designed the experiments with Taguchi's design of experiment, L8 orthogonal array, 7 parameters and 2 levels. The experimental results were analyzed by ANOVA for optimum conditions such as minimum surface roughness and maximum cutting speed. Mohammadi et al. [22] used WEDM for machining of precise cylindrical forms on hard and difficult-to-machine materials An L18 (21×37) Taguchi standard orthogonal array was chosen for the design of experiments. The signal-to-noise (*S/N*) ratio analysis is employed to find the optimal conditions.

Based on the literature review, it is understood that authors have considered the values of machine settings like wire tension, wire speed, wire diameter, wire material and dielectric conductivity as follows.

Gap voltage: 90V
Wire diameter: 0.25mm
Wire tension: 80N
Wire speed: 4.7 m/min
Dielectric conductivity: 48 mhos
Flushing pressure: 1.5kN/mm^2
Spark on time: 5µs

III. EXPERIMENTAL SET UP

Fig1.Shows the schematic view of the experimental set up.

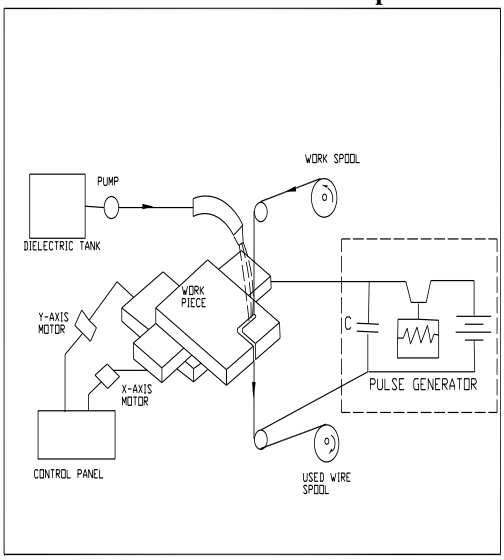

Figure 1: Schematic view of experimental set up

The HC-HCr steel specimens of 20mm x 40mm size on thicknesses 5, 7.5, 10, 12.5, 15, 17.5,20,25, 30, 35, 40, 45, 50, 55, 60, 65, 70, 75 and 80mm are prepared. The experiments are conducted on the work piece of every thickness by cutting L shape and "[“shape by varying the machining current from a lower value to a value where the machining is in consistent in 5 steps. At every machining current, I value the machining criteria is measured. The machining current, I value at which the machining is consistent with continuous cutting, better finish with least wire rupture is selected as optimal. The cutting speed is noted from the machine display, surface finish is measured on "[“cut using Talysurf. The cutting width is measured on L cut with shadow graph and checked with microscope. The spark gap (wire off set) is calculated from cutting width.

The optimum values of machining current, cutting speed, spark gap and MRR for every thickness are used for plotting the curves and best fit curve is selected using the software. The mathematical relation is generated for this best fit curve and statistical analysis is performed to find the fitness of the curve.

IV. RESULTS AND DISCUSSIONS

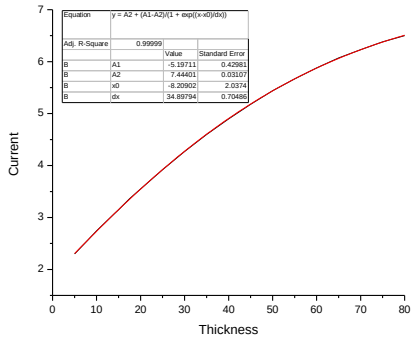

Figure 2: Effect of thickness on current

The variation in the discharge current with the increase in work piece thickness is shown in Fig.2. For a specified set of machining conditions it is observed that with increase in thickness, the required discharge current also increases. This is attributed to the high amount of energy required for high thickness job in which machining is possible only by increasing the current. This plot is useful to extract suitable minimum discharge current required for machining of any thickness HC-HCr steel work piece with in the machine working range. By regression analysis of the data the equation for the best fit curve is obtained as

$$I=7.444 - [12.641/ \{1+\exp [(T+8.209)/34.898]\}]$$

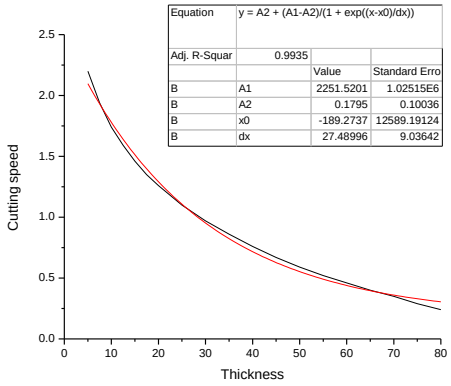

Figure 3: Effect of thickness on cutting speed

Fig.3 shows the effect of thickness on cutting speed for various sizes of the work pieces. The plot indicates that as thickness of the work piece increases the cutting speed decreases rapidly. For thickness beyond 70mm the cutting speed almost remains constant. If the thickness increases, the volume of metal to be removed increases which demands more energy and it may become a machine constraint. At the same time the spark is jumping to the sides of the wire causing more width of cut, reducing the cutting speed. The data thus obtained is subjected to regression analysis and mathematical correlation for the best fit curve is obtained as

$$C_s = 0.1795+ [2251.34/ \{1+\exp [(T+189.27)/27.49]\}]$$

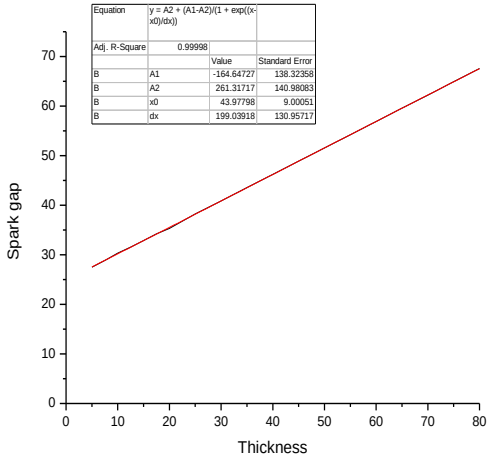

Figure 4: Effect of thickness on spark gap

The variation of spark gap occurred with change in workpiece thickness is shown in fig.4. The plot shows a steep increment in spark gap with increase in current. This may be due to higher energy in sparks at higher current. High energy spark jumps longer causing more material to melt and evaporate, creating wider cut. Spark gap should be as low as possible, as it causes unnecessary machining in width direction, reducing the cutting in length direction. The length of cut decreases with increase in spark gap, even though MRR is increasing. The spark gap value must be known in before hand for programming the cutter path other wise accuracy of machining will be lost. The best suitable curve is drawn and statistical analysis (ANOVA) is carried out. The mathematical correlation obtained is

$S_g = 261.31 - [425.95/ \{1+\exp [(T-43.98)/199.03]\}]$

The statistical analysis shows the values of $R^2 = 0.9998$ and standard deviation as 0.0034. The correlation is useful in finding the spark gap in turn cutting width, to compute the MRR and program the wire off set during CNC part programming, and hence higher accuracy can be achieved.

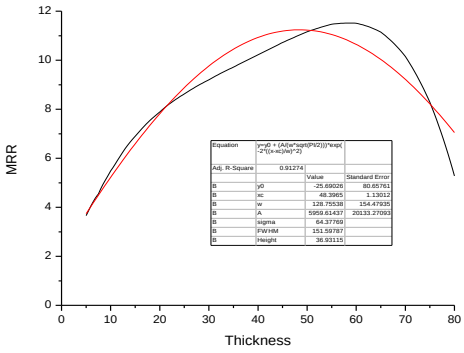

Figure 5: Effect of thickness on MRR

Fig.5 depicts the change in MRR with increase in workpiece thickness The experimental observations reveal that beyond 60mm thickness, the machining is not stable and wire rupture occur frequently. MRR is a calculated value obtained as a product of cutting speed, cutting width and thickness of the work piece. The plot shows a constant rise with a positive slope up to 50mm thickness and then falls. This is due to the increase in thickness, decrease in cutting speed and increase in cutting width. Beyond 40mm thickness the cutting speed is decreasing drastically contributing to lower the MRR. However in this process, cutting speed is an important factor as the machining is a through and through cutting operation.

V. CONCLUSIONS

The influence of parameters like discharge current, job thickness, on the machining criteria such as cutting speed, spark gap, material removal rate are determined. The results are useful in setting the parameters required for quality cuts on HC-HCr die steel. Suitable parameters can be selected for machining with the 0.25mm diameter wire. The mathematical relations developed are much more beneficial for machine settings, to estimate the cutting time, cost of machining and accuracy of cutting for any size of the job within machine range. The maximum error obtained in the calculated values and experimental values are less than 2%. These results will be useful to make the Wire EDM system to be efficiently utilized in the modern industrial applications like die & tool manufacturing units for parametric setting, machining time, cost calculations and also for process planning.

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Analysis of Optical Time Division Multiplexing Using Packet Interleaving Scheme

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Abstract- Optical time-division multiplexing (OTDM) expands and extends the well-known technique of electrical time-division multiplexing into the optical domain. In OTDM, optical data streams are constructed by time-multiplexing a number of lower bit-rate optical streams. In this paper we present an experiment work in optical time-division multiplexing and simulating a packet-interleaved operation using OptiSystem.

Index Terms- Compression stage, Multiplexer, Packet interleaving scheme

I. INTRODUCTION

Functionally, optical TDM (OTDM) is identical to electronic TDM. The only difference is that the multiplexing and demultiplexing operations are performed entirely optically at high speeds. At the inputs to the network, lower-speed data streams are multiplexed optically into a higher-speed data streams, and at the output of the network, the lower-speed data streams must be extracted from the higher-speed data streams optically by means of demultiplexing function. Optical signals representing data streams from multiple sources are interleaved in time to produce a single data stream. The interleaving can be done on a packet by packet basis as shown in the Figure 1: If the packets are of fixed length, the recognition of packet boundaries is much simpler i.e., fixed length packets are used. In packet-interleaved case, framing pulses can be used. These framing pulses mark the boundary between the packets.

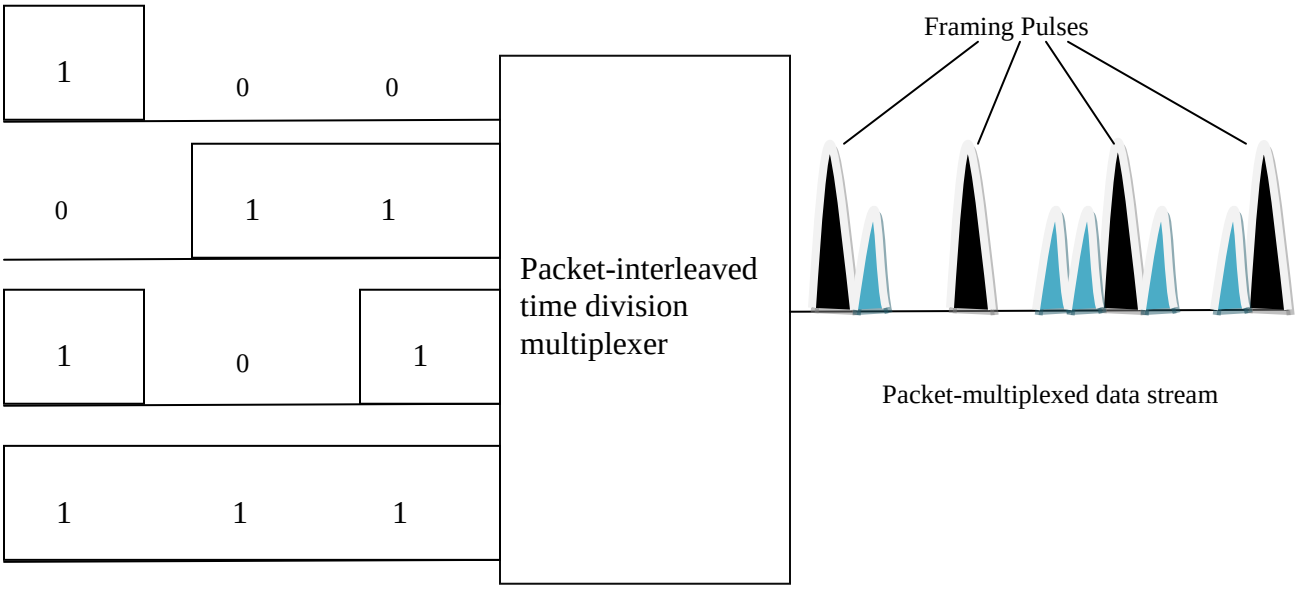


Figure 1: Function of a packet-interleaved optical multiplexer.

Assume if n data streams are to be multiplexed and the bit period of each these streams is T. Also framing pulses are used then the inter pulse width is $\tau = T/n+1$ because n+1 pulses must be transmitted in each bit period. There is a guard time between the successive pulses. One purpose of this guard time is to provide some tolerance in the multiplexing and demultiplexing operations.

II. EXPERIMENTAL SETUP

In this packet interleaved case, the data stream externally modulates a periodic stream of narrow pulses. If the bit interval is T the separation between successive pulses is also T. In this way, we have to reduce the interval between successive pulses to τ, a high-rated multiplexed data signal. This is done by passing through the initial sequence through a series of compression stages as shown in Figure 2: If the size of each packet is l bits, the output goes through [$k=\log_2 l$] compression stages. In the first compression stage, bits 1,3,5,7…are delayed by T-τ. In the second compression stage, the pair of bits (1,2),(5,6),(9,10)…. are delayed by 2(T-τ). In the third compression stage, the bits (1,2,3,4),(9,10,11,12)….are delayed by 4(T-τ). Design of compression stages by using OptiSystem.

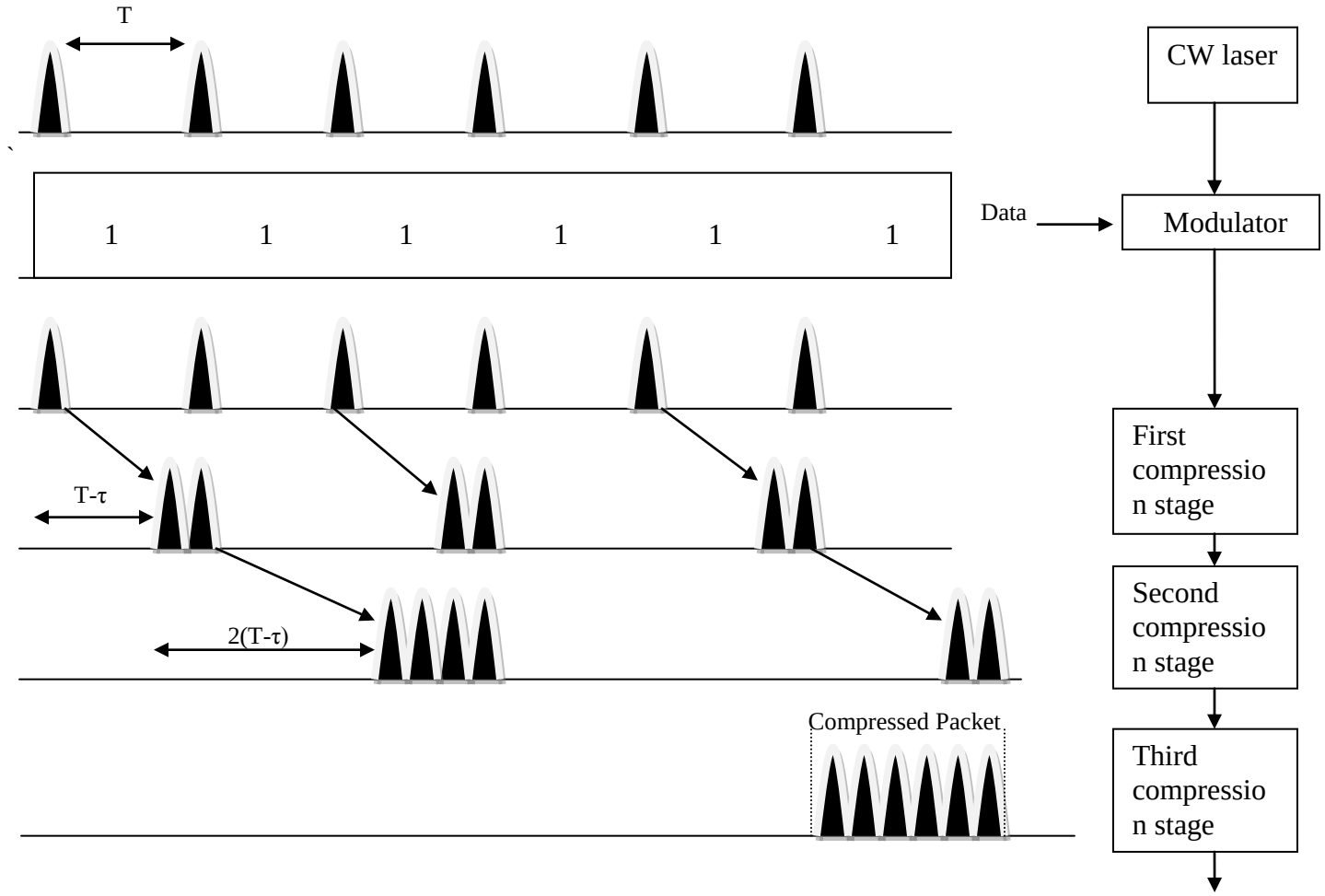


Figure 2: An optical multiplexer to create a packet-interleaved TDM stream

A block schematic of the four channel system is shown in Figure 3: A common 10-GHz clock drives the four transmitters via a series of micro wave delay lines that are adjusted to provide correct timing of the optical pulses. Four channels at wavelength 1550nm are RZ modulated at each 10 Gbit/s data are delayed by ¼ of the time window and then multiplexed. Total power of all channels set to -12dBm NRZ coding is a line code in which binary value '1' is represented by positive voltage and '0' is represented by negative voltage. The pulses have more energy than others. It requires only half the bandwidth than other coding. In RZ coding, a binary 1 is represented by first half of the bit duration, during the second half the level return to zero. Absence of a pulse represented a binary 0. Twice the bandwidth is required for RZ coding.

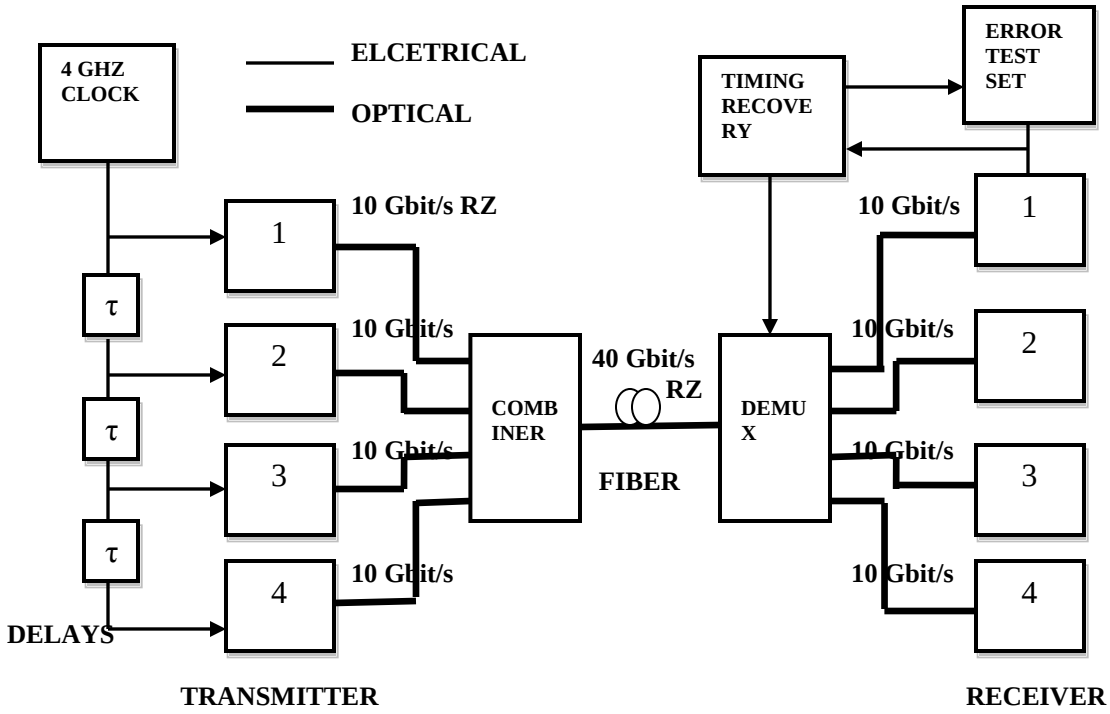


Figure 3: Block schematic of 4-channel OTDM system.

To demonstrate the packet interleaving, created a sample in which packets with 6 bits length are created. The Figure 4: shows the layout with three compression stages.

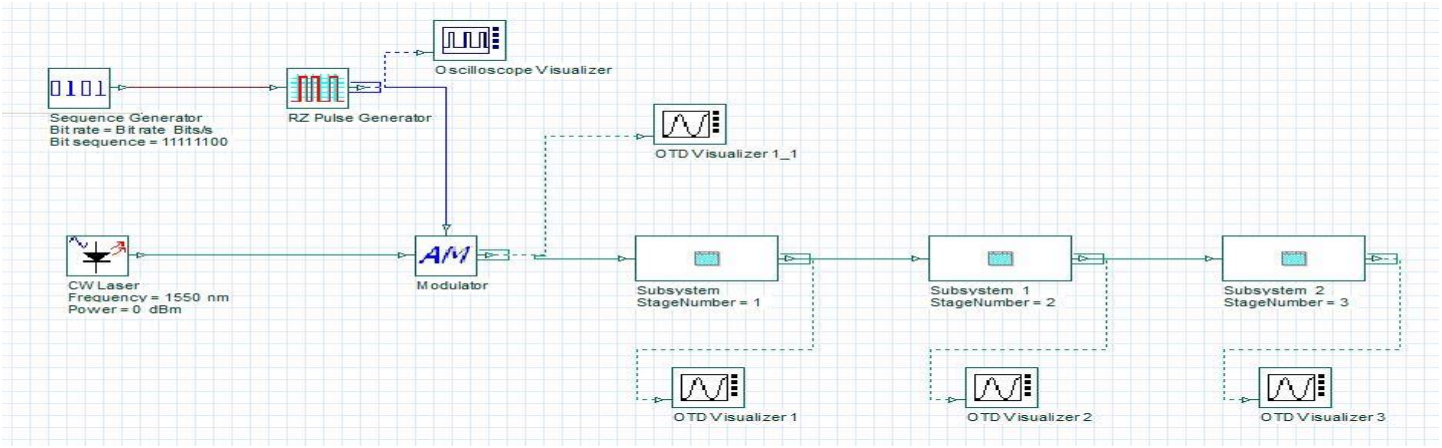


Figure 4: Packet interleaving with 6bits/packet

The stage was implemented in a subsystem because each compression stage has the same layout as shown in Figure 5: consist of a PRBS generator with bit rate of 10 Gbit/s and a RZ pulse generator which generates the continuous electric pulses. A light coming from a CW (Continuous Wavelength uses a Conventional band or C-band of 1550nm wavelength) laser is modulated by using AM modulator.

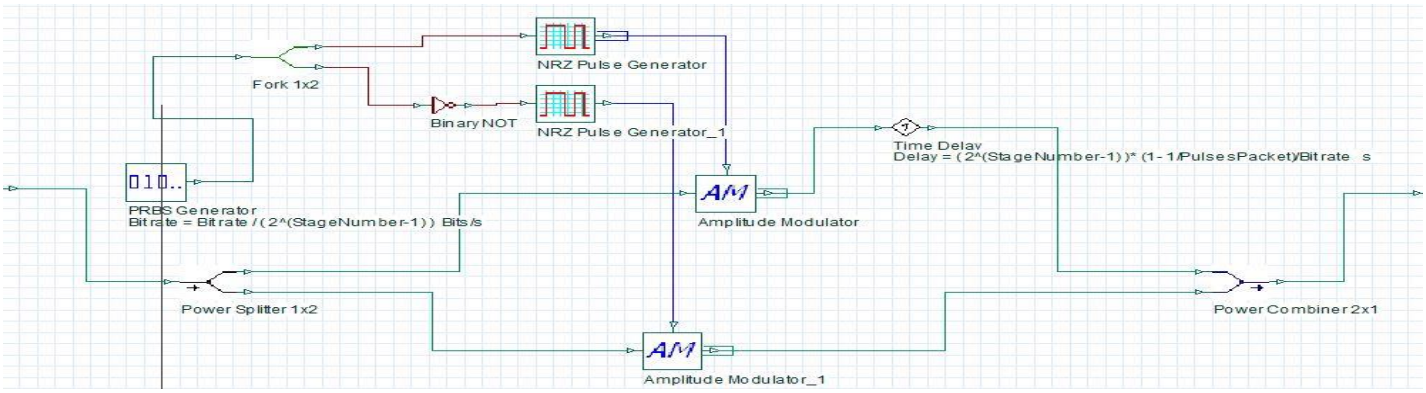


Figure 5: Compression Stage

There are three compression stages, the only difference between these compression stages were some parameters such as Bit rate, PRBS (Pseudo Random Bit Sequence) Generator and the Delay at the time component. The Bit rate is calculated by using Bitrate/($2^{stagenumber-1}$) bits/sec and Delay is calculated by using ($2^{stagenumber-1}$)*(1-1/pulses packet)/Bitrate.

Pulse time τ (τ specifies the pulse width) = 0.06bit

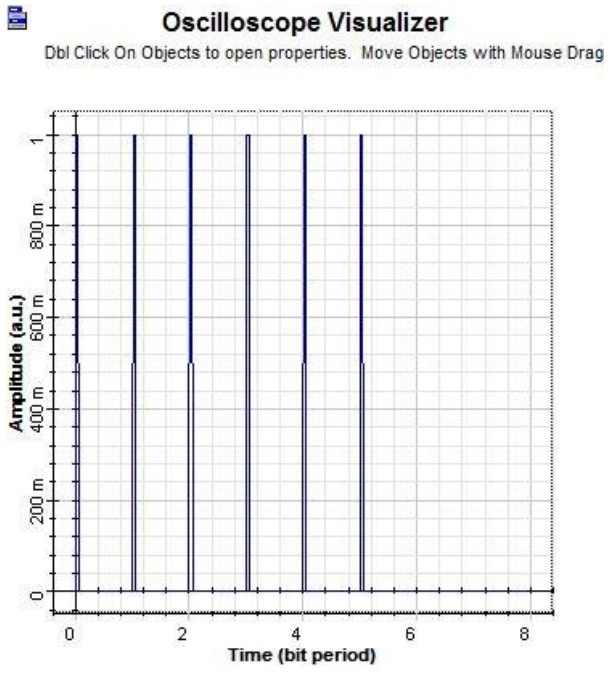
Pulses packet (specifies the number of bits per packet) = 6bits

Bit interval T = 1bit

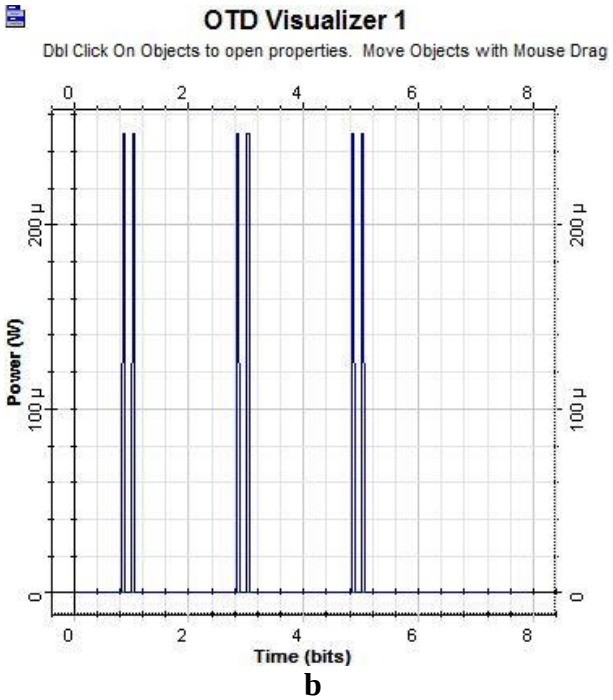
Therefore, T-τ = 0.94 2(T-τ) = 2(0.94) 4(T-τ) = 4(0.94)……..

III. RESULTS AND DISCUSSIONS

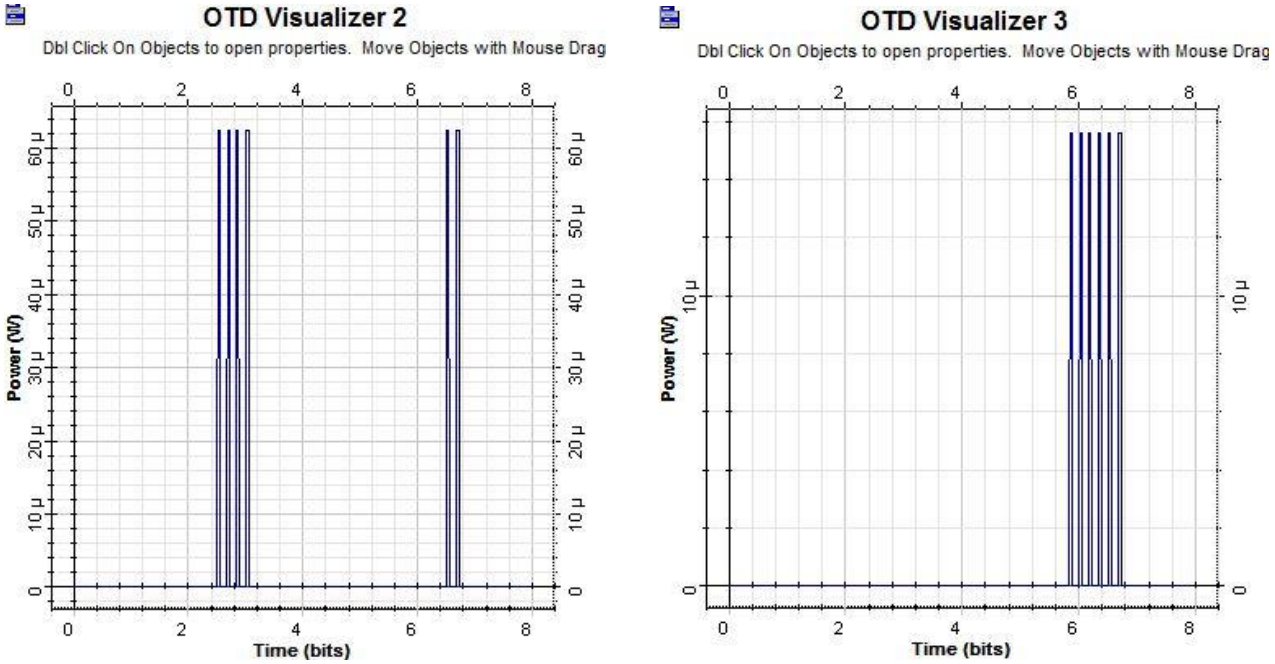
The initial sequence is 11111100 and the behavior of the optical signal along the transmission can be shown in Figure 6:



a



b



c **d**

Figure 6: (a) initial bit sequence (b) after 1st compression stage (c) after 2nd compression stage (d) after 3rd compression stage.

Here the optical pulses are compressed at each stage as following Figure 6: In the first compression stage the bits 1,3,5,7 are compressed by T-τ , similarly the bits (1,2),(5,6), and (1,2,3,4) are compressed by 2nd and 3rd stages. As the use of compressing bits the bandwidth is reduced for transmitting bits.

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Multiple Antenna & Diversity: Smart Antennas

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Abstract- In communication systems, we have to increase the reliability of the communication operation between transmitter and receiver while maintaining a high spectral efficiency. The ultimate solution relies in the use of diversity, which can be viewed as a form of redundancy [1]. Multiple-input multiple-output systems include a variety of techniques capable of not only increasing the reliability of the communication but also impressively boost the channel capacity. In addition, smart antenna system can increase the link quality and lead to appreciable interference reduction. Smart antenna technology is one of the most vital developments in mobile communication. The signal that is been transmitted by a smart antenna cannot be tracked or received by any other antenna thus ensuring a very high security of the data transmitted.

Index Terms- Adaptive antenna array, Diversity techniques, MIMO, Smart antennas, Spatial diversity

I. INTRODUCTION

Multiple antenna technology proposed for communication is an important means to improve the performance of wireless systems. In a system with multiple transmit and receive antennas (MIMO channel), the spectral efficiency is much higher than that of the conventional single antenna channels. Smart antenna technique can significantly increase the data rate and improve the quality of wireless transmission, which is limited by interference, local scattering and multipath propagation. A smart antenna is an array of elements connected to a digital signal processor. Smart antenna was discovered in early 1990s when well developed adaptive antenna arrays originated from Radar system. Recently, smart antenna technique has been proposed as a promising solution to the future generations of wireless communication systems and broadband wireless access networks.

II. MIMO DIVERSITY

Diversity schemes provide two or more inputs at the receiver such that the fading phenomena among these inputs are uncorrelated. There are many diversity techniques that can be applied to communication systems – time diversity, frequency diversity and spatial diversity or any combination of these three diversities. [2] In time diversity, the same information-bearing signal is transmitted in different time slots where a good gain can be achieved when the duration between the two slots, in which the same symbol is transmitted, is greater than the coherence time of the channel as per fig.1

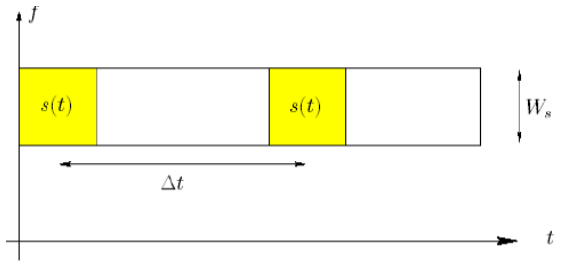

Fig.1: Time diversity

In frequency diversity, the same information-bearing signal is transmitted on different subcarriers where a good diversity gain can be achieved when the separation between subcarriers is greater then the coherence bandwidth as per fig.2.

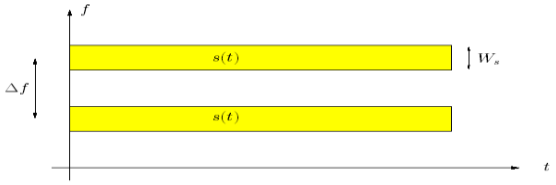

Fig.2: Frequency diversity

In spatial diversity, the same information-bearing signal is transmitted or received via different antennas where the maximum gain can be achieved when the fading occurring in the channel is independent. In receiver space diversity, M different antennas are used at the receiver to obtain independent fading signals as in fig.3.

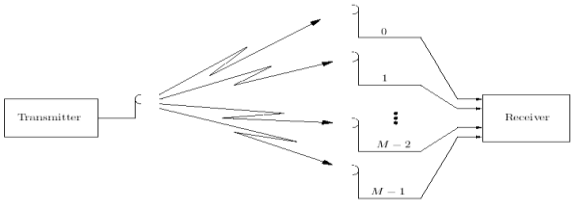

Fig.3: Space diversity in receiver

In transmitter space diversity, M different antennas are used at the transmitter to obtain uncorrelated fading signals at the receiver as in fig.4.

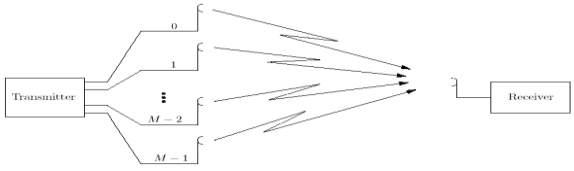

Fig.4: Space diversity in transmitter.

III. SMART ANTENNAS

The smart antenna works as follows. Each antenna element sees each propagation path differently, enabling the collection of elements to distinguish individual paths to within a certain resolution. As a result, smart antenna transmitters can encode independent streams of data onto different paths or linear combinations of paths, thereby, increasing the data rate and providing diversity gain.[3] No manual placement of antennas is required. The smart antenna electronically adapts to the environment. Smart antenna system can be categorized into three main groups – Phased antenna array system, switched beam systems and adaptive antenna array system. Phased antenna array is a group of antennas in which the relative phases of the respective signals feeding the antennas are varied in such a way that the effective radiation pattern of the array is reinforced in a desired direction and suppressed in undesired directions. [4] In switched beam approach, the sector coverage is achieved by multiple predetermined fixed beam patterns with the greater gain placed in the centre of a beam. [4] In practice, switched beam system can simply replace conventional sector antennas without requiring significant modifications to the radio base station antenna interface or the baseband algorithms implemented at the receiver. Fig 6 illustrates the produced antenna pattern with 4 antennas.

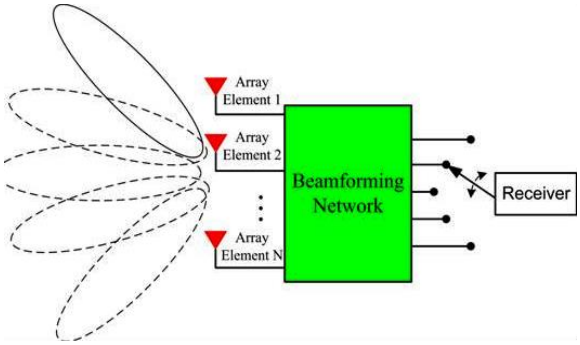

Fig.5: Functional block diagram of switched beam antenna.

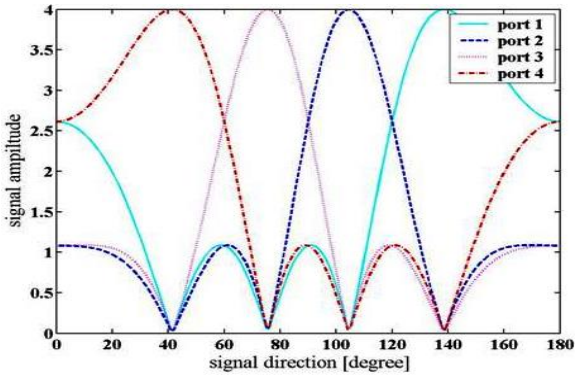

Fig.6: Functional block diagram of switched beam system.

Adaptive antenna array beam forming techniques have been employed to remove unwanted noise and jamming from the output, mainly in military applications. With powerful digital signal processing hardware at the base-band, algorithms could control antenna beam patterns adaptively to the real signal environment, forming beams towards the desired signals while forming nulls to co-channel interferers. Thus the system performance is optimized in terms of link quality and system capacity. [5] Adaptive antenna array can be utilized in the transmitter side, which is known as transmit beamforming as in fig.7 or in the receiver side, which is called receive beamforming.

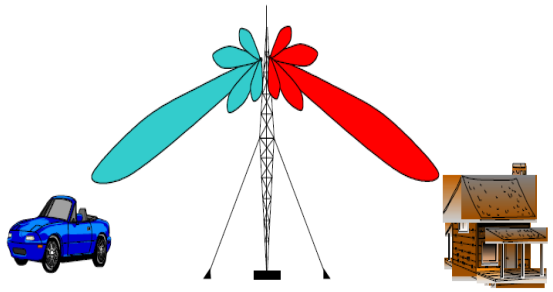

Fig.7. Illustration of transmit beamforming.

Multipath propagation in mobile radio environment leads to inter-symbol interference. Using transmit and receive beams that are directed towards the mobile of interest reduces the amount of multipath and inter-symbol interference. [6]

IV. CONCLUSION

This paper introduced the multi antenna technologies which can be considered as one of the most vivid area of research. Diversity techniques are used to improve the performance of the radio channel without any increase in the transmitted power. The use of multiple antenna system increases link reliability, increases channel capacity and reduces interference in both uplink and downlink. Further, smart antenna techniques are introduced which are intelligent devices whose radiation pattern can be varied without being mechanically changed. General conclusions about how best to upgrade a network are not easily made because every operator has a unique set of constraints such as what frequencies are available, what interfaces pre-exist and what network, antenna and user equipment has been previously deployed. Even so, the wide variety of antenna types and processing techniques makes impressive improvements in the capacity and coverage of emerging 4G networks.

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AN INTELLIGENT TECHNIQUE FOR UNIFIED POWER FLOW CONTROLLER WITH MATRIX CONVERTER

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Abstract - The Matrix Converter is an array of controlled semiconductor switches that connects directly the three phase source to three phase load. In the Matrix Converter there is no dc link for ac/ac conversion. MC-based UPFC (MC-UPFC) has reduced capacitor loss, volume, cost. sliding mode control techniques is used in the principles of direct power control (DPC) which is established for an MC-UPFC model as consists of input filter. Simulation results of DPC controllers for MC-UPFC provide no cross coupling in dynamic and steady state response, fast response time and decoupled active and reactive power control. The Proposing System matrix converter- based UPFC using a direct power control approach (DPC-MC) based on an MC-UPFC dynamic model. The design UPFCs, presenting robust behavior to parameter variations and to disturbances.

Index terms — Direct power control (DPC), Matrix converter (MC), MATLAB/ SIMULINK, Unified power-flow controller (UPFC).

NOMENCLUTURE

K_p,K_q	proportional gains
e_p	active power error
e_q	reactive power error
V_d,V_q	matrix converter voltage in dq components
i_d,i_q	input current in dq component

I. INTRODUCTION

In recent years, with ever-increasing demand for electricity, the power transfer grows, the power system becomes increasingly more complex to operate and the system can become less secure for riding through the major outages. [1]. These days, UPFCs are one of the most versatile and powerful flexible ac transmission systems (FACTS) devices. The UPFC results from the combination of a static synchronous compensator (STATCOM) and a static synchronous series compensator (SSSC) that shares a common dc capacitor link. UPFCs can occur over line impedance due to sending and receiving end voltage amplitudes and phase differences. Unified power-flow controllers (UPFC) enable the operation of power transmission networks near their maximum ratings, by enforcing power flow through well-defined lines[1]. In dc-link capacitance of the UPFC increase its losses, weight, cost, and volume and decreasing the lifetime.

In the last few decades, an increasing interest in new converter types, are performing the same functions but with reduced storage needs, has arisen. Matrix converter is highly applicable for ac to ac conversion, allowing bidirectional power flow, assurance of near sinusoidal input and output currents, voltages with variable amplitude, and adjustable power factor [8].These minimum energy storage ac/ac converters have the capability to allow independent reactive control on the UPFC shunt and series converter sides, while guaranteeing that the active power exchanged on the UPFC series connection is always supplied/absorbed by the shunt connection.

Conventional UPFC controllers do not guarantee robustness. Obtained results show that DPC is a strong nonlinear control candidate for line active and reactive power flow. It ensures transmission-line power control as well as sending end reactive power or power factor control. Presented simulation and experimental results show that active and reactive power flow can be advantageously controlled by using the proposed DPC. Simulation and experimental results confirm the performance of the proposed controllers, showing no cross-coupling, no steady-state error (only switching ripples), and fast response time for different changes of power references. In dc-link capacitance of the UPFC increase its losses, weight, cost, and volume and decreasing the lifetime. These matrix converters are capable of performing the same ac/ac conversion, allowing bidirectional power flow, assurance of near sinusoidal input and output currents, voltages with variable amplitude, and adjustable power factor. these minimum energy storage ac/ac converters have the capability to allow independent reactive control on the UPFC shunt and series[8].

II. DIRECT POWER CONTROL OF MC-UPFC

A. Line Active and Reactive Power Sliding Surfaces

In DPC method have the steady state, v_d is imposed by source v_s. the transmission-line currents can be considered as state variables with first-order dynamics dependent on the sources and time constant of impedance L_2 / R_2 Therefore, transmission-line active and reactive powers present first-order dynamics and have a strong relative degree of one , since from the control viewpoint, its first time derivative already contains the control variable (the strong relative degree generally represents the number of times the control output variable must be differentiated until a control input appears explicitly in the dynamics).

In the DPC control ,sliding mode control theory is used, to control the P and Q variables with a relatively strong degree of one can be obtained considering proportionality to a linear combination of the errors of the state variables .Therefore, define the active power error e_P and the reactive power error e_Q as the difference between the power references the actual transmitted powers p_{ref}, Q_{ref} and P,Q respectively.

$$e_p = P_{ref} - P \qquad (1)$$

$$e_Q = Q_{ref} - Q \qquad (2)$$

Then, the robust sliding surfaces $S_p(e_p,t)$ an $S_Q(e_Q,t)$ must be proportional to these errors, being zero after reaching the sliding mode

$$S_p(e_p,t) = K_p(P_{ref} - P) = 0 \quad (3)$$

$$S_Q(e_Q,t) = k_Q(Q_{ref} - Q) = 0 \quad (4)$$

converter sides, while guaranteeing that the active power exchanged on the UPFC series connection is always supplied/absorbed by the shunt connection .The proportional gain K_p and k_Q are chosen to impose appropriate switching frequencies.

B. Line Active and Reactive Direct Switching Loss

Direct power control must ensure that the sending end power follows power references .The nonlinear law, based on errors e_Q and e_P is selected in the real time by using DPC in the matrix converter switching states. High control speed is achived because where there are no modulators and pole zero-based approaches. To maintain stability for active and reactive power controllers, the conditions of sliding-mode stability (5) and (6) should be verified

$$S_P(e_P,t)\dot{S}_P(e_P,t) < 0 \qquad (5)$$

$$S_Q(e_Q,t)\dot{S}_Q(e_Q,t) < 0 \qquad (6)$$

These conditions mean that if $S_P(e_P,t) > 0$.then the $S_P(e_P,t)$ value must be decreased meaning that its time derivative should be negative $S_Q(e_Q,t) < 0$. According to (3) and (5), the criteria to choose the matrix vector should be

1. .If $S_P(e_P,t) > 0 \Rightarrow \dot{S}_p(e_p,t) < 0 \Rightarrow P < P_{ref}$ then choose a vector suitable to increase P

2. If $S_P(e_P,t) > 0 \Rightarrow \dot{S}_p(e_p,t) < 0 \Rightarrow P > P_{ref}$ then choose a vector suitable to decrease P

3. If $S_P(e_P,t) = 0$ then a vector is chosen for the real power to not be changed.

The same procedure is applied for the reactive power error. The six vectors of group I cannot not be used to design the DPC control system, that is why they require extra algorithms to calculate their time-varying phase [6]. From group II, the variable amplitude vectors, only the 12 highest amplitude voltage vectors are certain to be able to assure the previously discussed required levels of V_{Lp} and V_{Lq} required, to satisfy their conditions. In group III the lowest amplitude voltages vectors, and three null vectors, could be used for near zero errors[2].

If the control errors ep and eq are quantized using two hysteresis comparators, each with three levels (-1, 0 and +1), combinations of nine output voltage error are obtained. The shunt reactive power is controlled by using two-level comparator (-1 and 1) then 18 error combinations will be defined, then enabling the selection of 18 vectors. selecting one out 18 vectors is sufficient so for the three zero vectors have a minor influence on the shunt reactive power control.

Using the same reasoning for the remaining eight active and reactive power error combinations and generalizing it for all other input voltage sectors, Table I is obtained .These P, Q controllers were designed based on control laws not dependent on system parameters, and dependent only on the errors of the controlled output to ensure robustness to parameter variations, operating conditions, allow system order reduction, and minimizing response time.

III. DIRECT CONTROL OF MATRIX CONVERTERS INPUT REACTIVE POWER

The minimum or a certain desired reactive power is obtained at the matrix converter UPFC can be controlled to ensure a at the matrix converter input. In the same way to the previous considerations, since the voltage source input filter dynamics has a strong relative degree of two, then a suitable sliding surface $SQ(eQ,t)$ will be a linear combination of the desired reactive power error $eQi = Qiref - Qi$ and its first order time derivative The time derivative can be approximated by a discrete time difference, a suitable switching frequency is given by choosing KQi, as stated

$$S_{Qi}\left(eQi,t\right) = \left(Q_{iref} - Q_i\right) + k_{Qi}\frac{d}{dt}\left(Q_{iref} - Q_i\right) \qquad (7)$$

Supposing that there is enough i_q amplitude, (7) are used to establish the criteria (8) to choose the ad- equate matrix input current vector that imposes the needed sign of the matrix input-phase current i_q related to the output-phase current by

1. If $S_{Qi}\left(e_{Qi},t\right) > 0 \Rightarrow S_{Qi}\left(e_{Qi},t\right) < 0$ then choose a vector current $i_q < 0$ to increase Q_i
2. If $S_{Qi}\left(e_{Qi},t\right) < 0 \Rightarrow S_{Qi}\left(e_{Qi},t\right) > 0$ then choose a vector current $i_q > 0$ to decrease Q_i

Before, this sliding surface needs to be quantized only in two levels (-1 and +1) using one hysteresis comparator. The sliding mode is reached when vectors applied to the converter have the necessary iq current amplitude to satisfy stability conditions. Therefore, to choose the most adequate vector in the chosen dq reference frame, it is necessary to know the output currents location since the iq input current depends on the output currents (Table I). Considering that the dq–axis location is synchronous with the input voltage (i.e., dq reference frame depends on the input voltage location), the sign of the matrix reactive power Qi can be determined by knowing the location of the input voltages and the location of the output current.

IV. MATRIX CONVERTER

The matrix converter (MC) is an AC to AC conversion and development of the fully controlled converter based on bi-directional fully controlled switches, PWM voltage control, as mentioned earlier. With the initial progress made by Venturing it has received considerable attention in recent years as it provides a good alternative to the double-sided PWM voltage source rectifier inverters having the advantages of being a single stage converter with only nine switches for three-phase to three-phase conversion and inherent bi-directional power flow, sinusoidal input/output waveforms with moderate switching frequency, possibility of a compact design due to the absence of dc link reactive components, and controllable input power factor independent of the output load current. They presented the power circuit of the converter as a matrix of bidirectional power switches and they introduced the matrix converter and the low-frequency modulation matrix concept. In their modulation method, as the direct transfer function approach, the output voltages are obtained using multiplication of the *modulation matrix with the input voltages*

A.. Input Filters

Filters must be used at the input of the matrix converters to reduce the switching frequency harmonics present in the input current. The requirements for the filter are as follows: to have a cutoff frequency lower than the switching frequency of the converter; to minimize its reactive power at the grid frequency; to minimize the volume and weight for capacitors and chokes to minimize the filter inductance voltage drop at rated current in order to avoid a reduction in the voltage transfer ratio. One should be noticed that this filter does not need to store energy coming from the load. Some filter configurations like simple LC and multistage *LC* have been investigated . It has been shown that simple *LC* filtering is the best alternative considering cost and size.

B. Venturini Method

Given a set of three-phase input voltages with constant amplitude Vi and frequency f= ωi/2∏, this method calculates a switching function involving the duty cycles of each of the nine bi-directional switches and generate the three-phase output voltages by sequential piecewise sampling of the input waveforms. These output voltages follow a predetermined set of target voltage waveforms and with a three-phase load connected, a set of input currents Ii and angular frequency ωi should be in phase for unity.

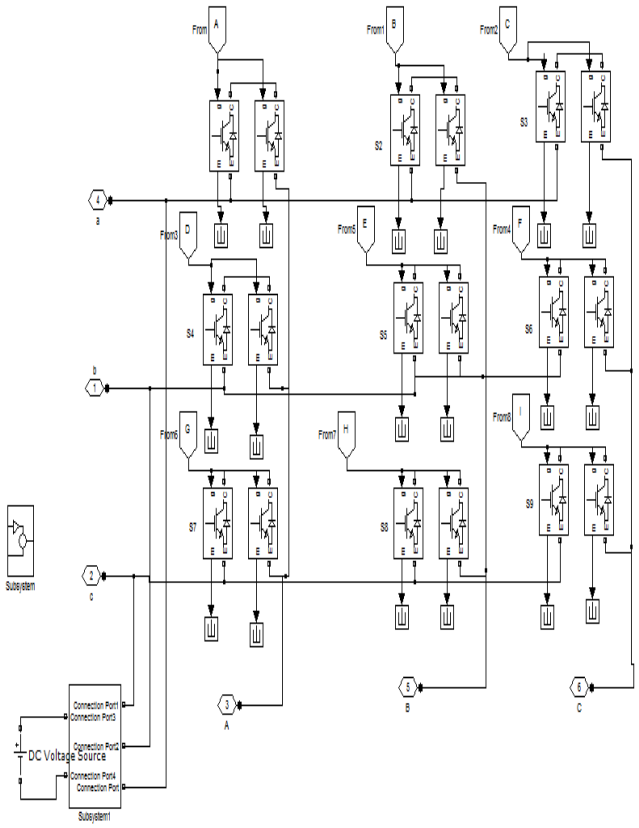


Fig 1. Modeling of matrix converter

Some problems appear during the power-up procedure of the matrix converter in the *LC* configuration of the input filter. *LC* circuit can create overvoltage during transient operation. The connection of damping resistors, to reduce over voltages is proposed in [5]. when the converter is running the damping resistors are short circuited. The use of damping resistors connected in parallel to the input reactors is proposed in [5].

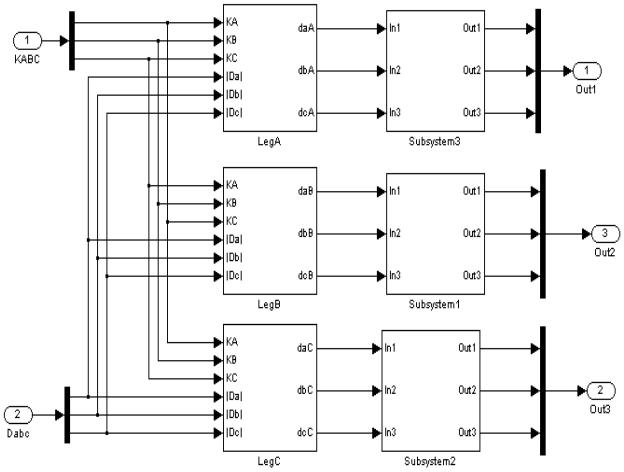


Fig 2. Modeling of vector selection block

The control of the instantaneous active and reactive powers requires the measurement of voltages and output currents necessary to calculate sliding surfaces. The output currents measurement is also used to determine the location of the input currents component. The control of the matrix converter input reactive power requires the input currents measurement to calculate . At each time instant,

the most suitable matrix vector is chosen upon the discrete values of the sliding surfaces, using tables derived from Tables I and II for all voltage sectors.

TABLE I : State-Space Vectors Selection, For Input Voltages Located At Sector

C_α	C_β	*Sector* $I_0 12 ; I_0 1$		$I_0 2 ; I_0 3$		$I_0 4 ; I_0 5$		$I_0 6 ; I_0 7$		$I_0 8 ; I_0 9$		$I_0 10; I_0 11$	
		C_{Qi}		C_{Qi}		C_{Qi}		C_{Qi}		C_{Qi}		C_{Qi}	
		+1	-1	+1	-1	+1	-1	+1	-1	+1	-1	+1	-1
-1	+1	-9	+7	-9	+7	-9	+7	+7	-9	+7	-9	+7	-9
-1	0	+3	-1	+3	-1	-1	+3	-1	+3	-1	+3	+3	-1
-1	-1	-6	+4	+4	-6	+4	-6	+4	-6	-6	+4	-6	+4
0	+1	-9	+7	-9	+7	-9	+7	+7	-9	+7	-9	+7	-9
0	0	-2	+2	+8	-8	-5	+5	+2	-2	-8	+8	+5	-5
0	-1	-7	+9	-7	+9	-7	+9	+9	-7	+9	-7	+9	-7
+1	+1	-4	+6	+6	-4	+6	-4	+6	-4	-4	+6	-4	+6
+1	0	+1	-3	+1	-3	-3	+1	-3	+1	-3	+1	+1	-3
+1	-1	-7	+9	-7	+9	-7	+9	+9	-7	+9	-7	+9	-7

V. SIMULATION RESULTS

The performance of the proposed direct control system was evaluated with a detailed simulation model using the MATLAB/Simulink Sim Power Systems to represent the matrix converter, transformers, sources and transmission lines, and Simulink blocks to simulate the control system. Ideal switches were considered to simulate matrix converter semiconductors minimizing simulation times.

Matrix converter was built by using three semiconductor modules from DANFOSS, each one with six 1200-V 25-A insulated-gate bipolar transistors (IGBTs) with an anti parallel diode in a common collector arrangement. To input filter parameter variation. The harmonics are nearly 30 dB below the 50-Hz fundamental for the line current, and 22 dB below the 50-Hz fundamental for the matrix converter current. The main harmonics are nearly 30 dB below the 50-Hz fundamental for the line current, and 22 dB below the 50-Hz fundamental for the Matrix Converter current. The power spectral density shows switching frequencies mainly below 2.5 kHz as expected. Simulation results confirm the performance of the proposed controllers, showing no cross-coupling, no steady-state error and fast response times for different changes of power references.

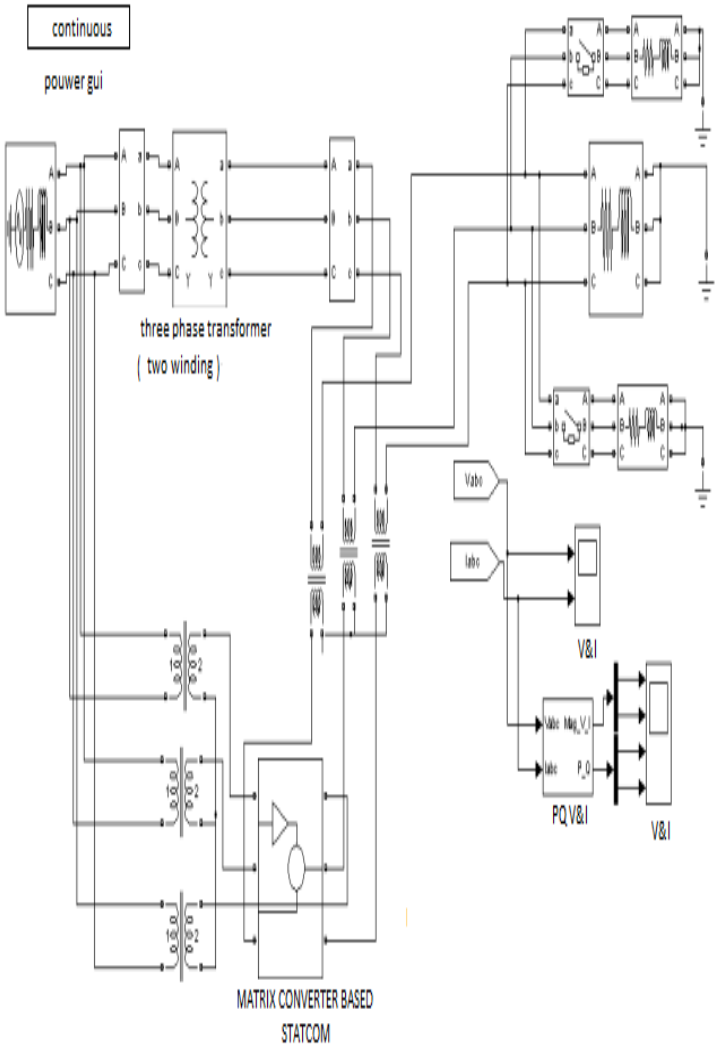
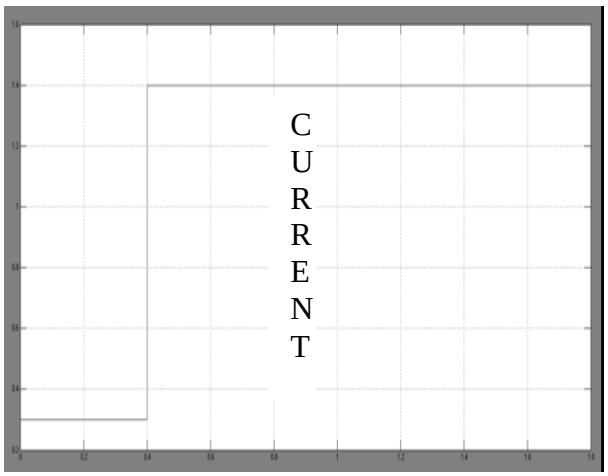


Fig 6.1 Modeling of UPFC with matrix converter

Digital Simulations are carried out in MATLAB 7.11.0 (R2010b) & was run for 10s with the controller. To get very accurate results, the step size for 3 simulations was taken to be very small. For the software implementation purposes, a 3 generator 9 bus system with 220 KV line and 100 MVA generators is considered.

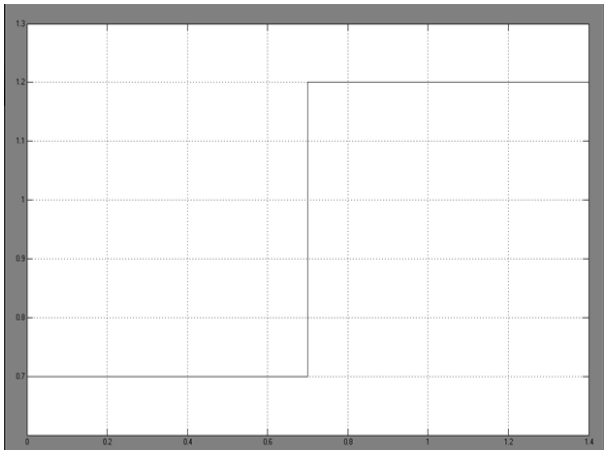
DPC controller ability to operate at lower witching frequencies, the DPC gains were lowered and the input filter parameters were changed accordingly to lower the switching frequency to nearly 1.4 kHz.



P

Time(s)

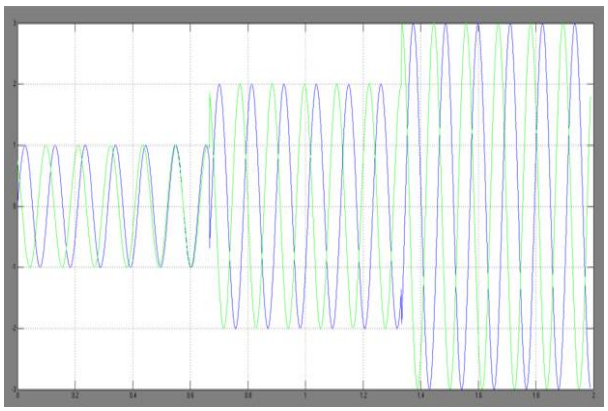
Fig 4. Active power response for Matrix converter.



Q

Time(s)

Fig 5. Reactive power response for Matrix converter



C
U
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T

Time(s)

Fig 6. Direct Power Controller

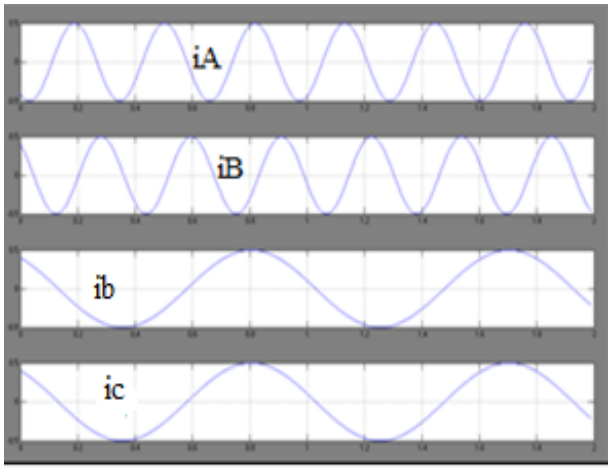

Fig 7. Line current (iA,iB) and input matrix converter current (ib,ic).

VI. CONCLUSION

In the present method the Matrix Converter based UPFC was implemented in the single bus system where as in this paper it is implemented in three bus system and derived advanced nonlinear direct power controllers were derived based on sliding mode control techniques, for matrix converters connected to power transmission lines as UPFCs. Presented simulation and experimental results show that active and reactive power flow can be advantageously controlled by using the proposed DPC. Results show no steady-state errors, no cross-coupling, insensitivity to non modeled dynamics and fast response time, thus confirming the expected performance of the presented nonlinear DPC methodology. The obtained DPC-MC results were compared to PI linear active and reactive power controllers using a modified Venturini high-frequency PWM modulator. Despite showing a suitable dynamic response, the PI performance is inferior when compared to DPC. Furthermore, the PI controllers and modulator take longer times to compute. Obtained results show that DPC is a strong nonlinear control candidate for line active and reactive power flow. It ensures transmission-line power control as well as sending end reactive power or power factor control.

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Mapping and Assessment of Land Use/ Land Cover Using Remote Sensing and GIS. Case study: Potential Area for Dates Palm in Al-Qassim Region, Central Saudi Arabia

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Abstract- Mapping of land use/land cover classes is important task to conserve natural resources and to put suitable management practices. Remote sensing techniques provide promising possibilities to map land use/land cover classes since remotely sensed data covers a large area with periodic synoptic view.

The Kingdom of Saudia Arabia contributed significantly to world date production. The Qassim region of central Saudia Arabia is considered as one of most famous producing area locally, regionally and internationally.

In this study Enhanced thematic satellite scene of path 168 and row 43 acquired on 28.08.1999 was used. Unsupervised classification approach was used. Three major land use/land cover classes were determined viz hilly area (1232566.06 ha), vegetated land (1663389.88 ha) and sand dunes (470068.08 ha). The vegetated land constitutes the first class, mainly the Wadis, and it is mainly devoted to dates production. Hilly unproductive areas constitute the second class. Finally, it is obvious that sand dunes cover a large proportion of the study area.

It is worth to mention that sand dunes surround the vegetated land, which in turn will lead to the loss of some of the productive lands through sand encroachment. This fact necessitates the adoption of suitable management practices to retard and, if possible, to stop sand encroachment.

Index Terms- Remote Sensing, Land use/Land cover, Saudia Arabia

I. INTRODUCTION

Remote sensing is becoming increasingly important for mapping land use and land cover due to specific characteristics of remotely sensed data e.g. large area coverage, good spatial resolution, accessibility to remote areas and faster interpretation with higher degree of objectivity and reproducibility. There are two broad types of image classification methods; namely, supervised and unsupervised classifications. In supervised classification approach, training samples for information classes are selected with aid of ground truth points and then are used to train the classifier, where in unsupervised classification approach the classifier determines dependently the spectral classes with the image and finally the image analyst determines the related information classes to these spectral classes. However, combined hybrid (supervised and unsupervised) classification approach is also used for land use and land cover classification (Sabins, 1996, Dafalla and Csaplovics, 2005).

The difficulties concerning land use / land cover classification by means of remote sensing in arid and semi-arid regions are well known. Since vegetation-soil-patterns in arid and semi-arid zones are characterized by a sparse distribution of non-photo synthesising vegetation (NPV) its spectral behaviour interferes with spectral signatures of bare soil patterns (Schmidt and Karnieli, 2000, Khiry et. al., 2006). Moreover, the spatial heterogeneity at pixel level strongly affects systematic separation between dominant land uses. Therefore, many studies have recommended subpixel unmixing analysis as a suitable method to overcome such constrains (Elmore, *et al.,* 2000), but still there are many difficulties to be overcome such as unavailability of spectral libraries for dominant plant species and soil types.

Mapping of land use/land cover classes is an important task to conserve natural resources and to recommend suitable management practices. Remote sensing techniques provide promising possibilities to map land use/land cover classes since remotely sensed data cover a large area with periodic synoptic view. The goal of this study is to determine the major land use/ land cover classes in this arid zone of Saudi Arabia by using remote sensing techniques with more emphasis on the effect of sand encroachment on date palm production.

II. METHODOLOGY

The Kingdom of Saudi Arabia contributes significantly to world dates production. Date Palm constitutes about 90% of the cultivated land (Shahean, 1990). The average dates production is in the range of 600-650 tons/acre (Al-Suliem, 2003). The study area, Qassim Region of Central Saudi Arabia, is considered as one of the most famous dates producing area locally, regionally and internationally. The region is known to be the alimental basket of the country because of its agricultural assets. Al-Qassim (26° 18´ N, 43° 58´ E) has a land height of 600-750 m above sea level generally gliding from west to east. The climate

is a typical desert climate with cold rainy winter and hot summer with low relative humidity (http:en.wikipedia.org/wiki/al_Qasim.).

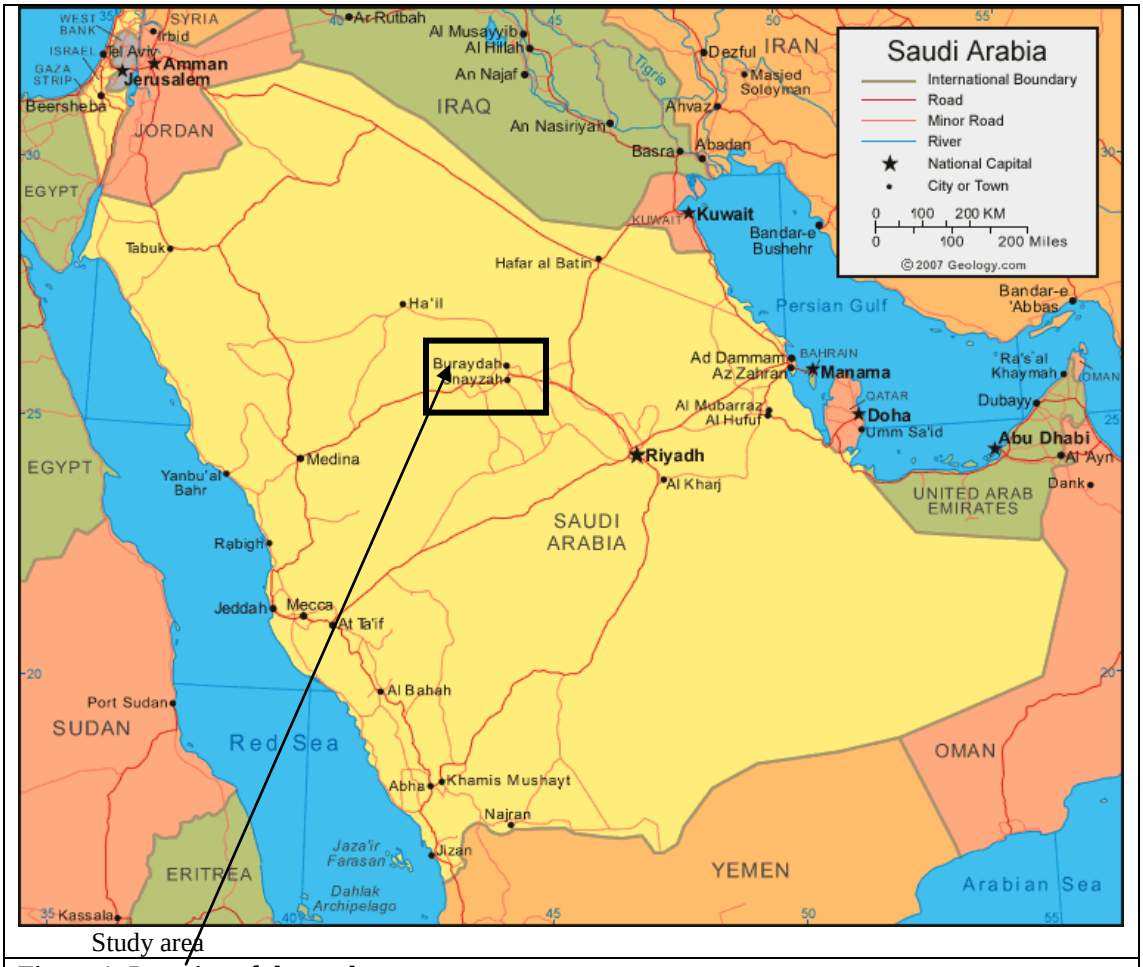

Figure 1: Location of the study area
Source: http://geology.com/world/saudi-arabia-satellite-image.shtml

III. METHODOLOGY

In this study geometrically and radiometrically corrected Enhanced Thematic Mapper Plus (ETM+) scene of path 168 and row 43 acquired on 28/08/1999 was used. The image was freely downloaded from Global Land Cover Facility (http://glcf.umiacs.umd.edu/data/). Visible bands 1, 2, in addition to near infrared band 4, 5, 7 were stacked to form false composite image (fig. 2). Visual interpretation was carried out to determine the major apparent land use/land cover patterns. Three patterns, hilly area, *Wadi* vegetation and sand dunes, were recognized and then used as inputs during the process of unsupervised classification.

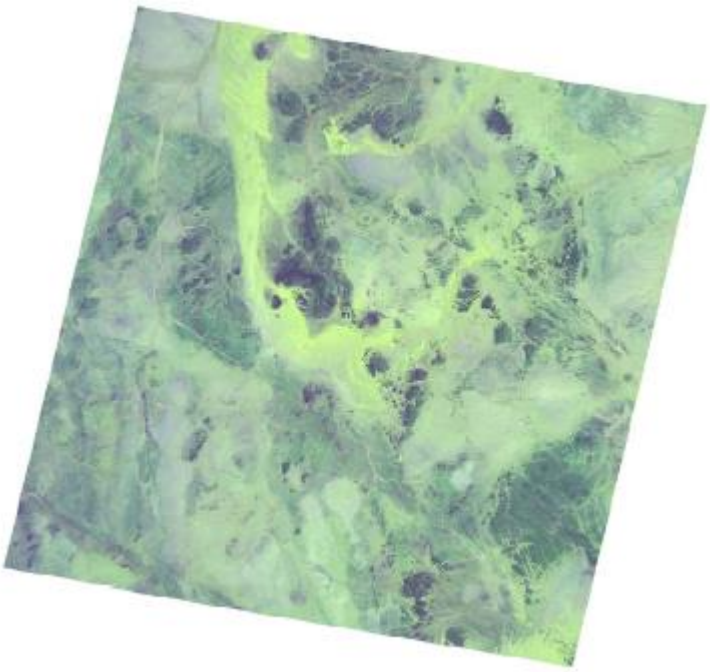

Figure 2: False composite image (RGB of bands 4,7 and 1) of the study area

IV. RESULTS

Figure (3) and Table (1) showed that three major land use/land cover classes were determined viz hilly area (1232566.06 ha), vegetated land (1663389.88 ha) and sand dunes (470068.08 ha). The vegetated land constitutes the first class, mainly the *Wadis*, and it is mainly devoted to dates production. Hilly unproductive areas constitute the second class. Finally, it is obvious that sand dunes cover a large proportion of the study area.

It is worth to mention that sand dunes surround the vegetated land, which in turn will eventually lead to the loss of some of the productive lands through sand encroachment. This fact necessitates the adoption of suitable management practices to retard and, if possible, to stop sand encroachment through sand dunes fixation, establishment of shelter belt and wind breakers.

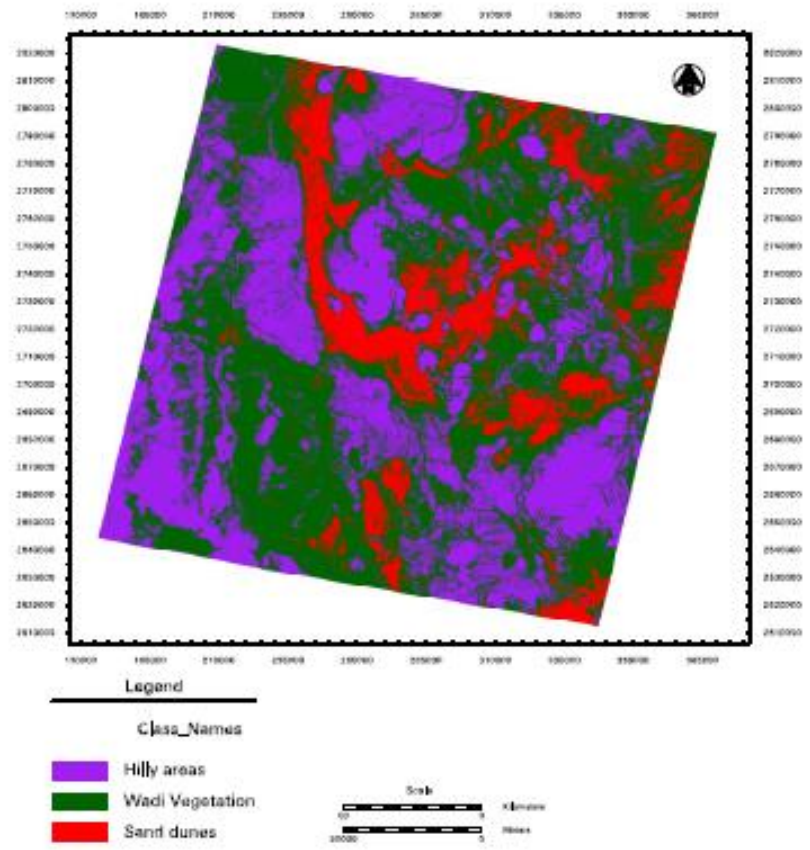

Figure 3: Major land use/land cover classes in the study area

Table 1: Areas and percentages of the major land use/land cover classes in the study area

Class name	Area (ha)	%
Wadi Vegetation	1663389.88	49.42
Hilly areas	1232566.06	36.62
Sand dunes	470068.08	13.97

V. CONCLUSION AND RECOMMENDATION

This study ends up with the following conclusions:

1. Remote sensing techniques can be used efficiently as time- and cost-efficient tool in land use/land cover mapping and assessment.
2. The Wadi vegetation covers the large portion of Al-Qassim region.
3. Sand dunes threaten the wadi vegetation in the Al-Qassim region.

This study stipulates the following recommendation:

1. Use of remotely sensed data and remote sensing techniques in monitoring natural resources in the regions, Saudi Arabia has many advantages, since the Kingdom had launched its satellite and host a receiving station for the American land resource observation satellites.
2. Adoption of sand dunes fixation ranging from sand dunes fixation to establishment of wind breakers.
3. Encourage of in depth detailed studies related to simulation of future changes in the land use/land cover in the region.

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Survey on Recent Clustering Algorithms in Wireless Sensor Networks

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Abstract- The use of wireless sensor networks (WSNs) has grown enormously in the last decade, pointing out the crucial need for scalable and energy-efficient routing and data gathering and aggregation protocols in corresponding large-scale environments. To maximize network lifetime in Wireless Sensor Networks (WSNs) the paths for data transfer are selected in such a way that the total energy consumed along the path is minimized. To support high scalability and better data aggregation, sensor nodes are often grouped into disjoint, non overlapping subsets called clusters. Clusters create hierarchical WSNs which incorporate efficient utilization of limited resources of sensor nodes and thus extends network lifetime. The objective of this paper is to present a survey on clustering algorithms reported in the literature of WSNs. This paper presents taxonomy of energy efficient clustering algorithms in WSNs.

Index Terms - Clustering algorithms, Energy efficient clustering, Network lifetime, Wireless sensor networks.

I. INTRODUCTION

A wireless sensor network is a collection of nodes organized into a cooperative network [4]. Each node consists of processing capability (one or more microcontrollers, CPUs or DSP chips), may contain multiple types of memory (program, data and flash memories), have a RF transceiver (usually with a single omnidirectional antenna), have a power source (e.g., batteries and solar cells), and accommodate various sensors and actuators. The nodes communicate wirelessly and often self-organize after being deployed in an ad hoc fashion. Systems of 1000s or even 10,000 nodes are anticipated. Such systems can revolutionize the way we live and work.

As sensor networks have limited and non-rechargeable energy resources, energy efficiency is a very important issue in designing the topology, which affects the lifetime of sensor networks greatly.

II. CLUSTERING

The grouping of sensor nodes into clusters has been widely pursued by the research community in order to achieve the network scalability objective. Every cluster would have a leader, often referred to as the cluster-head (CH). Although many clustering algorithms have been proposed in the literature for ad-hoc networks, the objective was mainly to generate stable clusters in environments with mobile nodes. Many of such techniques care mostly about node reachability and route stability, without much concern about critical design goals of WSNs such as network longevity and coverage. Recently, a number of clustering algorithms have been specifically designed for WSNs. These proposed clustering techniques widely vary depending on the node deployment and bootstrapping schemes, the pursued network architecture, the characteristics of the CH nodes and the network operation model. A CH may be elected by the sensors in a cluster or pre-assigned by the network designer. A CH may also be just one of the sensors or a node that is richer in resources. The cluster membership may be fixed or variable. CHs may form a second tier network or may just ship the data to interested parties, e.g. a base-station or a command center.

III. THE CLUSTERING PROBLEMS

Assume that N nodes are dispersed in a field. Our goal is to identify a set of cluster heads which cover the entire field. Each node $u_{i,}$ where $1 \le i \le N$, is then mapped to exactly one cluster c_j. where $1 \le j \le N_c$, and N_c is the number of clusters ($N_c \le N$). The node can directly communicate with its cluster head (via a single hop).

The following requirements must be met:

1) Clustering is completely distributed. Each node independently makes its decisions based on local information.

2) Clustering terminates within a fixed number of iteration (regardless of network diameter).

3) At the end of each T_c. each node is either a cluster head or a non-head node (which we refer t_n as regular node) that belongs to exactly one cluster.

4) Clustering should be efficient in terms of processing complexity and message exchange.

5) Cluster heads are well-distributed over the sensor field.

IV. CLUSTERING ALGORITHMS

A. Single-Level Clustering Algorithm

Each sensor in the network becomes a clusterhead (CH) with probability *p* and advertises itself as a clusterhead to the sensors within its radio range. We call these clusterheads the volunteer clusterheads. This advertisement is forwarded to all the sensors that are no more than *k* hops away from the clusterhead. Any sensor that receives such advertisements and is not itself a clusterhead joins the cluster of the closest clusterhead. Any sensor that is neither a clusterhead nor has joined any cluster itself becomes a clusterhead; we call these clusterheads the forced clusterheads. Because we have limited the advertisement forwarding to *k* hops, if a sensor does not receive a CH advertisement within time duration *t* (where *t* units is the time required for data to reach the clusterhead from any sensor *k* hops away) it can infer that it is not within *k* hops of any volunteer clusterhead and hence become a forced clusterhead. Moreover, since all the sensors within a cluster are at most *k* hops away

from the cluster-head, the clusterhead can transmit the aggregated information to the processing center after every *t* units of time.

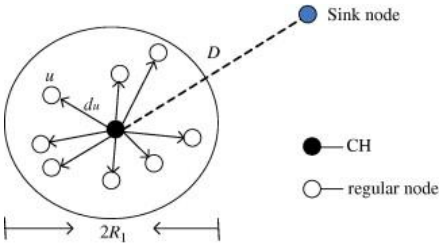

Fig. 1. A level-1 cluster with radius R_1 and number of nodes N_1. In the clustering scheme, CH locates in the center of this cluster. *D* is the distance between CH and sink node.

The limit on the number of hops allows the cluster-heads to schedule their transmissions. Note that this is a distributed algorithm and does not demand clock synchronization between the sensors. The energy used in the network for the information gathered by the sensors to reach the processing center will depend on the parameters *p* and *k* of our algorithm. Since the objective is to organize the sensors in clusters to minimize this energy consumption, we need to find the values of the parameters *p* and *k* of our algorithm that would ensure minimization of energy consumption. We observe[3] that the algorithm leads to significant energy savings. The savings in energy increases as the density of sensors in the network increases.

B. Hierarchical Clustering Algorithm

In previous section, we have allowed only one level of clustering; this algorithm allows more than one level of clustering. Assume that there are *h* levels in the clustering hierarchy with level 1 being the lowest level and level *h* being the highest. In this clustered environment, the sensors communicate the gathered data to level-1 clusterheads (CHs). The level-1 CHs aggregate this data and communicate the aggregated data or estimates based on the aggregated data to level-2 CHs and so on. Finally, the level-h CHs communicate the aggregated data or estimates based on this aggregated data to the processing center. The cost of communicating the information from the sensors to the processing center is the energy spent by the sensors to communicate the information to level-1 clusterheads (CHs), plus the energy spent by the level-1 CHs to communicate the aggregated information to level-2 CHs, …, plus the energy spent by the level-h CHs to communicate the aggregated information to the information processing center.

Algorithm

The algorithm [3] works in a bottom-up fashion. The algorithm first elects the level-1 clusterheads, then level-2 clusterheads, and so on. The level-1 clusterheads are chosen as follows. Each sensor decides to become a level-1 CH with certain probability p_1 and advertises itself as a clusterhead to the sensors within its radio range. This advertisement is forwarded to all the sensors within k_1 hops of the advertising CH. Each sensor that receives an advertisement joins the cluster of the closest level-1 CH; the remaining sensors become forced level-1 CHs.

Level-1 CHs then elect themselves as level-2 CHs with a certain probability P_2 and broadcast their decision of becoming a level-2 CH. This decision is forwarded to all the sensors within k_2 hops. The level-1 CHs that receive the advertisements from level-2 CHs joins the cluster of the closest level-2 CH. All other level-1 CHs become forced level-2 CHs. Clusterheads at level 3,4,...H are chosen in similar fashion, with probabilities $P_3,P_4,...P_h$ respectively, to generate a hierarchy of CHs, in which any level-i CH is also a CH of level (i-1), (i-2),…, 1.

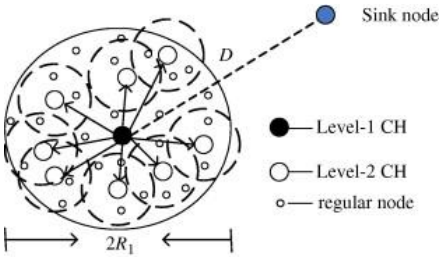

Fig. 2. A level-1 cluster with radius R_1. Each level-2 cluster is approximated as the circle-shaped region, and CH locates in the center of that cluster. *D* is the distance between level-1 CH and sink node. In this level-1 cluster, the number of level-2 CHs is denoted as N_{CH2}.

The experimental results[3] of this algorithm shows that in networks of sensors with higher communication radius, the distance between a sensor and the processing center in terms of number of hops is smaller than the distance in networks of sensors with lower communication radius and hence there is lesser scope of energy savings. The energy savings with increase in the number of levels in the hierarchy are also observed to be more significant for lower density networks. This can be attributed to the fact that among networks of same number of sensors, the networks with lower density has the sensors distributed over a larger area. Hence, in a lower density network, the average distance between a sensor and the processing center is larger as compared to the distance in a higher density network. This means that there is more scope of reducing the distance traveled by the data from any sensor in a non-clustered network, thereby reducing the overall energy consumption.

Since data from each sensor has to travel at least one hop, the minimum possible energy consumption in a network with n sensors is *n*, assuming each sensor transmits 1 unit of data and the cost of doing so is 1 unit of energy. The density of sensors and their commun-ication radius. Hence, if one chooses to store the numerically computed values of optimal probability in the sensor memory, only a small amount of memory would be needed.

C. Low Energy Adaptive Clustering Hierarchy(LEACH)

LEACH [5] minimizes energy dissipation in sensor networks due to constructing clusters. This protocol does not consider node's residual energy in the clustering process.

LEACH operation is done in two phases, setup phase and steady state phase. In the setup phase, a sensor node selects a random number between 0 and 1. If this number is less than the threshold T(n), the node becomes a CH. T(n) is computed as:

$$T(n) = \begin{cases} \frac{p}{1-p\times\left(r \bmod \frac{1}{p}\right)} & if\ n \in G \\ 0 & otherwise \end{cases}$$

where r is the current round; p, the desired percentage for becoming CH; and G, is the collection of nodes that in the last 1/p rounds have not been elected as a CH. After electing CHs, every CH announces all sensor nodes in the network that it is the new CH. When each node receives the announcement, it chooses its desired cluster to join based on the signal strength of the announcement from the CHs to it. So, the sensor nodes inform their appropriate CH to join it. Afterwards, the CHs based on a TDMA approach assign the time slot to each node so that a member can send its data to its CH in this period. The sensor nodes can initiate sensing and transmitting data to the CHs during the steady state phase. The CHs also aggregate data received from the nodes in their cluster before sending these data to the BS via a single hop fashion.

D. Hybrid Energy Efficient Distributed Clustering[HEED]

Younis and Fahmy [1] proposed an iterative clustering protocol, named HEED. HEED is different from LEACH in the way CHs are elected. Both, electing the CHs and joining to the clusters, are done based on the combination of two parameters. The primary parameter depends on the node's residual energy. The alternative parameter is the intra cluster "communication cost". Each node computes a communication cost depending on whether variable power levels, applied for intra cluster communication, are permissible or not. If the power level is fixed for all of the nodes, then the communication cost can be proportional to

(i) node degree, if load distribution between CHs is required, or

(ii) 1/node degree, if producing dense clusters is required.

In this approach, every regular node elects the least communication cost CH in order to join it. On the other hand, the CHs send the aggregated data to the BS in a multi hop fashion.

HEED periodically selects cluster heads according to a hybrid of their residual energy and secondary parameter, such as node proximity to its neighbors or node degree. HEED does not make any assumptions about the distribution or density of nodes, or about node capabilities, e.g., location-awareness. The clustering process terminates in O(1) iterations, and does not depend on the network topology or use. The protocol incurs low overhead in terms of processing cycles and messages exchanged. It also achieves fairly uniform cluster head distribution across the network. A careful selection of the secondary clustering parameter can balance load among cluster heads. HEED outperforms weight-based clustering protocols in terms of several cluster characteristics. HEED prolongs network lifetime, and the clusters it produces exhibit several appealing characteristics. HEED parameters, such as the minimum selection probability and network operation interval, can be easily tuned to optimize resource usage according to the network density and application requirements. HEED can also be useful in multi-hop networks if the necessary conditions for connectivity (the relation between cluster range and transmission range under a specified density model) hold to the network density and application requirements. HEED can also be useful in multi-hop networks if the necessary conditions for connectivity (the relation between cluster range and transmission range under a specified density model) hold.

HEED (hybrid energy-efficient distributed clustering) is an iterative clustering protocol that uses information about the nodes' remaining energy and their communication costs in order to select the best set of cluster head nodes. During the clustering process, a sensor node can be either a tentative cluster head, a final cluster head, or it can be covered (meaning that it has heard an announcement message from a final cluster head node). At the beginning of the clustering phase, a node with higher remaining energy has a higher probability CH_{prob} of becoming a tentative cluster head. If the node becomes a tentative cluster head, it broadcasts a message to all sensor nodes within its cluster range to announce its new status. All nodes that hear from at least one tentative cluster head choose their cluster head nodes based on the costs of the tentative cluster head nodes. For this purpose, the authors in[7]define the average reachability power(AMRP), which is a cost metric used to ''break ties" in the cluster head election process. The AMRP of a node u is defined as the mean of the minimum power levels required by all M nodes within the cluster range to reach the node u

$$\mathrm{AMRP}(u) = \frac{\sum_{i=1}^{M} \mathrm{Min}Pwr(i)}{M}.$$

During each iteration, a node that is not ''covered" by any final cluster head can elect itself to become a new tentative cluster head node based on its probability CH_{prob}. Every node then doubles its CH_{prob} and goes to the next step. Once the node's CH_{prob} reaches 1, the node can become a final cluster head, or it can choose its cluster head as the least cost node from the pool of final cluster head neighbors. If the node completes HEED execution without selecting its final cluster head, then it considers itself uncovered and becomes a final cluster head for the upcoming round. Once the clusters are formed, all sensors send their data to the cluster head, where the data are aggregated into a single packet. The cluster head nodes form a network back-bone, so packets are routed from the cluster head nodes to the sink in a multi-hop fashion over the cluster head nodes.

E. Hausdorff Clustering

It involves three parts. First, nodes organize themselves into several static clusters by the Hausdorff clustering algorithm [7] based on node locations, communication efficiency, and network connectivity. Second, clusters are formed only once, and the role of the cluster head is optimally scheduled among the cluster members. We formulate the maximum lifetime cluster-head scheduling as an integer programming problem and propose a greedy algorithm for its solution. Third, after cluster heads are selected, they form a backbone network to periodically collect, aggregate, and forward data to the base station using minimum energy (cost) routing. This method can significantly lengthen the network lifetime when compared with other known methods.

With increasing node density, the energy consumption per node increases because there is more need for the local exchange of messages and for the radio channels to compete. WuLi and Hausdorff clustering are seen to be very energy efficient.

The duration of WAF clustering is much longer than that of the others because WAF requires nodes to sequentially communicate through a ranking order. In addition, WAF requires significant information exchanges between potential gateways and cluster heads for backbone formation. The clustering times for the other three protocols are all modest. It should be noted that Hausdorff clustering is initiated by one node and extended to other nodes (i.e., sequential operation). However, the clustering procedure is carried out only once. Overall, Hausdorff clustering offers a good compromise between the conflicting requirement of smaller number of clusters, energy consumption per node, and average clustering time.

V. CONCLUSION

In this paper we have examined the current state of proposed clustering protocols. In wireless sensor networks, the energy limitations of nodes play a crucial role in designing any protocol for implementation. In addition, Quality of Service metrics such as delay, data loss tolerance, and network lifetime expose reliability issues when designing recovery mechanisms for clustering schemes. These important characteristics are often opposed, as one often has a negative impact on the other.

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Direct and Derivative Spectrophotometric Determination of Cobalt (II) using 3,4-DihydroxyBenzaldehyde-1-(3-Chloro-2-Quinoxalinyl)Hydrazone(DHBCQH) in Presence of Micelle Medium

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Abstract- A rapid and sensitive method has been developed for the determination of Cobalt(II) based on complexation reaction between the metal ion and 3,4,-dihydroxybenzaldehyde-1-(3-Chloro-2-Quinoxalinyl)Hydrazone (DHBCQH) in the presence of non-ionic surfactant Tween-80. The important parameters affecting the analytical procedure were optimized. Absorption maximum for a ternary complex was noted at 450 nm. The reaction was found to be rapid at room temperature and absorbance remained constant for more than 24hs. The method obeys Beer's law in the range 29.46 to 132.59ng/ml. The apparent molar absorptivity of 4.45×105 L mol-1 cm-1 and Sandell's sensitivity 0.13ng/ml. The effect of foreign ions was tested by taking a constant concentration of metal ion and determining its concentration in the presence of $\geq$ 100 folds in excess of foreign ions. The method was successfully used in the determination of Cobalt(II) in Biological samples. Second order derivative spectrophotometric method were developed at λmax = 482nm for the determination of Cobalt(II), which was more sensitive than the zero order method.

Index Terms- Spectrophotometric Determination, Cobalt (II), 3,4-DHBCQH, Surfactant Tween-80, Biological sample, Alloy sample..

I. INTRODUCTION

Dissolved cobalt occurs in the environment at concentrations ranging from 0.5 to 12µg/L in the sea water up to 100 µg/L in wastewater [1,2] . Calculation of the inorganic complexation of cobalt using an ion-paring model and stability constants [3] shows that it is weakly complexed by inorganic ligands, the predominate inorganic species being Co(II) and its chloride complexes. On the other hand, there is evidence that cobalt is strongly complexed by organic ligands especially in burdened waters and solids [4, 5].

A literary survey reveals that there are several techniques and methods for the determination of cobalt in different samples including atomic fluorescence [6], atomic absorption [7, 8], Chromato graphy [9, 10], x-rays fluorescence [11], inductively coupled plasma –atomic emission spectrometry [12] and spectrometry [13, 14]. Among the most widely used analytical methods are those based on the Uv-visable spectrophotometric techniques due to the resulting experimental rapidity, simplicity and wide application. 3,4-Dihydroxybenzaldehyde-1-(3-chloro-2-quinoxalinyl) hydrazone (DHBCQH) is a new chromogenic regent recently synthesized and has been applied to the spectrometric determination of trace Cobalt(II) . it is shown that DHBCQH reacts with Cobalt(II) to form a stable , water-soluble ,positively charged, 1:2 binary complex. Water –miscible organic solvents such as N,N-dimethyl formamide (DMF),aceton , dioxide and their aqueous solution (higherthan 20%) may be used for solubilizing the chromogenic reagent and increasing the sensitivity of the chromogenic reaction. The recent literature on the analytical Applications of the entitled reagent 3,4–dihydroxybenzadehyde-1-(3-Chloro-2-Quinoaxalinyl)Hydrazone abbreviated as (3,4 –DHBCQH) has revealed no study on the use of reagent for Cobalt(II) determination. Therefore, the goals of the present manuscript are focused on the synthesis and spectroscopic characterization(Uv-Vis IR and H NMR, Mass) of the DHBCQH reagent. Moreover the stiochiometry of the formed cobalt (II)-DHBCQH chelate was elucidated in an attempt to develop and accurate method for the analysis of cobalt(II) in Biological samples.

Recently enhanced sensitivity in spectrophotometer was achieved by utilizing the ability of certain surfactants to sensitize the binary complexes of the metal ion with chromogenic ligands[15,16] sensitization are result of the replacement of acidic protons of the liganded dye molecule by surfactant[17] (or) adsorption of the metal reagent complex on the micelles of the surfactant [18] cationic, anionic and nonionic surfactants are often used to sensitize the metallochromic indicators . In this context, updated surfactant – sensitized reaction has recently been developed in spectrophotometer [19-23].

The present study was hence planned to suggest a very simple and reasonably good method for determination cobalt (II) ions at low concentration, using the DHBCQH reagent. DHBCQH as binary complex and sensitizing the reagent with Tween- 80 as ternary complex using spectrophotometer which is still frequently used because of its low cost and simplicity.

II. EXPERIMENTAL

2.1. Apparatus

The Absorbance and pH measurements were made on a Shimadzu UV-visible spectrophotometer (Model UV -160A) fitted with 1cm Quartz cells and Philips digital pH meter (model L1 613 respectively.)

2.2. Reagents and solution

All chemicals used were of analytical reagents grade or the highest purity available. Double distilled deionized water, which is non-absorbent under ultraviolet radiation, was used throughout. Glass vessels were cleaned by soaking in acidified solutions of $KMnO_4$ or $K_2Cr_2O_7$, followed by washing with concentrated HNO_3 and rinsed several times with deionized water.

2.2.1 Tween-80 solution 1%. . A 100 mL of Tween-80 solution was prepared by dissolving 1 mL of pure tween-80 in 100 mL if doubly distilled deionized water, sonicated for 15 min and diluted with deionized water when it became transparent.

2.2.2. 3,4-Dihydroxybenzaldehyde-1-(3-chloro-2-quinoxalinyl)hydrazone (DHBCQH) (1 X 10^{-2}mol/dm^3): A 25 mL solution was prepared by dissolving 0.0783 g of recrystalised sample in dimethylformamide and it was suitably diluted to get the required concentration.

2.2.3. Cobalt(II) standard solution (1 X 10^{-2}mol/dm^3): A 100 mL stock solution (1 mg/mL) of divalent mercury was prepared by dissolving 0.291 g of cobalt nitrate (Merck, Darmstadt) in deionized water containing 1 – 2 mL of nitric acid (1 + 1). More dilute standard solution was prepared from this stock solution, as and when required. .

2.3. Procedure

2.3.1. Direct spectrophotometry

In each of set of different 10 mL volumetric flasks, 5 ml of buffer solution (pH 5.0),1 mL of DHBCQH(1 x 10^{-3}mol/dm^3) and various volumes of 1x 10^{-6}mol/dm^3 cobalt(II)finally added 1 mL of 1% Tween-80 and f solution were taken and made up to the mark with double distilled water. The absorbance was measured at 450nm against the reagent blank. The calibration plot was prepared by plotting the absorbance against the amount of cobalt(II).

2.3.2. Second order derivative spectrophotometry

For the above solutions, second order derivative spectra was recorded with a scan speed of fast (nearly 2400 nm min^{-1}); slit width of 1 nm with nine degrees freedom, in the wavelength range 390-510 nm. The derivative amplitude measured at wavelength 482 nm and plotted against amount of cobalt (II) to obtain the calibration.

The calibration graph follows the straight line equation Y = aC +b [24]; where C is the concentration of the solution, Y is measured absorbance or peak or valley height and a and b are constants. By substituting the corresponding experimental data substituted in the above equation, the calibration equations were calculated as A_{450} = 0.065C + 0.051 for zero order method ,A_{482} = 0.0013C+0.002 for second order derivative method.

III. Results and Discussion

3.1. Factors affecting the absorbance

3.1.1. Absorption spectra

The absorption spectra of the cobalt (II)-3,4-DHBCQH system in a micellar medium is recorded using a spectrophotometer. The absorption spectra of the Co(II)-3,4-DHBCQH is a symmetric curve with the maximum absorbance at 450mm and an average molar absorption coefficient of 4.45 x 10^5L mol^{-1} cm^{-1} (Fig. 1). The reagent blank exhibited negligible absorbance, despite having a wavelength in the same region. In all instances, measurements were made at 450 nm against a reagent blank.

3.1.2. Effect of surfactant: Of the various surfactants [nonionic{poly oxyethylenedodecylether(Brij -35), Polyoxyethylenesorbitanmonoplamitate (Tween-40), Polyoxyethylenesorbitanmoni-oleate(Tween-80), Triton-X-100} ; cationinc[cetyltrimethylamoniumbromite(CTAB) } : and anionic { cetlylpridinium chloride (CPC) , sodium dodecyl sulfate SDS } studied Tween-80 was found to be the best surfactant for the system. In a 1%(V/L) Tween-80 medium however, the maximum absorbance was observed hence, a 1%(V/L) Tween-80 solution was used in the determination procedure.

Different volumes of 1%(V/L)Ttween-80 were added to a fixed metal ion concentration, and the absorbance was measured according to the standard procedure. It was observed that at 58.93 ng/ml Co(II)-chelate metal. 0.25-1.75 mL of 1%(V/V) Tween-80 produced a contestant absorbance of the Co(II)-chelate(Fig. 2). A greater Excess of Tween-80 was not studied. For all subsequent measurements, 1mL 1%(V/V) Tween-80 was added.

3.1.3. Effect of Acidity

The absorbance of the complex reaches a maximum over a pH range of 4.0 to 6.0 The higher pH values may lead to hydrolysis of Co(II) -3,4-DHBCQH system. The absorbance was at a maximum and constant when a 10 ml of solution (1mgL^{-1}; path length ,1) contained 1-7 mL (pH5.0) of acidic buffer at room temperature (25±5 ℃) . Outside this range of acidity , the absorbance decreased (Fig. 3) . For all subsequent measurements 5.0 ml (pH 5.0)acidic buffer was added.

3.1.4. Effect of time

The reaction is very fast. Constant maximum absorbance was obtained just after dilution to volume, and remained strictly unaltered 24hs.

3.1.5. Effect of reagent concentration:

Different molar excess of 3,4 –DHBCQH were added to a fixed metal ion concentration, and the absorbances were measured according to the standard procedure. It was observed that at 58.93 ng/ml Co(II) metal(optical path lenth , 1 cm) reagent molar ratios 1:10 and 1:100 produced a constant absorbance of the Co(II)-Chelate(Fig. 4) . A greater excess of the reagent was not studied. For all subsequent measurement 1mL of 1 x 10^{-3}mol/dm^3 3, 4 –DHBCQH reagent was added.

3.1.6. Calibration graph (Beers Law and sensitivity)

The Calibration curve for the determination of Cobalt has been constructed (Fig. 5) using the optimum experimental conditions. The straight line calibration curve indicates that Beers' law is obeyed at least over a range of 29.46 to132.59 ng of Cobalt (II) per10 mL. Linear regression analysis of the calibration curve gives a correlation coefficient of 0.9968. The apparent molar absorptivity(ϵ_{450}) calculated from the slop of regression line is 4.45 X 10^5L mol^{-1} cm^{-1} The Sandell's

sensitivity(concentration for 0.001 absorbance unit) was found to be 0.132 ng cm^{-2}.

3.1.7. Effect of foreign ions:

The effect of various foreign ions that are generally associated with cobalt (II) on the determination under optimum conditions developed was studied and the results are presented in Table 1. Cations like Ba(II), Sr(II) and U(VI) do not have any effect on the complex of Co(II)-DHBCQH, when present up to 4,900 µg, Zinc (II) and Tin (II) do not interfere in the determination of Cobalt(II), even present up to 4,700µg. Pb(II) and Mn(II) can be tolerated up to 3,600µg. Fe(III), Cu(II) and Hg(II) interference in the determination of Cobalt(II), even when present in trace amounts.

3.1.8. Composition of the Metal –Reagent complex

Jobs method of continuous variation (Fig.6) and the molar – ratio method were applied to ascertain the stoichiometeric composition of the complex. A Co (II) -3, 4-DHBCQH(1:2) complex was indicated by both methods. The stability constant was determined by Jobs method as $9.67X10^{6}$.

IV. Applications

The Present method was successfully applied to the determination of cobalt (II) in biological samples and alloy steels.

4.1. Determination of Cobalt (II) in biological samples (Tea leaf and vehicle exhaust)

The tea leaf samples were supplied by Andhrapradesh Agricultural research institute (APARI), Hyderabad (A.P.) India. The vehicle exhaust sample was collected from Environment protection training and research institute (EPTRI), Hyderabad (A.P.) India. A 0.1 g of tea leaf sample was taken in a beaker and dissolved in conc. nitric acid (5 mL) with heating. The solution was cooled, diluted and filtered. The filtrate was made up to 100 mL water in a calibrated flask. Vehicle exhaust particles (1g) were dissolved in a mixture of 18 mL of conc. nitric acid, 18 mL of conc. perchloric acid and 2 mL of conc. hydrofluoric acid in a 100 mL Teflon beaker, evaporated to a small volume, filtered through a filter paper and made up to 100 mL with distilled water. and the absorbance was measured at 450nm. The amount of cobalt was calculated from a predetermined calibration plot and the results are presented in Table 2.

4.2. Determination of Cobalt (II) in Alloy steel

A known aliquots of the sample solutions were taken in different 10ml volumetric flasks, 5 ml of buffer solution (pH 5.0). 1270µg of iodine (to mask Fe(II)) and 1860 µg ETDA (to mask Ni(II)) and 1ml of DHBCQH (8×10^{-3} M) were added to each flask. The contents were made up to the mark with distilled water and the absorbance was measured at 450nm. The amount of cobalt was calculated from a predetermined calibration plot and is given in Table 3.

4.3. Second order derivative method

The second order derivative curve recorded (Fig. 7) for experimental solution showed the derivative amplitude were measured at 482 nm (peak) for different concentrations of Cobalt(II) and plots were made between the amount of Co(II) and the derivative amplitude. The plots were linear and obeyed Beer`s law in the range 11.78-70.71 ng/mL at 482 nm respectively.

4.4. Effect of foreign ions

The effect of various cations and anions on the derivative amplitude was studied and it was noticed that all the ions that did not interfere in the zero order determinations of Co(II) also did not interfere in second order derivative method. The metal ions Cu(II),Hg(II) and Fe(II) were interfered in 3200 µg excess, in zero order method. But in second order derivative method. Cu(II) and Fe(II) were tolerate up to 3700µg excess.

V. Application

5.1. Analysis of alloys and steels

A known aliquot of the sample solution was taken in a 10ml volumetric flask containing 5ml buffer solution (pH 5.0) and 1270 µg ml^{-1} of iodide (to mask Fe(II)) and 1ml of DHBCQH ($8x10^{-3}$M) solution. The contents of the flask were made up to the mark with distilled water. The second derivative spectrum of the solution was recorded. The derivative amplitude at 482 nm was measured and the amount of cobalt (II) was computed from the predetermined calibration plot and presented in Table 4.

5.2. Comparison of results

The analytical characteristics of the zero and second order derivative methods in the present investigations for Cobalt(II) were compared and presented in Table 5. The results in this table reveal that second order derivative method is more sensitive and selective than the zero order method.

The sensitivity of the zero method in the present investigations was compared with those of some reported methods and presented in Table 6.

VI. Conclusion

This method offers several interesting features such as simplicity, rapidity, and low cost besides sensitivity. The number of associated elements does not interfere in the determination. The selectivity of the reagent is also improved by the use of suitable masking agents to suppress the interference of metal ions like Fe(III), Cu(II), and Hg(II). Hence the proposed method is recommended for the determination of Cobalt(II) with DHBCQH in the presence of micells by spectrophotometric method, at minor and trace levels, besides its use for analysis of real samples such as biological samples .

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Table 1. Effect of Foreign ions in (58.93 ng/mL) amount of Cobalt(II)

Ions	Tolerance limit(µg)
Ba^{+2},Sr^{+2},U^{+6}	4900
Zn^{+2},Sn^{+2}	4700
Pb^{+2},Mn^{+2}	3600
Fe^{+3},Cu^{+2},Hg^{+2}	3200
F^-,Cl^-,CH_3COO^-,	5600
I^-,So_4^{-2},Br^-,HCO_3^-,SCN^-	5100

Table 2. Determination of Co(II) in biological samples

Sample	Amount of cobalt (mg/g)	
	Certified value	Present method*
Tea leaves	0.12	0.10
Vehicle exhaust	3.3	3.25

*Average of five determinations

Table 3. Determination of Co(II) in Alloy steel samples

Sample	Amount of Co(II) (%)		Error (%)
	Certified	Found*	
Eligiloy M-1712 (40% Co, 20% Cr, 15% Ni, 15% Fe, 2% Mn, 0.05% C, 0.05%Be)	40.00	39.48	-1.3
BCS-483 (1.94% Co,10.8% W,3.21% Cr, 0.5%V,0.29% Mn,0.17% Mo,rest Fe)	1.94	1.93	-0.5

*Average of five determinations

Table 4. Determination of Co(II) in Alloy steel samples

Sample	Amount of Co(II) (%)		Error (%)
	Certified	Found*	
High speed steel (6-18% W, 5.5% Mo, 4.15% Cr, 0.4% Mn, 0.35% Si, 0.05% S, Rest Fe)	9.26	9.29	+0.32
Udimet-700 (15% Cr, 4.3%Al,5.21% Mo,0.08% C,0.003% B)	18.00	18.10	+0.55

*Average of five determinations

Table 5. Comparison of results of Cobalt(II)

Parameter	Zero order	Second derivative
Analytical wave length(nm)	450	482
Beer`s law range(ng/mL)	29.46-132.59	11.78- 70.71
Angular coefficient(m)	0.065	0.013
Y-ntercept(b)	0.005	0.002
Correlation coefficient(r)	0.996	0.998
Standared deviation(s)	0.023	0.028

Table 6. Coparison with other methods

Reagent	λ_{max}(nm)	Molar absorptivity(ε) (L mol^{-1} cm^{-1})	Ref.
3-hydroxy piconaldehyde thiosemicarbazone	450	0.78 x 10^4	25
Phthaldehyde thiosemicarbazone	385	0.56 x 10^4	26
Piconaldehyde thiosemicarbazone	410	0.74 x 10^4	27
Salicylaldehyde thiosemicarbazone	400	1.1 x 10^4	28
Acenaphthaquinone thiosemicarbazone	410	0.48 x 10^4	29
Biacetyl monoxime thiosemicarbazone	325	0.49 x 10^4	30
3,4-Dihydroxybenzaldehyde-1-(3-chloro-2-quinoxalinyl)hydrazone(DHBCQH)	450	4.45 x 10^5	**Present work**

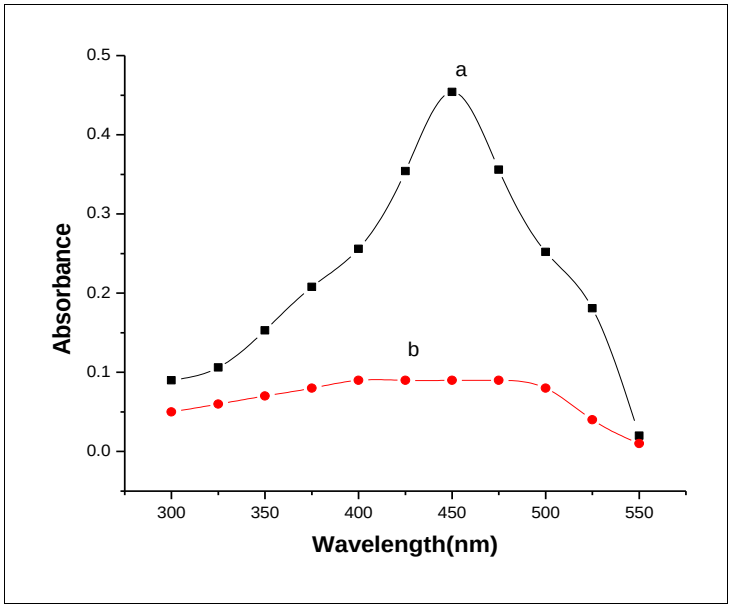

Fig. 1 a and b absorption spectra of Co(II)-DHBCQH and reagent blank system (λ_{max} = 450nm) in nonionic micellar medium

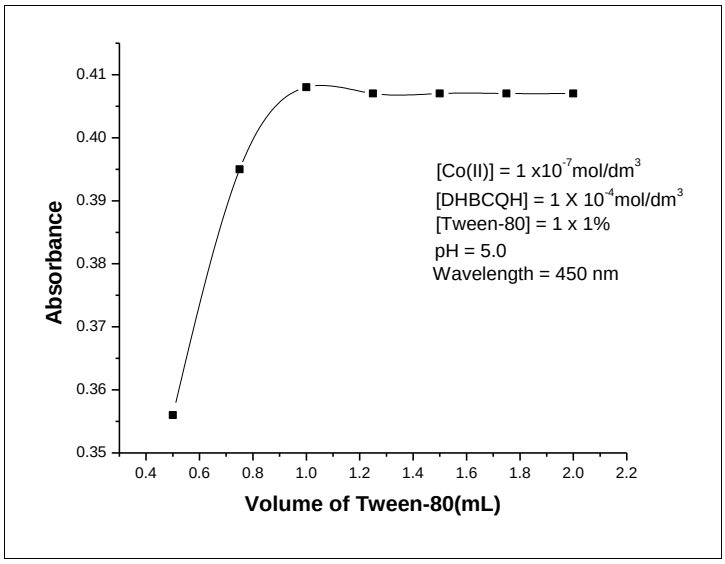

Fig. 2 Effect of surfactant on the absorbance of the Cobalt(II)-DHBCQH system

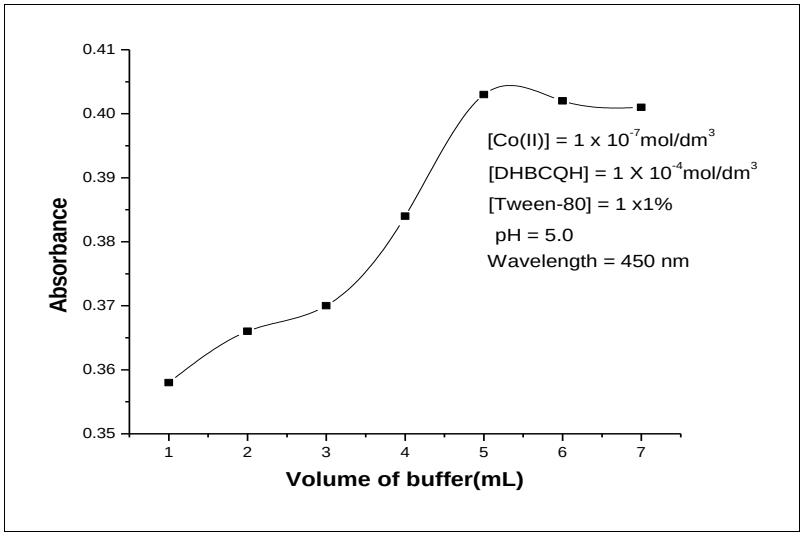

Fig. 3 Effect of Buffer solution on the absorbance of the Cobalt (II)-DHBCQH system

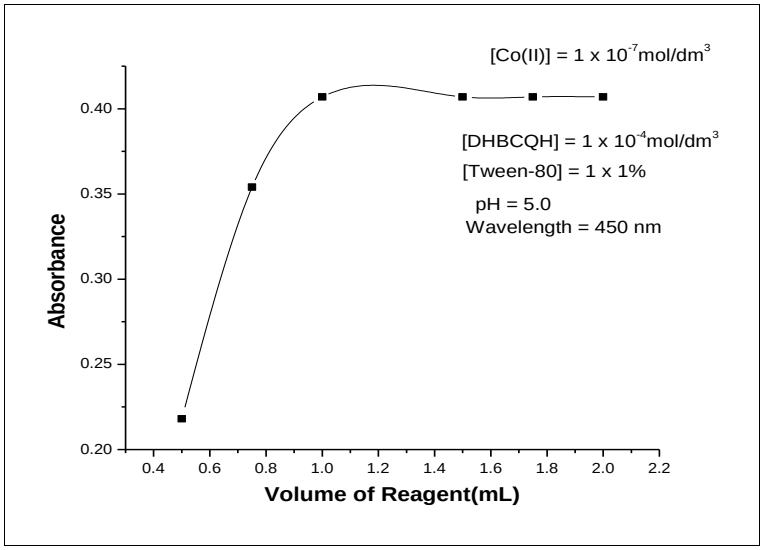

Fig. 4 Effect of reagent(DHBTSC) on the absorbance of the Cobalt(II)-DHBCQH system.

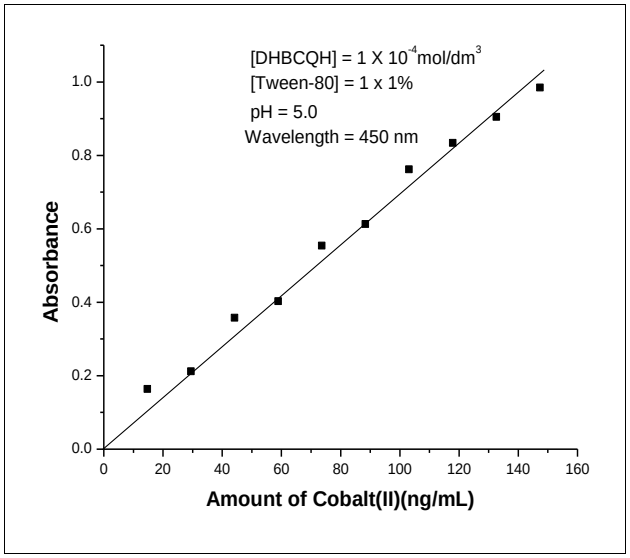

Fig. 5 Calibration graph

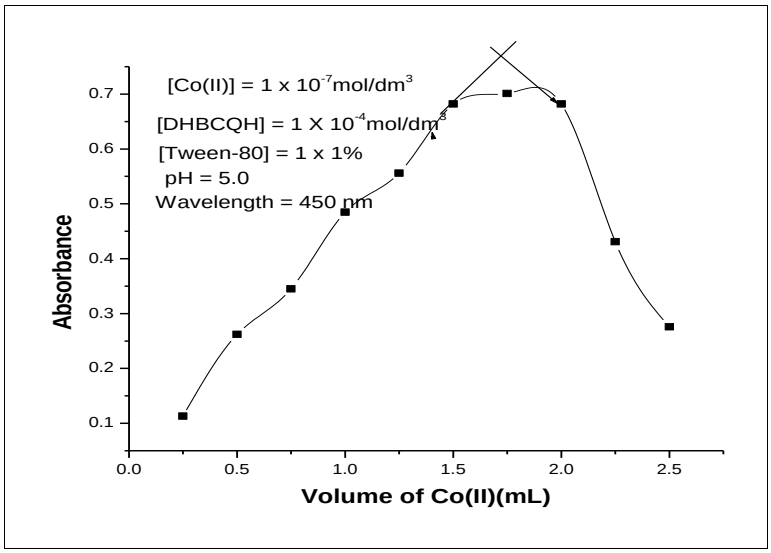

Fig. 6 Job`s curve

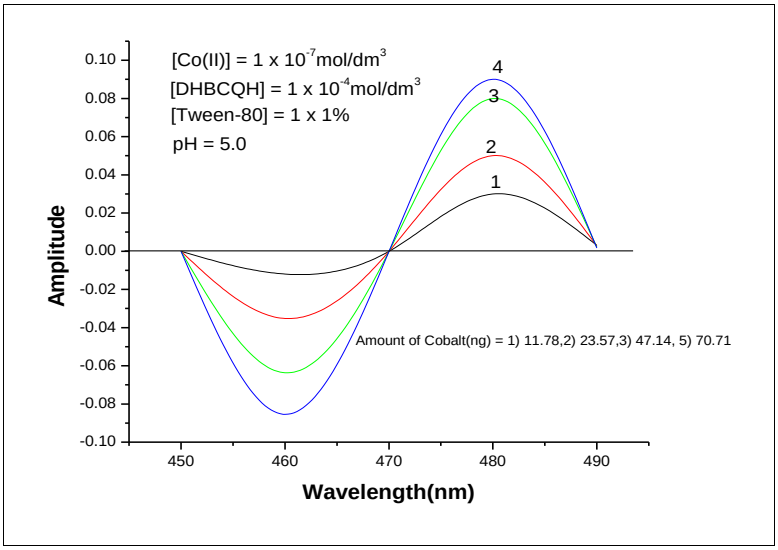

Fig. 7 Second derivative spectra of Cobalt (II)-DHBCQH Vs reagent blank

Palmprint Recognition using Phase Symmetry

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Abstract- This paper proposes an automated system for recognizing palmprints for biometric identification of individuals. Palmprint images are converted to the frequency domain using 2D DFT and thereafter bandpass filtered using a log-Gabor filter to extract the phase symmetry information. Classification is done on basis of correlation between training and testing set images. The approach is tested over a data set of 200 images divided into 10 classes and seen to provide 100% recognition accuracy. A contemporary technique is also implemented on the current dataset for comparison of accuracy results.

Index Terms- Palmprint recognition, Biometric identification, Phase symmetry

I. INTRODUCTION

Biometrics helps to identify individuals based on their physiological and behavioral characteristics, which can be used for their personal identifications. Various physical characteristics like retina patterns, iris patterns, fingerprint patterns, palmprint patterns, facial features etc. are utilized for such purposes. Palmprint recognition involves identifying an individual by matching the various principal lines, wrinkles and creases on the surface of the palm of the hand. The basis for using palmprints lies in the fact that since palmprint patterns are generated by random orientations of tissues and muscles of the hand during birth, no two individuals have exactly the same palmprint pattern. Advantages of using palmprints include the fact that such patterns remain more or less stable during one's lifetime and also that reliable images of the palm can be obtained quite easily using standard digital imaging techniques. Challenges in palmprint recognition include building a reliable data model to represent the randomly oriented lines of the palm with sufficient accuracy as also to find ways of comparing the models generated from different palmprint images with reliability. Added to this is the fact that different images can have different rotational variation, size, lighting conditions and resolution. This paper presents an efficient algorithm for palm print recognition by utilizing phase symmetry obtained from the images. The organization of the paper is as follows: section 2 provides an overview of the related works, section 3 outlines the proposed approach, section 4 details the experimentations done and results obtained, section 5 analyses the current work vis-à-vis contemporary works, section 6 brings up the overall conclusions and future scopes.

II. RELATED WORK

A number of techniques have been proposed for palmprint recognition over the years. Some of the earlier works are based on Gabor wavelets and transforms for feature extraction [1, 2]. Neural networks have been used for classification in some cases [3] while in others features like entropy have been combined with wavelet based features to improve recognition [4]. PCA and LDA based techniques have been proposed to reduce dimensions of wavelet based features [5, 6]. Another technique frequently used for plamprint recognition is the use of Zernike moments [7]. Neural networks have been used to classify Zernike based features [8]. Features based on a combination of Zernike moments and Gabor filters have also been proposed [9]. A number of approaches based on shape and contour based techniques for modeling the principal lines have been proposed. Modeling of principal lines using curvelet transforms [10], Radon transform [11], Fourier transform [12] and wavelet transforms [13] have been studied. Alignment techniques for the principal lines have been used in [14] to improve recognition rates. Image moments have been used to model palmprint textures in [15, 16] while in [17] features based on hand geometry have been used in combination with palmprint textures. Another class of works have used multiple features extracted from the palmprint lines and creases, often using edge detection techniques and thereafter correlation filters to classify them [18, 19]. In some works binary pattern vectors have been used for classification. A binary co-occurrence vector have been used in [20] while local binary patterns (LBP) have been proposed in [21, 22]. Other related techniques involve combination of features extracted from 2D and 3D images [23], using specialized symbolic representations [24], specialized distances [25], classifiers like SVM [26] and using hand shapes by truncating the fingers [27]. A comparative study of several palmprint recognition techniques is presented in [28].

III. PROPOSED APPROACH

A. Phase

The current work proposes a novel scheme for palmprint recognition by utilizing phase information inherent in images. If any image is converted from spatial domain to the frequency domain, it can be seen that the energy of the magnitude is basically concentrated only in the center, while the phase is distributed through all frequencies [29]. Furthermore, for a finite length signal, phase information alone might be sufficient to completely reconstruct a signal to within a scale factor. The images are converted to frequency domain, using the 2D Fourier Transform, which is the series expansion of an image function

over the space domain in terms of orthogonal basis functions. The definitions of the forward transform Γ and the inverse transform Γ^{-1} are given below :

$$\Gamma:\ F(u,v) = \sum_{x=0}^{M-1}\sum_{y=0}^{N-1} f(x,y).e^{-j2\pi\left(\frac{ux}{M}+\frac{vy}{N}\right)} \quad (1)$$

$$\Gamma^{-1}: f(x,y) = \frac{1}{MN}\sum_{u=0}^{M-1}\sum_{v=0}^{N-1} F(u,v).e^{j2\pi\left(\frac{ux}{M}+\frac{vy}{N}\right)} \quad (2)$$

Here $f(x,y)$ is the image and is real, but $F(u,v)$ is in general complex. Generally, $F(u,v)$ is represented by its magnitude and phase rather that its real and imaginary parts, where the magnitude $|F(u,v)|$ and phase $\varphi(F(u,v))$ are represented as:

$$|F(u,v)| = \sqrt{\{Re(F(u,v))\}^2 + \{Im(F(u,v))\}^2} \quad (3)$$

$$\varphi(F(u,v)) = tan^{-1}\left\{\frac{Im(F(u,v))}{Re(F(u,v))}\right\} \quad (4)$$

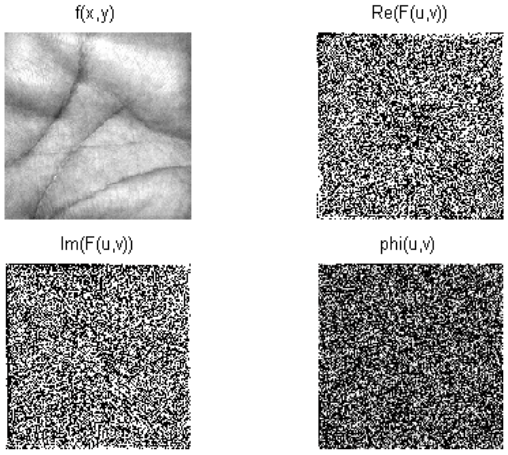

Figure 1: Frequency domain components of a sample image

Briefly, the magnitude depicts how much of a certain frequency component is present and the phase provides information as to where the frequency component is in the image. Figure 1 shows sample representations corresponding to quantities described above

B. Phase Symmetry

Symmetry is a useful mechanism for identification of objects in an image, without any prior recognition or segmentation of objects. An important aspect of symmetry is that objects exhibit periodicity in structure clearly noticeable in the Fourier series. Symmetric points are points with maximum or minimum values in all sine waves (phase 0 or π). Hence it is feasible to detect symmetry with phase information [30, 31].

To calculate phase symmetry, the frequency domain image is filtered using a bandpass log-Gabor filter. Log-Gabor filters are used as an alternative to standard Gabor filters used when broad spectral information are required, as maximum bandwidth of a standard Gabor filters are limited to about one octave. Suggested by Field [32], log-Gabor filters use a logarithmic frequency scale instead of a linear scale. The transfer function of a log-Gabor filter is of the form below where w_0 is the filter's center frequency.

$$G(w) = \exp\left[\left\{-log\left(\frac{w}{w_0}\right)^2\right\}/\left\{2log\left(\frac{w}{w_0}\right)^2\right\}\right] \quad (5)$$

The band-passed image in frequency domain is obtained by applying the bandpass log-Gabor filter to the frequency domain image.

$$IMF = F * G(w) \quad (6)$$

The filtered image is converted back to the spatial domain using eqn. (2). Let f represent the real part of the bandpass image in spatial domain. Then

$$f = Re\{\Gamma^{-1}(IMF)\} \quad (7)$$

For adaptive windowing of the log-Gabor bandpassed image, a monogenic filter H derived from the horizontal and vertical extents of the image, is convolved with the bandpassed image. Let h represent the window filtered image i.e.

$$h = \Gamma^{-1}(IMF * H) \quad (8)$$

The squared amplitudes of the filtered results are

$$A = \{Re(h)\}^2 + \{Im(h)\}^2 \quad (9)$$

The energy of the filtered image is calculated as the cumulative sum of the energy of the bandpassed image and the filter outputs over a number n of bandpassed versions of the image.

$$E = \sum_n \sqrt{f^2 + A} \quad (10)$$

The symmetry energy is given by

$$S = \sum_n (f - \sqrt{A}) \quad (11)$$

The normalized phase symmetry is then given by

$$P = \frac{S}{E} \quad (12)$$

The following figure shows pictorial representations of these quantities for the sample palmprint image, shown in Fig. 1.

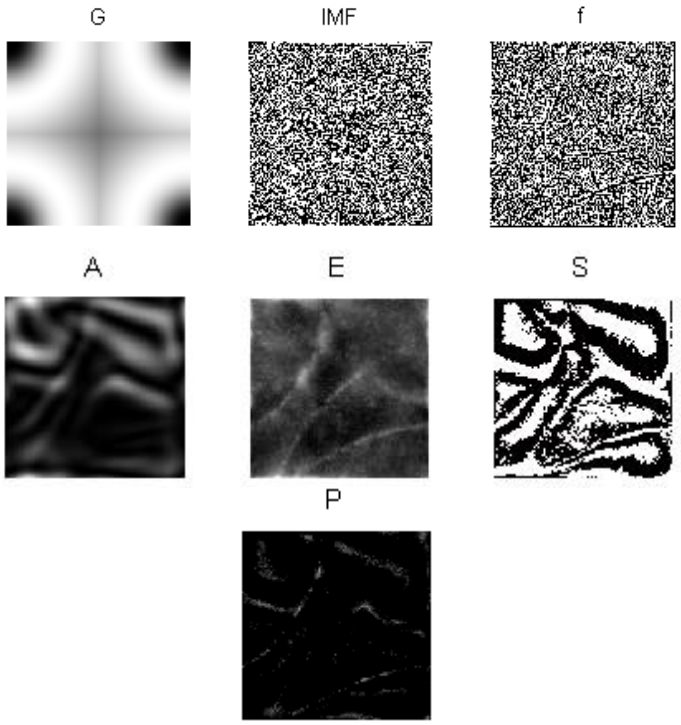

Figure 2: Phase symmetry components of a sample image

IV. EXPERIMENTATIONS AND RESULT

Experiments are conducted by using 200 palm images from the PolyU palm print database [33]. The dataset is divided into 10 classes, each class consisting of 10 training images and 10 testing images. The Region of Interest (ROI) from each image is extracted manually and resized into 160 by 160 pixels in dimension and stored in BMP format. Samples of training and testing ROI images are shown in Figures 3 and 4.

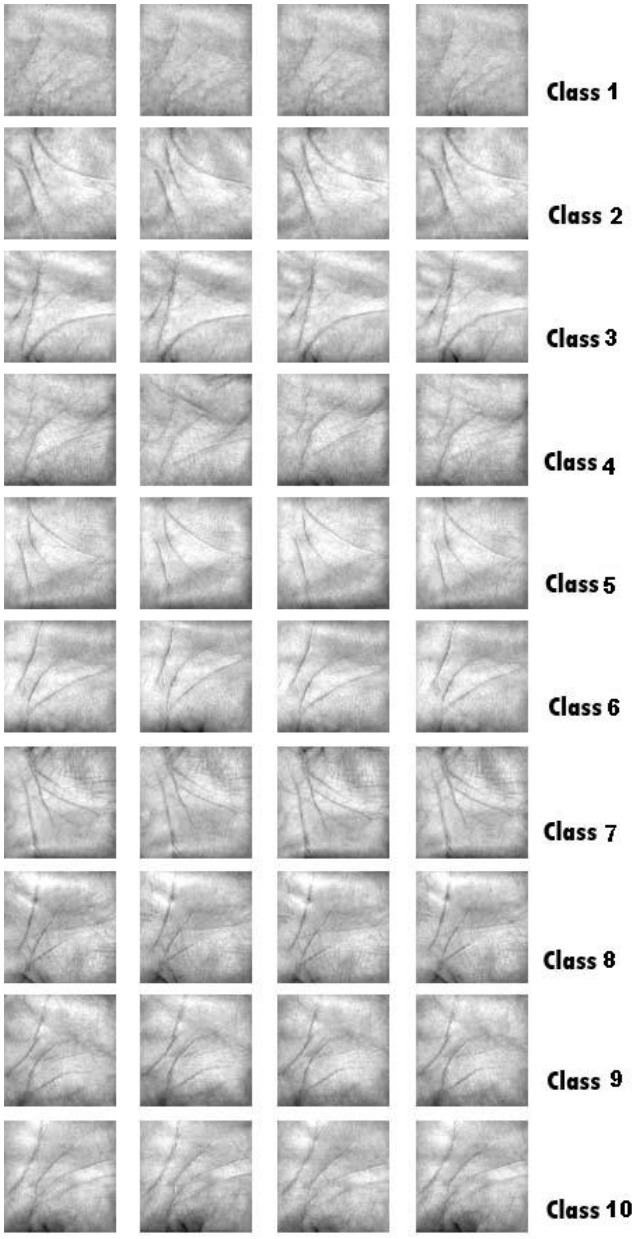

Figure 3: Samples of Training set images

The symmetry energy S and normalized phase symmetry P information is extracted from each image using the method outlined above. Number of bandpassed versions used is 5. The mean of the correlation value, between the phase and symmetry energy of the training and the testing datasets, are used for accuracy calculations. Test samples are assumed to belong to the class for which the correlation value is maximum. Let $S_{i,j}$ represent test sample j of class i. If $D_s(S_{i,j}, T_k)$ represent the mean differences in symmetry energy between test sample $S_{i,j}$ and all training samples of class k and $D_p(S_{i,j}, T_k)$ represent the same with regard to normalized phase symmetry, then to calculate recognition, the following matrices are generated for $1 \leq i \leq 10$.

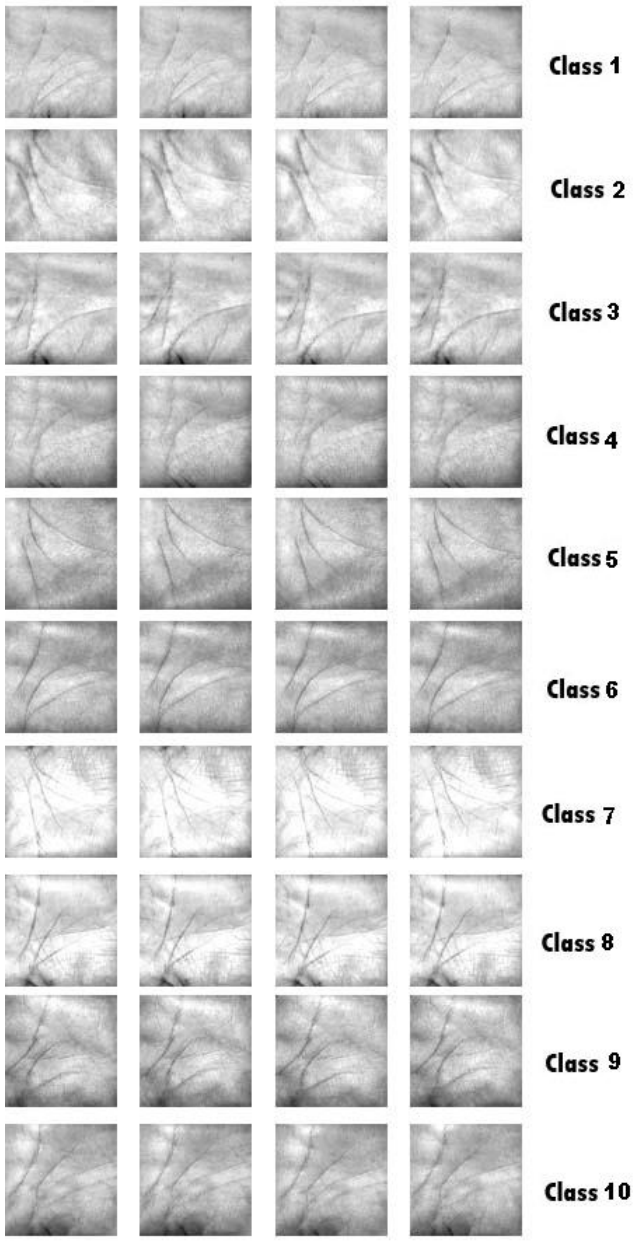

Figure 4: Samples of Testing set images

$$A_i = \begin{bmatrix} D_s(S_{i,1},T_1) & D_s(S_{i,2},T_1) & \dots & D_s(S_{i,10},T_1) \\ D_s(S_{i,1},T_2) & D_s(S_{i,2},T_2) & \dots & D_s(S_{i,10},T_2) \\ \dots & \dots & \dots & \dots \\ D_s(S_{i,1},T_{10}) & D_s(S_{i,2},T_{10}) & \dots & D_s(S_{i,10},T_{10}) \end{bmatrix}$$

$$B_i = \begin{bmatrix} D_p(S_{i,1},T_1) & D_p(S_{i,2},T_1) & \dots & D_p(S_{i,10},T_1) \\ D_p(S_{i,1},T_2) & D_p(S_{i,2},T_2) & \dots & D_p(S_{i,10},T_2) \\ \dots & \dots & \dots & \dots \\ D_p(S_{i,1},T_{10}) & D_p(S_{i,2},T_{10}) & \dots & D_p(S_{i,10},T_{10}) \end{bmatrix}$$

Since correlation between similar items is maximum, ideally row i for each of the above matrices should contain maximum values. A test sample $S_{i,j}$ is classified as belonging to class k if the k-row of A_i or B_i i.e. $D_s(S_{i,j},T_k)$ or $D_p(S_{i,j},T_k)$ is maximum for $1 \le k \le 10$.

$$S_{i,j} \rightarrow k, if\ D_s(S_{i,j},T_k) | D_p(S_{i,j},T_k)\ is\ max for\ i = k \qquad (13)$$

The classification plots for A_1 and B_1 for class-1 test samples are given below in Figure 5.

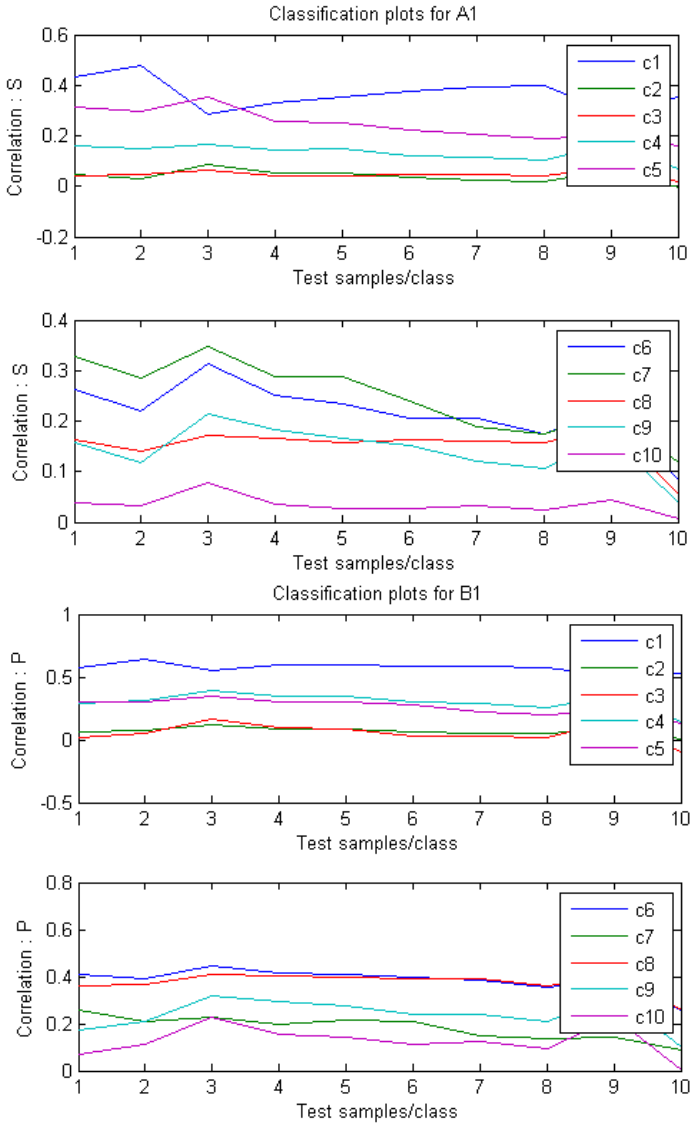

Figure 5: Classification based on S and P features for class-1

The percentage of the correct recognition rate for the testing samples of the 10 classes is shown in Table 1.

Table 1: Percent recognition accuracy using proposed approach

Class	Accuracy
Class 1	100%
Class 2	100%
Class 3	100%
Class 4	100%

Class 5	100%
Class 6	100%
Class 7	100%
Class 8	100%
Class 9	100%
Class 10	100%
Overall	100%

Thus the overall accuracy for the current dataset using phase symmetry method is 100%.

V. ANALYSIS

Automated discrimination of palmprint images belonging to 10 classes is done using phase symmetry method and the best accuracy of 100% is obtained. The CPU time required is 108 seconds for calculating accuracies of all 10 classes. To put the results in perspective with other contemporary techniques, the SAX (Symbolic Aggregate approximation) based method outlined in work [24] is also applied to the current dataset for a comparison. With this method the input of a time series is converted into a symbolic string representation. Each image in the training and testing dataset is represented with 8 characters. The SAX representation of training and testing dataset is shown below in Figure 6.

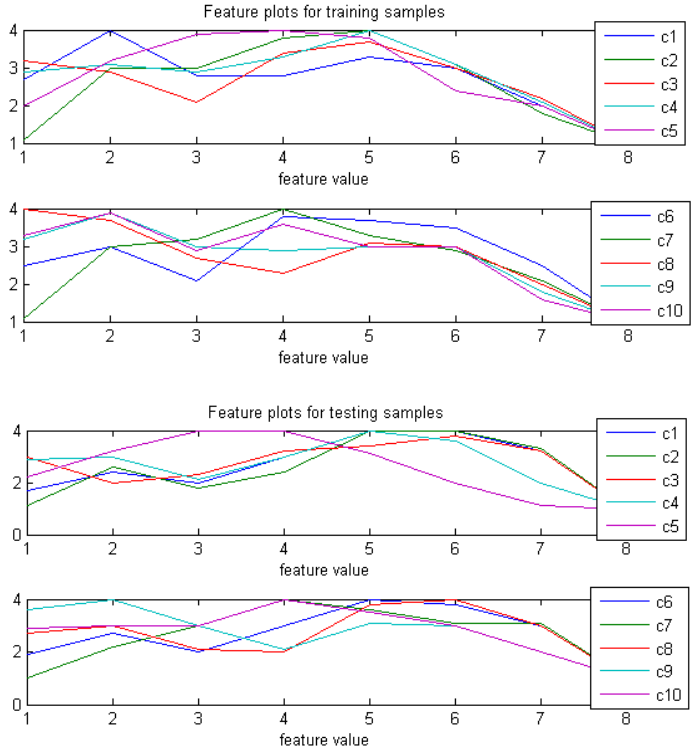

Figure 6: SAX [24] representation of training and testing samples

The minimum distance between two strings is used for comparisons between training and testing sets. The percentage of correct recognition of the testing samples using SAX on the current dataset is shown in Table 2.

Table 2: Percent recognition accuracy using SAX approach [24]

Class	Accuracy
Class 1	40%
Class 2	30%
Class 3	80%
Class 4	100%
Class 5	100%
Class 6	100%
Class 7	90%
Class 8	70%
Class 9	100%
Class 10	100%
Overall	81%

Table 2 shows that the overall accuracy using SAX representation on the current dataset is 81%. The proposed method is therefore seen to provide an improvement over it for the current dataset. Some of the best reported accuracies reported in contemporary works include 97% [1], 97% [3], 97% [4], 96% [5], 98% [6], 97.5% [8], 96% [9] etc.

VI. CONCLUSIONS AND FUTURE SCOPES

This paper proposes an efficient algorithm for palm print recognition using phase symmetry. Palmprint images are first converted to frequency domain using a 2D DFT, and then filtered using a bandpass log-Gabor filter. The bandpassed image is then convolved with a monogenic filter for adaptive windowing, and the symmetry energy is calculated from the windowed version of the image. The method is tested over a dataset of 200 palmprint images divided into 10 classes, and classification is done by using correlation values between training and testing images. A recognition accuracy of 100% is achieved for the current dataset, and CPU overheads are seen to be minimal. In comparison, accuracy values reported in contemporary literature mostly ranges between 97% to 98%. An existing technique (SAX) applied to the current dataset yields an accuracy of 81%. Hence the proposed method has potentials to be incorporated as a quick, simple and efficient palmprint classification algorithm within a biometric security system. Future work to improve the system would involve combining phase symmetry with other methods like Moment Invariance and Legendre Moments. Also statistical classifiers like neural network can be used for more robust classifications.

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Yellamma Cult and Divine Prostitution: Its Historical and Cultural Background

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Abstract- The worship of Renuka, the mother goddess can be seen in many places of India. In the southern parts of India, Renuka is worshipped in the name of Yellamma or Mariamma. Renuka, as mentioned in Puranas is the mother of Parasurama who according to Indian mythology is considered as the incarnation of Visnu and his cult also can be seen in many parts of India. But in the regions of Vindhya Mountains, worship of Renuka gets predominance. Although Renuka is considered as the mother of Parasurama in epics and Puranas, there are mythical stories, which made her a mother goddess. The rise of Renuka as a mother goddess was perhaps the result of a complex process including the merges of numerous personalities and myths. In south Indian regions like Andhra and Karnataka, there are many stories related to Renuka and she is being worshipped as Yellamma or Mariamma. The main thing to be noted in this cult is that Renuka has been more popular among the lower castes rather than among the Brahmanical groups. This may be due to the reason that the development of this cult largely occurred in the regions dominated by pastoral communities, who have several other deities with similar mythical structures. One of the interesting features in southwestern India is that the ritual performed by the community of Devadasis is associated with Renuka shrines. In this study an attempt is made to analyse how Devadasi system/prostitution is connected with the Renuka/Yellamma cult in and around Soundatti. The historical and cultural background of prostitution becoming divine in nature also will be examined.

Index Terms- Cult, Mariamma, Muttu, Parasurama, Prostitution, Renuka, Yellamma

I. Introduction

Followers of many religions in India as other civilizations worship many gods and goddess as a part of their culture. These gods and goddesses are mostly worshipped by a group of people who are gathered in a place by building an abode such as temples, shrines beneath any tree or above any platform with or without roof. These people consider the deity of the concerned place as their own guardian deity or the one who protects the village from illness and who helps them in their prosperity and goodness of their place.

Though in many places, male gods predominate the settlement, there will be a goddesses in most of the villages as the village deity, also can be called as local deity or *Gramadevata*. Though there will be 'great gods' such as *Visnu* and *Siva* as the deities in the villages, there will also be a *Gramadevata* who is considered as one who protect of the village from evil and bad happenings, often these village deities will share the names the deities from the Sanskritised form and will be identified with these deities. This is not necessarily meant that village deity or local deity does not bear any similarities to the 'great gods'[i]. In most cases where such identification exists the *Gramadevata* differs obviously from the 'Great' deity with whom it shares a name.

II. Village Goddesses of South India

A number of goddesses are worshipped in and around villages of South India. They all have several similar characteristics such as they mostly will be female who guard the village, represented by any uncarved stone, tree or a small statue in a small shrine, having direct association with the villagers, non Brahmin priests, curing disease, calming calamities etc. Mostly their names end with '*Amman*' or simply '*Amma'* meaning mother. Almost all these goddesses have same kind of myth, cult and worship in one or the other way. Even though there are many types of village goddesses worshipped in the villages of south India, all these seems to have a common base for their origin based on the *Renuka* myth described in *Puranas* or any other stories mostly having connection with this myth. It is also notable that, those goddesses who do not have direct connection will connection with it in the origin or that deity or having the same theme as explained in the *Renuka* myth.

III. Renuka Myth in Mahabharata and Other Puranas

Renuka is famous as the wife of *Jamadagni* and the mother of *Parasurama*. The earliest references to *Renuka* occur in *Mahabharata* at two places. One is in *Anusasanaparva*,where the origin of sandals and umbrella[ii] and another is in *Vanaparva* where *Renuka*'s death and her rejuvenation is explained.[iii] According to the second episode, *Jamadagni* had five sons named *Rumanvan*, *Susena*, *Vasu*, *Visvavasu* and *Parasurama*. When all these went outside, *Renuka* after getting permission of *Jamadagni* went to bath in a river. While returning, she saw *Citraratha* the king of *Martikavarta* sporting in the water along with his wife. This made her into a mood for sport. She became pale and shame after seeing this. In this condition she reached the hermitage. *Jamadagni* after noticing her and her situation knew what happened to her and came to a conclusion that she had done a sin. As the sons returned to hermitage, he ordered the sons one by one to kill *Renuka* by beheading her. The first four disobeyed. When *Parasurama* was ordered, he obeyed it and beheaded his mother. *Jamadagni* was satisfied. Pleased with

Parasurama, *Jamadagni* asked him to demand boons he desired. *Parasurama* asked six boons. 1)Mother should come back to life, 2)his brothers should be again made into human beings, 3)all should forget this incident happened, 4)he should be free from the sin of matricide, 5)no one should defeat him and 6) a long life. All this was granted by *Jamadagni*. The same incident is narrated in various Puranas too like as in *Brahmandapurana*, *Visnupurana*, *Skandapurana* etc but with some slight variations.

IV. Origin and Development of Renuka Cult

From the story mentioned above and myth narrated in many Puranas, it is clear that *Renuka* was just considered as a woman who became entirely submissive to her husband and as the mother of *Parasurama*. But now *Renuka* is worshipped as goddesses in many places especially in and around the part of Maharastra, Andhra, Karnataka and Tamilnadu. The reason behind this is clearly the mythical stories and narratives emerge or perhaps made in many places of South India.

In one of the famous story in south India is as follows. Here *Renuka* is a chaste woman and because of his power of chastity, she did not require any container or jar to fetch water. She could hold water in her palm like a ball and after her bath the cloth would flap and get dried in the air above her head, following her as she walked. Once after the bath in the river, she saw few *Gandharva*s sporting in the mid region in the water ball at her hand. Immediately she looked upon this for a moment. The next moment, the solid water ball started diminishing and got sprinkled on the ground the cloth overhead also stopped getting dry. She came to know that because she saw the *Gandharva*s sporting overhead and because of her desire, lost the power of chastity. When she reached the hermitage saw her husband with anger asked about the water ball missing and about the wet cloth. *Jamadagni* ordered *Parasurama* to take her to forest and cut her head. On the way *Renuka* saw an out caste woman passing nearby. She for protecting *Renuka*, helped her. *Parasurama* tried to separate them but was unable to do so. *Parasurama* gave a severe blow with his axe to cut his mother's head but due to strength both of the women's head got cut. After that he straightly went to *Jamadagni* and because of gladness he asked *Parasurama* to ask a boon. *Parasurama* asked the renewal of his mother's life. *Jamadagni* gave him some water to him and said that when the head was fixed to the body of the dead one, should sprinkle that water to the head and so *Renuka* will rejuvenate. *Parasurama* reached near her mother's corpse and in hurry he joined the heads wrongly i.e. his mother's head onto body of that other woman and the other woman's head onto body of his mother. They both got rejuvenated. When this news reached the people around the area, due to surprise, all of them started worshipping those women as goddesses. The one with the head of *Renuka* was named as Mariamma and other one came to be known as *Yellamma*.[iv] Another local version of this same story preserved by the devotees of *Renuka* is thus. Here a woman belonging to Mang community[v] took pity on *Renuka* and because of this she stopped *Parasurama* from cutting the head of *Renuka*. Here also *Parasurama* cut the heads of both the women. According to a myth famous in Maharastra, when *Kartaviryarjuna* killed *Jamadagni* and wounded his wife *Renuka*, *Parasurama*, before taking revenge on Arjuna, decided to take the dead body of his father for doing funeral rituals. He found that Mahur near Sahyadri ranges as a proper place. *Renuka* decide to burn herself in funeral pyre and asked her son to leave the place. He left the place but after crossing over some distance, came back to save his mother but found that the entire body of *Renuka* except the head had burn. In that place itself the head of *Renuka* was worshipped by the local people considering her as the Sati Devi.

Renuka is worshipped in different parts of south India having different names in different regions. She is worshipsped as *Yellamma*, Ekavira, Yamai, Mariamma and so on. As these goddesses became popular with a particular name and particular place or region, various myths and folk stories gradually developed around the deity in that particular area. Such stories are found in the traditionally preserved literature including folk songs.

Through these stories it is very clear that the *Renuka* becomes goddess or attain divinity. This is because of head change incident in the first myth and due to the head remaining incident in the second one. *Renuka* is worshipped as *Yellamma* and as *Mariamma* in south Indian regions.

V. Worship of Yellamma in Southern India

Yellamma is perhaps the most popular form of *Renuka* and is being worshipped throughout a very wide region covering Maharastra, Karnataka and Andhra Pradesh. Though the identification of *Yellamma* with *Renuka* is doubtful, *Renuka* Sahasranama mentions the name '*Ellamba'* as one of the names of *Renuka*. N D Kambaley opines that *Renuka Ellamba* is the Sanskritised form of the word *Yellamma*. Besides this, according to myths popular among the devotees of *Yellamma*, such a marriage with *Jamadagni*, her death by her own son *Parasurama* etc., clearly speak of her clear identity with *Renuka*[vi]. The origin of *Yellamma* is explained in many ways:

a) Once *Renuka* went to Soundatti a place where regional language is Kannada. The people asked her "*Yella Amma*". In Kannda, '*Yella'* means 'Where' and '*Amma'* means 'Mother.' *Renuka* was associated with these two words and later came to be known as *Yellamma*
b) In Karnataka, the seven deities form important part of pantheon. They are known as '*Yollumma'* (Seven Mothers or *Saptamatrkas*). In Kannada *Yollu* means seven. So *Yellamma* may be regarded as a corrupt form of *Yollumma* i.e. seven mothers.
c) In another way *Yella* means all and *Amma* means mother. The word thus means the mother of all. This may be the probable etymology. In Tamil *Yella* means border of a village or any place. The temples of *Yellamma* are generally constructed or situated in the boarders of the village and she is supposed to be the protection force of that village. So this goddess got the name *Yellamma*.
d) There is another belief also that *Yellamma* represents mother earth. There is a peculiar rite performed on the full moon days coming between December and January and between March and April. The former full moon day is considered to the day of widow hood and the later

to be of begetting marital status. This ritual represents that by harvesting of corps, the earth loses its prosperity and this condition is thought to be similar to widowhood. Again in the *Caitra* month the spring starts with fresh blossom everywhere and this is supposed to be regaining the marital status. This rite is performed by the *Devadasi*s in the areas of Andhra Pradesh and Maharastra. Thus *Yellamma* cult in these areas and *Devadasi* system in these areas and in many areas of Karnataka are interrelated. *Devadasi* system, prostitution or sex work thus is connected with the *Yellamma* cult in a peculiar way.

VI. Historical and Cultural Background of Yellamma Cult and Prostitution

The famous temple of *Yellamma* in Soundatti at Belgaum district in Karnataka deals with this cult and *Devadasi* system. According to history, Soundatti is a place ruled by a king belonged to Jaina sect. Before their rule this was under the local chiefs where worshipping *Yellamma* as their mother goddess and there Virgin women were appointed as the priests. Later by 8th century, when Jaina kingdom took over the rule, they started enrolling nuns to do rituals in this temple. But in 9th century, when this kingdom lost its power and power in the religion too, the *Saktas* took over the shrine of *Yellamma* once again. In 10th and 11th centuries, *Saiva* kings belonged to *Kapalika* sect got control over this area. During this time, female priests were replaced by male priests who also belonged to *Kapalika* sect. They started indulging and using earlier women priests of the shrine for sex.[vii] In 12th century when *Virasaivism* became dominant and these priests were replaced by *Jangama* priests.[viii] When *Virasaivas* got the power over here, they banned all the ill practices and introduced more refined rituals and belief systems.

During the late 12th century, *Virasaivism* started declining owing to religious revolt in it. Politically also Soundatti area came under the Vijayanagara rulers who were basically *Vaisnavites*. They again converted this shrine in a *Vaisnava* one by appointing Brahmana priests there. These priests installed the idols of *Vaisnava* gods such as *Jamadagni*, *Dattatreya* and *Parasurama*. Various myths were developed both to support their faith and to weaken the earlier deity and tradition also. However till the end of 15th century they continued the administration of the temple and responsibilities towards the devotees. At the beginning of 16th century, once again this shrine came under the local rulers and they slowly started appointing the persons of lower castes as the priests of the temple. They were also known as Joiggayya and *Joggamma*, who were the worshippers of goddess *Bhavani*. They made donations and offered protection to the shrine. It was during that period, attempts were made by Brahmanical priests to take over the temple from non Brahmanical priests. But as already noted, they just succeeded in installing and doing offerings to the idols of *Parasurama*, *Dattatreya*. They even installed *Laksmi* too. During this time *Yellamma*'s names changed as *Renuka*. It is also notable that many local myths are similar to *Renuka*'s myth as narratives in the Puranas. The myths were incorporated with the minor or local myths of *Yellamma* to make the *Yellamma* as same as *Renuka* itself.[ix]

There are some hints about the connection of prostitution with *Yellamma* cult. But a careful analysis is needful to fine out the fact. *Yellamma* became free from skin disease because of serving *Ekayya* and *Jogayya*. She also became a person having the same value as those holy persons. She also took a few women into her service, who had vowed to serve her by spreading her glory and collecting people to serve them. After Yellama's death may be she too became holy person and all believed that any vows taken in her name can solve the problems including diseases on humans. There are several types of mendicants attached to the shrine of *Yellamma*. They are male, female and both young and old ones. They are known to be *Yallappa* or *Yallavva*. Their main job is to spread the glory of *Yellamma*. They carry with them a few objects such as *Chowri* (bunch of hair), metal pot, basket, image of the deity etc. Also they took many vows which can be called as custom and not as tradition. Vows mainly contain three elements. They are praying to the deity to avoid or overcome difficulties or to grant a boon, promising the deity to offer something in return and fulfillment of vows. These vows can be again divided into two types, those involving offerings or gifts and those involving the punishment of the deity. The former is again divided into two those having permanent nature and having temporary nature. In this type of offering any gifts, offering girl to the deity also is included.

VII. Dedication to Divine Prostitution

By doing a vow to *Yellamma* the initiation of dedication starts. Sometime parents of an unborn child also decide to offer the baby to *Yellamma* after birth if that baby is female child as a divine prostitute. When one girl is dedicated to the shrine of *Yellamma* after many religious rituals, she is sent to a man who is waiting for her in the shrine of the temple for union. It is believed that from that day the family which has gifted the girl to the deity starts getting the grace of the deity and prosperity. The duration of this so called relationship depends on the person and the family. These are agreed by both the parties and the payments are also made. Few girls live with that man as 'wife' for life and others live together for a short duration, ranging from one night to a week, and at the most one month. Like this the life of the dedicated woman called as the divine prostitute continues. But the condition put forward is that woman has to observe mourning while she drops a man or is dropped by a man every time. According to the tradition, widowhood is being followed by her till she gets another man has her 'husband'. The thing is that for every rite, right from the initiation after dedication for every religious action or things done by this woman, the place of Sthanika[x] is must. Since their faith is unwavering, they observe these rites strictly and in the meantime, the priests and the Sthanika at the temple get their share of income also.

VIII. Different Types of Divine Prostitution

The main identity of the divine prostitutes is the necklace of beads they wear. *Gati Muttu*, *Sule Muttu* and *Jogati Muttu* are the three types. These represents three sections of prostitutes dedicated virgin, active living with a sigh and enjoying sexually pleasures and retired inactive sexually and active ritually. There

is yet another type of classification is of this type. Those wearing *Gati Muttu* type, by rule remain virgin offering their virginity to the deity, these woman are both sexually functional sometime. These groups sometimes after the vowed period undergo another ritual and go as sex partners of men who help them perform to the shrine and consummation rites and make payments to the shrine and the priests. Those who go as commercial prostitutes, undergo the initiation rite of the order of wearing *Sule Muttu* type. The dedication rites and offerings to the deity are the same in both cases. In the case those wearing the *Jogathi Muttu*, are religious mendicants. They constitute several types of men and women who are sexually a functional are also directly initiated into the service of the deity as religious mendicant functionaries. In addition to these categories there are also few men and women who belong to this tradition transact, wear the dress of opposite sex, i.e. men wearing the women's dress. They are also called as *Yellavva* and *Yellappa*. Perhaps this may be started when the female priests were replaced by the male priests, and due to pressure of the locals; this practice was followed by the male priests to influence the local devotees. As every Dasas these people also spread the glory of *Yellamma* by dancing and singing various traditional and folk songs related to *Yellamma* or *Renuka* and thus collecting money for running their life.

The other thing to be noted here is the prostitution came to be called as a divine one and it even becomes a religious practice. Though this was practiced a religious one, later this became a custom among the *Devadasi* group who were mostly from Mahar, Mang Dowani and Chambhar caste groups.[xi] This thus became a way of income for many families. The families trapped by poverty and often depended on this income supplied by their daughters. Girls dedicated to the goddess not only due to the request or command of rich people but due to number of reasons such as lack of male children in a family, increase number of daughters, mother being in the same profession, elder sister or any other member of the family often being a *Devadasi*, manifestation of 'signs' such as *Jata*, dry hair, white patch, leprosy and even mental problems have been traditionally interpreted as signs of the call of the goddess *Yellamma* to join her. It is also told that some families for gaining religious sanction for their need of desire of profit from their daughters earning as a prostitute make these signs artificially.[xii] The above mentioned dedication ceremony is financed by a wealthy man and after attainment of puberty this same man has the right to take that dedicated girl. This man pays a lump of money to the girl's family as long as she continues her work as *Devadasi*. A family of one *Devadasi* also receives gifts presents and cash on those days, which are especially dedicated to *Yellamma* from those who worship this goddess. There is another practice known as *Jogin* system in which a girl before puberty are married to the goddess and enter prostitution when they reach puberty. In the *Basavi* system, the girls are dedicated at pre puberty age to a number of deities. But these girls after attaining puberty, does not immediately enter into life careers as prostitutes and according to tradition by using begging baskets should beg for money and other things to serve *Yellamma* at least twice a week. Basavis usually turn to prostitution as begging does not usually generate sufficient income to them.[xiii]

The social need was of the emergence of the decadent *Devadasi* cult, a week remains of the young girls termed as Matangi, those were dedicated to the king, as the human god, and were believed to be endowed with miraculous power. These girls now dedicated to the goddesses got *Yellamma* as their goal and prototype. The goddess having head only and the goddess having the body is identified with *Renuka*. On the baisis of the popular myth of the latter's severed head, the *Devadasi* softly came to be associated with *Renuka*. The girls dedicated to *Yellamma* carried in their baskets the images of *Parasurama* and *Renuka* too. This shows how *Renuka* got to be associated with the customs which were, originally observed for the southern goddess *Yellamma*. A social custom was thus transformed into divine in nature and even as a religious practice of several communities which has to be done strictly. The prostitution here is changed into 'divine' and thus becomes a practice having divinity in it. This system that has existed throughout several places is recorded in the history of south India and entire culture. This system of divine prostitution can be said larger than the *Devadasi* system, because this kind of custom was functioning within the structure of a region and as a temple ritual. Even though this tradition has lost its socioeconomic base, the concept and thus the need of a divine prostitute is still very much alive.

The identification of *Yellamma* with *Renuka* can be formed in all these instances. The association of practices in connection with *Yellamma* to *Renuka* was naturally the next step. But we have also myths, originally in respect of *Yellamma* inserted onto *Renuka*. Thus it can be concluded that *Renuka* whose glories are mentioned in Mahabharata and many Puranas changes her identity into many goddesses and her identity is shared by these local goddesses for their influence among the locals. It is also possible that by incorporating *Renuka*'s myth to various minor myths of local deities, the myth makers made use of them to establish *Renuka*'s myth through the local divinities of a particular region.

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Conceptual Interrelatedness in Ksanikavada and Svabhavoparamavada

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Abstract- Ayurveda, the science of life identifies man as an embodiment of nature. It encloses the essential concepts in various streams of Indian philosophy. Theory of instantaneous or ksanikavada is a significant concept in Buddhism, according to which all the universal things are momentary. A similar thought can be seen in Ayurveda known as Svabhavoparama or natural destruction, this expounds that all the things are in a succession of natural destruction. According to this theory, the process of ending occurs without a specific reason. Even though there are numerous conceptual similarities in the Buddhist and Ayurvedic doctrines, the present study mainly concentrates on the analysis of Ksanikavada and Svabhavoparamavada.

Index Terms- pratityasamutpada, ksanikavada, svabhavoparamavada, similarity, Buddhism, Ayurveda

I. INTRODUCTION

Ayurveda and Buddhism are two important knowledge systems in Indian history. These two systems are developed side by side, and they shared many common concepts. Bhuddhist philosophy emerged as a way to find out a solution for sufferings in the world. In Indian context it had a deep influence unlike the orthodox systems as it open a new path to enlightenment for everyone, irrespective of their caste. Buddhism established a distinct perspective about the life unfamiliar to the then society, and recommended the eight fold path for a peaceful life. The teachings of Buddha have been a topic of discussion among the scholars for last two and a half millennia. Buddha's theory of *Pratityasamutpada*, developed as *Ksanikavada*-a well discussed doctrine in the later Buddhism. This theory implies that there is nothing in the universe that is stable. All the things are in a continuity of origin and decay. The universe is always moving and the momentary entities themselves act as their causes. This implies that the existence of the things means its efficiency to produce something. It means every object exists as itself as well as the cause of another one. Such an existence, which acts as cause and effect at the same time, is momentary. Causality, reality, efficiency and instantaneous are interchangeable terms in this system[1]. Causation or *Pratityasamutpada* is kinetic. What exists is always moving, it is an illusion that a thing exists without acting. This motion itself is causation[2]. Buddha explained *Pratityasamutpada* by using the example of river. The waves in a river are not same which are changing in every second and different from one another. This is the same in the case of all universal things, which are momentary without exception.[3] In a nutshell the idea of this doctrine is that everything in the universe is transient, ie., originating and ceasing constantly, and nothing exists in eternity or annihilation. This principle was later adapted as the doctrine of *Ksanikavada* by the *Sautrantrika* sect of Buddhism.

II. KSANIKAVADA

Ksanikavada or the theory of momentariness, the modified form of *Pratityasamutpada*, explains that the cause and effect are momentary. The cause does not exist when the effect has appeared, i.e. the cause and effect cannot be exists simultaneously. The concept that a thing ceases to be the same as soon as it originates is called momentary or *Ksanika*. This term refer to the thing that does not continue to exist after its origin. In other words the existence is identical with the moment i.e. existence is a point of instant[4]**.** The later philosophers like Dharmakirti, Kamalasila etc. hold the view that the real things are efficient to produce a purposive action i.e. arthakriyakari[5].This real is momentary, the succession of moments produce the illusion that the real or existence is constant and uninterrupted.

Nirnibhanda hi samagri svakaryotpadane yatha|
Vinasam prati sarvepi nirapeksasca janminah|| [6]

Thus Ksanikavada explained the same theory of Pratityasamutpada of Buddha. The major criticism against this principle is that if everything is in a succession and continually renewed, then how the recognition of objects is explained. The Buddhists answer that the things in two moments of cognition are only similar and we mistake them to be the same. All the recognition is erroneous since similarity is mistaken for its identity.[7] As most of the Sautrantika works are not available or lost, the details of Ksanikavada was accessible only through the purvapaksas of the scholars of Nyaya- Vaisesika, who were always dead against the Ksanikavada. In Ayurveda this theory is known as *Svabhavoparamavada.* According to *Carakasamhita* knowledge of this basic theory is essential for a doctor while treatment.

III. SVABHAVOPARAMAVADA

Carakasamhita explains Svabhavoparamavada in the context of defining the importance of treatment. The word Svabhavoparama means 'svabhavat vinasah', the natural end. In *Carakasamhita,* the sage Atreya elucidates Svabhavoparamavada and the purpose of treatment in Ayurveda as a response to purvapaksa[8]. He says that

Jayante hetuvaisamyatvisama dehadhatavah|
Hetusamyat samastesam svabhavoparamah sada||[9]

This implies that it is always the nature of the constitution of body to suffer decay or destruction and it never need a cause. Cakrapanidatta the commentator of *Carakasamhita*, says that 'svabhavat vinasakarananirapeksat uparamo vinasah svabhavoparamah'. This principle suggests that for the generation of all existent objects, there is an adequate cause, but for their ceasing there is no cause; it is natural.

Pravrttiheturbhavanam na nirodhesthi karanam|[10]

Here Cakrapanidatta explains that the harmonious condition and its imbalance of the body are of some reasons but it will cease in the next moment without any cause. He uses an example of candle to explain this. Oil, thread etc. needed for the enlightenment of a candle, but it turns off without any cause. All the constituent elements of the body have this feature of svabhavanasa. The imbalanced elements also have this nature. The whole body is always in a process of origin and ceasing. So it can be observed that all the things are momentary or Ksanika.
In *Carakasamhita* Agnivesa raises a question about the relevance of treatment thus: 'if all the things including the imbalanced elements are naturally ceases without any reason, then what is the purpose of treatment?' His teacher Punarvasu said that, for the destruction of existing objects there is no cause. The same happens with that of the time, which goes on ceaselessly the cause which brings about destruction cannot be ascertained. As the time passes out quickly, the existent objects also are destructed, even in the lack of causes.[11] Still the medical treatment never becomes inoperative.

Yabhihi kriyabhirjayante sarire dhatavah samah |
Sa cikitsa vikaranam karma tadbhisajam smrtam ||[12]

The treatment in Ayurveda aims at the 'dhatusamya' or harmony of elements in the body. When the elements become imbalanced, the Ayurvedic treatment tries to remove the cause of it. According to Svabhavoparama theory the imbalanced elements of the body ceasing in every moment and cause the generation of another imbalanced element. Since this is a continuous process, the physician treats for the harmonizing of imbalanced elements. It helps to destroy the cause of disease and make the successive elements in harmony. This implies that the medicinal treatment provides balance to the successive dhatus, by destructing the cause of visama or imbalance and serving the cause of samya.[13] This is the nature of the principle of Svabhavoparama is discussed in *Carakasamhita*. The other treatises on Ayurveda also speak of this principle indirectly.

IV. Conclusion

It can be assumed that at the time of Agnivesa itself the idea of Ksanikavada was established in the society. Ayurveda accepts this doctrine at the same time it declares the relevance and effectiveness of treatment. It opines that all the existing objects which are made up of pancabhuta suffer a series of natural decay. In *Nyayabhasya*, Vatsyayana presents an example of Ksanikavada doctrine as a purvapaksa, he says that the body is always in a process of growth and decay. The continuous process of growth generates new cells in the body and the natural decaying process causes the destruction of matured cells. It should be understand that as in the case of body all the things have the quality of Ksanika.[14]

All the above points lead to the conclusion that Svabhavoparamavada and the Ksanikavada hold the same principle. Ayurveda and Buddhism are closely related knowledge systems and are greatly influenced each other in their development. It may also be observed that the base of Svabhavoparamavada in Ayurveda is Ksanikavada, in other words Svabhavoparamavada is an extension or adapted version of Ksanikavada.

Notes

[1] Joshi,L.M, 1967, Studies in the Buddhist Culture of India, Motilal Banarsidass, Delhi, p.198.

[2] Ibid; p.197.

[3] Because of the peculiarity of Pratityasamutpada theory, Buddhism came to be designated as Madhyamamarga, for it never denies the existence of cause and the reality of effect. Unithiri,N.V.P, 2006, Indian Bhauthikavada Paithrikam, Progress publications, Calicut, p.148.

[4] Joshi,L.M, op.cit; p.197.

[5] Ibid;

[6] Quoted in Unithiri,N.V.P, op.cit; p.165.

[7] Hiriyanna,M, 1994, Outlines of Indian Philosophy, Motilal Banarsidas, Delhi, p.141.

[8] According to the followers of Ksanikavada the diseases are also Ksanikas the treatment is ineffectual. Sharma,T.Yadhav(Ed.), 1994, Carakasamhita, Chowkhambha Sanskrit series, Varanasi, p.97.

[9] Ibid;

[10] Ibid;

[11] Ibid; p.98.
Na nasakaranabhatbhavanÆ nasakaranam |
jnayate nityagasyeva kalasyatyayakaranam ||
sighragatvadyatha bhutastatha bhavo vipadyate |
nirodhe karanaÆm tasya nasti naivanyathakriya || (su.s 17, 32-33)

[12] Ibid; p.98.

[13] Ibid;

[14] Tyagatvisamahetunam samanam copasevanat | Visama nanubadhnanti jayante da¡tavah samah || (su.s.17, 36) Cakrapanidatta says that evam manyate –yadhyapi dhatuvaisamyam ksanikatvena vinasvaram, thatapi vinasyadapi tad dhatu vaisamyam svakaryam visamameva dhatumarabhate, evam sopyaparam visamamiti na dhatuvaisamyasantananivrtihi dhatusamya janakahetum vina; yada tu dhatusamyaheturpayukto bhavati, tada tena sahitam vaisamyasantati patitamapi karanm samamevadhatusantanamarabhate//

[15] Sastri, D.D(Ed.), 1986, Nyayabhasya, Sudhi Prakashana,

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[5] Sharma,T.Yadhav(Ed.), 1994, *Carakasamhita*, Chowkhambha Sanskrit Series, Varanasi.

[6] Sastri,Chandrasekara(Ed.), *Nyayabindu*, 1954, Kasi Sanskrit Series, Varnasi.

[7] Shastri,D.D(Ed.), 1986, *Nyayabhasya*, Sudhi Prakashana, Varanasi.

[8] Stcherbatsky,T, 1923, *Central Conception of Buddhism and the Word Dharma*, Asian Educational Services, Delhi.

[9] ~~~~~, *Buddhist Logic*, Part One, Reprint 2004, Kessinger Publishing, USA.

[10] Unithiri,N.V.P, 2006, *Indian Bhauthikavada Paithrikam*, Progress publications, Calicut.

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Survey of Adaptive On Demand Distance Vector Learning Protocol (AODV)

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Abstract: Mobile Ad-hoc Network (MANET) is an infrastructure less and decentralized network which need a robust dynamic routing protocol. In such an environment, it may be necessary for one mobile host to enlist the aid of others in forwarding a packet to its destination, due to the limited propagation range of each mobile host's wireless transmissions. Some previous attempts have been made to use conventional routing protocols for routing in ad hoc networks, treating each mobile host as a router. There have been several routing protocol proposed. In Manet protocols are classified in two categories Reactive and proactive protocol.The most efficient reactive protocol is Ad-hoc on demand distance vector (AODV) routing protocol. We examine AODV in term of advantage and limitation.

***Index Terms*-:** MANET**,** Proactive and Reactive routing protocols, AODV, DSDV

I. INTRODUCTION

A MOBILE ad hoc network (*MANET*) [3] consists of a many of wireless mobile nodes communicating with each other without any centralized control or fixed network infrastructure.MANET have been evolving to serve a increasing number of applications that rely on multihop wireless infrastructures that can be easily installed. The potential applications include emergency disaster relief, battlefield command and control, mine site operations, and wireless classrooms in which participants wish to share information.
Today, wireless technologies such as IEEE 802.11 [4], Bluetooth [1], and third-generation cellular have led to a proliferation of mobile devices. The number of mobile Internet devices is expected to reach a billion in the near future [2] and exceed the number of stationary nodes. Therefore, we need MANETs to be interconnected to the Internet in many applications.

II. ROUTING IN MANET

MANET routing protocols may be broadly classified into two major categories: Proactive and Reactive.

Proactive Routing Protocols: Proactive protocols continuously learn the topology of the network by sharing topological information among the network nodes. Thus, when route is requiredfor a destination, such route information is given immediately. If the network topology changes too frequently, the cost of maintaining the network could be very high. If the network activity is low, the information about actual topology could not be used.

Reactive Routing Protocols: The reactive routing protocols are based on query-reply dialog. Reactive protocols establish route(s) to the destination only when the need arises. They need not periodic transmission of topological information of the network.
Based on the method of delivery of data packets from the source to destination, classification of MANET routing protocols may be done as follows:

- Unicast Routing Protocols: This protocols consider sending information packets to a single destination from a single source.
- Multicast Routing Protocols: In this protocols, the delivery of information to a group of destinations simultaneously, using the most efficient strategy to deliver the messages over each link of the network only once, creating copies only when the links to the destinations split. Multicast routing use both multicast and unicast for data transmission.

PROACTIVE ROUTING PROTOCOLS:

- Dynamic Destination-Sequenced Distance-Vector Routing Protocol (DSDV)
- Wireless Routing Protocol (WRP)
- Cluster Gateway Switch Routing Protocol (CGSR)
- Global State Routing (GSR)
- Fisheye State Routing (FSR)
- Hierarchical State Routing (HSR)
- Zone-Based Hierarchical Link State Routing Protocol(ZHLS)
- Landmark Ad Hoc Routing (LANMAR)
- Optimized Link State Routing (OLSR)

REACTIVE ROUTING PROTOCOLS:

- Associativity-Based Routing (ABR)
- Signal Stability–Based Adaptive Routing Protocol (SSA)
- Temporarily Ordered Routing Algorithm (TORA)
- Cluster-Based Routing Protocol (CBRP)
- Dynamic Source Routing (DSR)
- Ad Hoc On-Demand Distance Vector Routing (AODV)

III. AODV

AODV Feature

- AODV does not need any central administrative system to handle the routing process. AODV reduce the control traffic messages overhead at the cost of increased latency in finding new routes.
- Tries to keep the overhead of the messages small. If host has the route information in the Routing Table about active routes in the network, then the overhead of the routing process would be minimal. The AODV has advantage in overhead over simple protocols which require to keep all the route from the source host to the destination host in their messages. The RREQ and RREP messages, whose responsibility for the route discovery, do not increase significantly the overhead from these control messages. AODV reacts relatively quickly to the topological changes in the network and keep update only the hosts that may be affected by the change, using the RRER message. The Hello messages, whose responsibility for the route maintenance, are also limited so that they do not create unnecessary overhead in the network.
- Loop free and avoids the counting to infinity problem, which were not easy to the classical distance vector routing protocols, by the usage of the sequence numbers.

IV. INTRODUCTION

A devices like laptop, mobile phone, Personal Digital Assistant (PDA), or similar devices, which can communicate directly with one another without a central administrator through MANET.A MANET, is an autonomous system of mobile routers and associated hosts connected by wireless links. MANET does not require a central administration infrastructure due to its wireless nature and can be deployed as a multi-hop packet network with low expense and rapidely[5].MANET has its own routing protocols which can be compromised with dynamic topology ,frequent route exchange, multi-hop routing, bandwidth constraint and A mobile ad hoc network routing protocol controls how nodes decide which way to route packets between computing devices [6]. The routing protocols in MANET are proactive (table driven), reactive (on demand) and hybrid routing protocols. Popular proactive routing protocols are highly dynamic DSDV (Destination-Sequenced Distance Vector) and Optimized Link State Routing protocol while reactive routing protocols include AODV(Ad hoc On demand Distance Vector) and Dynamic Source Routing (DSR).hybrid routing protocol is Zone Routing Protocol (ZRP).

The MANET requirements for multi-hop routing dynamic,self-starting between mobile nodes [7]. AODV protocol makes routes among nodes when required by source nodes thats why called on demand routing protocol. Maintenance of routes depend on the sources need [8].The mobile Nodes maintain a route cache and use a destination sequence number for each route entry. The node in AODV looking for information about the network only when needed reduces overhead since nodes do not have to maintain unnecessary route information while the use of a sequence number ensures loop freedom.

V. The Ad-hoc On-Demand Distance Vector Algorithm

The AODV algorithm is on-demand route request system; nodes that do not depend on active paths neither store any routing information nor take part in any periodic routing table exchanges. The AODV routing algorithm designed for mobile networks which is ad hoc[9] [10]. Unicast and multicast routing is supported by AODV [11]. AODV builds routes among nodes when only source nodes requires. Sequence numbers in AODV is to ensure the freshness of routes. AODV apply in large numbers of mobile nodes and based on loop free. [8].

AODV builds routes using a route request route reply query cycle. Source node broadcasts a route request (RREQ) packet across the network when it want a route to destination. The route tables has backward pointers for nodes which receive packet and source nodes and update their information for the source node. In addition to the broadcast ID ,source IP address, current sequence number the RREQ has updated sequence number for the destination.A route reply (RREP) is sent by node which receiving the RREQ ,node which is the destination or has a route to the destination with greater sequence number. the RREQ's source IP address and broadcast ID [11] is maintained by nodes. Already processed RREQ,is not forwarded.The RREP sends back to the source by Nodes.it set forward pointers. The source node forward data packets to the destination.if it receive RREP

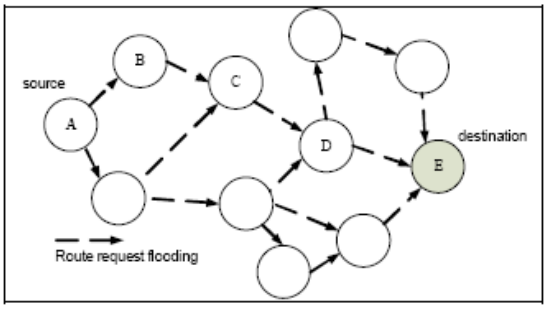

Figure 1: Route Request (RREQ) flooding

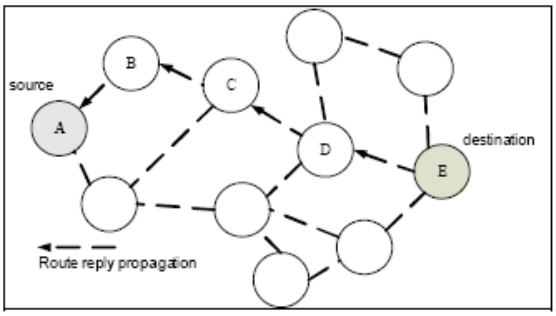

Figure 2: Route Reply (RREP) propagation

Route continue to be maintained if it is active. In case of link break when the route is active, the node along the route send a route error (RERR) message to the source node.If the source node apply route discovery procedure if it require after receiving the RERR.

Path Discovery

The Path Discovery process is generated whenever a source node has no routing information in its routing table.if it communicate with nodes.

Every node maintains two separate counters: a node sequence number and a broadcast id. The source node looks for route by broadcasting a route request (RREQ) packet to its neighbors. The RREQ contains the following fields:

< source_addr; source_ sequence #; broadcast_ id; dest_ addr; dest _sequence #; hop cnt >

The pair < source addr; broadcast id > uniquely identifies a RREQ. broadcast id is incremented whenever the source send a new RREQ. Each neighbor either accept the RREQ by sending a route reply (RREP) back to the source or rebroadcasts the RREQ to its own neighbors after increasing the hop count. Notice that a node may receive many copies of the same route broadcast packet from various neighbors. When RREQ is received by intermediate node, which has same source address and broadcast id, RREQ is deleted. If a node cannot accept the RREQ, it keeps track of the following information in order to implement the reverse path setup, as well as the forward path setup that will accompany the transmission of the eventual RREP:

- Destination IP address
- Source IP address
- Broadcast id
- Expiration time for reverse path route entry
- Source node's sequence number

The optimization of AODV is based on the recent draft of the AODV specification [4]. The essential functionality of AODV includes:

• *RREQ* and *RREP* messages (for route discovery)
• *RERR* messages, HELLO messages, & precursor lists (for route maintenance)
• Sequence numbers
• Hop counts
• Expanding ring search

The following fields exist in each route table entry of AODV:

• *Destination IP Address*:
• *Destination Sequence Number*: It is associated to the route.
• *Next Hop*: Either the destination itself or an intermediate node designated to
forward packets to the destination
• *Hop Count*: Between the source IP Address to the Destination IP Address the number of hops
• *Lifetime*: Consider the route is valid in Time for node that have RREQ
• *Routing Flags*: The state of the route; up (valid), down (not valid) or in repair.

Suppose S would like to communicate with D Figure 3, the node send out a RREQ to explore a route to the destination. S generates a Route Request with destination address, Sequence number and Broadcast ID and sent it to his neighbour nodes. Each node receiving the route request sends a route back (Forward Path) to the node.

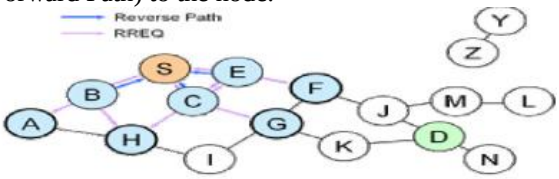

FIGURE 3: Path finding in AODV

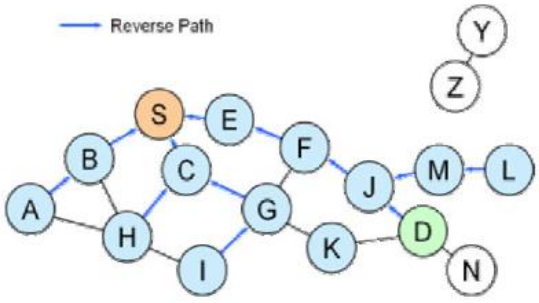

FIGURE 4: Path finding in AODV

The route is made available by unicasting a RREP back to D and is written in the routing table from S Figure 4. After receiving the route reply every node has to update its routing table if the sequence number is more recent.

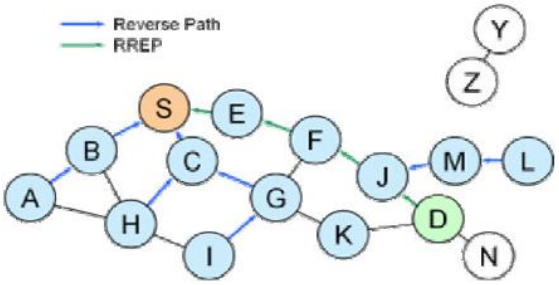

FIGURE 5: Path finding in AODV

Now node S can communicate with node D, Figure 5, 6.

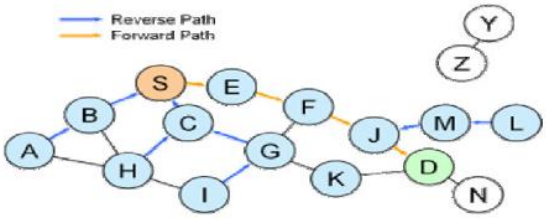

FIGURE 6: Path finding in AODV

A RERR message is sent to other nodes when active route has broken link Figure 7. If the nodes have a route in their routing table with this link, the route will be erased. Node S sends once again a route request to his neighbour nodes. Or a node on the way to the destination can try to find a route to D. That mechanism is called: *Local Route Repair.*

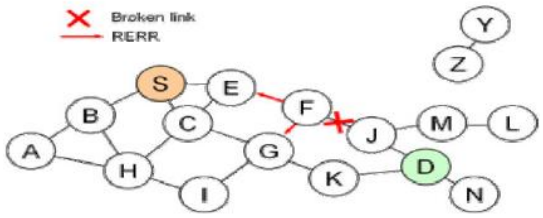

FIGURE 7: Path finding in AODV

Route Table Management

The source and destination sequence numbers, other useful information is also recorded in the route table entries. Reverse path routing entries is a timer, called the route request expiration timer. The work of this timer is to maitain reverse path routing entries from those nodes that do not have the path from the source to the destination.The expiration time depends upon the size of the adhoc network. Another important parameter with routing entries is the route caching timeout.A neighbor which sends minimum one packet to destination in active timeout period is considered active.. This information is maintained so that all active source nodes can be informed when a link along a path to the destination breaks. Active neighbors use route entry is considered as active.Active route entries has a path from a source to a destination, which is followed by packets is called an active path.In DSDV, all routes with destination sequence numbers in the route table are marked, which guarantee that no routing loops can form, even if of out-of-order packet delivery and high node mobility conditions. A mobile node maintains a route table entry for each destination of interest. Each route table entry contains the following information:

- Destination
- Next Hop
- Number of hops (metric)
- Sequence number for the destination
- Active neighbors for this route
- Expiration time for the route table entry

When source to destination data transmission, the route entry timeout is reset to current time. The mobile node compares the destination sequence number of the new route to the destination sequence number for the current route if a new route is given to a mobile node. The route is selected that have the greater sequence number. If the node have same sequence numbers, then the new route is selected only if it has a smaller metric (fewer number of hops) to the destination.

Path Maintenance

Movement of nodes not lying along an active path does not deflect the routing to that path's destination.If the source node moves during an active session, it may reinitiate the route discovery procedure to form a new route to the destination. A RREP is sent to the affected source nodes when either intermediate node or the destination moves. Periodic hello messages is used to ensure symmetric links,as well as to detect link failures, Alternatively, and with far less latency, such failures would be detected by using link layer acknowledgments. A link failure is also showed if attempts to forward a packet to the next hop fail.Once the next hop becomes unreachable, the RREP with a fresh sequence number is forwarded by node and hop count of 1 to all active upstream neighbors. Those nodes subsequently propogate that message to their active neighbors and so on. When all active source nodes are informed then process stops.it ends because AODV maintains only loop-free routes and there are only a finite number of nodes in the ad-hoc network.Upon receiving notification of a broken link, source nodes can restart the searching if nodes require route to the destination.A node determine if route is still needed it inspect upper level protocol control blocks to see whether connections remain open using the indicated destination as well as the route has been used recently.

Local Connectivity Management

Nodes know their neighbors in one of two ways.Whenever a node receives a request from a neighbor, it updates its local connectivity information to ensure that it add this neighbor. In the event that a node has not sent any packets to all of its active neighbors within hello interval, it send to its neighbors a hello message (a special RREP), containing its identity and sequence number. The node's sequence number is not changed for hello message transmissions. The hello messages that have less TTL value of 1 is not rebroadcasted outside the neighborhood of the node .Neighbors update their local connectivity information to the node if receive packet.Failing to receive hello messages from not active neighbors does not activate any protocol action. If hello messages are not received from the next hop along an active path, the active neighbors using that next hop are sent notification of link failure We have determined the optimal value for allowed hello loss is two.

The hello messages can also be used to maintain that neighbors of nodes that have bidirectional connectivity are considered. For this purpose, a node lists the nodes from which it has heard hello messages. Each node uses only routes to neighbors that have heard the node's hello message

Limitations of AODV

- Need on broadcast medium: The algorithm requires that the nodes in the broadcast medium can detect each others broadcasts.
- Overhead on the bandwidth : Overhead on bandwidth will be occurred compared to DSR. When an RREQ travels from node to node in the process of discovering the route info on demand,it setsup the reverse path in itself with the addresses of all the nodes through which it is passing & it carries all this info all its way.
- No reuse of routing info: AODV lacks on efficient route maintenance technique. The routing info is always obtained on demand including for common case traffic.
- It is vulnerable to misuse: The messages can be misused for insider attacks including route disruption, route invasion, node isolation, and resource consumption.AODV lacks support for high throughput routing metrices.AODV is designed to support the

shortest hop count metric.This metric favors long,low bandwidth links over short, high bandwidth links.

- High route discovery latency:AODV is reactive routing protocol.This means that AODV does not discover a route until a flow is initiated.This route discovery latency result can be high inlarge-scale mesh networks.

VI. CONCLUSION

We have studied in AODV that the route maintenance technique is not efficient in AODV. The routing info is always obtained on demand including for common case traffic.The nodes can detect each others broadcasts in the broadcast medium This is the requirement of broadcast medium.Overhead on bandwidth will be occurred compared to DSR. When node to node travels of an RREQ for the route information on demand, the reverse path is setup in itself with the addresses of all the nodes through which it is passing & it carries all this info all its way. The frequent route repair causes higher route cost and delay this is due to breaking of link.The simple algorithm neighbor change ratio resolve the faults. But the problem of the method is that the nodes are not forwarding control packets or data packets which have the neighbor messages of their neighbors but still need to send Hello packets periodically. We have concluded that the neighbor stability algorithm can incorporate piggyback mechanism to resolve the issue of longer delay and periodically Hello packet sending.

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Production and Partial Purification of Protease by Actinomyces Species

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Abstract- Enzymes are biocatalysts obtained from plants, animals and microorganism. Microbial enzymes are becoming important for its technical and economical advantages .Various kinds of microorganisms in nature degrade different type of proteins indicating the Proteolytic capabilities of microorganism.

About 80% of enzymes produced annually are simple hydrolytic enzymes, of which 60% are proteases. Different bacteria, fungi and actinomycetes are the major sources of microbial proteases. Extracellular proteases have high commercial value and multiple application in detergent, food, diary, pharmaceutical, leather, diagnostic, waste management and silver recovery industries.

Here we consider the extra cellular protease which is secreted by Actinomycetes Sp. This enzyme is free from most of the toxic agents. And it is considered as a safe/suitable protease for the oral medication purposes. Here we find the best suit Actinomycetes strain which yield high protease, and its suitable Ph and the medium (Plate assay method). Then we find the high Protease production in different SSF and LSF. Thereafter we discuss the protease isolation methods, conformation and their application in the medical and industrial fields.

Index Terms- Actinomycetes, Extra Cellular enzyme, Hydrolytic Enzyme, Protease production

I. INTRODUCTION

Protease constitutes one of the major important groups of industrial enzymes; about 80% of enzymes produced annually are simple hydrolytic enzymes, in which 60% are proteases. Different bacteria, fungi and actinomycetes are the major source of microbial proteases. Extracellular proteases have high commercial value and multiple application in detergent, food, diary, pharmaceutical, leather, diagnostic, waste management and silver recovery. The microbial protease represents 60% of the worldwide market of industrial enzyme. They find commercial application for toothpastes as antiplaque and antitartar, cosmetics and for the recovery of silver from used X-ray films (Ishikawa *et al*.,1993).

The wide diversity and specificity of protease is used to develop effective therapeutic reagents. Proteases are used to treat burns and wounds. Oral administration of proteases produces an anti-inflammatory response in burned patients and speeds up the healing process. Trypsin and chymotrypsin acts as anti-inflammatory and antioxidant agents on burn wounds. (Bitange et al., 2008) Protease deficiency leads to arthritis, osteoporosis and other calcium deficient diseases. Because protein digestion occur and digested protein converted into glucose. This leads to hypoglycemia; resulting in moodiness, mood swings and irritability.

The microbial enzyme (Protease) secretes in the phase of their growth. For the commercial requirement they may obtained as a by-product of antibiotic production.

II. METHODOLOGY

The marine filamentous bacterium *Actinomycetes* was originally isolated from marine sources.

The actinomycetes were streaked on 0.3% casein, gelatin and skim milk. The plates were incubated at room temperature (28°C) for four days. Then the plates were flooded with 15% mercuric chloride solution and 20%HCl for 5min. Then the plates were observed for the development of clear zone. Then the Actinomycetes were grown in broth culture media. The culture was inoculated with 3% of seed culture and incubated in a rotary shaker with the speed of 200rpm. The medium was harvested and collected after 7 days and the amount of protease was estimated.
Actinomycetes from culture medium was poured and streaked aseptically in substrate amended agar plates and incubated in room temperature for 10 days. The enzyme activity was visualized as a clear zone of substrate utilization after flooding the plates.

The culture was harvested after 6days. The harvested culture was filtered through whatmann No.1 filter paper. Then the free cells were separated by centrifugation. (10,000rpm for 15 minutes) Supernatant was fractionated by precipitation with ammonium sulfate 80% of saturation. Another harvested culture was suspended by centrifugation, and the supernatant was collected and measured. Fractionated enzyme was precipitated by adding half the amount of ice cold acetone.

The collected protein was dialyzed for two days with 0.1m phosphate buffer. The dialyzed fraction was collected and stored. Protein pellet was dissolved in 0.1M phosphate buffer, (P^H7) and dialyzed. The protein content of the culture was estimated by binding method of Bradford (1976).

The molecular mass of purified protease was determined by SDS-PAGE. Purified protein samples were loaded on SDS-PAGE with the standard protein ladder. (Consists 100 kDa) After the separation, the gels were stained with silver nitrate.

III. RESULTS

A clear zone was observed and the protease Production (by *streptomyces* sp.) was high in casein. The maximum zone was observed on the 6^{th} day.

After 3days a clear zone was measured in the quantitative medium. It shows that the Actinomycetes produce high amount of protease on third day.

The 6th day culture was collected, filtered, and the filtrate was precipitated by 80% ammonium sulfate saturation or acetone. Proteases were harvested by centrifugation at 5000rpm for 25mins.Then the protease was dialyzed for 2days against phosphate buffer and the crude protease was obtained.

Partially purified proteases were separated by using 10%acrylamide in SDS-PAGE. Protease bands were visualized, when the gel was stained with silver nitrate solution. The partially purified protease produces a band on polyacrylamide gel.

Proteases were separated on 10% acrylamide under non-denaturing conditions. Bands were visualized on native gel when stained with Coomosive Brilliant Blue and silver nitrate.

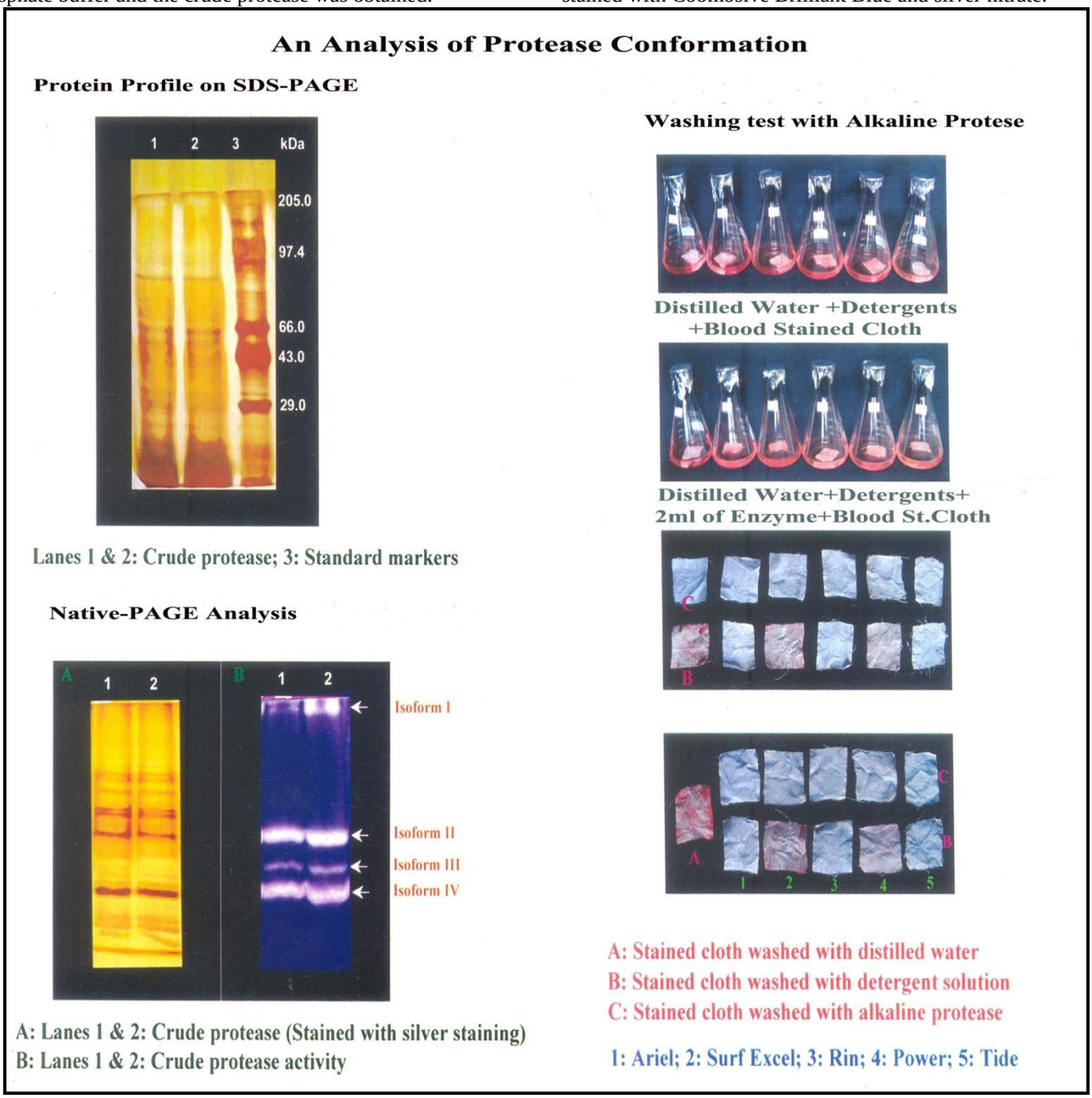

Figure 1: An analysis of Protease conformation

IV. DISCUSSION

Actinomycetes strain isolated from marine sediment were identified as they produce alkaline protease. The proteolytic activity of the marine streptomyces sp.MML1614 was detected, in which the casein and gelatin forms a clear zone around the colonies.Also streptomyces lividans 1326 produces significant clear zone when it was grown on milk agar plates. The formation of clear hydrolytic zone indicates that the marine strain of

streptomyces M11ViLL1614 produce extracellular protease. (Dekleva *et al*. 1985).

Extracellular production of protease is influenced by various physical factors, such as media composition, pH, temperature and incubation time The marine streptomyces sp.NIIvIL 1614 was exhibite good growth and high protease production at pH 7, and 28°C. Maximum growth of streptomyces sp.NIIvIL 1614 was observed in in 40°C. It had favored the maximum rate of enzyme production. (Nehete *et al*., 1985)

Secretion of alkaline protease dependents on the available sources of carbon and nitrogen in the medium.Various nitrogen sources, including amino acids are found to repress the production of enzyme at certain concentrations. The marine streptomyces sp. NUvIL1614 produce high amount of protease when glucose was used as carbon source (Gibb and Strohl, 1988) Nitrogen source plays a key role in protease production.Protease production in streptomyces sp.MML1614 was increased with the addition of peptone (upto 2%) but beyond a certain level the production rate of protease was repressed. (Kanekar *et al*.,2002)
In general,Liquid-state Fermentation is more convenient than submerged cultivation. Hence, it is feasible to apply the solid-state cultivation for producing agricultural by-products. In LSF, microorganisms secrete the necessary enzymes to degrade the available substrates to fulfill their nutritional requirements. (Tunga *et al*., 1998) Protease production medium (PPM) was prepared to inoculate actinomycetes. Then the enzyme was collected on 6^{th} day by filtring the medium

Agro-industrial wastes/substrates (defatted soybean cake, gram bran. wheat bran, rice bran, banana waste, etc) can de used for the production of Protease enzyme. Solid state fermentation is a simple process and requires less pre-processing energy than submerged fermentation. (SMF) Other advantages include superior productivity, low waste water Management, and improved product recovery (Tunga et *al*., 1998).

As our protease produced by marine Actinomycetes, it can able to withstand wide range of pH and temperature. Also it shows the compatibility to various commercial detergents; that it can be used as an additive detergent. These additive proteases improve the cleaning performance of the detergent. Additions of protease enzyme in surf excel results in increased blood stain removal ability of it. Protease plays a vital role in dehairing process of leather industries.

V. CONCLUSION

Protease plays a vital role in various industrial applications, and the demand for protease is increasing day-by-day. Therefore, it is expected that hyperactive strains will emerge and the enzymes produced by these Actinomycetes could be used in different industries. In order to meet the ever-growing demand for this enzyme made us to search for new protease producing microorganism.

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Extraction and Characterization of Water Soluble Chitosan from Parapeneopsis Stylifera Shrimp Shell Waste and Its Antibacterial Activity

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Abstract- Preparation and characterization of water soluble chitosan was examined for their antibacterial activity from P. stylifera. The yield of crude chitosan and water soluble chitosan was 54.3 % and 87.8%. The FT-IR spectrum of chitin, chitosan and water soluble chitosan also determined and characterization was done and compared with standards. Compare to other bacterial strains S.auerus (18.3mm) having more potential antibacterial activity in crude chitosan as well as water soluble chitosan. Both chitosans might have the antibacterial activity which would be used in novel drugs from the shrimp shell waste.

Index Terms- Shrimp shell waste, Water soluble chitosan, FT-IR and Antibacterial

I. INTRODUCTION

Chitin is a natural polysaccharide synthesized by a great number of living organisms and functions as a structural polysaccharide[1].Chitosan is the only pseudonatural cationic polymer which has many potential biomedical and other applications. Chitosan has been proved usefully for would dressing and bone tissue engineering[2-3]. It shows good performance in drug delivery and analgesia[4]. Chitosan has some beneficial properties, such as antimicrobial activity, excellent biocompatibility and low toxicity that promote its applications in many fields including food industry and pharmaceutics[5-7].

Chitosan is natural, non toxic, copolymer of glucosamine and N-acetylglucosamine prepared from chitin by deacetylation, which in turn, is a major component of the shells of crustaceans. It is found commercially in the waste products of the marine food processing industry[8-9]. Various chemical modifications have been investigated to try and improve chitosan's solubility and thus to increase its range of applications[10-11]. Recent studies on chitosan depolymerisation have drawn considerable attention, as the products obtained are more water-soluble. Beneficial properties of chitosan and its oligosaccharides include: antitumour[12]; neuroprotective[13]; antifungal and antibacterial[14-15]; and anti-inflammatory[16].

The antimicrobial activity of chitin, chitosan and their derivatives against different groups of microorganisms, such as bacteria, yeast, and fungi, has received considerable attention in recent years[17-18]. Two main mechanisms have been suggested as the cause of the inhibition of microbial cells by chitosan. One means is that the polycationic nature of chitosan interferes with bacterial metabolism by electrostatic stacking at the cell surface of bacteria[19-20]. The other is blocking of transcription of RNA from DNA by adsorption of penetrated chitosan to DNA molecules. In this mechanism the molecular weight of chitosan must be less than some critical value in order to be able to permeate into cell[21]. The antimicrobial activities of chitosan are greatly dependent on its physical characteristics, most notably molecular weight (Mv) and degree of deacetylation (DD). Chitosan with a higher degree of deacetylation tends to have a higher antimicrobial activity[22]. Chitosan is more effective than chito-oligosaccharides (COS) in inhibiting growth of bacteria; for example, water insoluble chitosans exhibited higher antimicrobial effect against E. coli than COS[23]. The preparation and characterization of chitosan and its biomedical applications are still limited. In this study, the antibacterial activities of water soluble chitosan against urinary tract infection bacterial suspension (*Escherichia coli, Pseudomonas aeruginosa, Klebsiella oxytoca, Staphylococcus aureus, Streptococcus pnemoniae, Klebsiella Pneumoniae, Salmonella typhi* were compared to chitosan prepared from shrimp shell waste (*Parapeneopsis stylifera*). Hydrogen peroxide was used to degrade the chitosan into water-soluble chitosan. The long term aim of this work is to increase the novel drug application from chitosan and water soluble chitosan in the medical industry.

II. MATERIALS AND METHODS

2.1. Chemicals

Hydrogen peroxide, acetic acid, hydro chloric acid and sodium hydroxide and all the other chemicals and reagents are purchased from Sigma Chemical Co.

2.2. Extraction of chitin from shrimp shells:

The *P. stylifera* shrimp shell wastes were collected from the Versova landing centre, Mumbai. Shells are removed and thoroughly washed with running tap water with sample care so as to remove sand adhered to it, the exoskeleton were subjected to shade drying for 2 days and then placed in hot air oven for at 60^0C for 24 hours. The preparation of chitin from shrimp shell followed by[24] with some modification. Diluted HCl solution was used for demineralization. One hundred grams of shrimp shell powder was immersed in 1000 ml of 7% (w/w) HCl at room temperature (25°C) for 24 h. After filtration with mid speed filter paper, the residue was washed with distilled water to neutral.

Then the residue was immersed in 1000 ml of 10% (w/w) NaOH at 60°C for 24 h for deproteination. The proteins were removed by filtration. Distilled water was used to wash the residue to neutral. Then the shrimp shell residue was subjected to the above program for two times. 250 ml of 95% and absolute ethanol were sequentially used to remove ethanol-soluble substances from the obtained crude chitin and to dehydrate. An air oven was taken to dry the chitin at 50°C overnight.

2.3. Preparation of chitosan and water soluble chitosan:

The preparation of chitosan and water soluble chitosan followed by[24] with some modification. The chitin (10g) was put into 50% NaOH at 60°C for 8h to prepare crude chitosan. After filtration, the residue was washed with hot distilled water at 60°C for three times. The crude chitosan (4.1g) was obtained by drying in an air oven at 50°C overnight. One gram of crude chitosan was added into 20 ml of 2% (w/w) acetic acid in a water-bath shaker. The conditions were set as follows: H_2O_2 level (4%), time (4 h) and temperature (60°C). After reaction, 10% NaOH was used to adjust the solution to neutrality. The residue was removed by filtration, while twofold volumes of ethanol were added to the filtrate. The crystal of water-soluble chitosan was liberated after incubation at ambient condition overnight and dried in an air oven at 50°C. The recovery (%) was calculated as (the weight of water-soluble chitosan/the weight of crude chitosan) ×100.

2.4. Fourier Transform - Infra Red spectroscopy (FT-IR):

The chitin, chitosan, water soluble chitosan, standard chitin and chitosan were determined using FT-IR spectrometer (Bio-Rad FTIS-40 model, USA). Sample (10 µg) was mixed with 100 µg of dried Potassium Bromide (KBr) and compressed to prepare a salt (10 mm diameter).

2.5. Assay of antibacterial activity of crude and water-soluble chitosan:

This assay was done according to the method of[25] with some modifications. 50 µl of urinary tract infection bacterial suspension (*Escherichia coli, Pseudomonas aeruginosa, Klebsiella oxytoca, Staphylococcus aureus, Streptococcus pneumoniae, Klebsiella pneumoniae,* and *Salmonella typhi)* was inoculated in a petri dish with Muller Hinton agar medium. After incubation at 37°C for 24h, the diameters of inhibition zones (in mm) were measured. Sterilized distilled water was used for control. All the Pathogenic bacterial strains were obtained from Raja Muthiah Medical College, Annamalai University. The concentrations of crude chitosan and water-soluble chitosan used in this assay were 500µg and 1mg respectively. The positive control was used as streptomycin and negative control was sterile double distilled water.

III. RESULTS

The yield of chitin and chitosan from *P. stylifera* shrimp shell waste was 32% and 54.31%, respectively. Chitin was prepared by using acid and alkaline treatments; the yield of chitin was 32% in the total weight of the dried *P. stylifera* shells, after N- acetylation, the yield of chitosans were in the range of 54.31%. Whereas in the case of water soluble chitosan obtained from the chitosan of *P. stylifera* was 87.8%.

Infrared spectroscopy of the structure changes of initial chitin, chitosan and water soluble chitosan were confirmed by FTIR spectroscopy with standard chitin and chitosan (Fig: 1-5). The FT-IR spectrum of chitin revealed that the peak 3293 cm^{-1} indicates the presence of OH stretching coupled and 2961 cm^{-1} indicates the presence of NH stretching. Compare to standard chitin this stretching wave number was more or less same. The wave number 2933 cm^{-1} characteristic of asymmetrical stretching of CH_2, whereas 1214 cm^{-1}, 1138 cm^{-1}, 933 cm^{-1} and 743 cm^{-1} positions of the spectrums are the characteristic C=O stretching, $CN_3H_5^+$, COH, CH, C-O and Skeletal stretch respectively (Table-1). These asymmetrical stretching, bending and skeletal stretch indicated that the presence of the chitin.

The standard chitosan peaks, six were found to be prominent and were representing chitosan (Structural unit - 3436cm^{-1}, (-NH2) Amide II 1636cm^{-1}, PO3 4^- 1322cm^{-1}, (NH) Amide III 894cm^{-1} and NH-out of plane bending 778cm^{-1}. The peak of crude chitosan and water soluble chitosan peak stretching was near by the standard chitosan wave number absorption only. This wave number absorption implies the substantiation of the chitosan and water soluble chitosan from the *P. stylifera* shrimp shell waste (Table 2).

In-vitro antibacterial screening of chitosan and water soluble chitosan from *P. stylifera* against selected clinical isolates were performed and zone of inhibition were given in Table 3. The concentration of chitosan and water soluble chitosan were 500µg and 1mg/ml respectively. All the experiment was done as a triplicate. The maximum inhibition zone (18.3 mm) was observed against the *S. aureus* in water soluble chitosan (1mg/ml). Compare to positive control streptomycin (11.6 mm), water soluble chitosan zone of inhibition was high. The range of inhibition in crude chitosan 1.4 mm to 8.9 mm. highest zone of inhibition was observed in *S.aureus* followed by *E.coli*, and *P.aeuroginosa*. The water soluble chitosan zone of inhibition range was high compare to crude chitosan as well as concentration wise also higher activity observed from the water soluble chitosan. Both crude and water-soluble chitosan showed higher inhibition activity against *S. aureus*, compared with the other bacteria tested. This indicated that both chitosans might have the antibacterial inhibition mechanism.

IV. DISCUSSION

The yield of chitin was 32% in the total weight of the dried *P. stylifera* shells, after N- acetylation, the yield of chitosans were in the range of 54.31%.[26] reported that, the crude polysaccharide was obtained as a water soluble dust-coloured powder from plant root of *B. chinense* by hot water extraction. The total yield of crude water-soluble polysaccharides was 6.5% of the dried material. The cuttlebone of *Sanguisorba officinalis* was found to be 20% of chitin[27-28], whereas in general, the squid/ cuttlefish reported 3-20% of chitin[29]. One of the major problems related to the preparation of pure chitins is keeping a structure as close as possible than the native form is to minimize the partial deacetylation and chain degradation caused by demineralization and deproteinization applied during process of the raw materials. Shrimp chitin showed no color and odor. Chitin occurs naturally partially deacetylated (with a low content of glucosamine units),

depending on the source[30]; nevertheless, both α - and β - forms are insoluble in all the usual solvent, despite natural variation in crystallinity. The insolubility is a major problem that confronts the development mechanisms and uses of chitin. But present study in the case of water soluble chitosan we obtained 87.8%. The β- chitin is more reactive than the α- form, an important property with regard to enzymatic and chemical transformations of chitin[31].

[32]observed that IR spectrum of chitosan oligomers showed peaks assigned to the polysaccharide structure at 1155, 1078, 1032, and 899 cm−1, and a strong amino characteristic peak at around 3425, 1651, and 1321 cm−1 were assigned to amide I and III bands, respectively. The peak at 1418 cm−1 is the joint contribution of bend vibration of OH and CH. [33] reported that IR spectrum of water soluble polysaccharide from *Bupleurum chinense* revealed a typical major broad stretching peak at 3411 cm^{-1} for the hydroxyl group, and a weak band at 2919 cm^{-1} showed C–H stretching vibration. The broad band at 1610 cm^{-1} was due to the bound water. The band at 842 cm^{-1} and 877 cm^{-1} indicated a- and b-configurations of the sugar units simultaneously existing in the polysaccharide. In the present study crude chitosan and water soluble chitosan observation band also similar to the following wave number such as chitosan 3429 cm^{-1}, 1568 cm^{-1},1559 cm^{-1}, 1405 cm^{-1}, 1105 cm^{-1} and 929 cm^{-1}. The water soluble chitosan stretching peak at 3399 cm^{-1} and 1654 cm^{-1},1647 cm^{-1}, 1078 cm^{-1} and 644 cm^{-1}.

The antimicrobial activity of chitin, chitosan, and their derivatives against different groups of microorganisms, such as bacteria, yeast, and fungi, has received considerable attention in recent years. Two main mechanisms have been suggested as the cause of the inhibition of microbial cells by chitosan. The interaction with anionic groups on the cell surface, due to its polycationic nature, causes the formation of an impermeable layer around the cell, which prevents the transport of essential solutes. It has been demonstrated by electron microscopy that the site of action is the outer membrane of gram negative bacteria. The permeabilizing effect has been observed at slightly acidic a condition in which chitosan is protonated, but this permeabilizing effect of chitosan is reversible[34].

Chitosan has been confirmed to possess a broad spectrum of antimicrobial activities[35]. However, the low solubility of chitosan at neutral pH limits its application. In this study H_2O_2 was taken to degrade the chitosan into water soluble chitosan. Several studies prove that an increase in the positive charge of chitosan makes it bind to bacterial cell walls more strongly[36-37]. [38]have mentioned that molecular weight is the main factor affecting the antibacterial activity of chitosan, from the results obtained. In contrast, some authors have not found a clear relationship between the degree of deacetylation and antimicrobial activity. These authors suggest that the antimicrobial activity of chitosan is dependent on both the chitosan and the microorganism used[39-40]. [41]studied the antimicrobial activity of hetero-chitosans with different degrees of deacetylation and Molecular weight against three Gram negative bacteria and five Gram-positive bacteria and found that the 75% deacetylated chitosan showed more effective antimicrobial activity compared with that of 90% and 50% deacetylated chitosan. In the present study 87.8% deacetlated water soluble chitosan showed higher antibacterial activity against *S.auerus* than crude chitosan. This indicated that both chitosans might have the antibacterial activity which could be used in pharmacological research.

V. CONCLUSION

We deduce that, the continuing and overwhelming contribution of water soluble chitosan to the development of new pharmaceuticals are clearly evident and need to be explored. After taking in to consideration the immense side effects of synthetic drugs, great attention has to be paid for the discovery of novel drugs from marine crustaceans waste.

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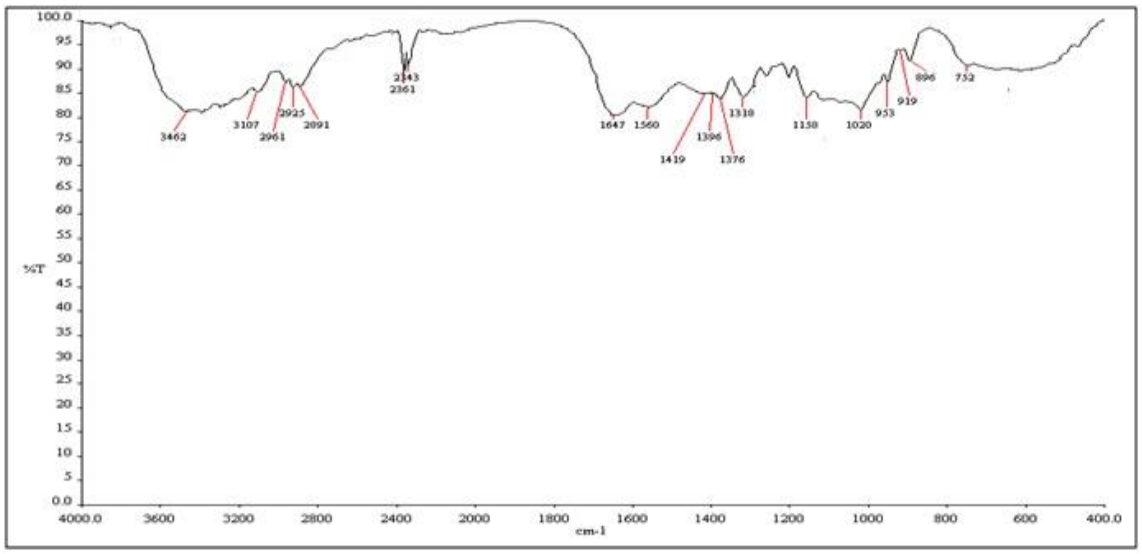

Fig: 1 FT-IR spectrum of standard chitin

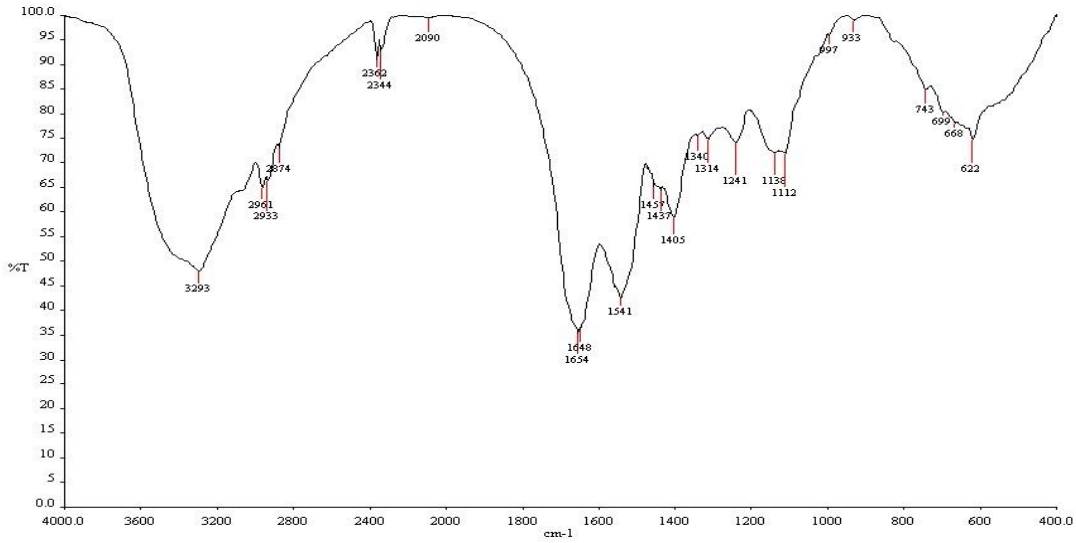

Fig: 2 FT-IR spectrum of chitin from *P. stylifera* shrimp shell waste

Table-1: Main bands observed in the FT IR spectra of standard chitin and *P. stylifera* shrimp shell waste

Vibration mode (Pearson et al., 1960)	**Std. Chitin (α-chitin) (cm^{-1})**	**Chitin from *P. stylifera* (cm^{-1})**
OH stretching	3462	3293
NH stretching	3107	2961
Symmetric CH3 stretching and asymmetric CH2 stretching	2925	2933
Amide I band	1647	1654 and1648

Amide II band	1560	1541
CH2 bending and CH3 deformation	1419	1437 and 1405
Amide III band and CH2 wagging	1318	1314
Asymmetric bridge O2 stretching	1150	1214
CO-stretching	1020	1138
CH3 wagging alone chain	953	933
NH-out of plane bending	752	743

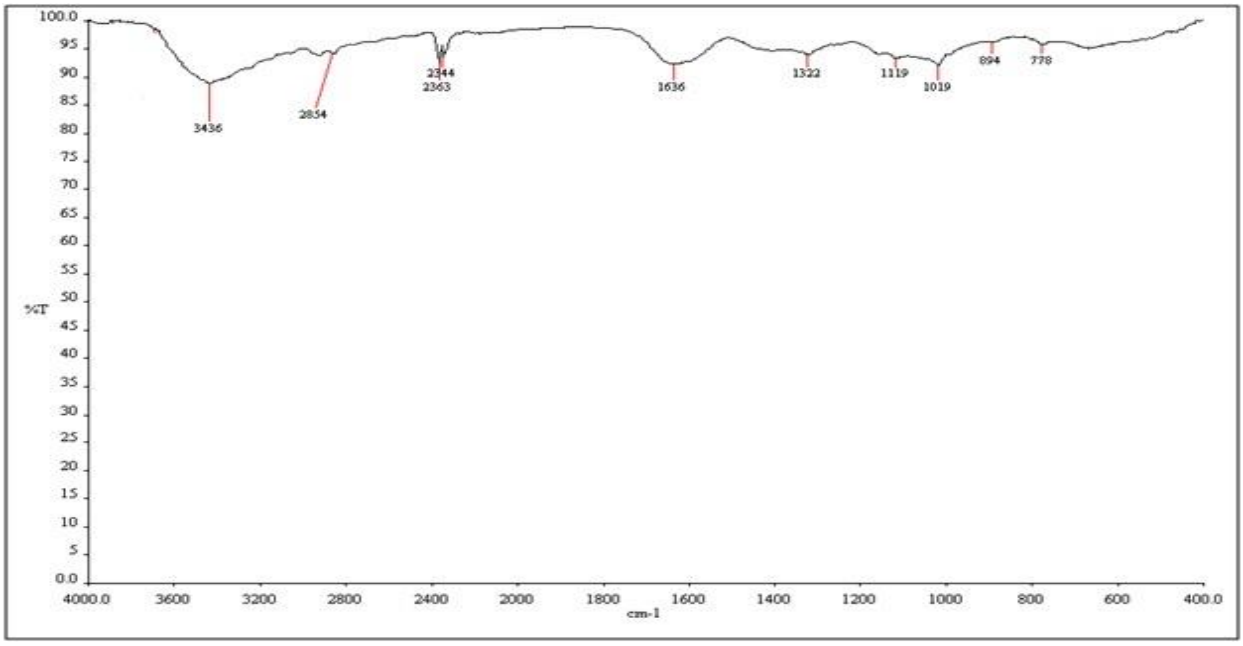

Fig: 3 FT-IR spectrum of standard chitosan

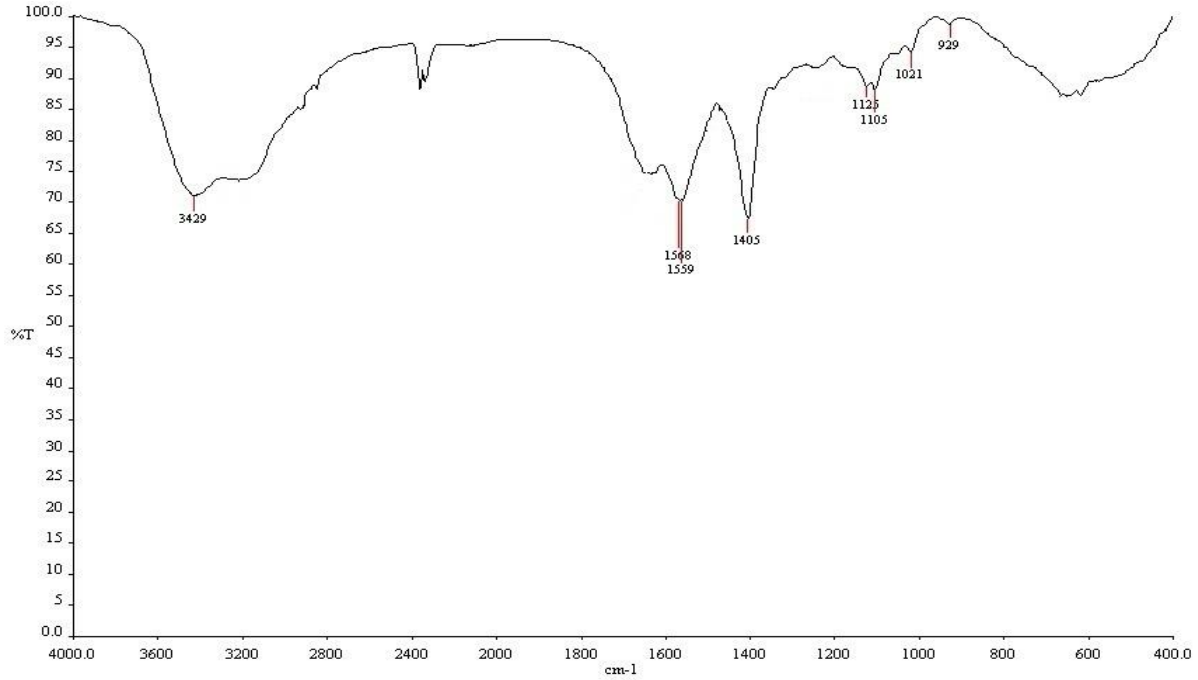

Fig: 4 FT-IR spectrum of crude chitosan from *P. stylifera* shrimp shell waste

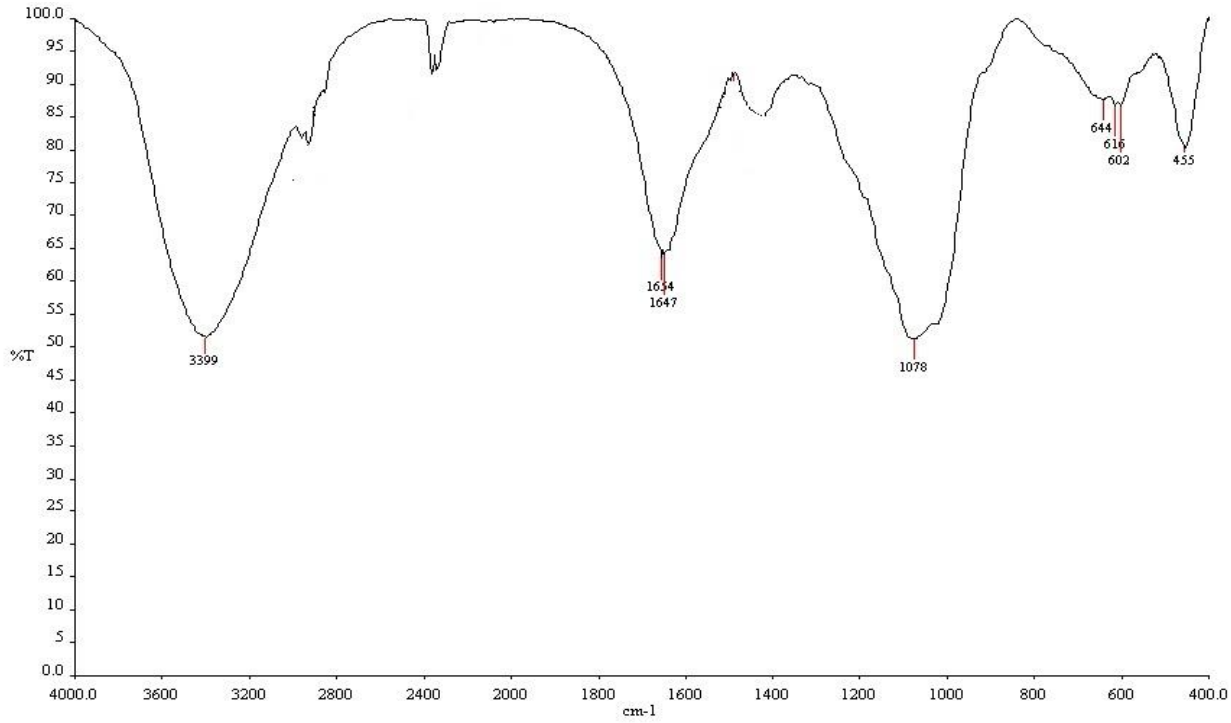

Fig: 5 FT-IR spectrum of water soluble chitosan from *P. stylifera* shrimp shell waste

Table-2: Wave length of the main bands obtained from the standard chitosan and Water soluble chitosan from *P. stylifera* shrimp shell waste

Vibration mode Chitosan Shell	**Std. chitosan**	**Crude Chitosan**	**Water soluble chitosan**
Structural unit	3436	3429	3399
(-NH2) Amide II	1636	1568 and 1559	1654 and 1647
$PO3\ 4^-$	1322	1405	---
$PO4\ 3^-$	1019	1105 and 1021	1078
(NH) Amide III	894	929	-
NH-out of plane bending	778	--	644

Table-3: Antibacterial activity of the crude chitosan and water soluble chitosan from *P. stylifera* shrimp shell waste:

	Inhibition Zone (mm)					
Microorganisms	Crude chitosan		Water soluble chitosan		Positive control	Negative control
	500µg/ml	1mg/ml	500µg/ml	1mg/ml		
E. coli	5.2	8.4	7.3	10.4	10	-
P. aeruginosa	4.3	6.1	7.5	8.4	6	-
K. oxytoca	-	3.2	4.0	7.3	5	-
S. aureus	6.4	8.9	10.2	18.3	17.6	-
S. pneumoniae	2	4.3	5.1	6.2	6	-
K. pneumonia	-	-	-	4.4	4.5	-
S. typhi	1.4	4.2	4.7	6.6	6.8	-

-, No activity was observed

A Multi-stage Learning Method with the Selection of Training Data for the Layered Neural Networks

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Abstract- This paper proposes an efficient learning method for the layered neural networks based on the selection of training data. The multilayer neural network is widely used due to its simple structure. When learning objects are complicated, the problems, such as unsuccessful learning or a significant time required in learning, remain unsolved. The aims of this paper are to suggest solutions of these problems and to reduce the total learning time.

Focusing on the input data during the learning stage, we undertook an experiment to identify the data that makes large errors and interferes with the learning process. Our method devides the learning process into several stages.

Computational experiments suggest that the proposed method has the capability of higher learning performance and needs less learning time compared with the conventional method. And, compares the QPROP and RPROP methods by computation experiments using function approximation problems as an example and demonstrates the efficacy of integrating the proposed method with either method.

Index Terms- Multi-layer Neural Network, Data Selection, Multi-stage Learning, Integrating the Proposed Method with Either Method

I. INTRODUCTION

Majority of neural network research focuses on complicated networks such as pulsed neural networks [1]. However, in reality, multilayer neural networks (NNs) that are composed of sigmoid units using back propagation (BP) are widely used because of their simple structures.

In our previous paper [2], we proposed a comprehensive learning method that features the categorization of training data based on the ease of learning prior to initiating a learning session. Multi-stage learning based on the dynamic adjustment of the learning rate in response to error size and input characteristics to the output layer unit (effectively utilizing amplitude diminishing conditions and desired value acquisition conditions) was applied in complex learning cases that require the use of large quantities of learning data.

Proactive use of the oscillation characteristic of BP through the proposed method [2] is expected be effective when integrated with the Quick-Prop method (QPROP method) [4], an extension of BP, or when integrated with the elastic/resilient BP method (RPROP method) [5]. The QPROP method makes adjustments that are as large as possible to the learning rate while also adding an inertia term to the weight coefficient variation to suppress oscillation. To avoid oscillation due to sudden changes in the weight coefficient, it also applies a maximum variation limit. Generally, three coefficients—the learning coefficient, the maximum variation, and the weight renewal suppression coefficient—are required. Furthermore, the elastic BP method [5] confirms the weight renewal code and suppresses oscillation to accelerate learning. Normally, it is a method that learns by adjusting five parameters such as the learning coefficient in response to current conditions. In both methods, learning occurs with acceleration parameters while oscillation is suppressed.

In this study, we report the effects of the integration of the proposed method with QPROP and RPROP on error and learning time.

The remainder of this paper is composed as follows. Section 2 discusses the QPROP and RPROP methods. Section 3 compares the QPROP and RPROP methods by computation experiments using function approximation problems as an example and demonstrates the efficacy of integrating the proposed method with either method. Section 4 summarizes our findings.

II. EFFICIENT LEARNING METHODS

Paper [2] discusses the dynamic adjustment of the learning coefficient in response to error, input characteristics to the output layer unit, the selection of learning data, and effective learning methods based on these properties. Here we will briefly discuss the QPROP and RPROP methods.

2.1. QPROP method and RPROP method

The QPROP method [5] is one of the BP methods where learning speed is enhanced while learning coefficients are made as large as possible and inertial terms are added to the amount of change of weight coefficients. The inertial terms are where the weights are updated by considering the previous amount of updated weight as well as the present amount of updated weight. That is, when the amount of corrected weight at the time k is obtained, the formula will be as follows by considering the amount of corrected weight of (k-1):

$$\Delta W_{ji}(k)=\rho^{k}S_{ji}(k)+\alpha_{ji}^{k}\Delta W_{ji}(k-1) \quad (1)$$

But,

$$S_{ji}(k)=\frac{\partial E(k)}{\partial w_{ji(k)}}+\lambda\Delta W_{ji}(k) \quad (2)$$

and the weight is updated with the Formula (k=1,2,…) where the subscript i is a middle layer unit i, and the subscript j means the output layer unit j. The mark λis a suppressing coefficient of updating weights. Let the learning coefficient be ρ^{k}=ρ when either of the conditions of ρ^{k} be $\Delta W_{ji}(k-1)=0$ or $S_{ji}(k)\,\Delta W_{ji}(k-1)>0$ is met and let the learning coefficient be ρ^{k}=0 when none of the conditions is met. For the inertial term α_{ji}^{k}, calculate $\tilde{\alpha}_{ji}^{(k)}=\frac{S_{ji(k)}}{S_{ji(k)}-S_{ji(k-1)}}$ and let α_{ji}^{k}=μ when either of the conditions of $\tilde{\alpha}_{ji}^{(k)}>\mu$ or $S_{ji}(k)\tilde{\alpha}_{ji}^{(k)}\Delta W_{ji}(k-1)<0$ is met and let $\alpha_{ji}^{k}=\tilde{\alpha}_{ji}^{(k)}$ when the answer does not meet any of the conditions.

The RPROP method [6] is where suppressed-vibration learning is carried out by adjusting five parameters in accordance with conditions, while the previous amount of updated weight should be memorized to suppress vibrations and a sign of the previous amount of updated weight and that of the current amount of updated weight are taken into consideration. Let $m=\frac{\partial E(k-1)}{\partial w_{ji}(k-1)}*\frac{\partial E(k)}{\partial w_{ji}(k)}$

and the amount of updated weight varies depending on signs of m. Basically, weights are updated based on the three formulas indicated below. The formulas feature that Δ_{max} is set so that the updated amount will not be so large and Δ_{min} is set so that the updated amount will not be so small.

m>0

$$\Delta_{ji}(k)=\min(\Delta_{ji}(k-1)*\mu^{+},\Delta_{max})$$
$$\Delta W_{ji}(k)=-\text{sign}\left(\frac{\partial E(k)}{\partial w_{ji}(k)}\right)*\Delta_{ji}(k) \quad (3)$$
$$W_{ji}(k+1)=W_{ji}(k)+\Delta W_{ji}(k)$$

m<0

$$\Delta_{ji}(k)=\max(\Delta_{ji}(k-1)*\mu^{-},\Delta_{min})$$
$$W_{ji}(k+1)=W_{ji}(k)-\Delta W_{ji}(k-1) \quad (4)$$
$$\frac{\partial E(k)}{\partial w_{ji}(k)}=0$$

m=0

$$\Delta W_{ji}(k)=-\text{sign}\left(\frac{\partial E(k)}{\partial w_{ji}(k)}\right)*\Delta_{ji}(k) \quad (5)$$
$$W_{ji}(k+1)=W_{ji}(k)+\Delta W_{ji}(k)$$

Table 1 Learning Functions

(a) **Rastrigin Function**
$f(x,y)=x^2-10\cos(2\pi x)+y^2-10\cos(2\pi y)$
$-0.8 \leq x \leq 0.8,\ -0.8 \leq y \leq 0.8$)
$-20.0 \leq f(x,y) \leq 20.5$
(b)**Ridge Function**
$f(x,y)= 2x^2+2xy+y^2$
$-6.0 \leq x \leq 6.0,\ -6.0 \leq y \leq 6.0$
$0.0 \leq f(x,y) \leq 180.0$

III. Computational Experiments Using Function Approximation Problems as Examples

To demonstrate the efficacy of the multistage learning method proposed in this study, we perform computational experiments using function approximation problems as an example. These problems have been chosen, because the difficulty of learning can be configured with the selection of different functions, and experiments of unlearned data can be verified easily. A number of useful functions are the well-known Rastrigin and Ridge functions for optimization problems. These are given in Table (1).

3.1. Training Data

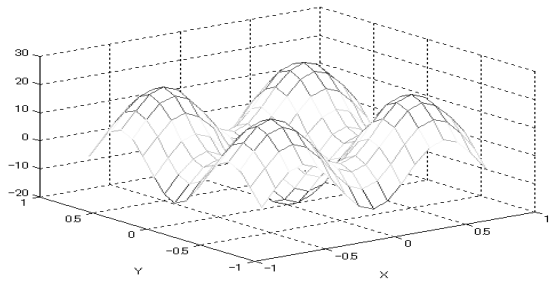
Table (1)(a) is shown in Figure 1. In Figure 1, adjacent training data are shown joined by a solid line. The region to learn in Table (1)(a) is given as $-0.8 \leq x, y \leq 0.8$ for the Rastrigin function. The training data sets the x, y direction stride as 0.1 and partitions the region into a 17 × 17 (289 point) grid given as training data set $\boldsymbol{D}$.

Table (1)(b) gives the region to learn as $-6.0 \leq x, y \leq 6.0$ for the Ridge function. The training data sets the x, y direction stride as 1.2 and partitions the region into an 11 × 11 (121 point) grid given as the training data set $\boldsymbol{D}$.

3.2. Training Data Selection

Here we discuss the method used to select the training data in Table (1)(a) for many experiments due to the difficulty of learning. The selection of the training data in table(1)(a)can be substituted with values obtained by the partial differentiation of functions $f_x(x, y)$ or $f_y(x, y)$ in both x and y directions.

Figure 1. An example of characteristics of the target function



Next, we divide the training data into three steps (s = 3) and perform multistage learning. $\boldsymbol{D}_l^1 \cup \boldsymbol{D}_d^1$ uses approximately 30% of all training data. $\boldsymbol{D}_l^1$ and $\boldsymbol{D}_d^1$ (selected using the values of $f_x(x, y)$ or $f_y(x, y)$ are important, because they grasp the outline of the function. The final number of data included in $\boldsymbol{D}_l^1 \cup \boldsymbol{D}_d^1$ was 31.1% (90 pieces). Within this data, 24 pieces fulfilled $|f(x, y)| \geq 0.90$, 34 fulfilled $|f_x(x, y)| \geq 0.3$, and 34 fulfilled $|f_y(x, y)| \geq 60.3$. Two pieces were repeated; thus, the total was 90 pieces.

Structure	Initial weight vector is modified 5 times. 2 input layer units. 9 middle layer units. 1 output layer unit. Updating a weight vector should be done every 1 epoch.
Learning data	Learning domains are $-6.0 \leq x \leq 6.0$, $-6.0 \leq y \leq 6.0$ Learning data is conditioned equally as the

	reference table(1)(c). All the domain is divided into 11×11 grid. All the learning data(***D***) is divided into 3 parts before learning. $0<D^i<1$, $D^i \subset \boldsymbol{D}$ ($D^i = f(x_i, y_i); i=1,\ldots,121$)
1st step	$\boldsymbol{D}_d^1 = \{f(x_i,y_i) \mid \lvert f_x(x_i,y_i)\rvert \geq 20 or \lvert f_y(x_i,y_i)\rvert \geq 20\}$ (38 learning data), $\lvert \boldsymbol{D}_l^1 \cup \boldsymbol{D}_d^1\rvert$=6(6 overlapped) 32 learning data are selected. (26.4%)
2nd step	$D_d^2 = \{f(x_i,y_i) \mid \lvert f_x(x_i,y_i)\rvert \geq 16 or \lvert f_y(x_i,y_i)\rvert \geq 16\}$ (70learning data) $\lvert D_l^2 \cup D_d^2\rvert$=20(20 overlapped) 50 learning data are selected. (about 41%).
3rd step	121 learning data are selected. (100%)
Each step	Every step are learned for 2,334 times.
Conventional method	Learn 7,000 times with all the learning data and learning coefficient 0.8. Initial weight vector is modified 5 times.

Table 2 : The method of Learning(Ridge Function)

In step 2, the final number of data included in $\boldsymbol{D}_l^2 \cup \boldsymbol{D}_d^2$ was 58.1% (168 pieces). Within this data, 52 pieces fulfilled $|f(x, y)| \geq 0.80$, 68 pieces fulfilled $|f_x(x, y)| \geq 60.1$, and 68 pieces fulfilled $|f_y(x, y)| \geq 60.1$. Twenty pieces were repeated; thus, the total was 168 pieces.

In $\boldsymbol{D}_d^3$ in step 3, all data was used. Thus, 31.1% of all data was used in step 1, 58.1% of all data was used in step 2, and 100% of all data was used in the final step. Table (2) shows this information for Table (1)(b). As a feature of the Ridge function shown in Table (1)(b), because $|f(x_i, y_i)|$ becomes larger when the values of $|f_x(x_i, y_i)|$ or $|f_y(x_i, y_i)|$ increase, the training data was selected on the basis of the values of $|f_x(x_i, y_i)|$ and $|f_y(x_i, y_i)|$.

3.3.Experiment Settings for Methods Used in Comparison

To verify the efficacy of the proposed method, the results of computational experiments were compared with those of existing methods. The NN used for the proposed method, existing method, QPROP method, RPROP method, and learning method, with the integrated proposed method, is a three-layer feed-forward network using sigmoid elements. This is composed of two input-layer elements, nine middle-layer elements, and one output-layer element. The number of middle-layer elements was determined by preliminary experiments. All weight coefficients were renewed once per epoch in each method. The learning count for the proposed method was 2333 epochs for step one, 2333 epochs for step two, and 2334 epochs for step three for a total of 7000 epochs. The two existing methods used all training data for 7000 epochs of learning. Computations were performed on a 3.0GHz Pentium 4 PC with 2GB RAM operating on Windows XP. The basic learning coefficient for the proposed method and existing methods was η = 0.8 (standard). To evaluate the learning results, the RMSE values for the training data and average weight sets for four initial weight coefficient values were used. Weight coefficient initial values were set randomly in a range from −0.01 to 0.01.

By a number of preliminary experiments, the QPROP parameters were set as follows. The weight variation suppression coefficient λ = 0.005, maximum change amount μ = 0.95, learning coefficient $\eta 1$ = 0.80, and by adjusting the direction of weights, the learning coefficient $\eta 2$ = 0.12. For RPROP experiments conducted, the following parameters were used: Δ_{max} = 5.0, Δ_{min} = 0.0025, η^+ = 0.97, η^- = 0.89, η^0 = 0.61.

Figure 2. Input characteristics for the +QPROP methods in the output unit(Ridge function, $\eta = 0.8$)

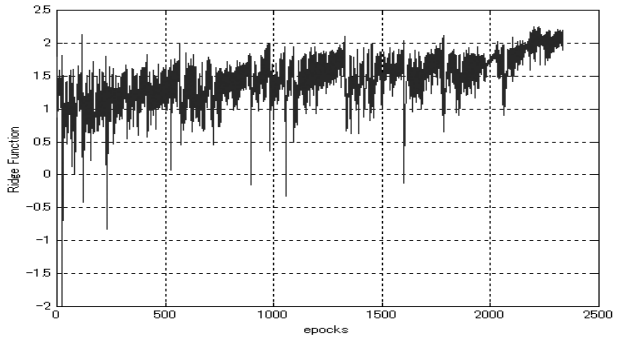
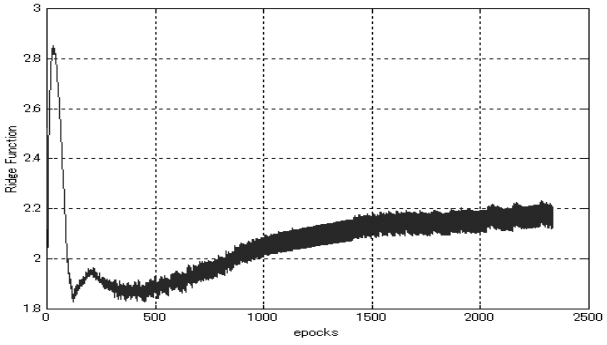


Figure 3. Input characteristics for the +RPROP methods in the output unit(Ridge function, $\eta = 0.8$)



3.4. Computational Experiment Results

Table 3 displays the learning results using the proposed method and the QPROP and RPROP methods independently, as well as the results of using the QPROP and RPROP methods with the proposed method integrated. The values shown in Table 3 is the averages for five experiments conducted with five different initial weight values selected randomly.

As shown in Table 3, integrating the proposed method with the RPROP or QPROP method improved the RMSE (mean) value and decreased the learning time when learning with the Ridge function (Figures 2 and 3).

As shown in Table 4, when the proposed method is applied to the Rastrigin function shown in Table (1)(a), all RMSE values are below 0.1 when the learning coefficient (the standard when using the proposed method) is between 0.9 and 0.2. When the learning coefficient decreases to 0.1 or 0.05, the RMSE values increase above 0.1, but the useful range of the learning coefficient value is extremely wide.

Conversely, as shown in Table 5, learning was carried out while changing the learning coefficient in the range from 0.9 to 0.02 for the two existing methods as well. From 0.9 to 0.3, no learning occurred with the RMSE values above 0.1. From 0.2 to 0.04, the RMSE value decreased below 0.1(Mean learning time = 28 minutes 17 seconds). Conversely, when reduced to 0.02, the RMSE value increased above 0.1.

When learning using existing methods with a constant learning coefficient of 0.04, oscillation occurs in the final 2334 repetitions, and the RMSE value is 0.053. These results are shown in Table 5. The input characteristics to the output layer unit are shown in Figure 4.

	RMSE		
	Mean value	**Maximum value**	**Minimum value**
Proposed method	0.107	0.235	0.038
Learning time	Mean 6 minutes 59 seconds		
Proposed method+QPROP	0.058	0.113	0.024
Learning time	Mean 7 minutes 10 seconds		
QPROP method	0.172	0.241	0.108
Learning time	Mean 10 minutes 24 seconds		
RPROP method	0.050	0.080	0.023
Learning time	Mean 13 minutes 52 seconds		
Proposed method+ RPROP method	0.038	0.051	0.030
Learning time	Mean 7 minutes 31 seconds		

3.5. Considering the Experimental Results

Here we consider the performance and calculation time of learning by integrating the proposed method with the QPROP and RPROP methods on an NN and discuss the necessity of oscillation based on the computational experiment results.

Integrating the Proposed Method into QPROP and RPROP Methods

Figure 4. Input characteristics for the traditional methods in the output layer.

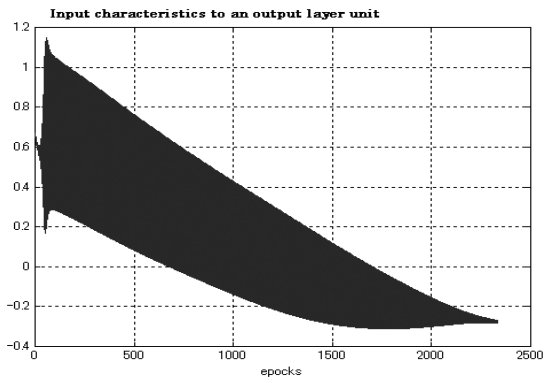


Table 3: RMSE for Proposed Methods and Traditional Methods in Learning of the Ridge Function

As shown in Table 3, by integrating the proposed method with the QPROP and RPROP methods, the RMSE average when learning with the Ridge function reduced in comparison with that before integration. In independent learning before integration, average error values for the RPROP method, proposed method, and QPROP method were 0.050, 0.107, and 0.172, respectively. However, by integrating the proposed method into the RPROP method, error reduced to 0.036. Integration with the QPROP method reduced the error from 0.172 to 0.058. When the QPROP and RPROP methods were used independently, the input characteristics to the output element unit showed almost no oscillation, but integrating the proposed method added a small amount of oscillation.

Table 4: RMSE for Proposed Methods in Learning of the Rastrigin Function

Proposed method	**RMSE**
Learning time	**Mean 15 minutes 56 seconds**
Learning coefficient	RMSE value
0.9	0.050
0.8	0.126
0.7	0.037
0.6	0.027
0.5	0.026
0.4	0.039
0.3	0.081
0.2	0.096
0.1	0.138
0.05	0.135

Table 5: RMSE for Traditional Methods in Learning of the Rastrigin Function

Proposed method	**RMSE**
Learning time	**Mean 28 minutes 17 seconds**
Learning coefficient	**RMSE value**
0.9	0.618
0.8	0.233
0.6	0.185
0.5	0.199
0.4	0.183
0.3	0.162
0.2	0.045
0.1	0.025
0.05	0.035
0.04	0.053
0.02	0.112

IV. Conclusion

This study discussed a comprehensive learning method for neural networks that categorized training data based on the ease of learning before learning, performed multistage learning based on those categories, and dynamically adjusted the learning coefficient in response to error size. Experiments were conducted by integrating the method with QPROP and RPROP methods. As a result, the RMSE value decreased, and the learning time improved.

In future studies, we will actively use oscillation characteristics, extend the proposed method to evaluate effects on pattern recognition problems, and investigate further decreases in error.

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Thermodynamic Studies on Activity and Stability of Immobilized *Thermomyces lanuginosus* in Producing Fatty Acid Methyl Ester (FAME)

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Abstract- The effect of temperature on enzymatic transesterification was studied by many researchers for process optimization purposes. The optimum operating temperature was defined on the basis of high FAME productivity or lipase activation without considering the lipase denaturation factor. High thermal energy was favored to accelerate transesterification reaction but lipases for enzymatic transesterification reaction were susceptible to denaturation even at moderate temperature operation. In this work, studies of thermal effects on lipase kinetics were carried out to propose the most desirable operating temperature that achieves high FAME productivity while preserving lipase catalyzing activity. Catalyzing activity of Lipozyme TL IM increased with temperature up to a threshold at 40°C and successive fell beyond this value have explained the occurrence of reversible biocatalyst inactivation. The reaction rates obtained for experiment under different heat treatments have confirmed the deactivation process of Lipozyme TL IM follows first-order kinetics pattern. The most desirable operating temperature for transesterification reaction is 40°C that leads to the highest productivities, 100 % FAME yield at 4 hrs while preserving acceptable stability levels of 1.09% activity lost after 1 hrs.

Index Terms **--** Biodiesel, Heat treatment, Lipase,Temperature effect, Thermodynamics

I. INTRODUCTION

Crude palm oil transesterification mediated by immobilized *Thermomyces lanuginosus* or Lipozyme TL IM in tert-butanol solvent system, have been attempted in previous study with encouraging FAME yield and initial reaction rate being achieved [1]. According to the principles of Arrhenius and Vant Hoff, transesterification reaction rate can be accelerated by either increasing operating temperature or increasing concentration of the reactants that helps to increase the rate of reactants collision between each other. The reaction is commenced in condition the colliding reactants possess sufficient activation energy to overcome the energy potential barrier. The free energy of activation, ΔG thus acts as a potential-barrier for the reaction to take place. Activation energy can be obtained from experimental studies with different reaction rate being determined at the studied ranges of operating temperature [2].

Nevertheless, transesterification reaction catalyzed by immobilized lipases under high temperature tend to expose Lipozyme TL IM to the conformational changes risk. Since lipases were easily denatured at high temperature in nature, the most operational working temperature thus must represent significantly high enzyme activity as a whole but to the extent that the lipases still retain good enzyme stability. Good thermal stability of lipase is of paramount important in transesterification to preserve high catalyzing activity for continuous process and for enzyme reuse in batch system. Thus, expensive cost of lipase production for transesterification can be circumvented by its reusability and long shelf life. Consequently, many researchers were concerned for enzyme stability under different thermal conditions [3, 4, 5].

Effects of temperature on stimulating enzymatic transesterification were initiated by many researchers for process optimization purpose. In fact the optimum operating temperature was defined on the basis of high FAME productivity or lipase activation without considering the lipase denaturation rate. In this work, thermodynamic studies were carried out as to propose the most desirable operating temperature that achieves high FAME productivity as a whole besides preserving lipase catalyzing activity. Hence, activation energy and reversible deactivation energy without thermal treatment was determined to detect the most productive temperature condition. Whereas irreversible deactivation energy together with half life time for Lipozyme TL IM were clearly defined after Lipozyme TL IM was preincubated in 45, 50, 55 and 60°C with different time periods (1 - 4 days). The desirable working temperature for the system was then justified from the FAME production rate and the extent of thermal resistance of the lipase towards temperature variations.

II. MATERIALS AND METHOD

A. Biocatalyst and Chemicals in Transesterification

Lipase *Thermomyces lanuginosus* immobilized on silica gel (Lipozyme TLIM) with catalytic activity of 170 IUN/gl converting 0.01% tristearin per minute under standard assay conditions was purchased from Novozymes (Bagsvaerd, Denmark). The unrefined crude palm oil (CPO) for transesterification reaction is procured from M.P. Mathew Palm Oil Mill Sdn. Bhd. located in Sungai Bakap, Malaysia. Analytical grade of tert-butanol and methanol in 99.0% purity are obtained from J.T. Baker and Merck respectively.

Chromatographic grade of lauric acid methyl ester, palmitic acid methyl ester, stearic acid methyl ester, oleic acid methyl ester, linoleic acid methyl ester and heptadecanoic acid methyl ester were purchased from Sigma.

B. Crude Palm Oil Transesterification

Unless otherwise stated, the detection of transesterification activities of Lipozyme TL IM when under different temperatures and heat treatments were carried out by monitoring FAME produced from the reaction between crude palm oil and methanol. The typical enzymatic transesterification reactions were carried out by adding 6.65 % enzyme loading in optimum reaction mixtures. From preliminary study, the optimum amount for crude palm oil, methanol and tert-butanol were 11.84, 3.2 and 16.96 ml respectively. 32 ml of reaction mixture in 100 ml of conical flasks were maintained in water bath shaker at 150 rpm mixing intensity and at 30°C. For thermodynamic studies, 50 µl aliquots of sample were withdrawn at 5, 10, 15 and 30 min for initial reaction rate determination. Product FAME achieved at various process conditions was analysed by gas chromatography. Each experimental run was performed in duplicates and the results were expressed as mean values $\pm$ standard deviation. The standard deviations for FAME were approximately $\pm$ 5% of the mean values.

C. Thermodynamic Studies on Lipozyme TL IM

Experimental works on thermodynamic studies for transesterification reaction were divided into two parts, mainly on the effects of temperature on lipase activity and stability. First, the effects of temperature in accelerating FAME production rate were implemented at various reaction temperatures of 30, 35, 40, 45, 50, 55 and 60 °C without imposing heat treatment on lipase. Second, the irreversible denaturation rate for Lipozyme TL IM when under different thermal treatment was studied to determine half-life time as well as the thermal resistance strength. The kinetics of irreversible denaturation was studied by incubating Lipozyme TL IM in the crude palm oil suspension without methanol at temperatures of 45, 50, 55 and 60 °C for variable times of 1, 2, 3 and 4 days. The Lipozyme TL IM residual activity was then examined by implementing typical transesterification reaction as described in section *B. Crude Palm Oil Transesterification*. The catalyzing activity for the lipase under thermal treatment was compared with the catalyzing activity for the lipase not subjected to thermal treatment. The comparison is necessary to determine the extent of heat destruction on lipase catalyzing activity.

D. FAME Determination with Gas Chromatography

Heptadecanoic acid methyl ester served as the internal standard. To quantify FAME produced, 0.4 µl sample was injected into Perkin Elmer Clarus 500 gas chromatography equipped with programmed split/ splitless injector (PSS) and flame ionization detector (FID). The split ratio was defined as 20:1. Nukol^{TM} fused silica capillary column with dimension of 0.53 mm i.d. x 15 m length x 0.50 µm film thickness (Supelco, USA) was used. The injector and detector temperature were 220°C and 250°C respectively. The column temperature was maintained at 110°C for 0.5 min, increased to 200°C at 10°C/min and maintained at this temperature for 10 min. Helium was employed as carrier gas with flowrate 1.2 ml/min.

III. RESULTS AND DISCUSSION

A. Effect of Temperature on Lipozyme TL IM- catalyzed Transesterification

The initial reaction rate of FAME yield under variation of temperature from 30 to 60 °C was graphed in Fig. 1. The graph clearly displayed two different temperature boundaries according to FAME produced. At temperature ranges of 30 – 40 °C, the FAME production rate increased linearly with temperature. Whereas temperature beyond 40 °C and thereafter was considered a non-productive region due to FAME production rate decreased with increasing temperature. Therefore, the catalyzing efficiency was largely enhanced with increasing temperature up to a threshold of 40 °C optimum point. The successive decreased in FAME production rate beyond 40 °C have prescribed the existing of irreversible denaturation phenomenon. The thermal inactivation of lipozyme TL IM might be due to the antagonistic interacting effect between solvent molecules with the "membrane-lipase" system that reversibly produces a conformational change on active structure of the lipase [6]. The positive relationship defined between FAME reaction rate and temperature in the study was in agreement with Arrhenius model. Thus, the activation energy and reversible unfolding energy were estimated from the semilog plot versus the reciprocal temperature. In the enzymatic transesterification with different types of acyl alcohols like methanol, 1-propanol, 2-propanol, catalyzing activity of Lipozyme TL IM were largely propagated at 40 °C as reported in literature [7, 8]. Thus, Lipozyme TL IM was most favour to produce FAME at 40 °C regardless of oil source and acyl acceptors.

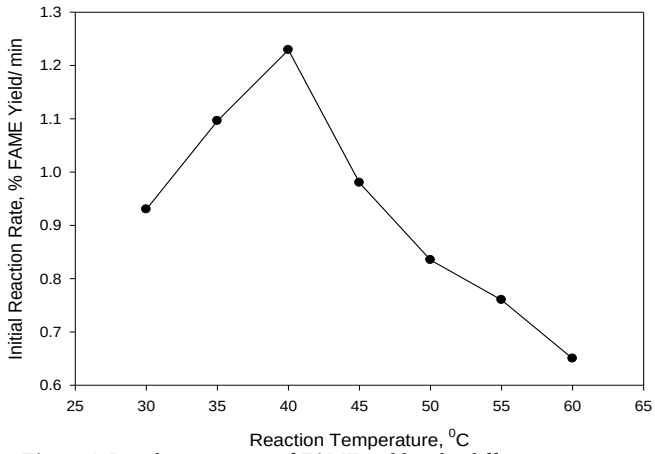

Figure 1. Initial reaction rate of FAME yield under different operating temperature.

B. Effect of Heat Treatment on Lipozyme TL IM- catalyzed Transesterification

The irreversible denaturation rate of Lipozyme TL IM was reflected by the residual activity of lipase after subjected to different thermal treatment conditions. In Fig. 2, FAME yield

reaction rate were decreasing with the extended period of heat treatment. Lipase either incubated at 45 or 60 °C were experienced almost similar denaturation rate for 1 day heat treatment or 4 days heat treatment. The result enlightened the proportional decreased of lipase activity with temperature. As clearly observed from the Fig. 2, the exponential decreased in lipase catalyzing activity was observed in 1 days of heat treatment and the rate was subsequently arrived into plateau after 4 days of heat treatment. The exponential decay of lipase activity with thermal treatment period agreed that lipase denaturation is a first-order process. Since the denaturation constant, K_d is a kinetic parameter and temperature dependent, hence the variable can be also represented by an Arrhenius-type equation.

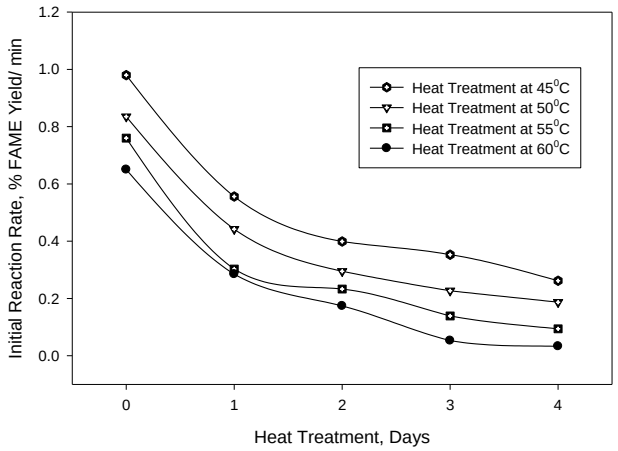

Figure 2. Initial reaction rate of FAME yield when subjected to different heat treatment conditions: 45 – 60 °C under 1 – 4 days.

C. Thermodynamic Studies on Activity and Stability of Immobilized Thermomyces lanuginosus

The dependent of initial reaction rate or enzyme activity on temperature variation is a first order process. During linear regression analysis, the relationships have been verified with high correlation coefficients of 0.9832 and 0.9913 being obtained for activation slope and denaturation slope respectively as shown in Arrhenius plot (Fig. 3). The activation energy that calculated from the negative slope of Arrhenius plot was 22 kJ/ mol whereas the reversible denaturation energy estimated from the positive slope was 26.5 kJ/ mol. The activation energy of 22 kJ/ mol for transesterification reaction suggested a low energy barrier was required for catalysis. Slightly higher reversible denaturation energy than activation energy was in agreement that lipase required much more energy to unfold lipase active conformation. This figure has been reported for immobilized *Candida antarctica* [9] lipase B and mycelium-bound carboxylesterase from *Aspergillus oryzae* [10].

The strength of lipase resistance towards heat was carried out by incubating the reaction mixture under variation of heat treatment conditions as described in materials and methods. The extent of deactivation was measured by the apparent first-order rate constant of enzyme denaturation. It was shown that denaturation constant that estimated from the slopes of plot ψ versus time, *t* (Fig. 4) progressively increased with increasing temperature from 45 to 60 °C. Consequently, the thermodynamic energy of irreversible denaturation that estimated based on the denaturation constants obtained from 45 to 60 °C was 45.18 kJ/ mol. It can be concluded that energy barrier as high as 45.18 kJ/ mol was required to impose thermal inactivation effect on Lipozyme TL IM because the binding of lipase to the carrier materials have greatly enhance lipase stability.

Activity of Lipozyme TL IM reached climax at 40 °C reaction temperature and beyond 40 °C, the lipase activity inactivated at faster rate with time. Therefore, the desirable working temperature for enzymatic transesterification reaction mediated by Lipozyme TL IM must be at 40 °C or below. The half life time for 30 and 40 °C operating temperatures were 112.87 and 63.23 hrs respectively. Even though Lipozyme TL IM gain higher life time when under 30 °C compared to 40 °C but the 40 °C gave 100 % FAME yield at 4 hrs reaction time. The specific lipase activity for 40 °C, 35.08 % FAME yield/ g lipase. hr was nearly 3 fold higher than the 30 °C of only 11.93 %FAME yield/ g lipase. hr. Therefore, 40 °C was selected as the optimum reaction temperature for the enzymatic transesterification reaction mediated by Lipozyme TL IM.

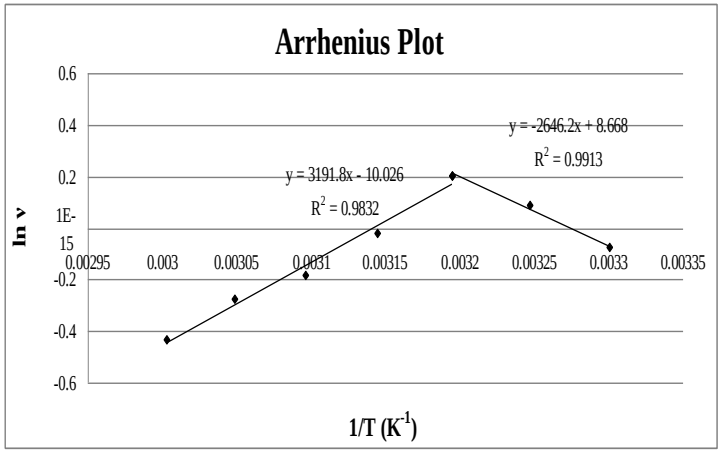

Figure 3. Arrhenius plots for the estimation of the activation energy and reversible denaturation energy.

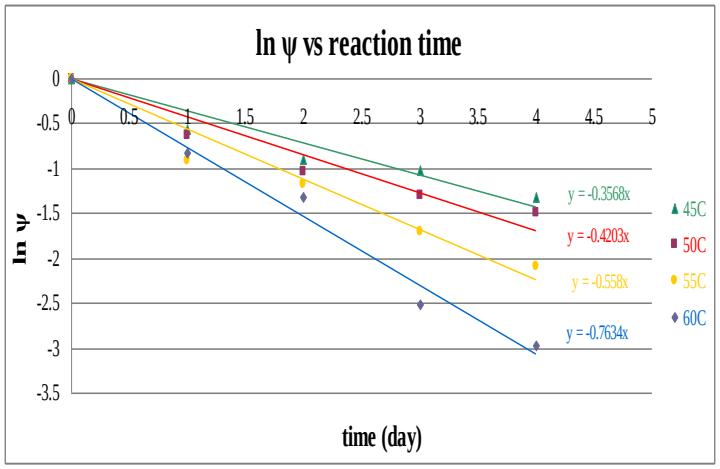

Figure 4. Semilog plots of irreversible denaturation of Lipozyme TL IM.

IV. CONCLUSION

The thermodynamic studies which involve simultaneous determination of catalyzing activity and stability of Lipozyme TL IM were particular useful to enhance production cost of enzymatic transesterification reaction. Expensive production cost of lipase for biodiesel production have resulted the lost of attention in commercializing enzymatic transesterification reaction. Thus, the appropriate operating temperature was particularly important to ensure FAME production rate generated by enzymatic reaction can compete with chemical methods besides preserving lipase activity for extensive period of lipase reuse. In this study, 40 °C was the most desirable operating temperature for enzymatic transesterification reaction where the process attained 100 % FAME yield at 4 hrs and retained 98.91 % residual activity after 1 hrs reaction time.

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An MS Excel Based Utility for Assigning Letter Grades

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Abstract- The main function and the duty of the teacher is to educate the next generation and prepare them for the challenges of the future. The task of teaching does not end by delivering the lectures, descriptions, demonstrations, or additional material that are designed to help the student follow the course. An integral part of teaching is assessment where the teacher is given the responsibility of assuring that the students have achieved the intended knowledge and skill targets prescribed in the course or the training program. Most teachers also find this as the most difficult component in their teaching career.

This software utility was developed to offer a convenient platform for the teacher to assign grades once the final marks have been computed for a certain course. It was developed using MS Excel 2007 environment which has improved flexibility in its functionality compared to its earlier versions. The teacher is given the space to transfer the final marks to the program and manipulate the grades division points in following any grade assignment protocol adopted or recommended by the educational institution. During the grade assignment procedures the program itself will provide graphical outputs of marks distribution plots in real time, to assist the teacher in his / her task. The final marks sheets can also be produced and printed with ease once the assignment is completed.

The program was given to the academic staff of the Faculty of Agriculture for their use, and a fully on-line survey was conducted where they were asked to provide their feedback on a 5 point Likert's scale. The summary of the responses were very positive in accepting the program as an efficient means of performing the grades assignment task. The design and implementation of different functions of the program were highly appreciated and accepted by the users.

Index Terms- Assessment, Letter Grades, Grade Assigning, MS Excel

I. INTRODUCTION

Teaching being the main job function, all teachers are tied with the responsibility of imparting knowledge, skills, and attitudes, and moreover, assessing how much of it is retained by the student. Even though testing and assessment has its own drawbacks and that alone is incapable of assuring the final performance ability of a graduate, it has inevitably become an integral part of teaching.

Nevertheless, assessment also has its use in ranking students within their population indicating their relative performance abilities, making it possible for the teacher as well as the society to identify those who perform better than others. Therefore, every teacher whether university or otherwise has not only the responsibility of teaching as well as identifying and highlighting the relative talents of the student population.

In order for the teacher to do this it is necessary that marks be compiled, and grades be assigned so that a numeric indicator for the rank each student has earned could be computed. In modern days, the educational institutions are adopting uniform grading and ranking procedures in order to make the final figures compatible and comparable across other academic institutes. This process comprises two main steps: (1) a weighted average is computed using the numerical scores obtained for different course components over the course of a grading period; and (2) assign a letter grade based on the final term average [1]. This inevitably requires the teachers to adopt certain mathematical and statistical operations on a routine basis, some of which may not be very familiar to the individual teachers depending on their area of expertise.

This effort was to investigate the possibility of using common application software to help the teachers to produce grades to students using what is known as the distribution based grading system, a procedure not as common and straightforward as the direct assignment of grades. The ability of using the MS Excel package to perform the statistical operations required was examined, and a ready to use flexible and automatic letter grade assigning software system was developed to facilitate any teacher to use his / her final marks lists to generate grades and assign them based on the marks distribution within the group / class. The inherent capability of storing, analyzing and presenting data within the MS Excel spreadsheet is advantageous since it prevents the redundant data entry. This paper describes the features of the software developed, the statistical background, and the excel routines incorporated. The system developed was tested by distributing it among the staff of the Faculty of Agriculture in the 2ndsemester 2012. Their feedback on the operation was also considered in producing the final refined version described herein. A comparative analysis was performed

through a survey to assess the effectiveness of the new software compared to individually performing the prescribed computation and analyses to generate the student's grades.

II. AN OVERVIEW OF DIFFERENT GRADING SYSTEMS

Grading systems used in conducting examinations in academic work are mainly two fold while there may be similarities and different types of blends of the same that can be adopted by some institutions and personnel. These are namely; Criterion-referenced systems and Norm-referenced system.

1. *Criterion-referenced systems*:

These tests are designed to measure the level of performance of individual students against a set of pre-determined performance levels. This provides a direct measure of their achievements in mastering the knowledge and skills during the course, and those who have achieved a predetermined minimum can be allowed to proceed or considered to have 'passed' the course. Assigning intermediate grades is also done in the same manner by comparing the individual level of achievement against pre-set standards for each letter grade (e.g., 95-100 = A, 86-94 = B). However, a peculiar characteristic of this grading system is that since the standard in this grading system is absolute, it is possible that all students could get A grades or all students could get D grades [2].

One of the disadvantages of criterion-referenced grading systems is identified as its deficiency in handling the variability among different batches of students, and the variability among different tests and teachers. If the course and the tests are handled by very experienced teachers, and the student cohort appears to be of the same standard, criterion referenced systems are considered to perform better, and can be considered to produce a fair and repeatable grading outcome. However, teachers find it difficult to design and conduct tests that would produce final marks that fall in the same range that could be matched with a fixed set of criteria for the purpose of assigning grades as described above.

2. *Norm-referenced systems*:

One of the ways of addressing the above problem is to adopt norm-referenced grading systems in processing final marks. In norm-referenced systems, students are evaluated in relationship to one another (e.g., the top 10% of students receive an A, the next 30% a B). This grading system rests on the assumption that the level of student performance will not vary much from class to class. In this system, the instructor usually determines the percentage of students assigned each grade, although it may be determined (or at least influenced) by departmental policy [2].

Advantages of norm-referenced systems are their convenience especially in differentiating among students and limiting the numbers reaching different standards. They are however, considered to be more appropriate in large courses which do not encourage cooperation among students.

One objection to using norm-referenced systems is that an individual's grade is influenced by the performance of others, and inherently, this type of grading will not yield all As or all Ds.

A second objection to norm-referenced grading is that it promotes competition rather than cooperation. When students are pitted against each other for the few As to be given out, they're less likely to be helpful to each other. Therefore, it is a method that is not recommended for small classes, generally less than 40.

Pogge [3] describes the method adopted for assigning grades, or curving, where he highlights the necessity to investigate intricate details of the distribution and compare with past performance of the same class in order to make it a fair judgment. Godolphin [4] highlights the necessity to adopt criterion referenced systems in order to assess the performance of the individual rather than influencing it with the performance of the others.

As it can be seen norm-referenced grading system has its own inherent disadvantages. However, it is more likely to be a fair way of assigning grades where the student cohort and the exam standard are bound to vary. Instead of adhering to fixed percentiles in allocating student numbers to different grades, a modification adopted by the Faculty of Agriculture in using the norm-referenced system is to examine the distribution of marks, and decide on the cutoff points for each grade by identifying groups that are markedly separated in the marks distribution.

It is clear that independent of the grade assignment system adopted by the faculty or the teacher, it undoubtedly requires intricate examination of the marks distribution and adjustment of grade break points in order to make it fair to the best possible level. This operation undoubtedly requires a plot similar to the 'Dot Plot' or the 'Histogram' be generated, and the teacher examining it in

intricate detail to identify 'groups' within the distribution. In addition, there must be a facility to assign grades to each group identified with flexibility in changing the group boundaries based on the apparent break points in the histogram.

Reliability

Once a students' grade assigning method is selected it should be characterized by key functional attributes: (1) face and content validity, and (2) reliability, and realistic expectations. The first one denotes that a grading procedure must have clear and suitable matrices to measure the degree of relevance of the evaluation process and the method of evaluation to the course objectives, and such process must be transparent to students as well. An evaluation method for assigning students' grades is reliable if it would invariably produce, with little variations, the same results for the same students as denoted by the second point [5]. Kelly [6] also defines a test to be reliable if it allows for stable estimates of student ability. A reliable test is expected to provide similar results for students who have similar ability and knowledge levels.

The need

Ever since the Faculty of Agriculture adopted the above grade assigning convention, the staff members were required to produce the Dot Plot after finalizing their marks, and identify justifiable break points in the plot, and then assign grades to each sub group thus identified starting from the highest grade to the lowest. This however, posed additional problems to the teachers by having to go through fairly complicated analyses, some of which are not directly supported by common data analysis packages such as MS Excel (neither the Dot Plot, nor the Histogram are available in MS Excel through a single icon or as a menu selectable option).

Besides, the grade allocations made by the staff at times were questionable during the results board meetings and the relevant teachers were requested to re consider their grade allocation to make it appear more balanced. This undoubtedly required the academic staff to be able to perform the statistical analyses to generate the Dot Plot (or the Histogram), and be able to change the cutoff points for each grade based on the plot. This may be less strenuous for a small group whereas, it could be a straitened job if the student number is large and a considerable amount of time will be wasted on prohibitive-drudgery of overwhelming numeric processing.

Therefore, the development and adapting the uses of an automated grade assigning system would certainly be a relief for the educators. It is believed that if the mechanistic aspects of the evaluative process are included within an automated and user-friendly procedure, educators' efforts would then be more meaningfully focused on ensuring more reliable and fair evaluation of students' mastery of the designated learning outcomes in a given course [5]. The software utility presented herein was developed to address this need, and was made available to the academic staff. The responses from the academic staff after using the utility software were compiled to identify possible bugs and places for improvement. The final revision and its features together with the latest web based survey results are presented herein.

III. THE SOFTWARE DEVELOPMENT

Based on the background described above and the problems faced by the teachers concerned, it was decided to embark on this development with the following objectives.

Objectives

The objective of this effort was to develop an easy to use software utility for the purpose of assisting the teachers in the process of assigning letter grades based on the marks distribution in the class. The software was to be developed within the MS Excel environment as it is one of the most common data analyses packages used by the staff for handling student's marks and grades. Besides, the MS Excel 2007 contained advanced features which aptly supported the intended task.

Development and testing process

The worksheet was divided into several regions for accepting user input, which consists of only the index numbers of students and final marks. Provision was made to handle different types of index number formats, and also entries in the marks column which may include 0, Ab or Absent, I, and other similar indicators. Automatic search routines were also added to identify whether any invalid number in entered by mistake into the marks column (numbers less than 0 or greater than 100 generated error signals to draw the attention of the user). The software is also capable of identifying missing values and informs the user about them (if any) in the marks list entered as the input.

Prime attention was paid to provide a work platform which is easy to use while ensuring that no errors occur due to lack of experience in using the program. In order to assure that the worksheet performs the intended analyses and produces the expected results, all areas that did not require direct user manipulation were write-protected. This was an essential step in ensuring that the results are error free.

In addition, the user was provided with slider bars (scroll bars used in form control) to adjust the break points of grades, in parallel with a visual display of the marks distribution making it very convenient for the person to adjust them (Figure 1). The resulting grade distribution plot was also readily generated and made available for inspection by the user (Figure 2).

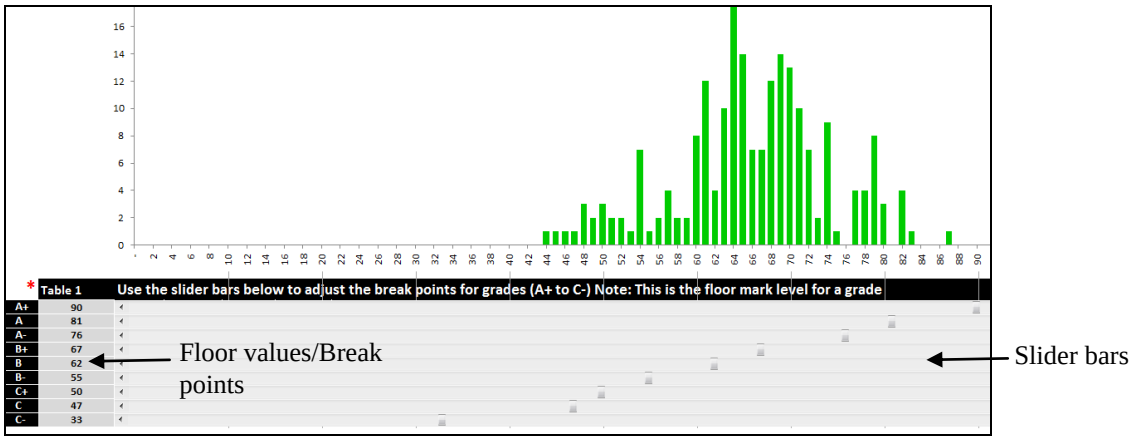


Figure 1: The histogram of final marks distribution

The letter grade floor value demarcation table and slider bars are also visible below the histogram.

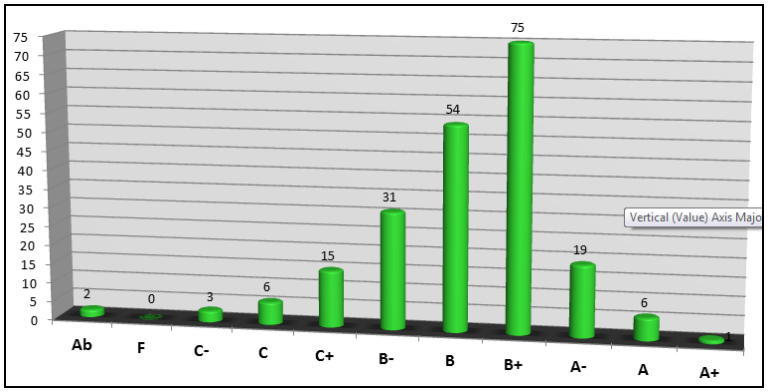
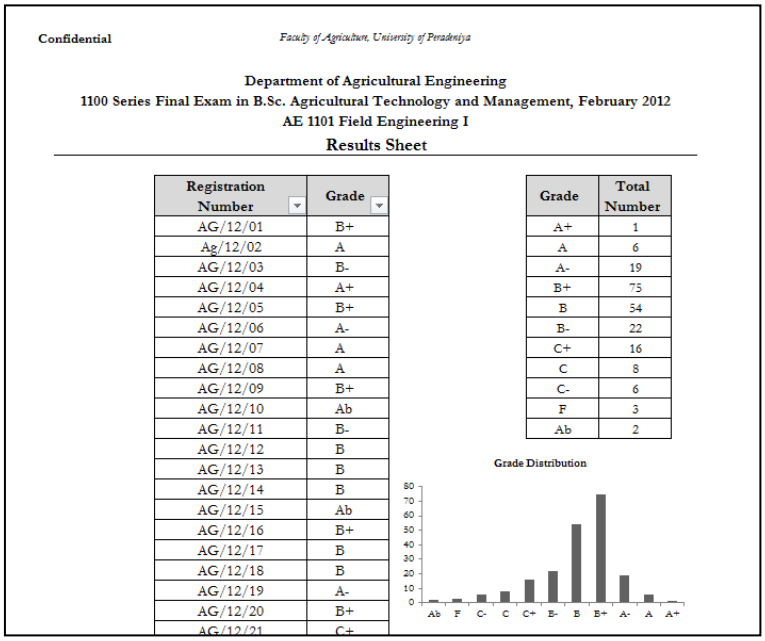


Figure 2: The grade distribution plot

Once the break points for each grade were determined, the user was given the provision to remove any blank lines that may remain in the output area using the Filter option available in MS Excel (Figure 3), and print the grade tables directly.

Error handling was mainly done by preventing the user from entering data into unauthorized tables, preventing accidental additions/deletions in the program area and also by generating error messages upon detection of an entry that lies outside of the standard range of magnitudes. The overall structure of the grading program is illustrated below in the Figure 4.



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Faculty of Agriculture, University of Peradeniya

Department of Agricultural Engineering
1100 Series Final Exam in B.Sc. Agricultural Technology and Management, February 2012
AE 1101 Field Engineering I
Results Sheet

Registration Number	Grade
AG/12/01	B+
Ag/12/02	A
AG/12/03	B-
AG/12/04	A+
AG/12/05	B+
AG/12/06	A-
AG/12/07	A
AG/12/08	A
AG/12/09	B+
AG/12/10	Ab
AG/12/11	B-
AG/12/12	B
AG/12/13	B
AG/12/14	B
AG/12/15	Ab
AG/12/16	B+
AG/12/17	B
AG/12/18	B
AG/12/19	A-
AG/12/20	B+
AG/12/21	C+

Grade	Total Number
A+	1
A	6
A-	19
B+	75
B	54
B-	22
C+	16
C	8
C-	6
F	3
Ab	2

Grade Distribution

Figure 3: A sample report generated using the software

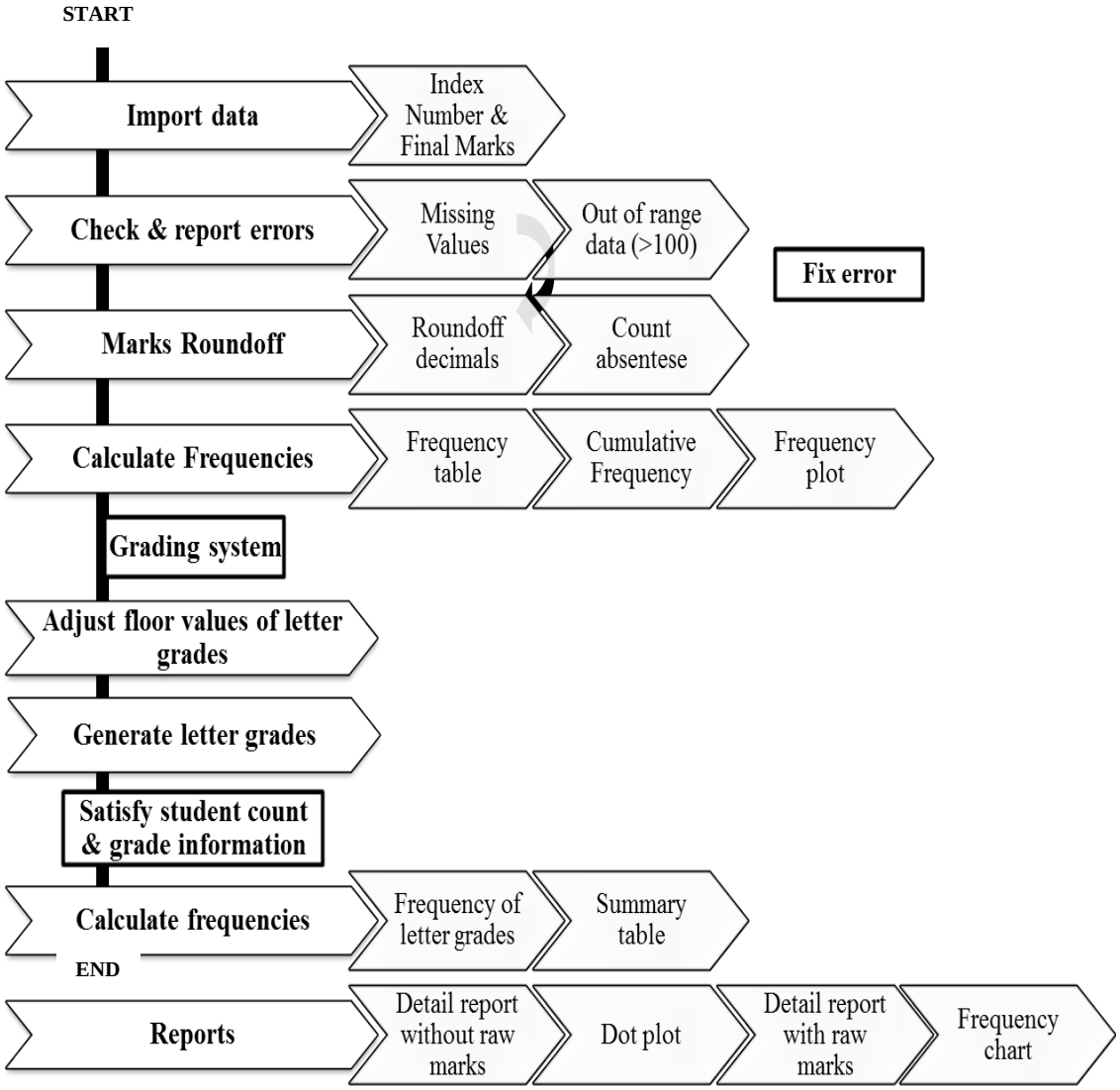

Figure 4: Schematic diagram of the program structure

User survey and feedback:

A fully on-line user survey was conducted after refining the program through several trials (the present version is named: Easy Letter Grade Assigner Utility – Ver. 1.5). The users were presented with 9 questions which required them to provide their feedback on a 5 point Likert's scale, and one descriptive type question to receive their comments. The survey was organized on a free web based survey site:www.surveymonkey.com(*http://www.surveymonkey.com/s/T8WHPYF*).

The survey format:

The users were presented with the rating scale as shown below (Figure 5). The other questions which are given subsequently were presented in the same format with individual rating scales.

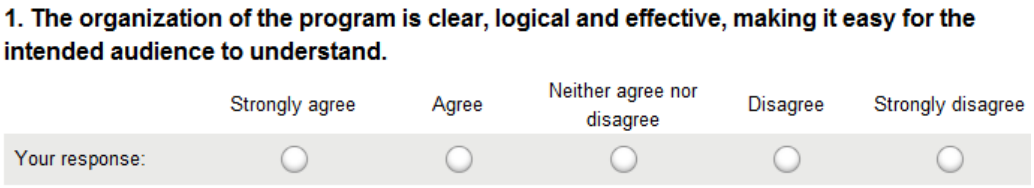

Figure 5: An example of the Likert's scale used in the online survey

2. The language in the program and in the user guide is clear to the intended audience.
3. The program and instructions are designed so that an individual can operate it independently.
4. The user is provided ample opportunity to control the decision making with regard to grade assignment.
5. Invalid commands and entries are handled in a safe and constructive manner.
6. The program is attractive and interesting. It motivates users to continue using the program.
7. The program achieves its purpose and reliable in normal use (no bugs / ambiguities were encountered).
8. Using the program to assign grades is much easier than doing the statistical analyses by myself.

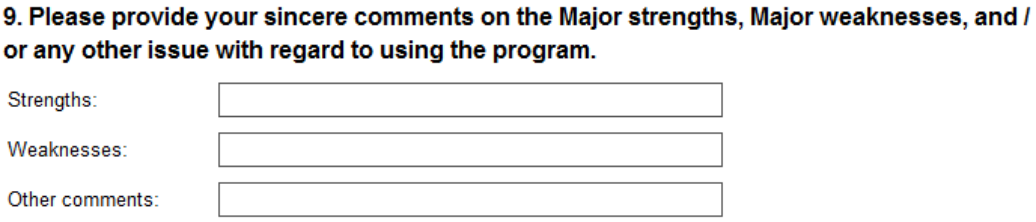

Figure 6: The open ended questions to receive comments

Summary responses:

A summary of the responses received at the web site is shown below in Table 1. It is clear that the participants of the survey, all members of the permanent academic staff of the Faculty of Agriculture, with sufficient experience in marking examinations and assigning grades, highly preferred the use of the new software utility.

Table 1: Summery of responses received

Question No:	Percentage Responses (%)		
	Positive	**Neutral**	**Negative**
1	100	0	0
2	94	6	0
3	78	17	0
4	94	6	0
5	61	39	0
6	83	11	6
7	78	22	0
8	94	6	0

IV. Results and Discussion

Internal Consistency of the Likert's scale responses was observed to be very high (Cronbach's Alpha = 0.947). Hence, the responses received by the questionnaire could be considered as highly reliable. As it is apparent from Figure 7 below, the users were very confident (100%) that the developed program provided a clear and easy to use work environment and was capable of making their task of analyzing marks easy and trouble free (94%).

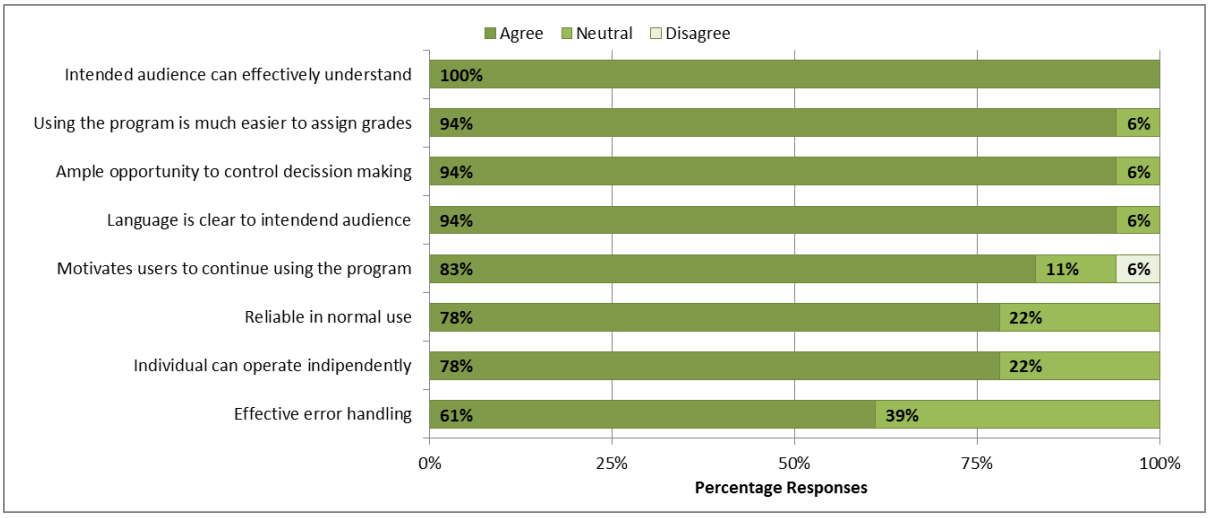

Figure 7: Perception of the academics about the software

Further, the independent work environment and flexibility offered by the program also received high acceptance scores (78%), while the way it provided the opportunity and the freedom for the user to decide on grade assignment was also rated high (94%). The participants rated the motivational aspect and the level of achievement of the main objective very high (83%, 78% respectively), and the convenience of performing the task using the developed program was undoubtedlyhigh (94%).All these responses were significant at 95% confidence interval.

The only negative comment was received for the question 6, which was mainly based on personal assessment of the motivational aspect offered by the program.

This program was used in assigning grades in the examinations held in February 2012 for the undergraduate students of Faculty of Agriculture across 8 academic departments with the direct involvement of a wide majority of the academic staff members. The software was proven successful saving significant amount of time and there were no complaints received by the authors about its consistency. Considering the overall responses, it was clear that the program performed well in offering the users with an efficient, easy to use environment for assigning grades based on the final examination marks. Since the grade assignment mechanism is flexible, this program could be used for distribution based (norm-referenced) grading or fixed scale (criterion-referenced) situations alike.

V. Conclusion

The user assisted interactive letter grade assigning software, designed based on MS Excel 2007 environment, could be effectively used to assign letter grades independent from the grading system adopted while saving time and effort significantly. It did not require any additional software installation or enabling macro operations but simply a computer installed with MS Excel 2007 or later version.

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Morphometric and Hydrological Analysis of North East Punjab Region: With Special Reference to Groundwater Management.

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Abstract- Water is extremely crucial for the human being and economy of the country. Almost every industry from agriculture, hydropower and industrial manufacturing to production of readymade food and tourism relies on water to grow. Continuous population growth and economic development has significantly increased the agricultural and industrial water demand. Watershed development and management is one of solutions to such problems. To prepare a comprehensive watershed development plan, it becomes necessary to understand the topography, erosional status and drainage pattern of the area. A comprehensive study of watershed for south west Punjab region was carried out. This watershed region has the total area of approximately 4161.132 km^2. The morphometric parameters are computed by using Geographic Information system (GIS). GIS was used in evaluation of linear and aerial aspects of morphometric parameters. The drainage patterns of the basin are dendritic and include a third order stream. In this study an attempt has been made to develop a recharge plan for the entire study area by locating the suitable site of recharge and recharge structures.

I. Introduction

In most of the regions of the world water is already over-appropriated. Approximately one-third of the world's population roughly 2.4 billion people, live in water-stressed countries and by 2025 the number is expected to rise to two-thirds (UN-FAO, 2007). In India Punjab region is one of the most agricultural productive region. It is located in the North West part of India and has an area of about 50,362 km^2 and having an annual rainfall of 780 mm. Entire Punjab state is underlain by thick quaternary alluvium. Due to continuous escalating population and water demand most of the surface and groundwater resources are depleting very fast. Groundwater tables and river levels are receding in many parts of the Punjab due to human water use. One of the major reasons for increasing the water scarcity in the region is changes in precipitation patterns and intensity. Drastically reduced precipitation in some regions has causes exponentially larger depletion in groundwater tables. Temperature increase changes like precipitation patterns, frequent severe weather events, and prolonged droughts are affecting the agriculture production.

In near future it will be difficult to satisfy the growing water and food demand because climate change is escalating problems like soil salinity, soil erosion, and desertification. Moreover continuous depletion in water table is recorded along with deterioration in water quality.

The decrease in the crop yield will further increase the already increased food inflation in India. This condition can only be improved by preventing the depletion of water resources and increasing the crop yield. In this study morphometric analysis of a minor watershed of North east Punjab region was carried out. This paper locates the suitable location of artificial groundwater recharge in the study areas to satisfy the demand of water for domestic, irrigational and industrial sector. The paper also suggests a road map for long-term and near-term mitigation and adaptation strategies for minimizing the impact of climate change on water resources.

II. Objectives

Some of the major objects of the study area are as follows:

- To derive the different drainage aspects of North east Punjab region and to understand the relationship of the drainage networking.
- To study the quantitative analysis of drainage system.
- To develop a recharge plan for entire watershed basin and describe different morphometric analysis.
- To develop a road map for long term and near-term mitigation and adaptation plan for water resource management of the region.

III. Watershed Area

Geologically most of the area of the Punjab state is represented by the Quaternary sediments differentiated into older alluvium and younger alluvium. The soils in the watershed area are micaceous. The average annual rainfall in Punjab is around 780 mm. Punjab is also called state of lakes and rivers. There are five rivers namely Beas, Sutlej, Ravi, Chenab and Jhelum in Punjab that drains the fertile lands. Despite plenty of water resources in Punjab the state fails to satisfy the growing demand of water. Figure 1 shows the location of selected water in the Indian continent. The watershed taken for the study covers an area of 4161.132 km^2.

IV. Methodology

For the morphometric analysis a minor-watershed basin area was taken. Watershed map was modified using GIS software

Rolta Geomatica 10.3.2 and stream orders were calculated by following standard drainage network analysis method described by laws of Horton (1945) and Strahler (1964). The quantitative analysis of the basin which include stream orders, stream numbers, stream lengths, bifurcation ratio, basin circularity, drainage density, drainage frequency, drainage texture, etc., have been analyzed through use of a Geomatica software that determines the geomorphic stage of development of the area. Figure 1 shows watershed map of the North east Punjab region and suitable location of different recharge structures.

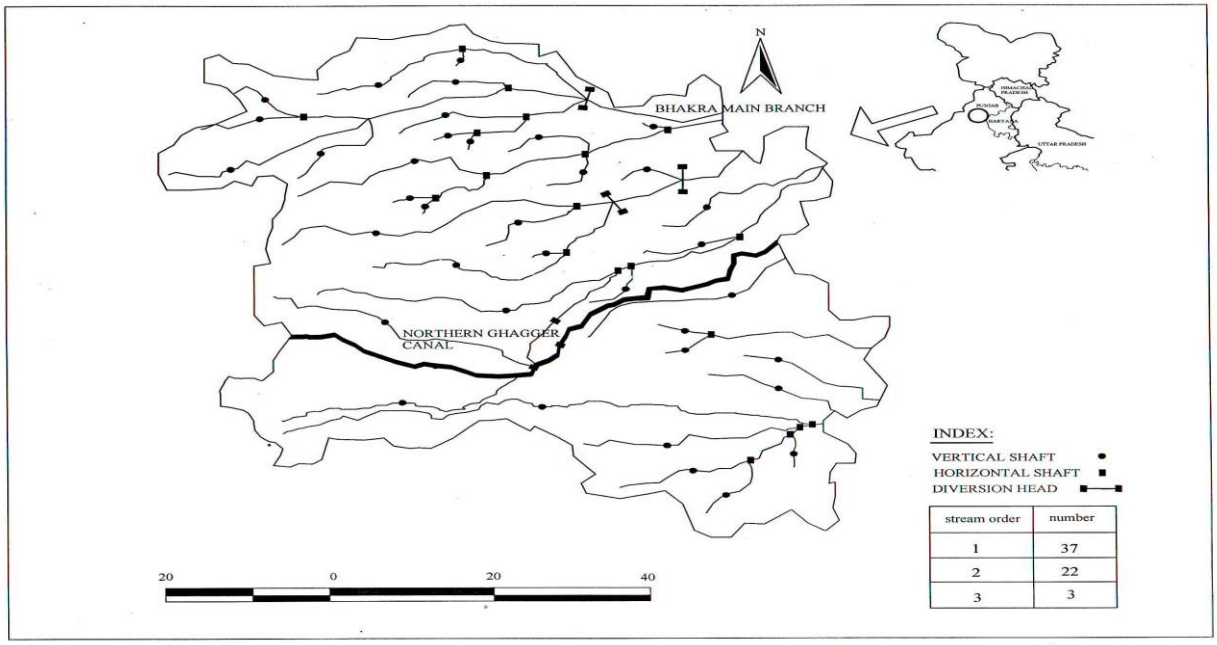

Figure 1. Drainage map of north east Punjab region, showing natural and man-made drainage (Source: Base maps Google Image).

V. SOME IMPORTANT ASPECTS OF MORPHOMETRIC ANALYSIS

Stream Order:

The streams of the basin have been ranked according to the method described by Strahler, when two first order streams join, a stream segment of second order is formed; when two second order streams join, a segment of third order is formed, and so on. The study area is a third order drainage basin.

Bifurcation ratio:

Bifurcation ratio is the ratio of the number of streams of an order to the number streams of the next higher order (Horton, 1945), Bifurcation ratios are related to the structural control on the drainage (Nautiyal, 1994; Strahler, 1964; Chow, 1964). The bifurcation ratio for 2nd order, and 3rd order, streams measured are 1.68 and 7.33 respectively.

Stream Length (Lu):

It is one of the most significant hydrological features of the basin as it gives surface runoff characteristics of streams. Longer lengths of streams are generally indicative of flatter gradients. Generally, the total length of stream segments is maximum in first order streams and decreases as the stream order increases. Total stream lengths of different streams are mentioned in table 1.

Basin area:

The basin area (A) is defined as the total area projected upon a horizontal plane contributing to cumulate of all order of basins. Basin area is extremely important from the hydrological point of view as it directly affects the size of the storm hydrograph and the magnitudes of peak and mean runoff. It is interesting that the maximum flood discharge per unit area is inversely related to size. The total area of the watershed was recorded 4161 km^2.

Drainage Density (D):

Drainage density (D) is an important indicator of the linear scale of land form elements in stream eroded topography (Horton 1932). It is the ratio of total channel segment lengths cumulated for all orders within a basin to the basin area. It indicates the closeness of spacing of channels, thus providing a quantitative measure of the average length of stream channel for the whole basin. The total drainage density (D) for watershed area was recorded around 0.122

Texture Ratio (T):

It is one of the important factors in the drainage morphometric analysis that depends on the underlying lithological conditions, porosity and permeability, infiltration capacity and topography of the terrain. In the present study the texture ratio of the watershed area was recorded 0.1436. Table 2

shows the outcome of different morphometric parameters recorded for the North-east Punjab region watershed.

Table 1. Outcome of measurement of total stream length and bifurcation ratio

STREAM ORDER	NO. OF STREAMS	NO. OF STREAM %	TOTAL LENGTH (Km)	MEAN LENGTH (Km)	BIFURCATION RATIO
1	37	59.68	318.17	8.60	_
2	22	35.48	158.93	7.22	1.68
3	3	4.84	30.56	10.19	7.33
TOTAL	62		507.66		

Table 2. Different morphometric parameters of North-east Punjab region watershed.

S.No.	Morphometric Parameters	Symbol / Formula	Values
1.	Area (sq. km)	A	4161.132
2.	Perimeter (km)	P	257.634
3.	Drainage density (km/sq. km)	D = Lu ÷ A	0.122
4.	Stream frequency	Fs = Nu ÷ A	0.015
5.	Texture ratio	T= N1 ÷ P	0.1436
6.	Basin length	Lb	65.88
7.	Elongation ratio	Re=(A/π)^0.5/Lb	0.552
8.	Circulatory ratio	Rc=4πA/(P)^2	0.78
9.	Form factor ratio	Rf = A÷(Lb)^2	0.95

VI. Availability of Water Resources in the Region

Punjab state is located just above the Haryana state and is mainly underlain by quaternary alluvium of considerable thickness. States average annual rainfall is approximately 780 mm. Punjab state is also called state of lakes and rivers. There are five rivers namely Beas, Sutlej, Ravi, Chenab and Jhelum in Punjab that drains the fertile lands. Despite the presence of several rivers and other water resources Punjab fails to satisfy the growing demand of water for domestic, irrigational and industrial purposes. To fulfill the escalating demand and preventing the continuous degradation of ground water resource in Punjab, ground water recharge and management is the only solution. Out of the total 138 of the state, 93 blocks have turned Dark (critical blocks) where the under ground water development has crossed 100%.

Table 3. Available water resources of Punjab district (Source: Jain, A.K. and Raj Kumar, 2010)

Annual canal water available at H/w	1.79 M ha-m
Annual canal water available at outlets	1.45 M ha-m
Annual ground water available	2.03 M ha-m
Total annual available water resources	3.48 M ha-m
Annual water demand	4.76 M ha-m
Annual water deficit	**1.28 M ha-m**

Table 4. Escalation of over-exploited, critical and semi-critical blocks in study area ((Source: Jain, A.K. and Raj Kumar, 2010).

CATEGORY	2000	2005	2010
Over-exploited (Dark)	73	103	110
Critical	11	5	3
Semi Critical	16	4	2
Safe	38	25	23

Punjab is one of the most intensively cultivated and an irrigated area of India with approximately 84% of the total geographical area under agricultural use. About 35% area is suffering with some kinds of soil degradation problems along with decreasing water table problems. In some areas especially in the central Punjab water table has gone down below critical depth of 10m with 3% in 1973, 25% in 1990 and 46% in 1994. It's now become imperative to take urgent steps to conserve the available ground water and increasing the recharge of ground water resources.

VII. Conclusion and Recommendation

Declining water quality is an acute problem with notable decrease in agricultural and industrial production, coupled with a lack of adequate wastewater treatment. Many rivers in the Punjab

are so badly polluted that not even industry can use the water. Rising water demand and the lack of water management schemes have badly affected the region. The watershed area shows that the entire catchment area of the watershed is the part of fluvial system and is dominated by 1st and 2nd order stream. The development of the stream segments in the watershed area is more or less affected by rainfall. The general topographic gradient of the entire area is towards north east.

The findings of the study demonstrated following points:

- In entire study area natural water storage capacity has decreased and subsequently long-term water availability has also reduced
- Entire watershed area is facing frequent droughts and flood situation. Water scarcity problems has increased manifold, due to changes in precipitation patterns and intensity. Both droughts and episodes of regional-scale flooding can both be linked to the large-scale atmospheric circulation patterns in the region and often occur simultaneously in different parts of the study area compounding their impact on human activities.
- Depletion of water resources has affected the agriculture production that has gone down and food inflation has escalated manifold.
- A comprehensive recharge plan is needed for entire Punjab region

The following recommendations are proposed to tackle the existing problems:

- More research is needed to improve existing capabilities of weather forecast related with short- and long-term drought conditions and to make this information more useful and timely for decision making.
- With time, significant land-cover changes have occurred in response to persistent droughts, and the role of land-cover changes in amplifying or damping drought conditions should be thoroughly analyzed.
- To increase the awareness about the ecological impacts and affect of over drafting of water withdrawal and discharge should be thoroughly studied.
- More studies are required for agricultural sector and adaptation of improved irrigation techniques to cope with drought, and new plant varieties which are resistant to drought or to salt water.
- Increase practice of recycling and reuse of wastewater especially for irrigational purposes. It should be implemented in the entire study area.
- Artificial recharge structures should be constructed. Awareness should also be raised among the people by taking the help of media and NGO's regarding climate and water conservation education and development
- There has been drastic change in the land use and land cover of study area. To minimize further changes in land-use pattern strict rules and regulations should be formed and adopted.

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"Training Status of Teachers Belonging to Higher Secondary Level" – A Case Study

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Abstract- The success of higher secondary education depends largely upon how well the schools are staffed with qualified and trained teachers. The present study is an attempt to study the real picture as regard to the training status of teachers belonging to higher secondary level under the Darrang District, Assam. Twenty five different higher secondary schools/Colleges (Govt. as well as private) of Darrang District, Assam have been selected for the present study. Four hundred samples of teachers have been selected from twenty five selected institutions. One questionnaire comprising of thirty seven questions are distributed among the selected teachers and data are collected. The result of the study revealed that the training status of teachers at the higher secondary level is not satisfactory in the Darrang District, Assam.

Index Terms: Attributes of teachers, Chi-square-test, Higher secondary, Training

I. INTRODUCTION

Education plays an important role in bringing about social change. In order to have the desired social change, a suitable system of education is required, but the success of any educational system depends on the teachers. Change in educational system demands for qualified and trained teachers. In order to improve the qualities of teacher turnout, the existing conditions of trained teachers should be brought to light. The role of teacher is most important in educational revival and he can perform his role properly only when he is given proper and efficient training. In the present study, trained teachers indicate the teachers having B.Ed. Degree.

The main objective of the study is to know the training status of teachers belonging to higher secondary level.

Present study will be based on the following hypotheses.

H_{01} : Attributes of the teachers and institutions are independent.

H_{02} : Attributes of the teachers and their sex are independent.

H_{03} : Attributes of the teachers and location of the institutions are independent.

For the present study, 25 institutions (13 higher secondary schools and 12 colleges) situated at Darrang District, Assam have been taken into account.

II. RESEARCH ELABORATIONS

The study was conducted on 400 teachers selected from 25 institutions (13 higher secondary schools and 12 colleges) situated at Darrang District, Assam. Purposive sampling technique was used to collect the sample (data). The tool used in present investigation was self-developed questionnaire comprising of 37 questions and were distributed among the teachers of the selected institutions.
The statistical measures used in the present study are chi-square tests (χ^2)

Chi-square (χ^2) test :

The χ^2 test statistic describes the magnitude of the difference between the observed and the expected value. It is defined by

$$\chi^2 = \sum_{i=1}^{n} \left\{ \frac{(O_i - E_i)^2}{E_i} \right\}$$

Where $O_i = i^{th}$ class observed value

$E_i = i^{th}$ Class expected value, n = number of classes

The χ^2 value ranges from 0 to infinity. If χ^2 value is zero, then the observed and the expected values completely coincide. If the χ^2 value is greater, then the observed and the expected values have greater discrepancy. The calculated value of χ^2 is compared with table value of χ^2 for given degree of freedom at a certain level of significance.

If the calculated χ^2 value is greater than the table value at 5% probability level of significance, then there is a significant difference between the observed values and the expected values. On the other hand, if the calculated χ^2 value is less than the table value at 5% probability level of significance, then there is no significant difference between the observed values and the expected values.

Results: The data analyzed described in the above have been presented in the following table.

Table No. – 1

Table showing No. of Teachers belonging to Higher Secondary Level According to Institutions, Sex, Location, Attributes, Degrees of Freedom, Calculated value of χ^2, Tabulated value of χ^2 (at 5% level of significance)

Number of teachers belonging to Higher Secondary Level									
Attributes of Teachers	Institution			Sex			Location of the Institution		
	Higher Secondary Schools	Colleges	Total	Male	Female	Total	Rural	Urban	Total
Trained	70	47	117	72	45	117	85	32	117
Untrained	80	203	283	203	80	283	170	113	283
Total	150	250	400	275	125	400	255	145	400
Degrees of Freedom (d.f.)	(2-1) x (2-1) = 1			(2-1) x (2-1) = 1			(2-1) x (2-1) = 1		
Calculate value of chi-square ((χ^2)	35.194			4			5.67		
Tabulated value of chi-square (at 5% level)	3.841			3.841			3.841		

Chi-square test to test the independence of attributes of the teachers and institutions :

Here the calculated χ^2 value is 35.194 and tabulated χ^2 at 5% probability level of significance is 3.841. Now since the calculated χ^2 value is greater than the corresponding tabulated χ^2 value at 5% probability level of significance, therefore we reject our null hypothesis H_{01} and may conclude that the attributes of the teachers are dependent on the institutions.

Chi-square test to test the independence of attributes of the teachers and their sex :

Here the calculated χ^2 value is 4 and tabulated χ^2 at 5% probability level of significance is 3.841. Now since the calculated χ^2 value is greater than the corresponding tabulated χ^2 value at 5% probability level of significance, therefore we reject our null hypothesis H_{02} and may conclude that the attributes of the teachers are dependent on their sex.

Chi-square test to test the independence of attributes of the teachers and the location of the institutions:

Here the calculated χ^2 value is 5.67 and tabulated χ^2 at 5% probability level of significance is 3.841. Now since the calculated χ^2 value is greater than the corresponding tabulated χ^2 value at 5% probability level of significance, therefore we reject our null hypothesis H_{03} and may conclude that the attributes of the teachers are dependent on the location of the institutions.

III. CONCLUSIONS

The present study reflects the training status of teachers belonging to higher secondary level. From the present study it is found that the training of teachers of higher secondary level is not satisfactory. All the aspects of the educational system at this stage have been changing as per recommendations of Commissions, Conferences and Committees. Teachers training in the changing system facilitates for better and effective management of education with better learning by students. The teachers being the key personnel to innovate new system of education with the changes in teaching methods, evaluation of students and introduction of new subjects or devices, training of teachers is an essential part of a stage of education.

Following are some suggestions from my point of view:

- There should be Government directions that all teachers should go through training.
- Provision of summer institutes, seminars, workshops should be made available for the teachers.
- The teachers need to be suitably trained and oriented to overcome their crisis of confidence, transaction deficiencies and lack of crucial competencies.
- Short-term courses should be organized for senior and experienced teachers who on account of some reasons have not been able to take admission in training colleges.
- The Government should depute the teachers democratically in order to experience, so that teachers can get chance for training.
- Privately managed schools/colleges should give the teachers leave to go for higher training. During training period the trainee should get some short of financial resistance.

- The training of the teachers should be residential type. There should be provision for pre-service training along with in-service training.

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Doubly Helical Antenna by Spiro Mode Maintaining Radiation

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Abstract- We are presenting by this paper on helical antennas which find important applications in communication systems, primarily because of their circular polarization and wide bandwidth. The helical antenna, first introduced by Kraus in 1946, The geometry of larger helical shape type antenna, which is a doubly helical structure, is fully described by five independent parameters, with the including two radial dimensions, two pitch angles, and the number of turns. Radiation properties of this antenna are examined both theoretically and experimentally. The obtain simulation result by the Numerical Electromagnetic Code is used. A large number of cases with different radii, pitch angles, and number of turns are investigated. Results for far-field patterns, gain, axial ratio, and bandwidth are presented. The influence of parameters on radiation properties are examined. Several prototypes of the antenna were constructed and tested using an outdoor antenna range. Measured far-field patterns are presented over a wide range of frequencies. The measured and computed radiation patterns are in good agreement. The results of this study indicate that the proposed antenna provides circular polarization and high gain over a wide frequency range. The one think is most important of this antenna which is size compared with a conventional helical antenna made of straight wire shaped into a helix.

Index Terms- MATLAB, HPBW, VSWR, SLH

INTRODUCTION

I. INTRODUCTION

Radiation Characteristics of a Cylindrical Helix

A conventional helix is formed by winding a wire on a cylindrical surface while maintaining a constant spacing between the turns. The geometry of a helix, shown in Figure , can be described using the following parameters:

D = diameter of helix
C = circumference of helix = pD
S = spacing between turns (center to center)
a = pitch angle = tan-1 (S/pD)
L0 = length of 1 turn rn = $C2 + S\ 2$
n = number of turns
L = axial length = nS
a = radius of helix wire conductor

Electromagnetic properties of a helix may be studied in two different modes of operation, known as the transmission mode and the radiation mode. The transmission mode corresponds to the propagation of an electromagnetic wave through an infinite helix. This can be viewed as a helix being a transmission line or waveguide [1]. The radiation mode, on the other hand, corresponds to the function of a helix as an antenna. Many different radiation modes are possible, but the focus is mainly on two specific[2] kinds; namely normal mode and axial mode. In a normal-mode helix, radiation is maximum perpendicular to its axis, while the axial-mode helix provides a main beam along the axis. The axial mode can exist over a wide range of frequencies and is also circularly polarized. Because of its broad bandwidth and circular polarization, the axial-mode helix is a widely used antenna. Below Figure describes these two modes graphically.

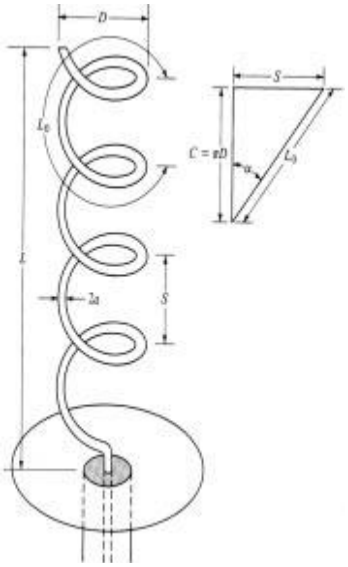


Figure :- Geometry of a conventional helix

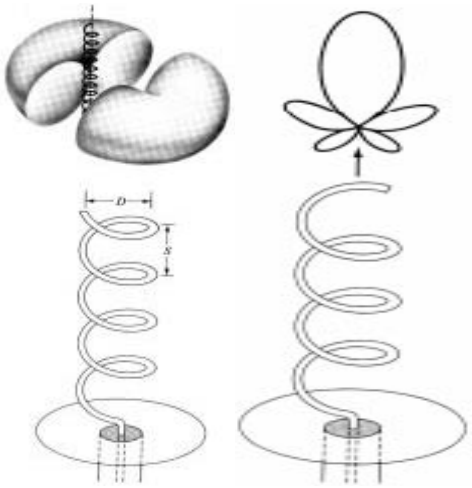


Figure :- Radiation pattern for (a) axial-mode helix ,(b) normal-mode helix

Axial-Mode Cylindrical Helix

The axial mode of operation occurs when the circumference of helix is on the order of one wavelength. Analyzing the current distribution on each turn of an axialmode [3] helix shows that the phase changes by 180° over one-half turn. Furthermore, the current direction is reversed over half a turn by the geometry; that is, another 180° phase shift. Thus, the current at opposite points on each turn will be in phase. hence the characteristic directive beam along the axis for the axial-mode helix. A helix operating in this mode is a traveling-wave antenna and is regarded [4] as a wideband antenna. Antennas whose radiation properties do not change significantly over a bandwidth ratio of upper to lower frequency about 2 are classified as wideband. It has been shown that the axial-mode is sustained over a range given by

$$\frac{3}{4}\lambda < C < \frac{4}{3}\lambda .$$

With these limits, a bandwidth ratio of

$$\frac{f_u}{f_l} = \frac{16}{9} = 1.78$$

is obtained. A unique property of the axial-mode helix is circular polarization of radiation along its axis. This is because the phase of the current changes by 90° over a quarter turn, which in turn leads to the orthogonal far-field components also differing in phase by 90°. The sense of circular polarization is dependent on the sense of the winding of the helix. When wound in the right-hand sense, a right-hand polarization exists, while winding in the left-hand sense results in a left-hand polarization [5]. Below Figure illustrates the difference between these two windings.

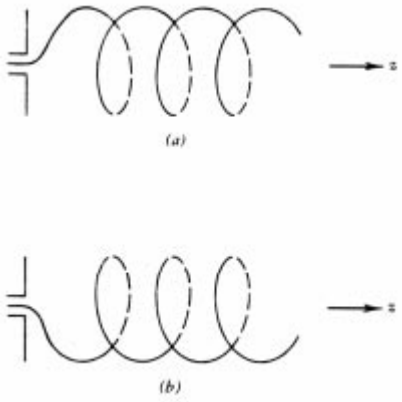


Figure (a) Right hand sense, (b) left hand sense

For the half-power beamwidth, an earlier empirical expression by Kraus and, a few decades later, a more accurate formula by King and Wong were developed. The results are

$$HPBW = \frac{52}{C_\lambda \sqrt{nS_\lambda}} \text{ degrees}$$

by Kraus and

$$HPBW = \frac{K_B \left(\frac{2n}{n+5}\right)^{0.6}}{\left(\frac{\pi D}{\lambda}\right)^{\frac{\sqrt{n}}{4}} \left(\frac{nS}{\lambda}\right)^{0.7}} \left(\frac{\tan\alpha}{\tan 12.5^\circ}\right)^{\frac{\sqrt{n}}{4}} \text{ degrees}$$

by Wong and King [6]. The input impedance of an axial-mode helix is largely resistive and insensitive to
frequency changes over the bandwidth of the antenna. A nearly pure resistive input impedance is attributed to the traveling-wave nature of the axial-mode helix [7]. Kraus's empirical formula for input impedance is

$$R = 140 \frac{C}{\lambda} \quad \Omega .$$

Examining the axial ratio, it has been found that the finite helix does not provide perfect circular polarization. The deviation from circular polarization is described in terms of an axial ratio defined as the ratio of the major axis over the minor axis of the polarization ellipse. The axial ratio for perfect circular polarization is 1, while for linear polarization it is infinity. The equation for the axial ratio developed by Kraus

II. MODIFICATION TO CYLENDRICAL HELIX

Many modifications have been made to the original helical antenna proposed by Kraus [8] in order to improve its radiation characteristics. These modifications have varied from geometric manipulations to changes in ground plane and feed.

Reflectors are referred to as the metal sections at the base of a single helix. A helix is backed by a reflector in order to improve gain and reduce back radiation [9]. Reflectors of various shapes, including planar, cylindrical, and conical, have been used. A helix with a conical reflector is referred to as the "helicone". The end of the helix can also be tapered in order to improve Voltage Standing Wave Ratio (VSWR), bandwidth, and axial ratio. The taper angle can be made the same as the pitch angle. Studies have shown that tapering leads to a more gradual helix-tofree- space transition and a reduction of reflected energy [10]. In addition to tapered ends, tapered feeds are also helpful in lowering axial ratio as well as VSWR. By using two different pitch angles within a single winding, a double-pitch helix is capable of improving gain, axial ratio and bandwidth. Two helices with different pitch angles are combined, end to end , to form a double-pitch helix. Helices designed with specific gain-frequency characteristics in mind utilize this design [11].

III. CONICAL HELICAL ANTENNA

There are certain similarities between the conventional helix and the conical helical antenna. It has been shown [8] that the conical helical antenna is able to provide unidirectional radiation (single lobe) toward the apex of the cone with the maximum occurring along the axis. A constant impedance and circular polarization are maintained over a large bandwidth, much like the conventional helix. A unique feature of the conical helical antenna is the control of bandwidth. If the conical helix is mounted on a ground plane, a reduction in bandwidth will occur.[19] By varying the cone angle q0 and the wrap angle d, the beamwidth can also be controlled.

IV. STUB-LOADED HELIX

The stub-loaded helix is a new design which was also recently developed at Virginia Tech [7]. The winding used for this antenna includes periodically placed loading stubs on the circumference of each turn. Each loading stub is directed along a radial path from the center axis to the helical winding. The stub-loaded helix (SLH) design produces a maximum gain when the pitch angle is about 8°. When compared to a conventional axial-mode helix with a pitch angle of 11° to 15°, the shallower pitch angle of the SLH results in a shortening of the antenna length by approximately one-half for the same number of turns. However, unlike the conventional helix, the SLH exhibits a complex input impedance which varies with frequency [12].Using a common polarization bandwidth of a 3 dB axial ratio or lower, the SLH exhibits an operating bandwidth which exceeds 20% of the center frequency [13].

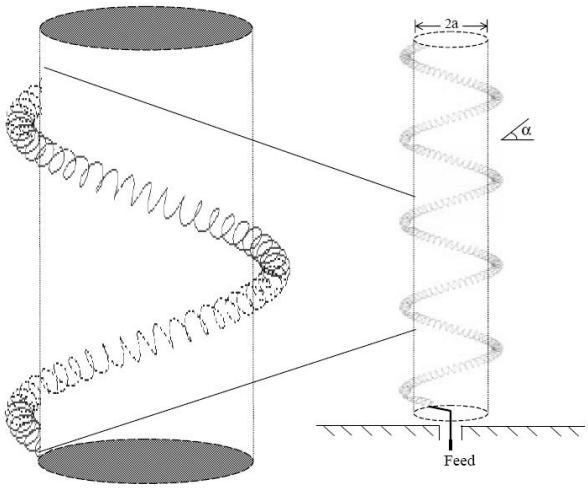


Figure: Spiro-helical antenna over a ground plane fed by a coaxial cable

might result in size reduction originated from the fact that helical structures, such as those used in microwave tube amplifiers, exhibit slow wave properties. However, a slower propagation velocity corresponds to a smaller wavelength as implied from the relationship l=u/*f*, where *f* is the frequency. Thus, intuitively, a helical antenna made of a spiral instead of a straight wire would allow smaller physical dimensions. An important purpose of the research work presented in this thesis was to verify this conjecture by examining radiation properties of the proposed doubly helical structure and comparing them with those of a conventional helix. Following the initial validation of expected properties of the spiro-helical antenna, a comprehensive investigation of the radiation characteristics was performed using both experimental and theoretical [14].

V. PARAMETRIC EQUATION OF THE SPIRO-HELICAL ANTENNA

In order to facilitate the numerical analysis of the spiro-helical antenna, a set of equations describing its geometry are needed. With the availability of these equations, the coordinates of an arbitrary point on the spiro-helical structure are readily determined in terms of the parameters *a*, *a*¢, a, a¢, and an axial dimension z*A*. Before embarking upon [14]. the derivation of equations for the spiro-helical geometry, we first examine the parametric equations for a simple cylindrical helix, such as the primary helix with radius *a*¢ and pitch angle a¢ shown in Below Figure . Furthermore, we use two sets of coordinates namely the primed Cartesian coordinates (x', y', z') , and cylindrical coordinates (r ¢ = *a*¢,j ¢, *z*¢)for the geometry of the primary helix, and the unprimed coordinates *(x, y, z*) and (r, j, *z*) for describing the geometry of the doubly helical structure.
The parametric equations of the primary helix are expressed as

$$x' = a'\cos\varphi'$$
$$y' = a'\sin\varphi'$$
$$z' = (a'\tan\alpha')\cdot\varphi'.$$

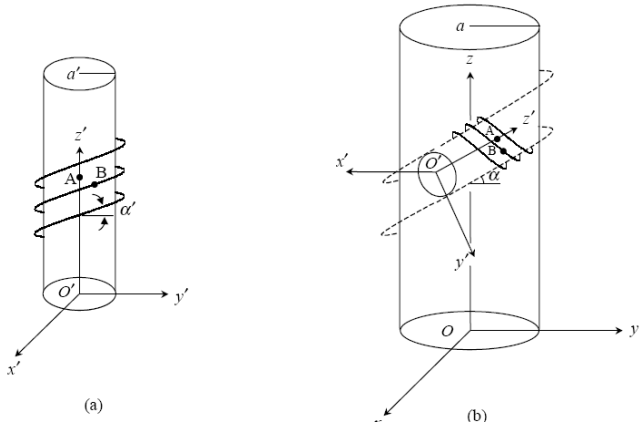


Figure: Geometry and coordinates for the derivation of parametric equations of the spiro-helical antenna. (a) primary helix, (b) spiro-helical geometry

VI. ANALYSIS OF THE HELICAL ANTENNA

Below Figure shows variations of gain versus frequency for the representative spirohelical antenna. A peak gain of 13.09 dB at 2.25 GHz and a minimum gain of 6.83 dB at 2.4 GHz is seen for q = 0°. Between 1.55 GHz and 2.3 GHz in Figure 4.2(a), the gain lies between the minimum and maximum values. It is clear that

the bandwidth of this antenna for q £ 20° is about 0.75 GHz for the gain. However, as will be seen later, the bandwidth for the axial-ratio is smaller and thus determines the overall bandwidth of the antenna. Comparison of these gain results indicates that gain increases with the number of turns as is the case with the conventional helix.

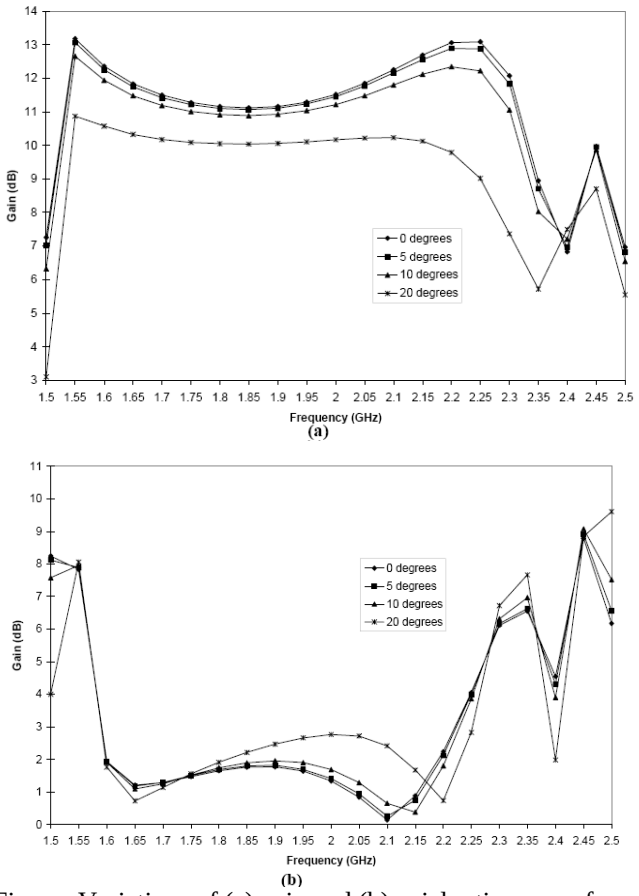


Figure**:** Variations of (a) gain and (b) axial ratio versus frequency for several values of q for the representative 10-turn spiro-helical antenna with a = 10°, a¢ = 30°, *a* = 16mm, *a*¢ = 2mm, and *r*o =0.2mm.

VII. RADIATION PATTERNS

The far-field patterns for the representative spiro-helical antenna are displayed in Below Figures These patterns show variations of normalized power density versus elevation angle q in the *y* = 0 plane. Examination of the patterns indicate that the half power [15]. beamwidth varies between 40° and 60° for the majority of frequencies between 1.5 GHz and 2.5 GHz. As seen in these figures, there is an increasing trend in the HPBW beginning at 1.5 GHz.

Starting at 44°, HPBW increases until 1.85 GHz where

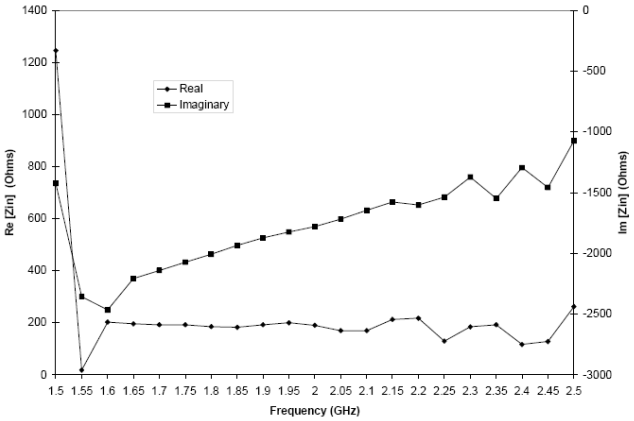


Figure: Input impedance versus frequency for the representative 10-turn spiro-helical antenna with a = 10°, a¢ = 30°, *a* = 16mm, *a*¢ = 2mm and *r*o = 0.2mm.

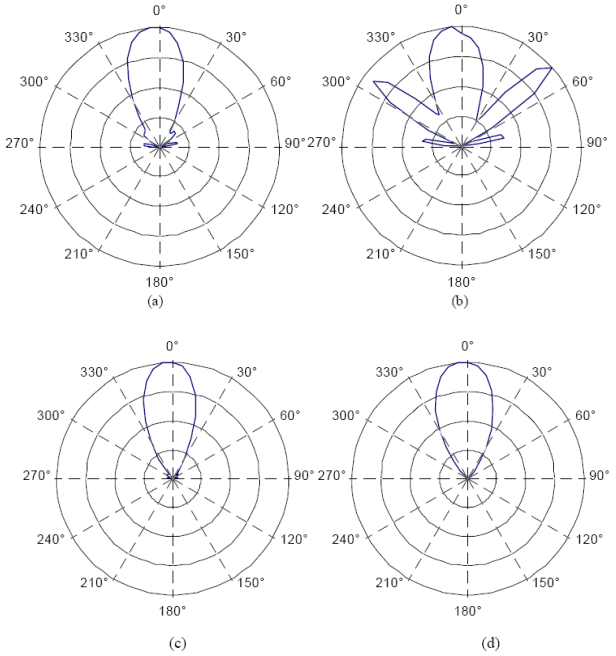
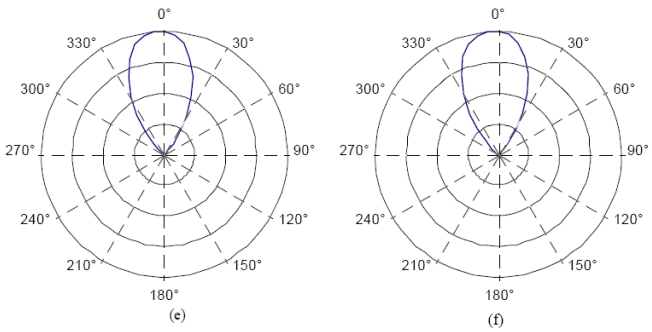


Figure**:** Computed power patterns for the representative case at (a) 1.5 GHz, (b) 1.55 GHz, (c) 1.6 GHz, and (d) 1.65 GHz.



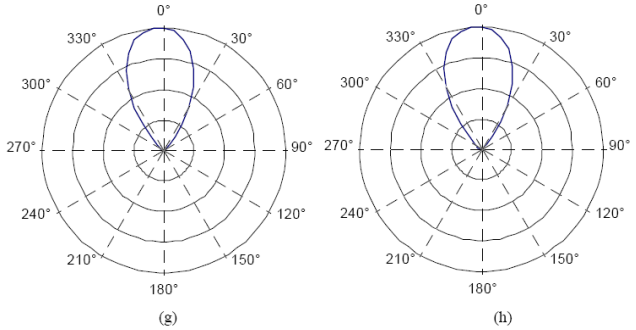


Figure**:** Computed power patterns for the representative case at (e) 1.7 GHz, (f) 1.75 GHz, (g) 1.8 GHz, and (h) 1.85 GHz.

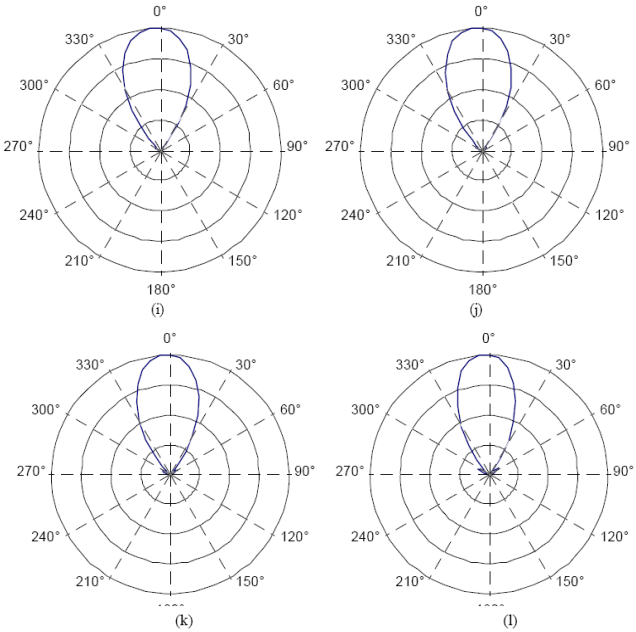


Figure**:** Computed power patterns for the representative case at (i) 1.9 GHz, (j) 1.95 GHz, (k) 2.0 GHz, and (l) 2.05 GHz.

VIII. EFFECT OF PARAMETERS ON RADIATION CHARACTERISTICS

The geometry of the spiro-helical antenna involves five parameters including the number of turns N, pitch angles a and a¢ and radii *a* and *a*¢. Each of these parameters influences the gain characteristics of the antenna and the overall bandwidth. which describes the effect of number of turns, in all other sections the number of turns is assumed to be 10. through 4.4.5, the radius of conductor wire is *r*o= 0.5 mm.

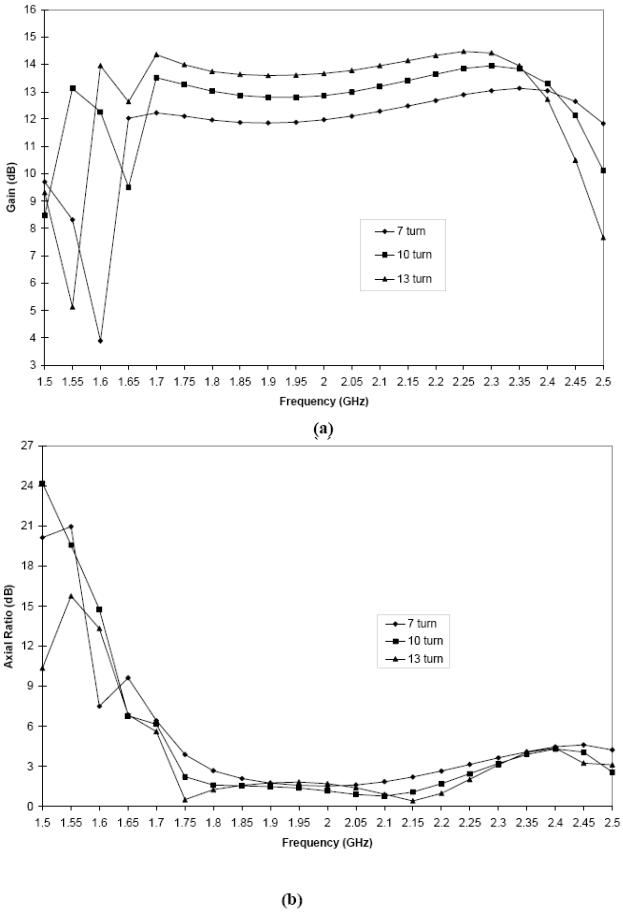


Figure: Variations of (a) gain and (b) axial ratio versus frequency for 7, 10, and 13-turn spiro-helical antennas with *a* = 16mm, *a*¢ = 2mm, a = 10°, and a¢ = 30°.

IX. Results

The important results of the investigation of spiro-helical antenna are summarized below:

The spiro-helical antenna is capable of maintaining radiation characteristics of a conventional helical antenna while occupying a smaller volume. The volume of a spiro-helical antenna can be 2.5 to 3.21 times smaller than a conventional helix while preserving circular polarization and comparable gain and bandwidth.

The peak gain occurs when the outer circumference 2p(*a* +2*a*¢) is about 0.96l. For a conventional helix of comparable gain, this peak occurs at a circumference of about 1.2l.; High gains and wide axial-ratio based bandwidths are obtained when the pitch angle a is about 7 degrees or 10 to 12 degrees. The gain based bandwidth is generally larger than the axial-ratio based bandwidth. The maximum gain, however, occurs when a = 10°; The gain and axial ratio do not vary significantly with the pitch angle a¢. However, a
maximum gain is achieved when a¢ = 30° and a = 10°; The gain increases with the number of turns, but the overall bandwidth is reduced;

X. CONCLUSIONS

A novel antenna that produces high gain and nearly circular polarization over a wide bandwidth and a broad beamwidth has been presented. This new antenna has radiation characteristics comparable to those of a conventional helix but occupies a much . smaller volume. It is made of a primary helix shaped into a larger helix, thus a structure with two degrees of helicity. This antenna is named spiro-helical antenna. A comprehensive numerical analysis of spiro-helical antenna has been carried out using the method-of-moments Far-field patterns, gain-frequency characteristics, and axial-ratio characteristics have been computed for numerous cases. The input impedance has also been examined, but the simulation results for the input impedance require a more in-depth analysis. Several spiro-helical antennas were made and tested. Generally, the measured and computed far-field patterns are in good
agreement. The reduced size of the proposed antenna by a factor of 2.5 to 3, compared to a conventional helix with about the same radiation properties.

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A Survey on Different Approaches for Efficient Mutation Testing

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Abstract- Before delivering the software to the customer the testing of the software is an essential task. Many testing have been used by the software testers in order to check the functionality and quality of software, one of these testing is Mutation testing. Mutation testing is the testing that has received a large amount of investment. In this paper we have reviewed various proposed methods that are capable of performing Mutation testing of software from the consumer's perspective and able to manage the quality requirement in the application.

Index Terms- Mutation Testing, Direct Graph, Regular Grammar, K-sequence Coverage, Mujava;

I. INTRODUCTION

In the recent years the use of computer aided software has become a common task. The people are becoming more dependent on the computer based software and computer application. In order to full fill the needs of people the software development companies are designing the various small sizes to big size and complex soft ware's. Since the use and demand of software is increasing day by day, this is the responsibility of the software development team to deliver efficient and right and testified software to their customer. Thus to achieve this testing of software is an important task. The testing of software is defined as the process of checking software that whether it is performing the function for which it has been designed, and as well as is it meeting the requirements of the customers, without testing the software cannot be delivered to the system. In short testing is a process of delivering error free software to the customer.

Now there are many software testing researchers who have proposed a number of different types of testing like white, black box and model testing. One of the proposed and most researched testing is Mutation Testing [2].

There are some terms in the testing that need to be defined before discussing any other thing about testing or mutation testing. These terms are test suite and coverage [10]. During the process of testing a number of test cases are designed to test the software, the collection of these test cases is known as Test Suite. The part of whole code of the software that a test suite can test is known as coverage. When the code become large and the results of test are not good then selecting the simple testing technique becomes invalid. In such case the tester cannot decide that whether there is any problem in code or the test suite selected is wrong. In such situation mutation testing can be used because mutation testing has been designed to overcome this problem only.

Mutation testing is a testing in which the original code is changed and the changed code is testified to check whether there is a problem in code or in test suites. This changed code is known as MUTANT. When mutant is tested under the selected test suite if the test suites are right then they will catch the error in the Mutant, it means that there is an error in the code and the test suites selected are fine but if the test suites cannot find the error in the mutant then it means that the selected test suites has some wrong test cases and those test cases need to be changed. When one of the test cases identifies the error in the mutant then it is said that mutant has been killed. This is the whole procedure of the Mutation Testing.

We can perform the mutation testing using various methods, the most common and easy way of performing mutation testing is the use of Directed Graph. In directed graph the whole program code is converted into the directed graph having a number of nodes and arcs but the Directed graph has some its own drawbacks during its use in the mutation testing.

Thus, in this paper we have discussed about the various proposed such techniques for performing the Mutation Testing and one of the tool used for performing mutation testing in various Object oriented programming language.

Rest of the paper is organized as: description of Directed Graph Model in section II, Regular Grammar Model as in section III and Semantic Mutation Testing is explained under section IV, Mutation testing in Java is summarized in V and MuJava tool in the section VI. Finally the future work and a conclusion of the discussion are provided in last section

I. DIRECTED GRAPH

It's the foremost preliminary step for proceeding with any research work writing. While doing this go through a complete thought process of your Journal subject and research for its viability by following means:

A directed graph (DG) has V set of finite nodes is a where tuple D = (V, A, S, F) , A has ordered pairs of elements of V which is represented as a finite set of directed graphs , S is a set of start nodes and F is a set of finish nodes. Both S and F belongs to V [3].

This is the definition of directed graph when it is used with the mutation testing. When a program code is converted into a directed graph the one statement of code is represented by the one node of the graph and the paths following that statement is shown by the arcs of the graph. While constructing a directed graph some of the nodes in the graph should be distinguished as start and finish node to clear the usefulness of DG and to show the start and end of the code. Therefore S and F are the Start and

Finish node in the Directed Graph. There are some definitions in context to directed graph in mutation testing.

In a directed graph a node v belonging to V is said to be useful only when there is a path from a start to the finish node [3].

For a given directed graph, a node v belonging to V, related to v strictly preceding nodes are those nodes which occurs in all the paths from these nodes to the finish nodes [3].

For a given directed graph, a node v belonging to V, related to v strictly succeeding nodes are those nodes that occur only in the paths from v to the finish nodes [3].

For implantation- oriented, white box testing nodes of the DG to be covered generally is represented by the statements of SUT (Software under Test) and with arcs sequence of statements is explained [5].
For specification, in the black box testing, behavioral events of SUT are represented by the nodes of DG and arcs represent the sequences of those events [6]. This is how the directed graph is constructed for a given code.

But in mutation testing negative testing is also performed it means it is tested that the software is not doing anything it is not supposed to do. For this [3] have proposed specific manipulation operators for the graphs that models the SUT. These operators are discussed as below:
The directed graph consists of nodes and arcs so it is obvious that the manipulation can be done either in nodes or the arcs. Therefore there are two types of the elementary manipulation operator that can be applied on arcs and nodes. These are insertion and omission. These are defined as below:

a) Node Insertion –

Here a new node v is added to the DG together by the operator with possibly nonzero numbers of arcs, and connecting this node with the remaining nodes. Both the sets A and V are updated after the node insertion.

b) Node Omission-

The existing node v is deleted by the operator from the DG along with arcs, ingoing and outgoing from the deleted nodes. After node omission both the sets A and V are updated.

c) Arc Insertion-

This operator inserts or adds a new arc to the DG, after this a new set A of arcs is formed.

d) Arc Omission-

This operator deletes an existing arc ‘a’ from the DG and the set of arcs A is updated. It may result in some nodes with no ingoing and outgoing arcs.

This is how manipulation is performed in the directed graph.
In every testing task there is a main focus on the coverage. So in context of directed graph coverage is defined as three practical coverage criteria and these criteria are defined as below:

1) Node Coverage:

Given a graph DG and a set of strings B, is said to cover a node v, if v occurs at least in one of the strings in B. If the set of string B covers all nodes in V, then it is said to achieve node coverage.

2) Edge Coverage:

Given a graph DG, a set of strings B is said to cover an edge (u, v) of A, if the sequence uv occurs at least in one of the string in B. If the set of string B covers all edges in A, then it is said to achieve edge coverage.

3) Path Coverage:

Given a graph Dg and a set of string B, is said to cover a path of length k if sequence u1u2---- uk occurs at least in one of the string sin B.

While performing manipulation in the model here it is directed graph one should keep in mind that manipulation operators should not invalidate the model and should take necessary required steps to convert invalid model into valid model. For directed graph model while constructing directed graph model of a SUT, considering the system semantics we should determine start and finish nodes. For doing this [3] have proposed two main approaches:

Fix the set of start nodes and finish nodes and do not allow any sequence of manipulation operations which violate the usefulness of any node in DG.

After performing some manipulation operations, if the resulting DG is invalid, then to validate usefulness of all the nodes select a new and different start and finish node and transform it to the valid one.

This is all about the Direct Graph model for performing mutation testing.

III. REGULAR GRAMMAR

In this section we will discuss that how regular grammar can be used for performing mutation testing and how a directed graph can be converted into a Regular Grammar. So in this section we are discussing the regular Grammar Model of performing Mutation Testing.

The Directed Graphs can only be used to model regular systems. The test generation algorithms based on DGs can be viewed as still being in their starting point means the existing few are relatively slow and they are memory consuming. To overcome such issues of Directed Graphs, regular grammar can be used to model the application, so here is an introduction of a new approach for modeling the mutation testing applications.

The basic definition of the grammar is a tuple (N, E, P, S) where, N is a finite set of non terminal symbols, E is a finite set of terminal symbols, P is a finite set of production rules and the S belongs to N and is a distinguished non terminal start symbols.

To construct the regular grammar model for mutation testing the Directed graph is converted to the Regular grammar. Since the FSA, DGs and RGs equivalently describe regular grammar, it is possible to construct regular grammars from the directed graphs. In the constructed regular grammar from a directed graph the production rule of grammar represents the arc of the graph and terminal and non terminal symbols represent the nodes of the graph. According to [3] the constructed grammar has the following properties that make it more efficient model than directed graph

- G is right RG (so it is unambiguous).
- nt(x) = Rx is a bijection to N\{S} via E.
- non terminal S appears only on the left side of the production rule.
- Each node in V corresponds directly a terminal in E.
- Each arc in V corresponds to a production rule in P (including the pseudo arcs used to mark start and finish nodes).
- All terminal symbols in the grammar are useful if and only if all nodes in the DG are useful.

Similar to the directed graph, the regular grammar also have some special definitions and the manipulation operators. These definitions are discussed as below:

- For a given grammar (N, E, P, S), a terminal symbol r and a non terminal symbol R. A terminal symbol r is said to be useful if it occurs in at least one string in L(G) and the non terminal symbol R is said to be useful if a rule of the form (N union E) * R(N union E) * …. Is used in a derivation S * … of at least one string in L (G).

- For a given grammar (N, E, P, S) and a terminal symbol r, then the strictly preceding terminal symbols for r are all those terminal symbols such that r occurs in all the strings in L (G) in which these terminal symbols also occur and r occurs after these symbols.

For a given grammar (N, E, P, S) and a terminal symbol r, strictly succeeding terminal symbols for terminal symbol r are those terminal symbols which only occur in the either one or all the strings where r also occurs and they occur after r.

Above we have discussed the special definitions of the regular Grammar, now further we will discuss the manipulation operators of the Regular Grammar Model.

a) Arc Insertion-

The arc insertion operator adds a new arc in the directed graph and in the grammar arc is represented by the production rule. Therefore when an arc is inserted in the DG then RG should be updated with a new production rule to show the insertion of this arc.

b) Arc Omission-

The arc omission operator deletes an existing arc from the directed graph and in the regular grammar an arc is represented by the production rule. Therefore when a arc is deleted from DG then a production rule must be removed from the RG. This operation may violate the usefulness of some terminal symbols so some extra care is needed while performing this operation.

Node Insertion-

The node insertion operator adds a new node in the Directed Graph and when a new node is added in the DG then new arc ingoing and outgoing form that node are also inserted in the DG. In Regular Grammar a node of DG is represented by the terminal and non terminal symbols, thus when a node is inserted in DG a new non terminal symbol and a new non terminal symbol is added to the RG. And due to the addition of terminal symbols and non terminal symbols in the RG the production rules are also added in the RG to show the arcs ingoing and outgoing from the nodes.

c) Node Omission-

The operation of arc omission deletes a nodes from the graph due to which the arc ingoing to and outgoing from this nodes are also deleted. Therefore the node omission operation is based on arc deletion of arcs from the graph.

Thus when deleting a arc from DG, a production rule is deleted from the RG and proper steps should be taken to first the arc is deleted to perform removal of the production rules, and then the isolated terminal symbols and non terminal symbols are removes from the grammar.

Now we will discuss the coverage criteria included in the Regular Grammar. The coverage criteria are discussed as below:

I. Terminal Symbol Coverage:

For a given grammar G and a set of sting A is said to cover a terminal symbol e, if e is present in at least in one of the string in A. if the string A contains all the terminal symbols of E then it is known as Terminal Symbol Coverage.

II. Production Rule Coverage:

For a given grammar and a set of strings A, then A is said to cover a production rule p, if p is used at least in one of the derivation of A. if the set of string A contains all the production rules belonging to P (set of production rules) then it is known as Production Rule Coverage.

This is all about the Regular Grammar model for Mutation Testing.

IV. SEMANTIC MUTATION TESTING

In this section we are going to introduce a new approach behind the mutation testing. The simple mutation testing about which we

have given a small description works on the syntax of the program or code of the software, means it focus on the code not the semantic of the code it only check the syntax not the semantic of the language used. In such condition the number of mutants gets increased in number which becomes difficult to execute practically. There is a need of a change in the standard mutation testing [4]. For this they have proposed a new mutation testing that is Semantic Mutation Testing. This testing focus on the semantic of the language used means it performs the operations of mutation on the semantic of the language used. In this section we are going to describe in brief why we need semantic mutation testing over standard mutation testing and what the basic concept behind the Semantic Mutation Testing is.

A. Need of Semantic Mutation Testing

When software is developed then it passes through various phases of software development Life cycle model, in each step of life cycle model the team members give the different description to the software and uses different languages in this situation the language vary from step to step. When the transformation of prototype model of software to target model of software take place the description again changes and there come different descriptions of software from this different misunderstanding regarding the actual description of software arises, in such case our traditional mutation testing cannot work efficiently in detecting the error in code and misunderstanding regarding the description of the software. To overcome this problem of Mutation Testing, a modified version of mutation testing "Semantic Mutation Testing" was proposed by [4]. This works on the semantic of the language used for the program. The semantic is modified to generate the mutant means mutant operations are applied on semantic rather than description. The working and detailed concept of Semantic Mutation Testing is described in the next part of this section.

B. Concept Behind Semantic Mutation Testing

In Standard Mutation Testing the mutation operators are applied at syntactic level in does not focus on the semantic misunderstanding generated by semantic mistakes as explained above. For this semantic mutation testing was introduced. In this testing mutation operations are applied on semantic and it explores the variance in the semantics.

The code which is under test is represented by a description. Take an example, a description D is written in language with semantic S, the behavior of the description is defined by an ordered pair (D, S). The standard mutation testing will apply the mutation operations on the D that is the description under test. Thus the mutant for (D, S) can be represented as (D, S) $\rightarrow$ (D', S) where D' is the change in the D.

On the other hand the Semantic Mutation Testing will apply the mutant operator in the semantic of the language that is S. thus the mutant obtained for the (D, S) will be represented as

(D, S) $\rightarrow$ (D, S') where S' is the semantic obtained after applying mutation operations.

Now how the Semantic Mutation Testing kills a mutant. For a given D of (D, S) after applying mutation operations on S there are two interpretation one its explanation under S and second its explanation under S' and we will call these as Ds, Ds' respectively. For a test case say t, Ds (t) will denote the behavior produced when applying t to D under semantics S and Ds' (t) will denote the behavior produced when applying t to D under semantics S'. Then mutant (D, S') is killed by test case t only when Ds (t) = Ds' (t).

Further, if for all t we have Ds (t) = Ds' (t) then this mutant (D,S') is said to be an equivalent mutant.

From here it is clear that, the notions of behavior, equality of behavior, and thus the language used by the testers will decide the killing of the mutant. Along with this, the semantic mutation made and the description under test are the factors that will decide the property of killing the semantic mutant [4].

V. MUTATION TESTING IN JAVA

In this section we are going to discuss a how mutation testing is different for the code written in language java. In the basic mutation system there are two main parts. In the first section the original code is converted into the mutants or also known as mutation programs, this is done by using the mutation engine that has a number of mutation operators. In the second section the test cases are generated and are run as input for the respective program [7]. Hence the accuracy of the test cases is decided if the test cases detect the fault in the mutant.

The major deficiency of the traditional mutation testing tool is that they lack in automation for major parts of the process [11]. The important tasks like entering test cases and then running the test case on original input and then checking the results, and after that running the test case on mutants and checking the results are very human-intensive [7]. In order to remove this lacking of tool, schema generator was introduced and the MuJava tool is based on this advancement. The concept behind schema generation is production of meta-mutant. It incorporates all the mutants on an original code in a single code/program.

In the mutation testing for code in object oriented language like java the mutation system has two types of mutation operators

a) Traditional mutation operators

b) Class level mutation operators

The traditional mutation operators are those which are generated from procedural language but the class level mutation operators' works on the features of OOP like inheritance, polymorphism, Data binding, overloading.

The table 1 shows some of the traditional as well as class level mutation operators [13].

TRADITIONAL OPERATORS	CLASS LEVEL MUTATION OPERATORS
ABS	AMC
AOR	PNC
LCR	OAN
ROR	JTD (java specific)
UOI	EOA

Table1. Different Mutation operators

For JAVA 24 class mutation operator were found by Ma, Kwon and Offutt for testing object oriented and integration issues [7]. The class mutation operators are applicable at different levels as in intra-method, inter-method, interclass and intra-class. In reference [7] author has proposed a mutation testing engine for the JAVA. In the next section we are going to discuss the MuJava tool for performing mutation testing in JAVA.

VI. MUJAVA TOOL

In this section we will discuss a tool "MuJava" which support the object oriented mutation testing for Java programs [7]. This system was generated by the Yu Seung Ma and Jeff Offutt initially it was known as JMutation, a mutation system for java programs but later it was changed to MuJava. The main concept of schema generator on which the MuJava is based is Schema Generation [14], in this the Meta mutant is generated which incorporates all the mutants on an original code in the one program. The MuJava does not generate test cases, it only generate the mutants and run the test cases supplied by testers [8].

The MuJava tool consists of three phases:

1) Mutant Generation:
In this phase the mutant operators are applied on the original piece of code to generate the mutants. In the tool interface there is an option of "Generate" button where tester can select the file for which they want to create mutants and can choose which mutant operator is to be applied on file.

2) Analyze Mutant:
n this phase mutants are analyzed. In MuJava tool there is an option "Mutant Viewer" to view the mutant created. There is an option where the user can analyze the generated mutant against the original code. When the class level mutant operators are applied MuJava generates mutants in two different ways (MSG and byte code translation)

3) Run Test Cases:
In this phase the test cases generated and supplied by the testers are run. To run the test cases in MuJava there is an interface that contains a button "Run" to execute mutants against the test set supplied by tester and shows the results in the form of a mutation score of test set. The mutation score decides the accuracy and correctness of test set. the information of live and killed mutant is shown on the lower right hand side of the interface.

This is how the MuJava tool supports the Mutation Testing for JAVA.

VII. CONCLUSION

In this paper we have discussed the various techniques for performing the mutation testing that focus on both the test suites used and codes of the application to satisfy the need of the customer and software requesters. All the discussed methods have their own field of use. In the future work we will make the use of these techniques and will make some results out of that. The use of these techniques in real scenario is a big research oriented work so in future we will focus on the implementation of these techniques.
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Advancement in Communication and Security with Quantum Physics and Mechanics

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Abstract- This paper presents the inventions and discoveries already made in quantum physics which are used now as the main base. It also contains the new technology that has come up by various scientists studying it and researching over it. The invention of quarks and neutrinos has not been an easy task. This paper proposes a new outlook with which we can start thinking in developing communication and technology.

Index Terms- Carbon nanotube, Schrödinger's Equation, De Broglie Equation, Uncertainty Principle

I. INTRODUCTION

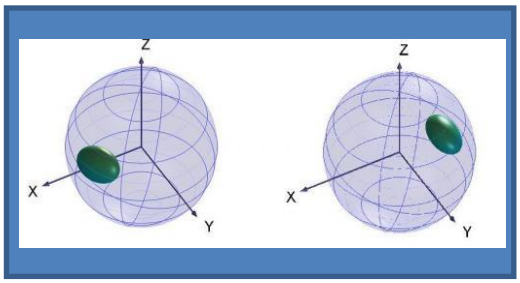
The world that had started with a small particle has grown to a giant mass. And all this was discovered by the people who think as to why. To start with, the thought that the world was made up of small particles was practically proved by Neil Bohr. His Copenhagen Interpretation that a particle sometimes acts as a particle and sometimes as a wave surprised everyone. Erwin Schrödinger created quantum equations based on wave mathematics, a mathematical system that corresponds to the world we know much more than the matrices. After the initial shock, first Schrödinger himself then others proved that the equations were mathematically equivalent. Bohr then invited Schrödinger to Copenhagen where they found that Schrödinger's waves were in fact not like real waves. Also, each particle that was described as a wave required three dimensions.

Schrödinger also said that particles still jumped from one quantum state to another; even expressed in terms of waves space was still not continuous. All this incomplete conclusions are still continuing today. Even today people try to imagine the atomic world as being a bunch of classical waves. The equations that Schrödinger gave were used even later for other observations. Even Bohr used those equations in his Copenhagen Interpretation. Werner Heisenberg proposed in 1925 that it is in fact *physically impossible* to do so. As he stated it in what now is called the Heisenberg Uncertainty Principle, if you determine an object's position with uncertainty x, there must be an uncertainty in momentum, p, such that $xp > h/4\pi$, where h is Planck's constant. Contradicting the Uncertainty Principle, in 1935, Einstein and two other physicists, Podolski and Rosen, presented what is now known as the EPR paper in which they suggested a way to do just that. The idea is this: set up an interaction such that two particles are go off in opposite directions and do not interact with anything else. Wait until they are far apart, and then measure the momentum of one and the position of the other. Because of conservation of momentum, you can determine the momentum of the particle not measured, so when you measure its position you know both its momentum and position. The only way quantum physics could be true is if the particles could communicate faster than the speed of light, which Einstein reasoned would be impossible because of his Theory of Relativity.

II. BRIEF DESCRIPTION OF LATEST TECHNOLOGY

A. ATOM'S SPIN

Researchers suggest one can affect an atom's spin by adjusting the way it is measured. All spin directions (represented by the spheres) collapse on one or the opposite direction depending on the measured photon polarization. One of the most basic laws of quantum mechanics is that a system can be in more than one state. It can exist in multiple realities – at once. This phenomenon, known as the superposition principle, exists only so long as the system is not observed or measured in any way. As soon as such a system is measured, its superposition collapses into a single state.



B. QUANTUM INFORMATION PROCESSING

Researchers from Yale, Surrey, and Paris have made an important breakthrough towards 'quantum information processing, which promises to lead to massive information technology advancement in the future. Quantum information allows for the ultimate information processing unit in terms of accuracy, distance of coverage, knowledge when eavesdropped, simplicity in design of parallel systems and potential bandwidth. In the current issue of Nature, researchers have demonstrated a new approach for the manipulation of quantum states of light. Photons need to interact with each other sufficiently strongly for this to be achieved. This was made possible by engineering a device where photons interact with each other even when only a few photons are present but at the same time the fragile quantum state is not destroyed by the environmental noise. Dr. Eran Ginossar, one of the authors who conceived the idea, based at the Advanced Technology Institute (ATI) at the University of Surrey said: "What makes this discovery so exciting is that up to date it

has been considered very difficult to engineer strong interactions between localised photons. This will open up a way of encoding quantum information directly to photons in one of the most promising architectures of quantum computing". Professor Ravi Silva, Director of the ATI, commented: "Previously envisaged bottlenecks to quantum information processing (QIP) over large areas and distances will now be enabled by this breakthrough. QIP is predicted to lead to enhance computing and communication even beyond the solutions to the spectrum crunch being resolved by the recently established 5G Innovation Centre at Surrey.

C. QUANTUM COMPUTERS COUNTING ON CARBON NANOTUBES

Using quantum mechanical phenomena, computers could be much more powerful than their classical digital predecessors. Scientists all over the world are working to explore the basis for quantum computing. To date most systems are based on electrically charged particles that are held in an "electromagnetic trap." A disadvantage of these systems is that they are very sensitive to electromagnetic interference and therefore need extensive shielding. Physicists at the Technische Universitaet Muenchen have now found a way for information to be stored and quantum mechanically processed in mechanical vibrations.
Carbon nanotubes can be used as quantum bits for quantum computers. A study by physicists at the Technische Universitaet Muenchen (TUM) has shown how nanotubes can store information in the form of vibrations. Up to now, researchers have experimented primarily with electrically charged particles. Because Nano mechanical devices are not charged, they are much less sensitive to electrical interference.

A carbon nanotube that is clamped at both ends can be excited to oscillate. Like a guitar string, it vibrates for an amazingly long time. "One would expect that such a system would be strongly damped, and that the vibration would subside quickly," says Simon Rips, first author of the publication. "In fact, the string vibrates more than a million times. The information is thus retained up to one second. That is long enough to work with."

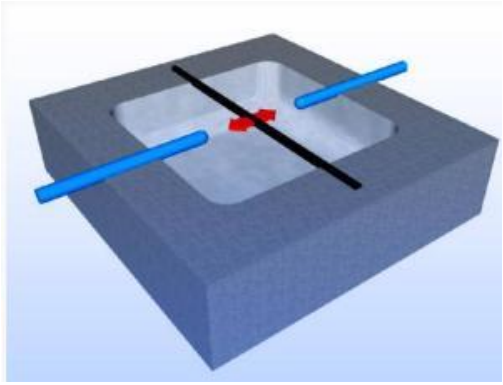


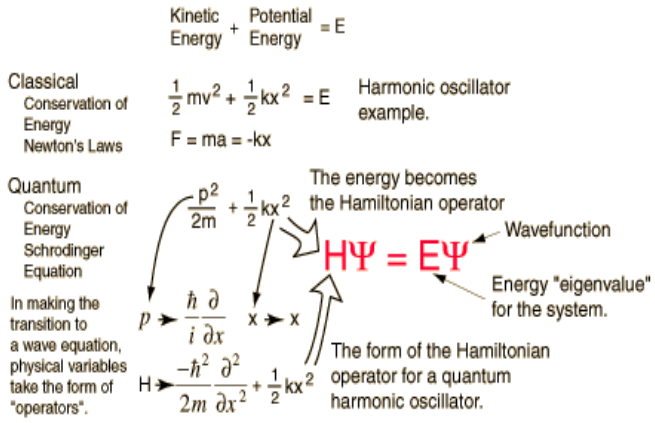
FIG: ***Like a guitar string nanotubes (black) can be clamped and excited to vibrate. An electric field (electrodes: blue) ensures that two of the many possible states can be selectively addressed. (Credit: M.J. Hartmann, TUM)***

Since such a string oscillates among many physically equivalent states, the physicists resorted to a trick: an electric field in the vicinity of the nanotube ensures that two of these states can be selectively addressed. The information can then be written and read opto electronically. "Our concept is based on available technology," says Michael Hartmann, head of the Emmy Noether research group Quantum Optics and Quantum Dynamics at the TU Muenchen. "It could take us a step closer to the realization of a quantum computer."

III. QUANTUM RELATED EQUATION

A. Schrödinger's Equation

The Schrodinger equation plays the role of Newton's laws and conservation of energy in classical mechanics - i.e., it predicts the future behaviour of a dynamic system. It is a wave equation in terms of the wave function which predicts analytically and precisely the probability of events or outcome. The detailed outcome is not strictly determined, but given a large number of events, the Schrodinger equation will predict the distribution of results.



The kinetic and potential energies are transformed into the Hamiltonian which acts upon the wave function to generate the evolution of the wave function in time and space. The Schrodinger equation gives the quantized energies of the system and gives the form of the wave function so that other properties may be calculated.

B. Particle in the box

The idealized situation of a particle in a box with infinitely high walls is an application of the Schrodinger equation which yields some insights into particle confinement. The wave function must be zero at the walls and the solution for the wave function yields just sine waves.

The longest wavelength is

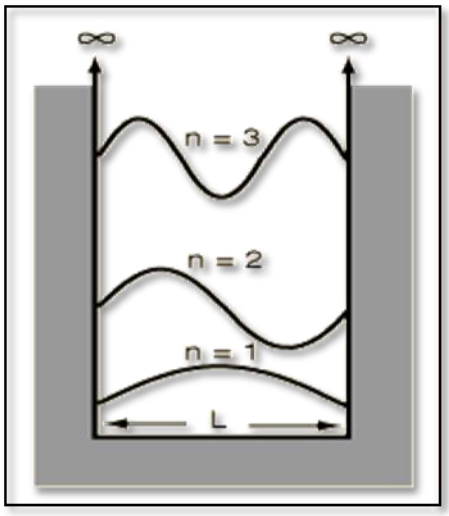
$$\lambda = 2L$$

and the higher modes have wavelengths given by

$$\lambda = \frac{2L}{n} \text{ where } n = 1,2,3,4 \ldots.$$

When this is substituted into the De Broglie relationship it yields momentum

$$p = \frac{h}{\lambda} = \frac{nh}{2L}$$



When the momentum expression for the particle in a box :

$$p = \frac{h}{\lambda} = \frac{nh}{2L} \quad \text{where n= 1,2,3,4 ....}$$

It is used to calculate the energy associated with the particles

$$\frac{1}{2}mv^2 = \frac{p^2}{2m} = \frac{n^2h^2}{8mL^2} = E_n$$ Energy for nth quantum state for particle in infinite box.

Though oversimplified, this indicates some important things about bound states for particles:

1. The energies are quantized and can be characterized by a quantum number n
2. The energy cannot be exactly zero.
3. The smaller the confinement, the larger the energy required.

If a particle is confined into a rectangular volume, the same kind of process can be applied to a three-dimensional "particle in a box", and the same kind of energy contribution is made from each dimension. The energies for a three-dimensional box are

$$E = \frac{\left(n_1^2 + n_2^2 + n_3^2\right)h^2}{8mL^2}$$

This gives a more physically realistic expression for the available energies for contained particles. This expression is used in determining the density of possible energy states for electrons in solids.

C. Carbon Atom

The carbon atom, in particular an atom of the isotope ^{14}C, has been chosen to illustrate the implications of quantum mechanics for the energy required to confine a particle to a region of space. Without recourse to the details of the nature of the fundamental forces, quantum mechanics gives the insight that it takes 10's of electron volts to contain electrons in atoms, and energies on the order of MeV's to contain protons in nuclei.

Taking the diameter of a carbon atom from the periodic table and calculating the minimum energy consistent with the uncertainty principle for a cubical volume of that dimension, we obtain a value of 10.4 eV. This compares with the observed value of 11.3 eV for the first ionization potential for the carbon atom. If we calculate the minimum confinement energy for the proton in a space the size of a carbon atom, we find that to be only 0.0056 eV. This extremely low energy can be compared with the average thermal energy of 0.04 eV at 300K! This implies that the proton, with only the energy provided by the internal energy of the normal environment can just wander in and out of an atomic space.

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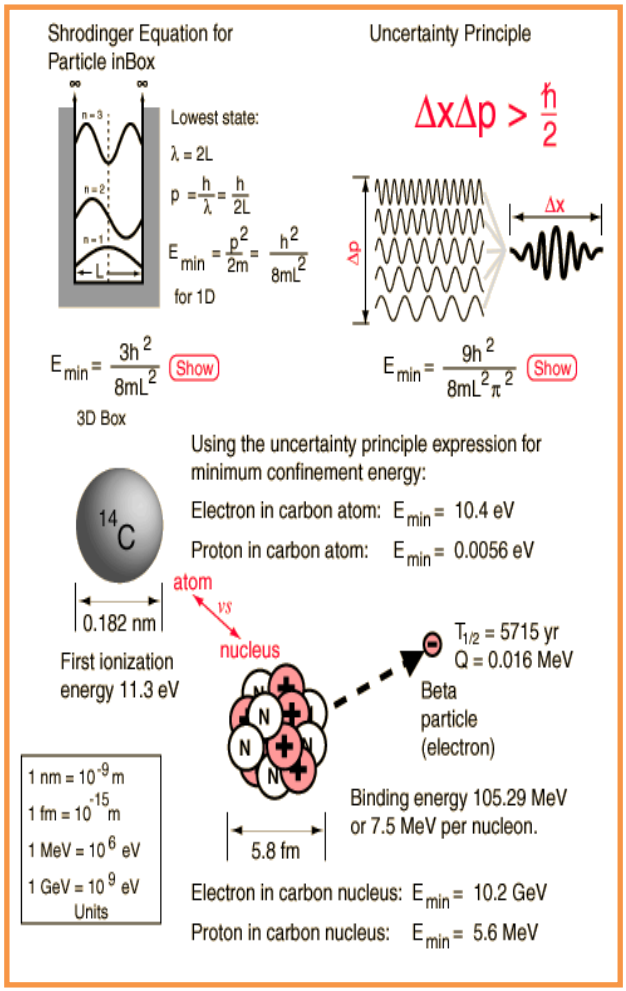
When we apply the minimum energy calculation to the carbon nucleus, the picture is very different. The confinement energy for keeping a proton inside the cubical volume of dimension equal to the nuclear diameter is 5.6 MeV.



This compares well in magnitude with the observed average binding energy of 7.5 MeV for nucleons in the carbon-14 nucleus. So with no tools other than the uncertainty principle, we have established the scale of energy for nuclear processes. Observed radioactive processes are in the range 0-10 MeV, and this is consistent with the uncertainty principle.

The story for the electron is more dramatic. We observe electrons being emitted from the carbon-14 nucleus (beta decay) with the relatively small energy of 0.016 MeV. Does that imply that there are electrons hanging around inside the carbon nucleus? Definitely not!! The minimum confinement energy for an electron in the nuclear volume is a preposterously high 10.2 GeV, a half-million times greater than the observed decay energy of the carbon-14 nucleus. The calculation of required confinement energy then implies that the observed electron has been created inside the nucleus as a part of the radioactive decay process rather than being simply an ejection of an electron which was already there.

IV. New Prospects

We can use the concept of quantum mechanics to find the distance between very small particles. We can analyse very tiny world. Scientist Werner Heisenberg found that subatomic particles can be present in more than one place at a time. In travelling from one place to another they don't even need to travel that distance. It is astonishing to know they can emerge out from any object and can also vanish when they want. Hence they are called as virtual particles.

By use this concept we can control a particular system which is far away from us. For example: when we travel in train we can know its working. How the operator gets the information that train must be stopped in a particular area. It is from the signal that he gets .Hence we can operate system by sending the information which can reach faster than the speed of light, as these are virtual particles. Through this we can make the communication faster and give the desire response. We know by now that there is a link between particles near and far. There is a link that synchronizes them. They acquire the same properties. If one particle is moving clockwise and has some speed then the other will also rotate with the same speed. But to maintain the law of nature, it will move anticlockwise. This concept can be used in running and controlling heavy machinery which are far and need vigilance from one specific point. We can make that link by making use of electromagnetic waves. We can send our data with very high speed to one place to another place. It can also be used for security purposes and the span of time with which we get the information will be very minute. In a fraction of seconds we will be able to detect the error at a place and the victim can be aided.

V. Conclusion

Quantum physics is such a vast topic, yet to be understood properly. But, scientists have contributed their whole strength and life in improving the technology to make the world comfortable. Everything that we observe has science hidden in that. Quantum has turned the thinking from macro to nanotechnology. The study of physics is interesting and mind boggling if one understands it and can relate it with the reality. So much improvement can be made still and hence we have to keep proposing ideas and accept those ideas from different individuals

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Reverse Engineering of Object Oriented System

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Abstract: In today's world, every aspect of business is driven by IT. Jobs in IT Industry are never static. They are most dynamic and people keep on switching from one industry to other. So this result in lack of continuity of same people working on same project tills its completion and hence at any time there is need to add new people to the project. So the most important thing for the new comers to any project is to understand the flow of entire project. Reverse engineering is the study of an application's code and behaviour, in order to better understand the system and its design. There are many existing tools that will assist the developer with this undertaking, such as Rational Rose®, jGRASP®, and Eclipse®. So there is necessity to provide a framework which would convert the given code into UML diagrams. So, our project idea is to develop a tool which would convert the given code into UML diagrams. Although, there are certain tools available in the market as given above which do this work, but all these tools are paid application and other which are open source lack behind in certain aspects which would be overcome by our tool.

Index Terms (Ref ACM Keywords):

D.2.2 Design Tools and Techniques- Object-oriented design method

Software Engineering– Design Tools and Techniques- Computer Aided Software Engineering (CASE), Object oriented modelling, Software libraries, Reverse engineering, Modules and Interfaces.

I. INTRODUCTION

In the world of computing applications, approximately 30-35% of the overall total lifecycle costs are devoted to helping the programmer understand the functionality of existing code. This is a necessary task, in order to correctly make required changes in response to new requirements, to resolve errors, or perform other changes [12]. Reverse engineering, analyses a system's code, documentation and behaviour to create system abstractions and design information [13]. Reverse Engineering is, essentially, the practice of examining existing systems, at any stage, to identify elements and dependencies. This information is then used to gain more knowledge about the design, the structure, system code, and functionality. The main objective of our project is to overcome the problem of code tracing. This can be achieved through a project code tracing software, which will give project flow in diagrammatic form by using standard UML notations which results in easy and quick understanding of the flow. Our project will accept the source code files and check if they are compliable. If yes then this files will go through a light weight parser developed by us. The parser will now extract the information necessary from our point of view to draw the uml diagrams and will store in it proper format in sqlite database.
Then the mapping function will convert this information into class diagram and activity diagram.

II. LITERATURE SURVEY

A. Multiplicity analysis of association:

Let us outline the approach for the associations:

1. Prior to the analysis, we have to interpret the notion of UML association and its multiplicity in a precise way:

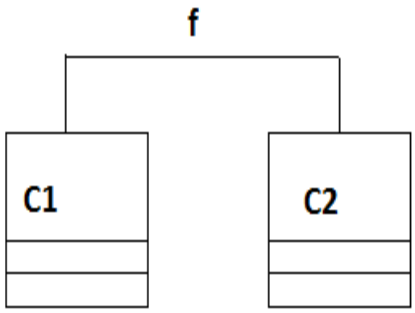


Fig. 1

An association f as shown in fig. 1 represents a field f in the definition of C1, defined either as C2 f; or as an array, C2 f[ ]; .The C2-end multiplicity of m::n is interpreted as the number of C2 objects f can reference during the lifetime of a C1 object.

2. We then make a key observation, which will be the underlying idea for the multiplicity analysis:
Every assignment to f, f = X, implies a local multiplicity. E.g., the intuition for f = new C2 would be a local multiplicity of 1::1.

3. Local cardinalities depend on where they are encountered. A constructor, called at most once, differs fundamentally from other methods which can be called an arbitrary number of times. Thus, local cardinalities have to be fine tuned" - in our example, f = new C2 would imply a local multiplicity of 0::* if encountered in an ordinary method, since we do not know how often that method might be called.

4. We store the data regarding association in the database and again show the association relationship in the class diagram where it exists.[14]

B. Reverse Engineering

Related Areas and Sub-Topics in Reverse Engineering

Reverse engineering is a broad subject area, which includes a variety of sub-topics and components. Many terms are used when discussing reverse engineering. Some of these terms include [15]:

• Forward Engineering - the process of starting at the gathering of requirements and then following through to design and finally to the implementation of the application.
• Design Recovery - gathering additional information, like domain knowledge, outside information, and deductive information for inclusion with other observations, to assist the professional in better understanding the system being studied.
• Restructuring - the movement from one form to another form at the same level of abstraction without changing the system's output. Essentially, it is changing code to put it in a more structured format.

Reverse Engineering Defined

With society's dependence on the Internet, many businesses need to modify their current applications, to make them web-based and move towards an electronic way of doing business. This trend has created more of an interest in code maintenance and evolution than in the past [15]. Thus, there is now a need for experts in older systems, as software maintenance and evolution is becoming more necessary. Roughly, one third of total life-cycle costs are used for the programmer to understand the functionality of the existing code [15]. Reverse engineering is the act of recognizing systems elements, along with their corresponding dependencies, to generate a variety of application abstractions and design data from these system elements

III. SYSTEM OVERVIEW

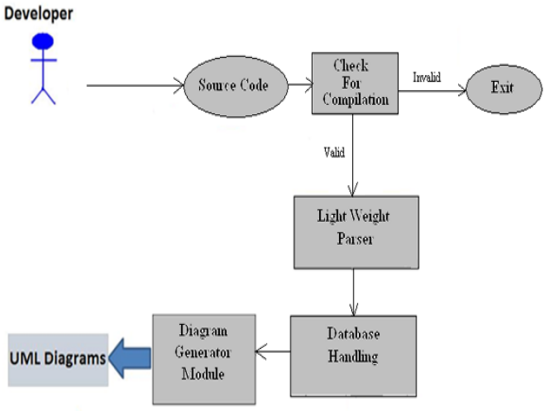


Fig. 2 System Architecture

The system overall consists of four major components. They are stated below in the order in which they will appear in the program structure:

-A Module “Check for compilation” that takes java source file as input and checks whether they are compliable.

-A Module “Light Weight Parser” that will take the valid files as input and extract information necessary for generation of class and activity diagram.

-A Module “Database Handling” that will store the retrieved data in the appropriate format in SQlite database.

-A Module “Diagram Generator” that will draw the class and activity diagram.

IV. DESIGN

Research for this thesis included examining various reverse engineering tools, such as those found in Rational Rose®, Eclipse®, NetBeans®, and jGRASP®, followed by comparing and analyzing the their outputs and methodologies. Our project includes accepting Java programs as input and determining the structural characteristics of the program. Our project includes creation of class diagram and activity diagram (flow diagram and object diagram).

A. Class Diagram

Our main objective while drawing the class diagram is to include the following relationships in the diagram:

1. Generalization
2. Association
3. Dependency

I. Generalization

The UML graphical representation of a Generalization is a hollow triangle shape on the superclass end of the line (or tree of lines) that connects it to one or more subtypes. The generalization relationship is also known as the inheritance or "is a" relationship. The superclass (base class) in the generalization relationship is also known as the "parent", superclass, base class, or base type. The subtype in the specialization relationship is also known as the "child", subclass, derived class, derived type, inheriting class, or inheriting type.

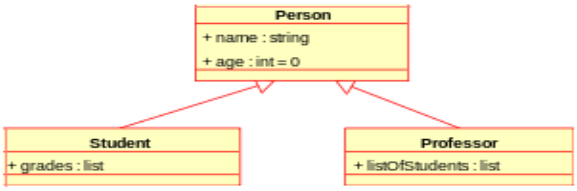


Fig. 3

In fig. 3 the classes student and professor are the base classes and the class person is the superclass.

1. An inheritance relationship exists within two classes of a given code when we encounter two keywords in the class definition :

1.Extends 2. Implements

2. So our project includes a parser which consists of such regular expressions which identify these key words in the class definition and retrieves the needed data in our database.

3. Our database consists of following tables that are used to store information needed to draw class diagram. The tables with the column names are given below :

I. Classes(class_id, class_name, static, final, abstract, access, parent)
II. Interface(class_id, interface_id, string)
III. Function(class_id, function_id, type, name, access, static, final, constant)
IV. Parameter(function_id, par_id, type, name)
V. Variable(class_id, var_id, type, name, access, static, final, abstract, const)
VI. Mapping(map_id, s_class, s_func, d_class, d_func)

4. The data from table Classes is used to provide class name and type, the table Variable provide the list and type of variables used within the class and Function provide the functions included in the class.

5. The column parent from table classes and Interface table provide the data required for showing generalisation relationship in the class diagram.

6. Finally, data is interpreteded by drawing a hollow triangle shaped arrow pointing from the base to the superclass.

- ALGORITHM:

1. Provide the parser with the source files whose class diagram is to be created.

2. Store the retrieved data in the necessary format in the database.

3. Use the the stored data to draw the class diagram.

II. Association

An association represents a family of links. Binary associations (with two ends) are normally represented as a line. An association can be named, and the ends of an association can be adorned with role names, ownership indicators, multiplicity, visibility, and other properties.

Fig. 4 Class diagram example of association between two classes

Our project uses the idea explained in literature survey 3.4 to find out multiplicity in association relation. The parser stores whether a instance variable of any class is defined within a class in the table Variable whose structure is as shown below:

Variable(class_id, var_id, type, name, access, static, final, abstract, const)

III. Dependency

In UML, a dependency relationship is a relationship in which one element, the client, uses or depends on another element, the supplier. We can use dependency relationships in class diagrams, component diagrams, deployment diagrams, and use-case diagrams to indicate that a change to the supplier might require a change to the client.

As the following figure illustrates, a dependency is displayed in the diagram editor as a dashed line with an open arrow that points from the client to the supplier.

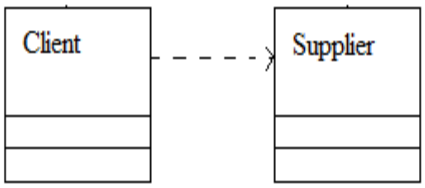


Fig. 5 Dependency

The dependency relationship indicates that the client class performs one of the following functions:

1. Temporarily uses a supplier class that has global scope

2. Temporarily uses a supplier class as a parameter for one of its operations

3. Temporarily uses a supplier class as a local variable for one of its operations

4. Sends a message to a supplier class

A dependency is used when the referenced class is not directly used inside the referencing class, but only passes through the class. For example:

A customerDAO would have a dependency to a customerTO, since the DAO uses to TO to store it in database, but it has no direct association with the TO (the TO is not a instance variable of the DAO).

The difference between dependency and association can be explained by the following example:

```
Ex:1
public class Flat
{
        void rent();
}

 public class Person
{
        flat f1;
        void name();

                }
 Ex:2
 public class Flat
{
        void rent();
}

 public class Person
{
        void name(Flat f1)
{
        f1.rent();
}

                }
```

Ex:1 shows association whereas ex:2 shows dependency. The difference is that in Ex:1 a instance of class Flat is made in the class Person whereas in Ex:2 an instance is used to call a method of class Flat.

- ALGORITHM:

1. Provide the parser with input files.

2. The classes in which only instance variables are created as per ex:1 are stored in table Variable.

3. Whereas the classes in which instances variables are used as per ex:2 are in separate dependency table in the database.

4. With this data retrieved data we can show dependency and association relationship in the class diagram.

B. Activity Diagram

Activity diagrams are graphical representations of workflows of stepwise activities and actions with support for choice, iteration and concurrency. In the Unified Modeling Language, activity diagrams can be used to describe the business and operational step-by-step workflows of components in a system. An activity diagram shows the overall flow of control.

Based on data flow models – a graphical representation (with a Directed Graph) of how data move around an information system

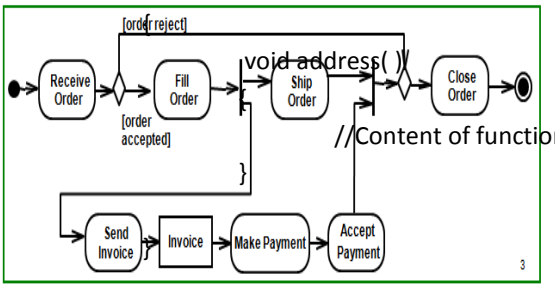


Fig. 6 Example of activity diagram

In our approach, the data which is retrieved from the java code and stored in the table Mapping is used to draw the activity diagram. The structure of the mapping table is as shown

Mapping (map_id, s_class, s_func, d_class, d_func)

map_id	s_class	s_func	d_class	d_func
1	Flat	rent	Person	name
2	Flat	rent	Apartment	address

Consider the following example,

```
public class Flat
{
        void rent( )
{
        Person P1=new Person();
        Apartment A1=new Apartment();
        P1.name();
        A1.address();

}
}

public class Person
{
        void name( )
        {
                //Content of function
        }

                }
public class Apartment
{
void address( )
{
//Content of function
}
```

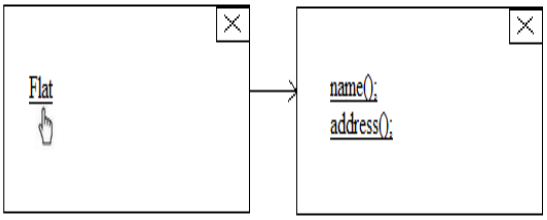
So when the above code is traced by the parser the following data is added in the mapping table

Now from this we come to know that from the function rent from class Flat we are calling two functions namely

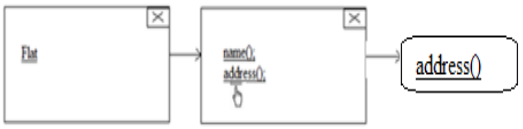
1. name from class Person
2. address from class Apartment

1. Now our application will first draw the source class Flat.

2. As two functions are been called from class Flat so those two called functions will be displayed in another window when you click on Flat.



3. Now as no other function has been called in name and address the functions would be displayed as action.



V. MATHEMATICAL MODEL

Let S be the system that creates class diagram and activity diagram for source code.

$S=\{I, O, F, S_c, F_c \mid \Phi s\}$
Where,
I → Input
O → Output
F → Functions
Sc → Success case
Fc → Failure case
Φs → Set of Rules

$I = \{J, J_g\}$

$J=\{J1, J2, \ldots\ldots, J_n \mid J_i \in \{\text{valid java files for desktop applications}\}\}$
$J_g=\{T1, T2, \ldots T_n \mid T_i \{\text{Valid java tokens}\}\}$

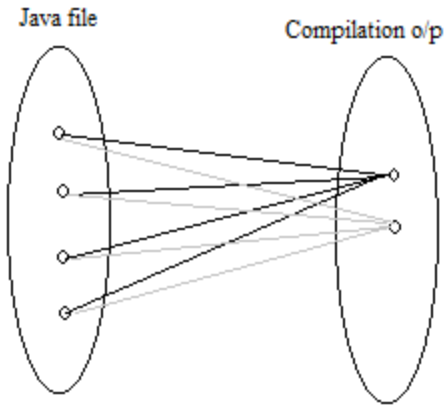
$n \in \{\{\text{method, class, interface}\} \varepsilon \text{ Java element}\}$

$O=\{Dm, Dg \mid \Phi s\}$

$Dm = \{Dm1, Dm2, \ldots.., Dm_n \mid Dm_i \text{ is mapping file}\}$
$Dg = \{Dg1, Dg2, \ldots\ldots, Dg_n \mid Dg_i \text{ is actual graphical representation}\}$
Where,
$Dg_i = \{U\}$
$U = \{U1, U2, \ldots\ldots, U_n\}$
Where,
U_i is UML notation.

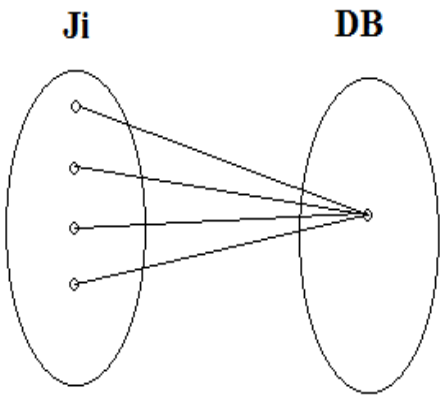
$F=\{F1, F2, F3\}$

F1: checkForCompilation()
$F1 : j_i \rightarrow b_i$
$b_i \varepsilon B$
$j \rightarrow J$
$B=\{True, False\}$

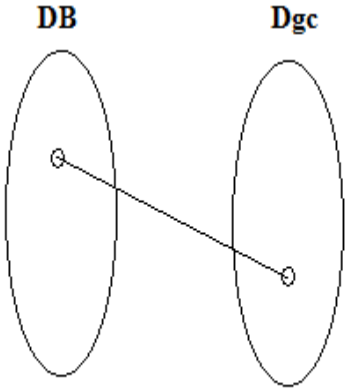


F2: javaPareser()
$F2 : j_i \rightarrow DB$

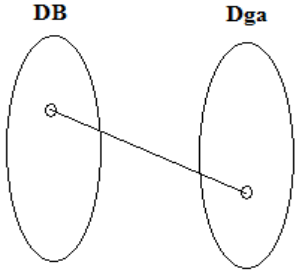
$j \rightarrow J$
DB= SQLite Database



F3: mapping_to_class_diagram()
F3 : (DB) → D_{gc}
DB→ Database File
D_{gc} → Class diagram



F4: mapping_to_act_diagram()
F4 : (DB) → D_{ga}
DB→ Database file
D_{ga} → activity diagram



Initial Condition :
$I \neq \{\phi\}$

Success Cases:
$S_c = \{S_{c1} \wedge S_{c2} \wedge S_{c3} \wedge S_{c4}\}$

S_{c1} → Class diagram generated correctly
S_{c2} → Sequence diagram generated correctly.
S_{c3}→ Successful parsing
S_{c4} →Successful database creation

Failure Cases:
$F_c = \{F_{c1}, F_{c2}, F_{c3}, F_{c4}\}$, $O= \phi$

F_{c1} → Input file is not java file
F_{c2} → Input file is not for java desktop application
F_{c3} → Java code having errors.
F_{c4} → Unhandled Java cases

VI. FUTURE SCOPE

Our future scope is to handle concurrency in activity diagram. Also this project can be used to support various OOP languages other than java.

VII. CONCLUSION

It is clear that our tool provides a clear view of class diagram and activity diagram. Our class diagram also includes dependency relation between various class which is not shown by various other tools. Our approaches analysis various suggested approach and implements a new approach for drawing class diagram and activity diagram.

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Design and Implementation of an Eight Bit Multiplier Using Twin Precision Technique and Baugh-Wooley Algorithm

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***Abstract*—** A novel technique for integer multiplication is implemented in this project. It is twin precision technique which is noteworthy for its low power dissipation. Multiplier is adapted to bit width of the operands to be computed to obtain the reduced power dissipation. The technique also results in an increased computational throughput, by allowing several narrow-width operations to be computed in parallel. Using Twin-precision technique with Baugh-Wooley algorithm, we achieve significant delay penalty and good power reduction.

Index Terms- twin precision, throughput, narrow-width operations.

I. INTRODUCTION

Multiplication is a complex arithmetic operation, which is reflected in its relatively high signal propagation delay, high power dissipation, and large area requirement. When choosing a multiplier for a digital system, the bit width of the multiplier is required to be at least as wide as the largest operand of the applications that are to be executed on that digital system. The bit width of the multiplier is, therefore, often much larger than the data represented inside the operands, which leads to unnecessarily high power dissipation and unnecessary long delay.

This resource waste could partially be remedied by having several multipliers, each with a specific bit width, and use the particular multiplier with the smallest bit width that is large enough to accommodate the current multiplication. Such a scheme would assure that a multiplication would be computed on a multiplier that has been optimized in terms of power and delay for that specific bit width. However, using several multipliers with different bit widths would not be an efficient solution, mainly because of the huge area overhead. It has been shown in many studies that more than 50% of the instructions are instructions where both operands are less than or equal to 16 bits. Such operations are called narrow-width operations.

This property has been explored to save power, through operand guarding. In operand guarding the most significant bits of the operands are not switched, thus power is saved in the arithmetic unit when multiple narrow-width operations are computed consecutively. It is shown that the power reduction of an operand-guarded integer unit was 54% to 58%, which accounts for a total power reduction of 5–6% for an entire data path. Narrow-width operands have also been used to increase instruction throughput, by computing several narrow-width operations in parallel on a full-width data path. It is showed a 7% speedup for a simple 4-bit ALU, which excluded the multiplier, in parallel with four simple 16-bit ALUs that share a 64-bit routing.

There have been several studies on operand guarding for multipliers. Two-dimensional operand guarding was introduced for array multipliers, resulting in a power dissipation that was only 33% of a conventional array multiplier. A similar investigation on a 32-bitWallace multiplier results in reduction of

the switching activity by 72% with the use of 16-bit operand guarding. While there has been a lot of work on simple schemes for operand guarding, work that simultaneously considers multiplication throughput is scarcer.

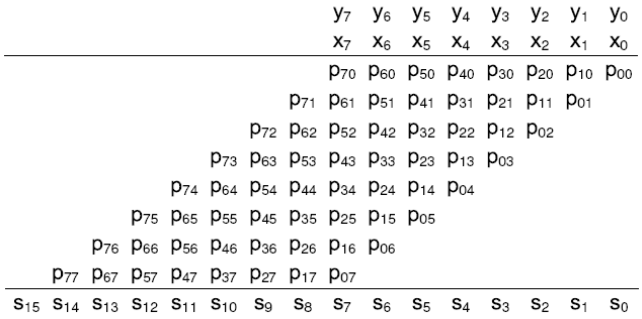


Figure 1: Illustration of an unsigned 8-bit multiplication

Achieving double throughput for a multiplier is not as straightforward as, for example, in an adder, where the carry chain can be cut at the appropriate place to achieve narrow-width additions. It is of course possible to use several multipliers, where at least two have narrow bit width, and let them share the same routing, but such a scheme has several drawbacks: The total area of the multipliers would increase, since several multiplier units are used. The use of several multipliers increases the fan out of the signals that drive the inputs of the multipliers. Higher fan out means longer delays and/or higher power dissipation.

There would be a need for multiplexers that connect the active multiplier(s) to the result route. These multiplexers would be in the critical path, increasing total delay as well as power dissipation. Work has been done to use 4:2-reduction stages to combine small tree multipliers into larger multipliers. This can be done in several stages, creating a larger multiplier out of smaller for each extra 4:2 reduction stage. The desired bit width of the multiplication is then obtained by using multiplexers, which has a negative effect on the delay.

We present the twin-precision technique that offers the same power reduction as operand guarding and the possibility of performing double-throughput multiplications. The twin-precision technique is an efficient way of achieving double throughput in a multiplier with low area overhead and with only a small delay penalty. We show how to apply the twin-precision technique on a un signed multipliers based on the regular High Performance Multiplier reduction tree. Baugh–Wooley algorithm is used for the multiplication.

II. PROPOSED TECHNIQUE

A. Twin-Precision Using the Baugh–Wooley Algorithm

TWIN-PRECISION FUNDAMENTALS:

We present the twin-precision technique using an illustration of unsigned binary multiplication. In an unsigned binary multiplication each bit of one of the operands, called the multiplier, is multiplied with the second operand, called multiplicand.

$$p_{ij} = y_i x_j$$

That way one row of partial products is generated. Each row of partial products is shifted according to the position of the bit of the multiplier, forming what is commonly called the partial- product array. Finally, partial products that are in the same column are summed together, forming the final result. An illustration of an 8-bit multiplication is shown in Fig. 1. Let us look at what happens when the precision of the operands is smaller than the multiplier we intend to use. In this case, the most significant bits of the operands will only contain zeros, thus large parts of the partial-product array will consist of zeros. Further, the summation of the most significant part of the partial-product array and the most significant bits of the final result will only consist of zeros. An illustration of an 8-bit multiplication, where the precision of the operands is four bits, is shown in Figure 2.

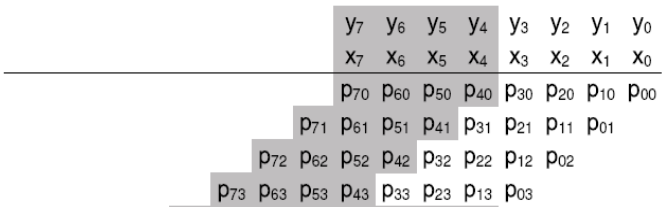


Figure 2: Illustration of an unsigned 8-bit multiplication where the precision of the operands is smaller than the precision of the multiplication. Unused bits of operands and product, as well as unused partial products, are shown in gray.

Figure 2 shows that large parts of the partial products are only containing zeros and are, thus, not contributing with any useful information for the final result. What if these partial products could be utilized for a second, concurrent multiplication? Since partial products of the same column are summed together, it would not be wise to use any of the partial products that are in the same column as the multiplication that is already computed. Looking closer at the 4-bit multiplication marked in white in Fig. 2, one can also observe that the column at position S7 should not be used either. This is because that column might have a carry from the active part of the partial-product array that will constitute the final S7. Altogether this makes only the partial products in the most significant part of the partial-product array available for a second multiplication.

In order to be able to use the partial products in the most significant part, there has to be a way of setting their values. For this we can use the most significant bits of the operands, since these are not carrying any useful information. By setting the other partial products to zero, it is then possible to perform two multiplications within the same partial-product array, without changing the way the summation of the partial-product array is done. How the partial products, shown in gray, can be set to zero will be investigated in the implementation section later on.

Assume, for now, that there is a way of setting unwanted partial products to zero, then it suddenly becomes possible to partition the multiplier into two smaller multipliers that can compute multiplications in parallel. In the above illustrations the two smaller multiplications have been chosen such that they are of equal size.

This is not necessary for the technique to work. Any size of the two smaller multiplications can be chosen, as long as the precision of the two smaller multiplications together are equal or smaller than the full precision (NFULL) of the multiplication, To be able to distinguish between the two smaller multiplications, they are referred to as the multiplication in the least Significant Part (LSP) of the partial-product array with size NLSP , shown in white, and the multiplication in the Most Significant part (MSP) with size MSP , shown in black.

NFULL = NLSP + NMSP

It is functionally possible to partition the multiplier into even more multiplications. For example, it would be possible to partition a 64-bit multiplier into four 16-bit multiplications. Given a number K of low precision multiplications their total size need to be smaller or equal to the full precision multiplication.

$$N_{FULL} \geq \sum_{i=1}^{K} N_i$$

For the rest of this investigation, the precision of the two smaller multiplications will be equal and half the precision (N=2) of the full precision (N) of the multiplier.

BAUGH WOOLEY ALGORITHMS:

Fig. 3 illustrates the flow chart for the algorithm of an 8-bit case, where the partial-product array has been reorganized according to the scheme of Hatamian. The creation of the reorganized partial-product array comprises three steps:

- The most significant partial product of the first N^{th} rows and the last row of partial products except the most significant have to be negated,
- A constant one is added to the n^{th} column,
- The most significant bit (MSB) of the final result is negated.

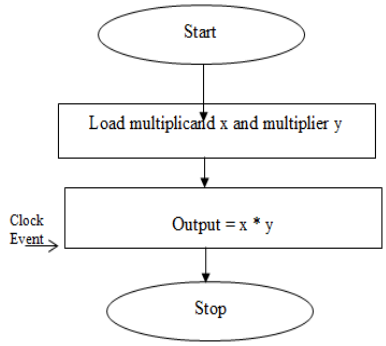
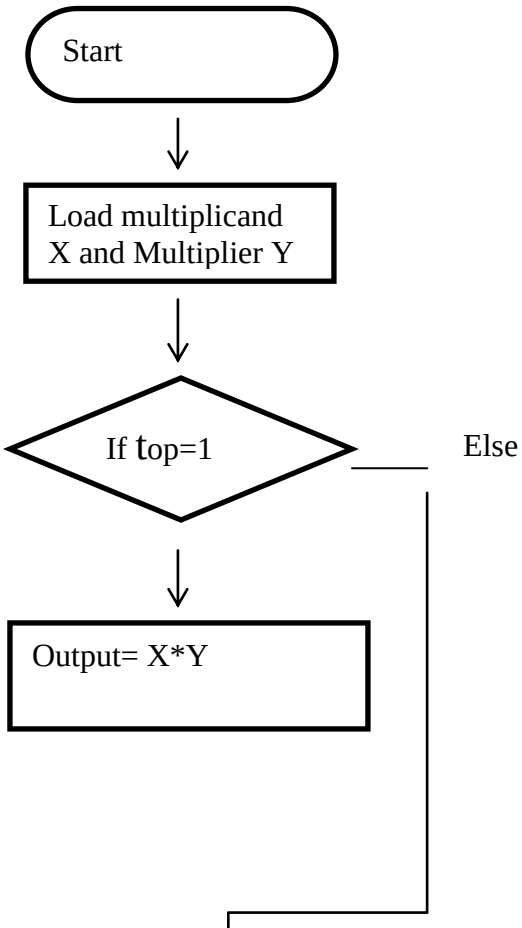


Figure 3: Flowchart for Baugh-Wooley algorithm

TWIN PRECISION USING THE BAUGH WOOLEY ALGORITHM:



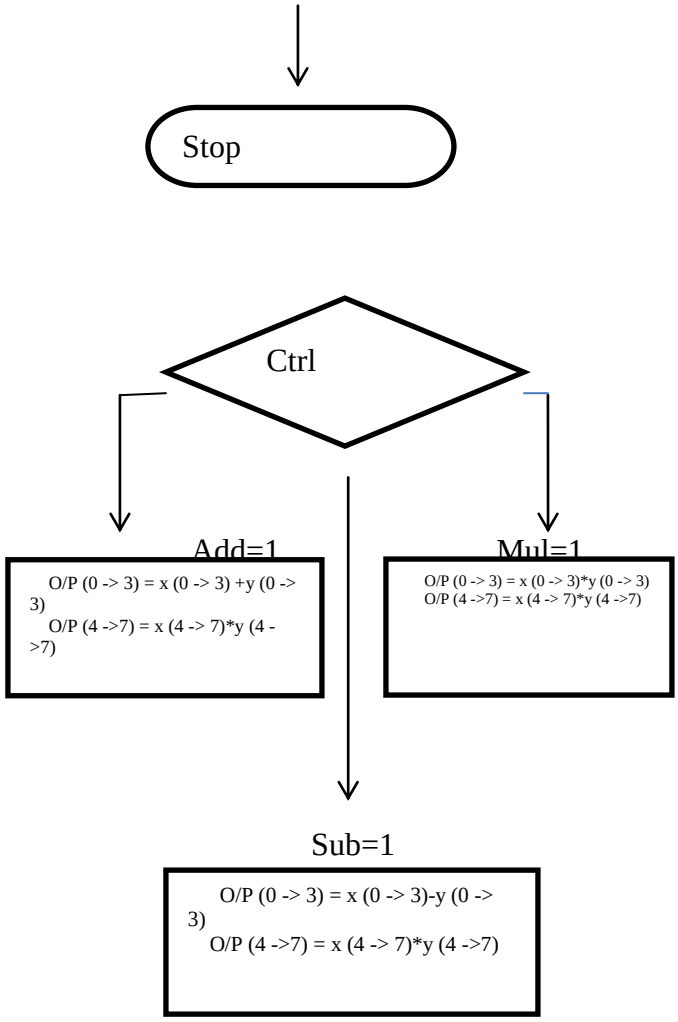


Figure 4: Flowchart for Twin precision Baugh-Wooley algorithm.

III. IMPLEMENTATION

It is not as easy to deploy the twin-precision technique onto a Baugh Wooley multiplication as it is for the unsigned multiplication, where only parts of the partial products need to be set to zero. To be able to compute two signed N/2 multiplications, it is necessary to make a more sophisticated modification of the partial-product array. Figure 5 illustrates an 8 bit Baugh-Wooley multiplication, in which two 4-bit multiplications have been depicted in white and black.

When comparing the illustration of Figure 5 with that of Figure 7, one can see that the only modification needed to compute the 4-bit multiplication in the MSP of the array is an extra sign bit '1' in column S12 . For the 4-bit multiplication in the LSP of the array, there is a need for some more modifications. In the active partial-product array of the 4-bit LSP multiplication (Shown in white), the most significant partial product of all rows, except the last, needs to be negated.

For the last row it is the opposite, here all partial products, except the most significant, are negated. Also for this multiplication a sign bit '1' is needed, but this time in column S4. Finally the MSB of the results needs to be negated to get the correct result of the two 4-bit multiplications. To allow for the full-precision multiplication of size to N coexist with two multiplications of N/2 size in the same multiplier, it is necessary to modify the partial-product generation and the reduction tree.

For the N/2 -bit multiplication in the MSP of the array all that is needed is to add a control signal that can be set to high, when the N/2-bit multiplication is to be computed and to low, when the full precision N multiplication is to be computed. To compute the N/2-bit multiplication in the LSP of the array, certain partial products need to be negated. This can easily be accomplished by changing the two-input AND gate that generates the partial product to a two-input NAND gate followed by an XOR gate.

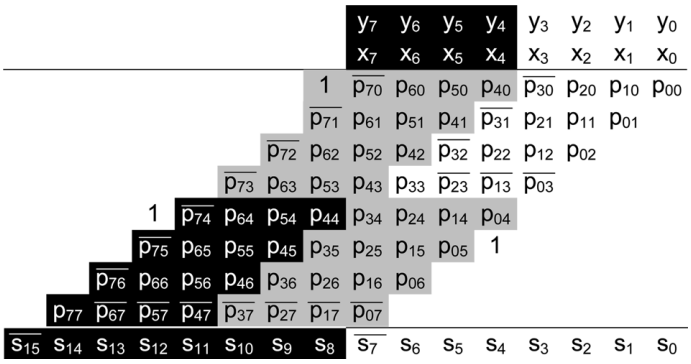


Figure 5: Illustration of a signed 8-bit multiplication, using the Baugh–Wooley algorithm, where one 4-bit multiplication, shown in white, is computed in parallel with a second 4-bit multiplication, shown in black**.**

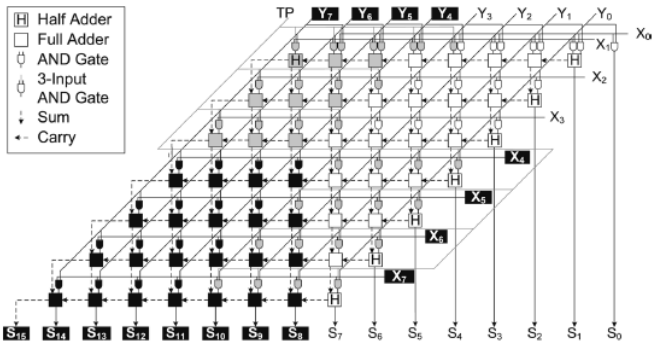


Figure 6: Block diagram of an unsigned 8-bit twin precision multiplier.

When computing the N/2 -bit LSP multiplication, the control input to the XOR gate is set to low making it work as a buffer. When computing a full-precision N multiplication the same signal is set to high making the XOR work as an inverter.

Finally the MSB of the result needs to be negated and this can again be achieved by using an XOR gate together with an inverted version of the control signal for the XOR gates used in the partial-product generation. Setting unwanted partial products to zero can be done by three-input AND gates as for the unsigned case.

Figure 6 shows an implementation of a twin-precision 8-bit Baugh-Wooley multiplier and it consist of three things:

- The half adders in column 4 and 8 have been changed to full adders in order to fit the extra sign bits that are needed
- For the sign bit of the 4-bit MSP multiplication there is no half adder that can be changed in column 12, so here an extra half adder has been added, which makes it necessary to also add half adders for the following columns of higher precision.

- And finally XOR gates have been added at the output of column 7 and 15 so that they can be inverted. The simplicity of the twin-precision BW implementation makes it easy to also compute unsigned multiplications. All that is needed is to set the control signals accordingly, such that none of the partial products are negated, the XOR gates are set to not negate the final result and all the sign bits are set to zero.

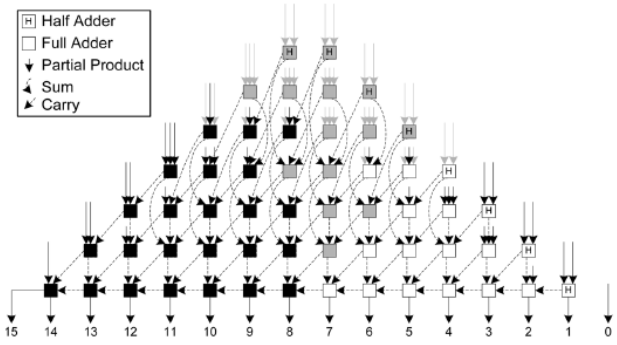


Figure 7: Block diagram of a signed 8-bit multiplication, using the Baugh–Wooley algorithm, where one 4-bit multiplication, shown in white, is computed in parallel with a second 4-bit multiplication, shown in black.

IV. SIMULATION RESULTS

TWIN BAUGH WOOLEY RESULT:

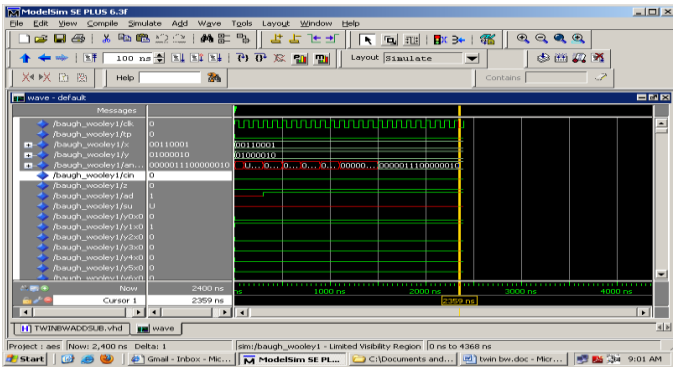


Figure 8: Simulation results for Twin precision Baugh Wooley algorithm.

SYNTHESIS RESULTS:

Following results are achieved in the synthesis of our process and showed the significant reduction in the power dissipation in the proposed multiplier.

Techniques	Execution Time	Power Consumption
Existing	7.675ns	16mW
Proposed	6.788ns	11mW

V. CONCLUSION

The twin precision multiplier presented in this paper offers a good tradeoff between precision flexibility, area, delay and power dissipation by using same multiplier for doing N, N/2 or two N/2-b multiplications. In comparison to a conventional 16-b twin precision multiplier has 8% higher transistor count and 9% longer delay. The relative transistor count overhead decreases for larger multipliers, since the number of AND gates needed to set the partial products to zero does not grow as fast as the number of adders in the tree.

We have shown that power cut-off techniques can be deployed in different regions of a twin precision functional unit, so that static leakage reduction can be effected not only when the entire unit is idle, but also when only parts of the unitary active, i.e. when the unit operates in half precision mode

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Application of Fuzzy Logic for Analysis of Vague Relational Database

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Abstract- The purpose of the proposed paper is that analysis of uncertainty and impression handling in fuzzy Relational Data Base by defining the fuzzy relationships between the existing database model to Fuzzy Vague Relational Database Model (FRDBMS) with the help of fuzzy membership function for analysis of the degree of uncertain information of existing data base. Some theoretical properties of the model can also be defined by the Slandered Relational Database Model (SRDBMS).

Index Terms- Crisp set, Fuzzy set, fuzzy subset, membership function, vague, attributes, set, fuzzy Relationship, data, metadata, logical adjunct, logical proximity, data security.

I. HISTORY

The present is often more meaning full when we have a better understanding of the past. Indeed many historians believe that one of the main characteristics of all the progressive civilization is their ability to produce and use of information effectively. In the Mesopotamian valley, civilization flourished as for back as 4500BC. An interesting point is that these civilizations kept fairly supplicated records on the clay tablet of various sizes well as shapes. These storage devices provided a great deal of information about receipts disbursement, loans, purchase and other transactions etc. The Egyptian was able to manage the complex pyramid building projects, because they also had advanced method of storing the data. More than 500 years ago the Inca Indians of south America developed the comprehensive information system, with data base and processing model composed the thousands of knotted strings some times called Queues i.e. quip is also called an accounting apparatus, an array of knots and different colors conveyed a combination of mnemonic, digits and narrative information the people who built these is some called **quipuamayus**. In the mid of eighteen century pressures to process and refinement of data is increased**.** The industry revaluated the basics of production from home to small shops. The development of large manufactures led to the development of service of markets and transportation of manufacturing units, the increased size and complexity of these originations made it impossible for any person to manage the above information to manage them effectively without enlisting the aid of data processing, moreover management needed more information for internal decision and more successfully running the origination.

II. INTRODUCTION

General data base were designed under a common assumption and based on the requirement with the various fields which is stored in a data base and the database is reliable. For example programmer used some numeric field then all information record in the required field, it ignore the facts with out comparing with the others, so some of the data of that field is not reliable for one user and rest of the data is reliable for others which forces the uncertainty i.e., indicate to the fuzzy approach for analysis and measuring the various applications of data and their reliability also. Form Codd introduce a Relational data base model by which he proven a useful model in information management system and applied in the wide spectrum application the inherent of the relational model is effective for precise and unambiguous data real world application, Dubois uncertainty and imprecision be the complementary aspect of imperfect information, which refers to degree of satisfactory and other trueness of data and. By the development of may be operation in the RDBMS, some researchers taken two kind of imprecise information null value and disjunctive information's in the RDBMS, generally the null value applicable (resp inapplicable) and disjunctive information is either inclusive or disjunctive function. How-ever these uncertain information arrives from Relational Database, some researchers gives mathematical model for defining the uncertainty by using the fuzzy approach i.e., fuzzy set theory and fuzzy probability theory. In the past decades fuzzy set technique have been used for modeling of uncertain database In 1982 Buckels and Petri proposed fuzzy Relational database for representing incorrect information in the real world problems in 1984, H.T. Parade defines a fuzzy database using possible distribution over the set of attribute domains. Since 1982 significant work has been done in inculpating the uncertainty management in relational database using fuzzy set theory. Many approaches found in the literature for processing the databases.

III. FUZZY LOGIC

Definition and design of Fuzzy Relational Maps (FRMs):In FCMs with the help of correlations between causal associations among concurrently active units. But in FRMs we divide the very causal associations into two disjoint units.

Definition 2.1: Let X be some set of objects, with elements noted as x. X = {x}.

Definition 2: A fuzzy set A in X is characterized by a membership function $\mu_A(x)$ which maps each point in X onto the

real interval [0, 1]. As $\mu_A(x)$ =1, the "grade of membership" or true membership function of x in A increases.

Definition2.7: Data are binary Computer representation of stored logical entities.

Definition 2.8: Index files and data dictionaries, store administrative information known ads meta data.

IV. Fuzzy Control Techniques

Various therapeutic situations are related to control problems. Although the early medical systems appeared at the same time as the article by Zadeh (1965), there has been little communication between the research fields, but recently this has changed due to the developments in computer systems, and rapid development of the literature searching methods motivated by the internet. Many systems are being developed which utilize fuzzy logic and fuzzy set theory.

Definition 3.1: A fuzzy relational R on a relational schema if A_i be the set of attributes and $R(A_1,A_2,\ldots\ldots,A_n)$ is the fuzzy sub set of the Cartesian product of Universe defined by $dom(A_1)\_dom(A_2)\_\ldots\ldots\ldots\ldots dom(A_n)$.According to the complexity of the $dom(A_i)$ the classical fuzzy relational can be classified to type1, fuzzy relational. In first type each attribute domain dom(Ai) can only crisp set or a fuzzy sub set so we can capture the impression of attributed value in a type 1 fuzzy relational allow each domain to be a crisp set, a fuzzy subset of fuzzy sets, and the second type is relation express the imprecision in the association among the attribute value.

Definition 3.2: Vague Set: Let U be the universe of discourse defined as U={u1,u2,……un} with a generic element of U denoted by ui, A Vague set is defined as a true membership function from universe to the interval defined as follows
μA:U [0,1] and a false membership function VA: U [0,1] where μ(Ai) (ui) is a lower bound on the grade of membership of ui derived by ui ,Vi(ui) is a lower bound on the negation of ui derived from the evidence against ui with μ(Ai) (ui) + VA (ui) 1 the grade of membership of ui in the Vague set A is bounded to a sub interval [μ(Ai) (ui) , 1- VA (ui)] of [0,1] . The Vague value [μ(Ai) (ui) , 1- VA (ui)] indicate the exact grade of membership ç A(ui) of ui may be unknown but bounded by [μ(Ai) (ui) çA"(ui) 1-VA(ui) with the property μ(Ai) (ui) + VA (ui) =1

General Vague Set:

Definition3.3: Let U be a nonempty set and A be a generalized set of U defined by $A=[\mu(A_i)_{(ui)}, 1- V(A)_{(ui)}]$ Where the membership function
μ_A:U→ [0,1] and false membership function VA: U→ [0,1] which satisfy the condition $[\mu(A_i)_{(ui)} =1- V(A)_{(ui)}]= 0.5$ for every $u_i \in U$ $\mu(A_i)_{(ui)}$ and $V(A)_{(ui)}$ are called the degree membership and non membership (u_i) to A.

Definition3.4:A generalized Vague relation from a non empty fuzzy X to Y is a sub set of generalized Vague sub set of X x Y than relation R be the Vauge relations defined by membership function R: X x Y→ [0,1] and false membership function VA: X x Y→ [0,1] where μ_R:(x,y) = 1- V_R (x,y)= 0.5 for all (x, y) ∈ {X x Y}.

Generalized Vague Relational database model:

A fuzzy relational R on a relational database is an extension of fuzzy relational database model proposed by Beaubouef and Petry . In this model a tuple it takes the form ($d_{i1},d_{i2},\ldots d_{im}$, $d_{i\mu},d_{iv}$) where dij be the domain is a domain value of a particular domain set D_j and $d_{i\mu},d_{iv} \in [o,1]$, the domain for truth membership values such that $d_{i\mu}$ d_{iv} =0.5. Let $P(D_i)$ denote any non-null value member of the power set of D_i.

Definition3.5: A generalized Vague relation R is a sub set of the set of cross proudest $P(D_1)$ x $P(D_2)$,…… $P(D_m)_x(1$ x 1). A generalized V ague tuple t is any member of R if ti is some arbiter tuple and defined as follows, ti= ($d_{i1},d_{i2},\ldots d_{im}$, $d_{i\mu},d_{iv}$) where d_{ij} D_j and $d_{i\mu}$,$d_{iv} \in 1$ such that $d_{i\mu}$ d_{iv} = 0.5.

Definition3.6: An interpretation of á= (a_i, a_i a_i……. a_m, a_μ a_v) of a vague rough tuple t_i= (d_{i1}, d_{i1} d_{i1} d_{i1}…….. $d_{i\,m}$, $d_{i\mu}$ d_{iv}) is any assignment such that aj ª dij for all j. The interpretation space is the cross product { (D_1 x D_2…….x D_m)} x{ (1 x 1)}, but it is limited for a given relation, R be the set of those tuples which are valid according to the underlying semantics of R. In an ordinary relational database because domain value are atomic, there is one possible interpretation for each tuple t_i . Moreover, the interpretation of 't_i' is equivalent to the tuple t_i. In the vague rough relational database, this is not always the case.

Definition3.7: Tuples t_i= (d_{i1}, d_{i1} d_{i1} d_{i1}…….. $d_{i\,m}$, $d_{i\mu}$ d_{iv}) and t_k= (d_{ik1}, d_{ik2} d_{ik3} d_{ik4}…….. $d_{i\,km}$, $d_{ik\mu}$, d_{ikv}) are the redundant if [d_{ij}] =[d_{kj}] for all j=I,2,3…n. If the relation contains those tuples of a lower approximation ie. Those tuples having the truth membership value1 and false value 0 The interpretation of á operation f at tuple is unique. This follows from redundant tuples. In vague rough relation there are no redundant tuples the merging process used in relational database operations removes duplicate since duplicate are not allowed in the sets, the structure upon which the relational model is used.

V. Generalized Vague SQL

There has been some studies which was discuss the some topic of fuzzy SQL quarries in fuzzy database which only cater for true membership i.e. combination of true and false membership, now describe the VSQL as an extension of SQL,it is powerful enough and retrieve any set of items of any degree of vagueness.

Model for VSQL:

We consider the data vagueness can occur in both relations and quarry expression. Thus we develops GVSQL and allow users to formulate a wide range of vauge quarries that occur in different modes of interaction between data and the queries we classified GVSQL as follows

Crisp Data	Generalized vauge data
Crisp data conventional SQL	GVSQL conventional SQL
Crisp data GVSQL conventional SQL	Generalized vague data GVSQL

I.Crisp Data Conventional SQL:

The first mode concern only conventional region of SQL where data value of GVSQL are both crisp This model is same as classical relational Database model so it is downloadable version of SQL. For example the following table shows the classical version of relational in which the data are crisp and the quarry defined as follows.

Ex.1: Find the item whose price = Rs-20

Product ID	Price	Weight
1	10	50
2	20	100
3	20	150
4	50	200
5	80	350

Select*From Product where Price=Rs.-20

Thus answer is given by following table it can be classified that by the true membership function it gives the answer shown as by the following table and by applying the false membership function the data is not considered and defined by the table number 2

True Answer:

Product ID	Price	Weight	Output
3	20	100	T
4	20	150	T

False answer:

Product ID	Price	Weight	Output
1	10	50	F
4	50	200	F
5	80	350	F

2 Mode: Vague Data Conventional:- The second mode concern the scenario that the data values are vague but queries are conventional . we allow classical SQL according to the vague table to be formulated and represents by the following example i.e. vague relational database table weight and price given as follows

Product ID	Price	Weight
1	10 Light	50
2	20 Light Middle	[0.6,0.8]/100
3	20, [1,1]/20 +[0.5,0.6]/50	[0.5,0,9]/150
4	50 Middle	[0.8,0.9]/200
5	80 Heavy	[0.7,1]/350

3 The vague product relation R_2:
Find the product which are the equal to
Rs-20/- SELECT*FROM Product WHERE

Price =Rs. -20/- We first transform 20 in to the vague set [1,1]/20 and that determine the SEQ BETWEEN THE PRICE VALUES R2 (Also in the form of vague rough set)

Example:

Consider the topple of ID 4 in R_2 then by degree of similarity SEQ (t_4[Price],[1,1/20] = SEQ (middle, [1,1]/200) = 0.617 then the rank of the touple by this SEQ value is given by the following table

Product ID	Price	Weight	Rank	Output
1	20,[1,1]/20+[0.5,0.6]/50	[0.5,0.9]/150	0.967	T
2	10 Light	[1,1]/50	0,624	F
3	20 [light, middle]	[0.6,0.8]/100	0.624	F
4	50 middle	[0.8,0.9]/200	0.617	F

5	80 Heavy	[0.7,1]/350	0.551	F

4 ModeIII: Crisp Data VSQL:

The third mode concerns the scenario that the data values are crisp but SQL are vague it is defined by the following example

Example1:

Find the product which are "high (Max.) in Price" SELECT*FROM Product WHERE Price =high(Max.) We transform the value in to the Price value is R1 in to the vague rough set and determine the rough set SEQ high for the given table ID 2 in R1 We Obtain the SEQ (t2[Weigh]) heavy by the previous table its rank is 0.551

Product ID	Price	Weight	Output	Rank
5	80 heavy	[0.7,1]/350	T	0.551
4	50 middle	[0.8,0.9]/200	F	0.617
3	20[1,1]/20+[0.5,0.6]/50	[0.5,0.9]/150	F	0.967
2	20 [light ,middle]	[0.6,0.8]/100	F	
1	10 light	[1,1]/50	F	

Example2:

Similarly we can determine the product of heavy weight SELECT*FROM Product WHERE Price =high (Max.)

Product ID	Weight	Price	Output	Rank
5	80 heavy	[0.7,1]/350	T	1
4	50 middle	[0.8,0.9]/200	F	0.750
1	10 light	[1,1]/50	F	0.591
2	20 [light ,middle]	[0.6,0.8]/100	F	0.591
3	20[1.1]/20+[0.5,0.6]/50	[0.5,0.9]/150	F	0.578

VI. Conclusion

We try the model of imprecision vagueness and uncertainty in database through an extension of the relational model of data to the generalized vague relational database. The new model is formally defined. Finally the generalized vague relational database model is and generalized vague query language is easy to understand and to use also. In addition it is more accurate model the uncertainty of real world enterprise than do conventional database through the vague membership function and non membership values by defining the suitable condition between these values.

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Fuzzy Based Hybrid Car Model (A Prototype)

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Abstract- In the present work we try to introduce new prototype hybrid car model with the help of fuzzy controller. Renewable energy was used for transportation long before any other energy sources thousands year ago the human used only their own energy to get around like walking and gaudily running ,later they are using animals like horses and carts, camels, elephants etc but all the energy source is food only and the food energy were comes out from sun air and water these are the forms of renewable energy, as soon as requirement is increasing the new invention comes than the variety of various energy sources comes out and now a days we are all dependent on all those product .

In this we utilize the conventional hybrid technology like solar and wind. Because from the conventional battery model having lot of draw back like difficulty of charging, limited speed due to using various number of batteries the weight is increased and consuming the large space and not secure too and the structure is also large which also increase lot of problems on the road. There are lot of drawbacks of individual one like limited driving speed, more charging time, Some of the other drawbacks are fixed distance; due to lack of sufficient wind and sun energy it can't give appropriate output. Due to all the reason which indicates a new way to introduce a new proto type model can easily designed with the help of fuzzy controllers has to be implemented for rectifying all sorts of drawbacks . Throughout this work we utilize three combined conventional energy sources and try to introducing new sophisticated models which can work efficiently.

I. INTRODUCTION

The Concept of Energy Generally energy is often defined as "the capacity to produce work". A better definition of energy is "the capacity to induce a change in that which inherently resists change". This capacity represents a combination of an effort, expended in overcoming resistance to a particular type of change, with the change it produces. The combination is called energy. The effort involved is measured quantitatively defined as a "driving force" in thermodynamics. A driving force is a property which both causes and also controls the direction of change in another property. The quantitative value of this change is called a "displacement". Renewable energy was used for transportation long before any other energy source. For hundreds of thousands of years, humans used only their own energy to get around, like walking and running. Later, they learned to use animals to get around like riding horses, camels, donkeys, or even elephants! People and animals get their energy from food. Since the energy in food comes from sunlight, food is a form of renewable energy. We eat plants, and plants are sometimes called biomass. Biomass is plants or garbage that can be used for energy. Biomass can also be used to make fuels to power our cars. A few thousand years ago, people discovered that they could use the wind to get around. Wind is another type of renewable energy. But less than 200 years ago, people started using fossil fuels like coal and oil The Energy is one of the most vital needs for human survival on earth. We are dependent on one form of energy or the other for fulfilling our needs. One such form of energy is the energy from fossil fuels. But the main disadvantages of these fossil fuels are that they are not environmental friendly and they are exhaustible. To deal with these problems of fossil fuels, we need to look at the .The product of a driving force and its associated displacement always represents a quantity of energy, but in thermodynamics this quantity has meaning only in relation to a specifically defined system. In any quantitative application of thermodynamics it is always important to make a careful distinction between energy changes within a system or within its surroundings and energy in transition between them. Energy is the ability to make things happen, cause changes and carry out work .In daily life energy is all around us in many different forms. Light and Sound energy is travel through the air as waves .Heat is a form known as thermal energy .Object even have energy because of their places or position .This is called Potential energy because the gravity tries to pull it down . Matter contains chemical energy, in the links of bonds of atoms the bond needs energy to form and they release the energy when they are broken .We are using chemical energy in fuels such as petrol, Diesel etc. The bonds break as the fuel burns and release the heat .The human body needs energy to drive its life processes like heartbeat breathing and movement. Energy can be changed or converted from one form to another form .But it never destroyed or created, lost or gained. It is converted the amount stays the same .At the end of the process or event the total amount of energy is the same as at the beginning. The principle of energy conservation means the total amount of energy in the universe is always the same. A similar process of changing matter in to energy happen naturally in the sun, Sun is made mainly by hydrogen. Tremendous temperature and pressure at its centre squeeze or fuses together the nuclei of the atoms

1. To form the nucleolus of a helium atom.
2. Vast amount of energy are given off.
3. This emerges from the sun mainly as light and heat.

"Natural energy" signifies the energy that drives, or activates natural phenomena. The term "energy from natural resources conversion system," has two meanings. The first is scientific and technological as in the example of"mechanical energy conversion to electric energy has high efficiency, whereas

thermal energy can be converted to mechanical energy with somewhat limited efficiency." The second meaning is economical and political as in the example of the term "conversion system" used in the present case has the second connotation. Putting this in concrete terms, "the present systems of energy carriers may be converted to advanced carrier systems the hydrogen energy systems in the twenty-first century." Energy is defined as a capacity for motion, which is sometimes equivalent to work. Work is scientifically defined by the physical quantity that is given to a body when it is moved a distance by a force. There are two kinds of energy dynamic and static. There are several kinds of energy carriers in present energy systems for example, gasoline, kerosene, electricity, CNG gas, and LPG. The Natural energy conservation law holds among the conventional energies, kinetic energy, potential energy, electric energy, chemical energy, photon energy, and thermal energy. Let us consider following conversion systems are realized:

1. Chemical energy to burner (heat).
2. From heat to heat energy.
3. From mechanical energy to electric generator.

It should be noted that "energy is conserved," but the useful available energy decreases at every step of energy conversion. Therefore, we obtain the second main principle: "available energy decreases whenever it is converted." This principle is called the second law of thermodynamics the law of increasing entropy. This is why energy is not always ability of work as indicated above

1.2 Wind Energy:

Most of sunshine is absorbed in the atmosphere and is converted to heat. The temperatures are not uniform, so that the air streams occur to give rise of wind blowing. Thus, large part of solar energy falling the earth converted to wind energy. The total amount of the wind energies on the earth's surface are huge, however it is distributed so widely that the energy density is not large, except for hurricanes, typhoons, and tornados. The wind energy is not always available (intermittent) and has moderate density

1.3 WATER Energy:

Water is one of the most abundant and important substance known to man. It is present in the air as water vapor and in the ground in the form of underground streams .Surface water in the form of oceans ,rivers and lakes covers about three-quarters of the earth 's crust .It is essential constituent of all animal and vegetable matter. The first is that the sources of free energy are:

1. Electrical energy
2. Chemical energy
3. Photon energy. Any of these can activate the binding electrons of water molecules and separate hydrogen and oxygen.

Among these three sources, electric energy can be considered as unstable in the natural world, chemical energy is stored in fossil fuels, while photon energy is plentifully available in sunlight. Therefore, it seems reasonable to utilize solar photon energy to split water, just as nature uses it in the photosynthesis of plants. The second fact is that if higher temperatures are available, then less free energy is needed.

II. WHY FUZZY MODELING

Fuzzy logic is a useful tool when dealing with uncertainty within data. Unlike Boolean logic, fuzzy logic has the power to deal with data that contains partial truth and ambiguity. Human beings deal with uncertainty in their everyday language. Since fuzzy logic is designed specifically to ad in data that is uncertain, it is an ideal tool for decision making involving real world applications. Fuzzy logic and set theory aid in uncertainty within data. Likewise an expression with a membership value of 1 would imply that the statement is absolutely true. In this IF and THEN rule is applicable ,because if solar panel unable to produce the sufficient energy then wind energy provided the support it is controlled by a fuzzy controller.

III. KNOWLEDGE BASED SYSTEM

Developer of knowledge based system (KBS) taken view to the problem solving task after require the use of domain is specific rule or thumb or heuristics. They provide technique for representing and reasoning with knowledge in a verity of the scheme including production rule or most common representation and allow one to heuristic representative with is contained from the express the consequent hold when some incident is true. The collection of rule for the knowledge based system of application and reasoning mechanism is called the inference engine and it can adopt various control strategies for select and applying rule, including proceeding from known data towards goal (forward changing) from goal towards data(backward changing) as well as mix mode and conflict resolution strategy. There is no sub statistical experience with development of knowledge based system and use of supporting development tool. Methodology of Simulation: Simulation process consists following eight rules

1. Identify and clarify the define problem.
2. List of statement in the objective of the problem.
3. Construct an appropriate model of the given problem.
4. Ensure that model represent the real solution.
5. Make Experiment with the model is constructed.
6. Analysis the simulation activity, make changes in the samples.
7. Tabulate the various values of the Decision criteria and choose the best fit.
8. Models include some elements of uncertainty.

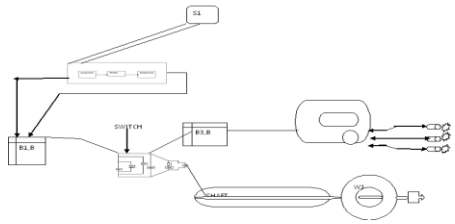
IV. SYSTEM DESIGN REQUIREMENT

This was divided following steps.

1. One prototype model
2. Three small wind mills.
3. Solar panel.
4. Batteries.
5. Two Sensors.

6. Two motors.
7. Switch.
8. Fuzzy Controller.
9. Wires
10. Triode, Capacitor, transistor
11. Connectors screws etc

4. Prototype Design:



Working Model Carbon Design

V. CONCLUSION

The Hybrid vehicle solves many problems related to the environment and is the best pollution free method. We need to make use of them so that we can reduce our dependence on fossil fuels. Hybrid vehicles do have some disadvantages like small speed range, initial cost is high. As this field of automobiles will be explored the problems will get solved. Prospective market and we should start using them in our day to day life. We have already completed making a solar vehicle prototype as our project and the vehicle is running successfully on solar and wind power .There are many hybrid cars available to buy today.

The proposed system is tested for practical purpose and found it having the following features through the testing design it is very easy in terms of installation, economically very cheap, comparatively very low in price no speed limits, no need for charging the batteries periodically.

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Study of Effectiveness of Using Concept Maps in Science among VI Grade Students

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Abstract- students of Faridabad district .The main findings of the study are:

(i) A significant difference has been found between the means scores of post test for the achievement test administered on the control group and experimental group both at 0.05 and 0.01 levels. The difference between the mean scores of the student s of the two groups are quite significant showing the effectiveness of the learning through Concepts Maps over Lecture method and proving that concepts Maps play positive role in enhancement of learning basic concepts of science.

(ii) It has been found that the correlation between Maps score and the scores on the maps related to multiple choice items on the unit test varied from 0.04 to 1, which inference that correlations are generally high. The present study was conducted to study the effectiveness of using concept maps in science among VI grade 100

Index Terms- Concept Map, Science, Effectiveness, Achievement

I. INTRODUCTION

Assessment of student achievement is changing largely because today's students' face a world that will demand new knowledge and abilities. In the present scientific and technological age, the conventional methods are not sufficient to arouse interest among the students and do not meet up to the intellectual, psychological and emotional needs of the students in new millennium. The method of teaching needs to be changed. The trend in education at present does not demand the text book learning alone. It aims at giving complete preparation to the pupils' .Since; science is a rapidly growing subject, its teaching demands continued re assessment and periodical review of the contents and the methods of teaching. Most new learning occurs through derivatives and correlative assumption of new concept meanings under existing concept or propositional frame work. Learning that is meaningful involves reorganization of existing beliefs or integration of new information with existing information. Cognitive structure is organized hierarchically, with new concept or concept meanings being subsumed under broader more inclusive concepts.

The theoretical frame work that supports the use of concept mapping is consistent with constructivist epistemology and cognitive psychology. Constructivism is a major influence in current science education.

Concept

Teaching concept in general science is not an easy task. It deals with the scientific objectives. The learning about the scientific objectives can be possible only by learning a concept. The general mental notion of things or events is arrived at by the students by processing of perception, classification and discrimination; it's used as a basis of thought and is expressed through symbolic language. The concept can be thought of as lower or higher order or as from the easily managed to the highly abstract. Concept formation is a process of developing concept, which was given by Piaget, Bruner etc. According to Gagne a concept is learned stimuli that belong to the same category or classification, but which can be done without applying such rule as "a chair must have four legs and a seat and a back." Recognition of the chair would be immediate upon presentation of an example just as rejection of a stool would be immediate and not subject to any logical application of rule.

A concept is defined as a class of stimuli or a coding system. According to Hung, Margin & Stone 1976, "A concept is a decision rule which, when applied to the description of an object specifies, whether or not a name can be applied."The concept centered teaching, a teacher – student interactive approach, following the constructivist approach, which helps in meaningful learning. It is a process of sequential instruction that applies theoretically, thousands of strategies for pupils understanding.

Constructivists make several assumptions about the learner, the context for learning, the learning process, and the outcomes. They believe that learning is a process of meaning making in which learners constantly explore their worlds, reflect upon their experience, articulate their thoughts and negotiate meaning with others. This process may start with a problem, a discrepant event or cognitive dissonance. On attempting to reconcile their conflicts, learners are able to create a shared ownership and meaning of the task, the concepts, the procedures and the strategies employed. Teaching is not regarded as a process of knowledge transmission. Rather, teachers should serve as a coach to facilitate learning and provide scaffolding and fading at appropriate junctures of the learning process.

Concept mapping fits well with the constructivist approach that learners "construct their own idiosyncratic understanding of concepts". The teacher can use a map as a basis for which to challenge student assumptions of how concepts are related maintain that most college faculty recognize that students will not remember specific facts from a course. What's more important is that students take away major themes or concepts and an understanding of how these concepts are related. Using a concept map to design a course can aid the teacher in guiding the

students to learn relevant concepts rather than trivial facts. Also, in knowing that students may perceive instruction differently from the way an educator intended, it can be helpful for the teacher to "construct a hypothetical model of the particular conceptual world of the students they are facing".

Concept mapping is a method to visualize the structure of knowledge. Science the knowledge expressed in the maps is mostly semantic. Concept maps are sometimes called semantic networks. Often it is claimed that concept mapping bears a similarity to the structure of long term memory. Instead of describing all concepts and their relation in text, one may choose to draw a map indicating concepts and relations in a graph or network. It the concept can be represented as eight era circle or a box .The relationship among concept and displayed using connecting link. Clarifying words and symbols are typically associated with linking lines to explain the specific relationships among the concepts.

II. NEED AND IMPOTRANCE OF THE STUDY

Knowledge of facts, terms and concepts is an important part of student's science knowledge in its own right and is also the basis for the development procedural knowledge while declarative knowledge may tend to be less valued.

The establishment of concept maps is a vehicle for assessing both declarative and procedural aspects of science knowledge. Concept maps monitor learning outcomes as instructional strategy emphasizes shift from basic to higher levels of performance. Concept mapping bears a similarity to the structure of long term memory Researches supports the idea that the use of concept mapping tool extend students learning in science and technology in important and unique ways. Concept maps may be most effective when use in science to stimulate and solve real problems, especially when scientific probes and technological tools are used to assist with lad experiments. Thus, in the present study investigator wants to study the comparative effectiveness of concept maps over traditional method

III. STATEMENT OF THE PROBLEM

"A Study of effectiveness of using concept maps in science among VI grade students"

OBJECTIVES OF THE STUDY

- To develop concept maps in science.
- To study the effect of using concept maps on achievement of VI grade student in science class.
- To study of effectiveness of using concept maps over traditional method in learning concepts of science among VI grade students.
- To study the correlation between concept maps and pupil's achievements.

HYPOTHESIS FORMULATED

- There is no difference between the two methods used traditional method and learning with concept maps.
- There is no difference between mean achievement scores in science of pre-test and post-test of VI class students.
- There is no difference between mean achievement scores in science of control group and experimental group.
- There is no relation in use of concept maps and pupil's achievement.

IV. VARIABLES

Independent Variable:

The independent variable that was used in the present study is **Concept Maps**. The variable was manipulated to study the effect on achievement in science. The experimental group was taught through **Concept Maps.**

Dependent Variable:

The dependent variable or the criterion variable that was use in the study is achievement in science. This variable may also be termed as moderator variable as they include the variable that could have a moderating affect on the treatment. The students were scored on this variable before and after the treatment in the group, this variable was measured twice during the study i.e. first before the beginning of the treatment (pre-test stage), then after completing the treatment. (Post-test stage)

Control variable:

Those factors which are controlled by the experimenter to cancel out or neutralize any effect they might, otherwise have on the observed phenomenon. In the present study nature of the school, grade level and administration are controlled.

V. DESIGN OF THE STUDY

The present study is of pre-test - post-test Equivalent Group of Experimental Group Design

RO_1 XO_2	**O_1**	**O_3**	=	**Pre-test**
RO_3 CO_4	**O_2**	**O_4**	=	**Post-test**

Here, the pre-tests are to be administered before the application of treatment on the Experimental and Control Group and Post –tests at the end of the treatment on both the groups.

Nature of the School

The sample was selected from Jagriti Vidhya Niketan School Faridabad, affiliated to C.B.S.E.

Grade level

Sixth class is selected for the present study and grade level is thus kept constant during the study.

Subject

The group was taught the one chapter of Science subject.

VI. SAMPLE

The present study was conducted on sample of **100** students (50 boys and 50 girls) studying in class VI of Jagriti Vidhya Niketan School.

VII. TOOL USED

(i)Concept maps (developed by the investigator)
(ii)Multiple choice test in science subject

VIII. STATISTICAL TECHNIQUES EMPLOYED

The statistical techniques employed for interpreting the data are:

1. Mean
2. Standard Deviation
3 critical ratio't'
4. Correlation (Karl Pearson)

IX. REVIEW OF THE RELATED LITERATURE

Kumar, Satheeshi (2010) Studied learning with multimedia a constructive co-operative approach in education. He found that they provide the students with ways to assemble knowledge constructively when students develop multimedia materials as a component of project based learning in cooperative groups.

Km, Ananta (2010) Studied the concept maps in teaching biology. He found that presenting the theoretical foundation and origin of concept maps graphical representation on the information and understanding.

Rai, Renu and Shweta (Aug2010) studied the concept mapping as a technology to facilitate teacher education programme and enhance learning. He found that new technology is capable of overcoming the barriers due to its importance and use in the field of education**.**

Antoine Laurent Laviosier (2009) Studied on concept mapping in elementary science. He found that it is essential that teachers use technology in a classroom to provide the richest education possible for our students.

David, Brown (2009) Studied the effect of individual and group concept mapping on students conceptual understanding of photosynthesis and cellular respiration in three different academic levels of biology class. He found that the constructing concept maps in small groups can significantly increase the knowledge gained by lower level students.

Hong Gao, E. Shen (2007) Studied on Collaborative Concept Mapping: What have we learned about the Technical and what is next. He observed that the concept mapping technique in combination with other instructional strategies will hopefully augment the positive result associated with the technique.

John, Marangos (2007) Studied effectiveness of concept mapping in economics: evidence from Australia & USA. He found that concept mapping was in corporate in the teaching material in both courses at different countries and, at the end of the semester; the students complicated a survey regarding the use, effectiveness and accessibility of concept maps.

Hong Gao, E Shen (2007) Studied collaborative concept mapping: An instructional strategy to foster both individual learning and group knowledge construction. He found that the collaborative concept mapping is a potential effective instructional strategy to facilitate learners in both group knowledge construction and individual learning.

Raghawan, Andal (1991) Studied concept mapping in learning physical science and its relation to scholastic performance, cognitive ability, attitude towards concept mapping and science interest among standard IX students. He found that the experimental and control groups of boys, girls and co-educational students were found to have no difference in post test scholastic performance scores in physical science.

Joshi, Anuradha (1993) conducted studies on concept Attainment Model on general ability. He found that the adjustment mean of the general mental ability scores of student taught through **CAM** was significantly different from those who were taught through traditional method when per general mental scores were taken as a covariate.

Table1.1INDEPENDENT DEPENDENT &CONTROL VARIABLES

Independent Variable	Dependent Variable	Control Variable	Control Employed
• **Concept Maps** (Instructional package by investigator & **Inspiration Programme**	• Achievement In Science	• Nature of the school • Grade level (only VI grade to be taught). • Administrative (one chapter of science subject in the group to be taught).	• Administrative (all schools working under C.B.S.E.) • Administrative. • Administrative (one chapter of science subject in the group to be taught).

X. DEVELOPMENT OF CONCEPT MAPS

There are 6 steps of creating Concept Map:-

1. Select key concepts: This is a recognition process that activates relevant knowledge and assists in topic identification.
2. Write the key concepts
3. Make an attribute list of the key concepts.
4. Relate key concepts in a special relationship.
5. Rearrange special representations.
6. Compare representation to the text.

XI. ANALYSIS AND INTERPRETATION OF THE DATA

Table 1.2 CALCULATION OF't' VALUE FOR THE CONTROL GROUP, PRE-TEST, POST- TEST

Mean, S.D, Std Error Mean and't' value of achievement in science of pre-test and post-test:

GROUP (Control)	M	S.D.	N	't' value	df	Remark
Pre-test	8.72	4.03	50	2.44	98	Significant at 0.05 level
Post-test	10.74	4.22	50			Significant at 0.01 level

From the Table 4.1, it is evident that the't'-value is 2.44 that is significant at 0.01 level that is 2.63 and at 0.05 level that is 1.98 with df = 98. It indicates that the mean scores of pre-test and post-test of control group differ significantly. Thus, **NULL** hypothesis that there is no significant difference between means scores of pre-test and post-test of control group is rejected. Further the mean scores of post-test of control group is as 10.74, which is significantly higher than the mean scores of pre-test as 8.72

Table 1.3 CALCULATION OF't' VALUE FOR THE EXPERIMENTAL GROUP, PRE-TEST, POST-TEST

GROUP (EXPERIMENTAL)	**M**	**S.D.**	**N**	**'t' value**	**df**	**Remark**
Pre-test	9.20	4.28	50	6.52	98	Significant at 0.05 level
Post-test	14.74	4.02	50			Significant at 0.01 level

From the table 4.2, it is evident that the't'-value is 6.52 that is significant at 0.01 level that is 2.63 and at .05 level that is 1.98 with df = 98. It indicates that the mean scores of pre-test and post-test of experimental group differ significantly. Thus, null hypothesis that there is no significant difference between means scores of pre-test and post-test of experimental group is rejected. Further the mean scores of post-test as 14.74, is significantly higher than the mean score of pre-test as 9.20.

Students' Concept Map Scores

12 concept maps were examined and student scores were recorded by using score. If one or more of the concepts were missing from the map, a score of 0 was given. To obtain a score other than 0 in this category, students had to have all essential stem and answer concepts somewhere on their maps. A positive score (+1) was given if the stem concept was linked accurately to the correct answer, and a negative score (-1) recorded if a linking error was made. The results of scoring for the maps of thirteen students are reported in Table 1.4

Conduction	0	+1	-1	+1	+1	0	+1	+1	+1	+1	+1	-1	-1	+1	-1	+1	-1	+1							

Table1.4 Students' Concept Map Scores

Concepts	S1	S2	S3	S4	S5	S6	S7	S8	S9	S10	S11	S12	S13	S14	S15	S16	S17	S18	S19	S20	S21	S22	S23	S24	S25
Heat	-1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	-1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1
Temperature	+1	+1	+1	-1	+1	+1	0	-1	+1	+1	0	+1	+1	+1	+1	+1	+1	+1	0	+1	0	-1	0	-1	+1
Unit of heat	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1
Unit of Temperature	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1
Kinetic Energy	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	0	+1	+1	+1	+1
Heat Energy	+1	0	-1	-1	+1	+1	+1	0	+1	+1	0	+1	+1	-1	+1	0	+1	-1	0	+1	+1	0	0	+1	+1
Electrical Energy	+1	-1	0	0	+1	0	+1	+1	0	0	+1	0	+1	0	+1	+1	-1	+1	+1	0	+1	-1	0	-1	-1
Heat Transfer	+1	0	-1	+1	+1	0	+1	+1	+1	+1	+1	+1	-1	-1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1
Convection	0	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	-1	+1	+1	+1	+1	+1	+1	-1	+1	+1	+1	+1
Conduction	+1	+1	-1	+1	+1	+1	-1	+1	+1	+1	+1	+1	+1	+1	-1	+1	+1	0	+1	+1	+1	+1	+1	0	+1
Radiation	+1	+1	+1	0	+1	+1	-1	+1	-1	+1	+1	+1	+1	+1	+1	-1	0	+1	+1	+1	+1	-1	+1	+1	-1
Reflection	+1	+1	+1	+1	+1	+1	+1	+1	+1	-1	-1	0	+1	+1	+1	+1	0	+1	-1	+1	+1	-1	-1	+1	+1
Matter	+1	+1	+1	+1	+1	+1	+1	0	+1	+1	-1	+1	+1	0	-1	+1	+1	-1	-1	0	+1	+1	-1	0	+1
Phase Changing	0	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	0	0	0	+1	+1	+1	+1	+1	-1	+1	+1	-1	+1
Boiling Point	+1	+1	+1	+1	+1	+1	+1	+1	0	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	-1
Expansion	-1	+1	0	+1	+1	+1	+1	+1	+1	0	+1	+1	+1	+1	+1	+1	-1	0	+1	+1	+1	0	0	+1	+1
Evaporation	+1	+1	-1	+1	+1	+1	0	-1	+1	+1	0	-1	+1	+1	+1	+1	+1	+1	0	-1	0	+1	0	-1	+1
Condensation	-1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	-1	+1	+1	+1	+1	+1	-1	+1	+1	+1	+1	0
Freezing	+1	+1	+1	0	0	+1	+1	+1	-1	+1	-1	+1	-1	-1	0	+1	-1	+1	-1	+1	+1	0	-1	+1	+1
Cooling	+1	-1	0	+1	+1	+1	-1	0	+1	+1	+1	+1	+1	+1	0	+1	0	+1	+1	+1	+1	+1	+1	+1	+1
Melting	-1	-1	0	+1	-1	+1	+1	+1	+1	+1	0	+1	+1	+1	+1	+1	+1	+1	0	0	+1	+1	+1	+1	+1
Fuel	0	+1	+1	0	+1	+1	+1	+1	+1	+1	-1	+1	+1	+1	+1	+1	+1	+1	-1	+1	+1	+1	-1	+1	+1
Total Score	15	17	13	15	19	20	17	17	18	19	14	20	19	14	16	20	16	18	14	17	17	15	13	16	18
%	68	77	59	68	86	91	77	77	82	86	64	91	86	64	73	91	73	82	64	77	77	68	59	73	82
Average														76											

+1: Correct -1: Incorrect 0: Not on the Map S1: Student 1 Contd…

Radiation	+1	+1	+1	+1	+1	+1	+1	0	+1	-1	+1	-1	+1	-1	0	+1	+1	+1							
Reflection	+1	-1	+1	+1	-1	0	+1	+1	-1	-1	+1	-1	+1	+1	+1	+1	+1	+1							
Matter	+1	0	-1	+1	+1	-1	0	+1	-1	+1	+1	+1	-1	+1	+1	+1	+1	+1							
Phase Changing	+1	+1	0	-1	+1	+1	+1	+1	+1	+1	+1	-1	-1	+1	0	-1	+1	+1							
Boiling Point	+1	+1	+1	+1	-1	+1	+1	-1	+1	+1	-1	+1	+1	+1	+1	+1	+1	+1							
Expansion	0	+1	+1	+1	+1	0	+1	+1	0	0	+1	0	+1	+1	+1	-1	+1	+1							
Evaporation	+1	+1	+1	+1	+1	+1	-1	+1	0	+1	+1	0	+1	-1	-1	+1	-1	+1							
Condensation	+1	+1	+1	+1	0	+1	+1	+1	+1	+1	+1	0	+1	+1	+1	+1	+1	+1							
Freezing	-1	+1	0	+1	+1	+1	-1	0	-1	0	+1	+1	+1	+1	+1	+1	+1	+1							
Cooling	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1							
Melting	0	+1	+1	+1	+1	+1	+1	-1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1							
Fuel	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1							
Total Score	-1	-1	-1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1							
%	73	82	73	86	73	64	73	73	59	68	59	50	73	73	68	82	82	95							
Average																									

+1 : Correct -1 : Incorrect 0 : Not on the Map S26 : Student 26

S29	S30	S31	S32	S33	S34	S35	S36	S37	S38	S39	S40	S41	S42	S43	S44	S45	S46	S47	S48	S49	S50
+1	0	-1	+1	+1	+1	+1	0	+1	+1	0	+1	+1	+1	+1	-1	+1	-1	0	+1	+1	+1
+1	-1	+1	0	+1	0	-1	-1	0	+1	+1	+1	+1	+1	+1	+1	0	+1	-1	+1	0	+1
+1	+1	0	+1	+1	+1	+1	+1	+1	-1	-1	+1	+1	+1	+1	+1	+1	+1	+1	-1	+1	+1
0	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	-1	+1	+1	+1	+1	+1	+1	+1	+1	+1
+1	+1	-1	-1	0	+1	+1	0	+1	0	+1	+1	0	+1	+1	+1	0	+1	+1	+1	+1	+1
+1	+1	+1	+1	+1	0	0	+1	0	+1	0	0	-1	+1	+1	+1	+1	0	+1	0	-1	0
-1	0	-1	+1	+1	0	-1	+1	0	-1	+1	+1	+1	+1	+1	+1	-1	+1	+1	+1	+1	+1
+1	+1	+1	0	-1	+1	+1	0	+1	+1	+1	+1	+1	+1	-1	+1	+1	+1	0	0	+1	+1
+1	+1	+1	+1	+1	+1	+1	+1	+1	+1	-1	+1	+1	+1	+1	+1	+1	+1	0	-1	+1	-1

Table 1.5 Comparison of answers of related multiple-choice items with concept map answer

	1		4		5		6		7		8		9		10		11		12	
	MC	CM	MC	CM	MC	CM	MC	CM	MC	CM	MC	CM	MC	CM	MC	CM	MC	CM	MC	CM
S1	1	+1	1	-1	1	+1	1	+1	1	-1	1	+1	0	-1	1	+1	0	-1	1	+1
S2	1	+1	1	+1	1	+1	1	+1	1	+1	0	0	0	+1	1	+1	1	+1	1	+1
S3	1	+1	0	+1	1	+1	1	+1	1	+1	1	+1	1	+1	0	0	1	+1	1	+1
S4	0	0	1	-1	1	0	1	0	0	-1	0	-1	1	+1	1	+1	1	+1	0	-1
S5	1	+1	1	0	1	+1	1	+1	1	0	1	0	1	0	1	+1	0	0	1	0
S6	0	+1	1	+1	0	+1	1	+1	1	+1	1	+1	1	+1	1	0	0	+1	1	+1
S7	1	0	1	+1	1	+1	1	+1	1	+1	0	+1	0	+1	1	+1	1	+1	1	+1
S8	1	+1	0	-1	1	0	1	0	0	-1	1	0	0	-1	0	0	1	-1	0	-1
S9	1	+1	1	+1	1	+1	1	+1	1	+1	1	+1	1	+1	1	+1	1	+1	1	+1
S10	1	-1	1	0	1	+1	1	+1	1	0	0	+1	1	0	1	+1	1	0	1	0
S11	1	+1	1	+1	1	+1	1	+1	1	+1	0	0	1	+1	1	+1	1	+1	1	+1
S12	0	+1	0	-1	1	0	1	0	1	-1	1	+1	1	-1	1	+1	1	-1	1	+1
S13	0	0	1	+1	0	+1	1	+1	1	+1	1	+1	0	+1	1	-1	1	+1	1	+1
S14	1	+1	0	+1	1	+1	1	+1	0	+1	1	+1	1	+1	1	+1	1	+1	1	+1
S15	1	-1	1	+1	1	-1	0	-1	1	+1	0	0	1	+1	1	+1	0	+1	1	+1
S16	1	+1	1	+1	1	+1	1	+1	1	+1	1	+1	0	+1	1	+1	0	+1	1	-1
S17	1	+1	1	+1	1	+1	1	+1	1	+1	1	+1	1	+1	1	-1	1	+1	1	+1
S18	1	-1	0	+1	1	+1	1	+1	1	+1	1	+1	1	+1	1	+1	1	+1	1	+1
S19	1	+1	1	0	1	-1	1	-1	0	0	0	-1	1	0	1	+1	1	0	0	0
S20	1	+1	0	-1	1	+1	1	+1	1	-1	1	+1	1	-1	1	+1	0	-1	1	+1
S21	0	+1	1	-1	1	+1	1	+1	1	-1	1	+1	1	-1	1	+1	1	-1	1	+1
S22	1	+1	1	-1	0	+1	1	+1	1	-1	1	+1	1	-1	1	+1	1	-1	0	+1
S23	1	+1	1	+1	1	+1	0	+1	0	+1	1	-1	1	+1	1	+1	1	+1	1	+1
S24	0	+1	1	+1	1	+1	1	+1	1	+1	1	+1	1	+1	1	+1	1	+1	1	+1
S25	0	+1	1	+1	1	+1	1	+1	1	+1	1	+1	1	+1	0	+1	1	+1	0	+1
Pearson r	0.4		0.4		0.9		0.5		1		0.4		0.4		0.9		0.4		0.4	

-1 0

Table 1.6Comparison of answers of related multiple-choice items with concept map answer

Map-Related Concepts

	1		2		3		4		5		6		7		8		9		10		11		12	
	MC	CM	MC	CM	MC	CM	MC	CM	MC	CM	MC	CM	MC	CM	MC	CM	MC	CM	MC	CM	MC	CM	MC	CM
S1	1	+1	0	0	1	+1	1	+1	1	-1	1	+1	1	+1	1	0	-1	+1	1	+1	0	-1	1	+1	1	+1	0	-1	0	-1
S2	1	0	1	+1	1	+1	1	0	-1	0	1	0	0	+1	1	1	+1	0	1	0	+1	0	0	1	0	0	+1	0	1	0	+1	1	+1	0	1	+1
S3	1	+1	1	+1	1	+1	1	1	0	+1	1	0	-1	0	+1	1	+1	-1	1	1	+1	-1	1	+1	1	+1	0	-1	0	0	1	+1	+1	1	1	+1
S4	1	+1	0	0	0	1	0	1	-1	-1	0	1	+1	-1	1	0	1	1	0	+1	0	1	-1	+1	0	0	+1	1	+1	1	+1	1	1	-1	+1	1	0	-1
S5	1	+1	1	+1	1	+1	1	1	+1	0	1	1	+1	0	1	1	+1	+1	1	1	0	+1	1	0	+1	1	0	+1	1	+1	0	1	0	+1	0	1	1	+1	0
S6	1	0	0	0	+1	1	0	0	+1	1	1	0	+1	0	1	+1	1	0	+1	1	+1	0	1	1	+1	0	1	+1	1	0	0	1	0	+1	+1	1	1	+1
S7	1	+1	1	0	1	0	1	0	+1	+1	0	1	0	+1	0	1	+1	0	1	0	+1	0	1	1	+1	1	0	+1	0	0	+1	1	0	+1	0	1	1	-1	+1	1	1	+1
S8	0	+1	1	+1	1	+1	1	1	+1	0	1	0	+1	-1	1	1	+1	0	1	0	0	+1	0	1	0	-1	+1	1	1	0	+1	1	0	-1	0	1	0	0	1	-1	-1	1	0	-1
S9	1	0	1	+1	1	+1	1	0	-1	+1	1	1	-1	+1	1	+1	-1	1	1	1	+1	-1	1	0	+1	0	-1	1	+1	0	-1	1	+1	+1	1	1	+1	+1	1	1	+1	1	+1
S10	1	-1	-1	1	+1	-1	0	1	+1	+1	1	1	+1	0	1	+1	+1	1	1	+1	+1	1	1	0	+1	0	+1	1	0	1	+1	1	1	+1	0	0	1	+1	0	1	0	1	+1	0
S11	0	+1	1	+1	1	+1	1	1	+1	0	0	1	+1	+1	0	1	+1	1	1	+1	+1	1	1	0	+1	0	0	+1	1	+1	1	1	+1	1	1	-1	+1	0	1	+1	1	1	+1	1
S12	0	+1	0	+1	0	+1	1	1	+1	+1	1	0	-1	+1	0	1	0	1	0	0	-1	1	1	-1	-1	1	+1	-1	1	-1	1	1	+1	1	1	1	-1	-1	1	1	+1	1	+1	1
S13	1	+1	0	0	1	0	1	1	-1	+1	1	1	+1	+1	0	1	+1	1	1	+1	+1	1	1	+1	+1	1	+1	1	0	+1	1	1	+1	1	1	+1	1	1	0	+1	1	1	+1	1
S14	1	0	1	+1	1	+1	1	1	+1	0	1	+1	1	+1	1	1	+1	+1	1	1	+1	+1	1	0	+1	+1	1	+1	1	+1	0	1	+1	1	1	-1	+1	1	1	+1	1	1	+1	1
S15	1	+1	1	-1	1	+1	1	0	1	+1	1	1	+1	-1	0	1	-1	+1	1	1	+1	+1	1	0	+1	0	1	+1	1	+1	1	0	0	-1	+1	1	1	+1	1	1	+1	1	+1	1
S16	1	+1	1	+1	1	+1	1	0	+1	+1	0	1	-1	+1	1	+1	-1	1	1	1	+1	-1	1	1	+1	-1	1	+1	1	+1	0	1	0	-1	+1	1	1	+1	1	1	+1	1	+1	1
S17	1	+1	1	0	1	+1	1	0	+1	+1	1	1	+1	+1	1	1	+1	+1	1	1	+1	+1	1	1	+1	+1	1	-1	1	1	+1	1	+1	+1	0	1	1	+1	1	1	+1	1	+1	1
S18	1	+1	1	-1	1	-1	1	1	+1	+1	0	0	+1	+1	1	1	+1	+1	1	0	+1	+1	1	1	+1	+1	1	+1	1	+1	1	1	-1	+1	1	1	+1	1	1	+1	1	1	+1	1
S19	0	+1	1	+1	1	+1	1	1	-1	-1	1	1	+1	0	0	1	-1	+1	1	1	-1	+1	0	1	0	+1	0	1	+1	1	+1	0	1	+1	0	1	0	+1	0	1	0	1	+1	0
S20	0	-1	1	+1	1	+1	0	1	+1	+1	1	0	-1	+1	0	1	+1	-1	1	1	+1	-1	1	1	-1	0	1	+1	0	1	0	1	0	0	-1	1	1	+1	1	1	+1	1	+1	1
S21	1	+1	0	+1	1	+1	1	1	+1	+1	0	1	+1	-1	1	1	+1	+1	1	1	+1	+1	1	0	-1	+1	1	+1	1	+1	1	1	0	-1	0	1	0	+1	1	1	+1	1	+1	1
S22	1	-1	1	+1	1	+1	1	0	+1	+1	0	1	+1	-1	1	0	+1	+1	1	1	+1	+1	1	0	-1	+1	1	+1	1	+1	0	1	+1	-1	1	0	+1	+1	1	1	+1	1	+1	1
S23	1	+1	1	+1	0	1	0	1	0	-1	1	1	-1	+1	1	0	1	+1	-1	0	1	+1	-1	1	1	+1	1	1	0	+1	0	1	+1	+1	1	1	0	+1	1	1	+1	1	+1	1
S24	1	+1	0	+1	1	+1	1	1	+1	+1	0	1	+1	+1	1	0	+1	+1	1	1	+1	+1	1	1	+1	+1	1	+1	1	+1	0	1	+1	+1	1	1	+1	+1	1	1	+1	1	+1	1
S25	1	-1	0	+1	0	+1	1	0	+1	+1	1	1	+1	+1	1	0	+1	+1	1	1	+1	+1	1	0	+1	+1	1	+1	1	+1	0	1	+1	+1	1	1	+1	+1	1	1	+1	1	+1	1
Pearson	1	0.4		0.4		0.4		0.5		0.4		0.6		0.4		0.9		0.5		0.7		1		0.5		0.4		0.7		0.4		0.5		0.9		0.4		0.4		0.4		

FINDINGS

- It has been found that the post-test achievement scores of students of control group were significantly higher than their pretest achievement scores.
- It has been found that the post-test achievement scores of students of experimental group were higher than their pretest achievement scores.
- A significant difference has been found between the means scores of post test for the achievement test administered on the control group and experimental group both at 0.05 and 0.01 levels. The difference between the mean scores of the students of the two groups are quite significant showing the effectiveness of the learning through Concepts Maps over Lecture method and proving that concepts Maps play positive role in enhancement of learning basic concepts of science.
- It has been found that the correlation between Maps score and the scores on the maps related to multiple choice items on the unit test varied from 0.04 to 1, which inference that correlations are generally high.
- It has been found that there is a significant difference between the mean scores of pretest and posttest for the achievement test administered on the experimental group is significant at 0.01 levels.

XII. CONCLUSION

The result revealed that the experimental group performed better than the control group in all achievement areas i.e. overall, by levels of cognitive domain and by the type of content .Students liked the Concept Maps program and benefitted from it. They found it to be better mode of instruction than the conventional method i.e. the lecture method.

XIII. EDUCATIONAL IMPLICATIONS

The present study has the implications for the science teachers and teaching of science.

- If the teaching is supplemented with Concept maps, it can prove to be more effective in enhancing achievement.
- Quality computer which include colorful animation, graphics display from a versatile and effective alternative change in instructional strategy.
- Concept Maps helps in professional development of teaching and learning and individuals involved in the programs for teacher education. It can be infused in the learning process so as to acquire the knowledge and skill efficiently.
- Using concept map as a tool in science classes will help student to develop better understanding of the important concepts, developing interrelationship and creating meaning schemes and constructing knowledge bases.
- The teachers to focus on making connections between facts and fostering new understanding in students by allowing , inspiring and guiding their students on their learning path.
- For leading the students on the path of acquisition of new learning, solving difficult and complex problem or discovering the facts related to a concept, it is quite advantageous to put students in situations that might challenge their previous conceptions and creating contradictions to encourage them on discovery path for their individual efforts or engage in discussion with their individual efforts or engage in discussion with their peers, teacher or elders.
- Since students have to construct their own meaning and discover the knowledge on their own, constructivist learning outcomes does not fit in the traditional evaluation pattern .It calls for flexible, comprehensive and continuous program to the constructive and create output of the learners for providing timely assessment, motivation and assistance to the learners in getting them engaged in the task of constructing and discovering the knowledge by themselves.

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New Subclasses of Meromorphically Multivalent Functions Defined by a Differential Operator

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***Abstract*-**In this paper authors introduced two new subclasses $\sum_{\square\square\square\square\square\square}(\square,\square,\square)$ and $\sum^{+}_{\square\square\square\square\square\square}(\square,\square,\square)$ of meromorphically multivalent functions which are defined by means of a new differential operator. By making use of the principle of differential subordination, authors investigate several inclusion relationships and properties of certain subclasses which are defined here by means of a differential operator. Some results connected to subordination properties, coefficient estimates, convolution properties, integral representation, distortion theorems are obtained. We also extend the familiar concept of $(\square,\square)$ −neighborhoods of analytic functions to these subclasses of meromorphically multivalent functions.

***Index Terms*-** Analytic functions, meromorphic functions, multivalent functions, differential operator, subordination, neighborhoods.

I. INTRODUCTION

Let $\square$ be the class of analytic functions in the unit disk $\square = \{\square \in \square: |\square| < 1\}$.

Consider

$$\square = \square\square \in: \square(0) = 0 \square\square\square |\square(\square)| < 1, \square \in \square \quad (1.1)$$

the class of Schwarz functions. For $0 \leq \square < 1$ let

$$\square(\square) = \square \in \square: \square(0) = 1 \text{ and } \square\square\square(\square) > \square, \square \in \square. \quad (1.2)$$

Note that $\square = \square(0)$ is the well-known Caratheodory class of functions. The classes of Schwarz and Caratheodory functions play an extremely important role in the theory of analytic functions and have been studied by many authors. It is easy to see that $\square \in \square(\square)$ if and only if $\frac{\square(\square)-\propto}{1-\propto} \in \square$. (1.3)

Using the properties of functions in the class P and the above condition, the following properties of the functions in the class $P(\alpha)$ can be obtained.

Lemma 1.1 Let $\square \in \square$. Then $\square \in \square(\square)$ if and only if there exists $\square \in \square$ such that $\square(\square) = \frac{1-(4(1-\square)\propto-1)\square(\square)}{1-\square(\square)}$ (1.4)

Lemma 1.2 (Herglotz formula) A function $\square \in \square$ belongs to the class $\square(\square)$ if and only if there exists a probability measure $\square(\square)$ on $\square\square$ such that

$$\square(\square) = \int_{|\square|} \frac{1-(4(1-\square)\propto-1)\square(\square)}{1-\square(\square)} \square\square(\square) \quad (\square \in \square). \quad (1.5)$$

The correspondence between $\square(\square)$ and probability measure $\square(\square)$ on $\square\square$, given by (1.5) is one-to-one.

If $\square$ and $\square$ are in $\square$. We say that f is subordinate to $\square$, written $f \in g$, if there exists a function $\square \in \square$ such that

$$\square(\square) = \square(\square(\square)) \quad (\square \in \square).$$

It is known that if $f \in g$,

then $\square(0) = \square(0)$ and $\square(\square) \subset \square(\square)$.

In particular, if $\square$ is univalent in U we have the following equivalence: $\square(\square) \prec \square(\square)$ $(z \in U)$ if and only if $\square(0) = \square(0)$ and $\square(\square) \subset \square(\square)$.

Let $\sum_{\square}$ denote the class of all meromorphic functions f of the form $\square(\square) = \square^{-\square} + \sum_{\square=1-\square}^{\infty} \square_{\square} \square^{\square}$ $(\square_{\square} \geq 0, \square \in \square)$ (1.6)

which are analytic and p-valent in the punctured unit disk $\square^{\ast} = \square\backslash\{0\}$.

Denote by $\sum_{\square}^{+}$ the subclass of $\sum_{\square}$ consisting of functions of the form $\square(\square) = \square^{-\square} + \sum_{\square=1-\square}^{\infty} \square_{\square} \square^{\square} \geq 0$ $(\square \in \square^{\ast})$ (1.7)

A function $\square \in \sum_{\square}$ is meromorphically multivalent starlike of order α $(0 \leq \alpha < 1)$ (see [2]) if $-\square\square \left\{\frac{1}{\square} \frac{\square\square'(\square)}{\square(\square)}\right\} > \propto$ $(\square \in \square)$.

The class of all such functions is denoted by $\sum_{\square}^{+}(\propto)$.

If $\square \in \sum_{\square}$ is given by (1.6) and $\square \in \sum_{\square}$ is given by $\square(\square) = \square^{-\square} + \sum_{\square=1-\square}^{\infty} \square_{\square} \square^{\square}$

then the Hadamard product (or convolution) of f and $\square$ is defined by $(\square * \square)(\square) = \square^{-\square} + \sum_{\square=1-\square}^{\infty} \square_{\square} \square_{\square} \square^{\square} \geq 0$ $(\square \in \square, \square \in \square^{\ast})$.

For a function $\square \in \square_{\square}$, we define the differential operator $\square^{\square}_{\square\square\square\square}$ in the following way:

$$\square^{0}_{\square\square\square\square}\square(\square) = \square(\square)$$

$$\square^{1}_{\square\square\square\square}\square(\square) = \square_{\square\square\square\square}\square(\square)$$

$$= (\square - \square)(\square - \square)\frac{[\square^{\square+1}\square(\square)]''}{\square^{\square-1}} + \frac{(\square-\square)-\square^{2}(\square-\square)}{\square}\frac{[\square^{\square+1}\square(\square)]'}{\square^{\square}} + \left[\frac{\square-(\square-\square)-\square^{2}(\square-\square)}{\square}\right]\square(\square). \quad (1.8)$$

And in general

$$\square^{\square}_{\square\square\square\square\square}\square(\square) = \square_{\square\square\square\square\square}\left[\square^{\square-1}_{\square\square\square\square\square}\square(\square)\right], \quad (1.9)$$

where, $\left(\begin{matrix} 0 < \varrho \le \frac{1}{2}, 0 \le \xi < 1, \tau \ge 1, 0 \le \varsigma < 1, 0 \le \eta < 1 \\ 0 \le \alpha < 1, 0 < \beta \le 1 \text{ and } m \in \mathrm{N} \end{matrix}\right)$.

If the function $\square \in \square_{\square}$ is given by (1.6) then, from (1.8) and (1.9), we obtain

$$\square^{\square}_{\square\square\square\square\square}\,\square(\square) = \square^{-\square} + \sum_{\square=1-\square}^{\infty} \square_{\square}(\square,\square,\square,\square,\square,\square)\square_{\square}\square^{\square}$$

(1.10) $(\square \in \square, \square \in \square, \square \in \square^{*})$, where

$$\square_{\square}(\square,\square,\square,\square,\square,\square) = \left\{1 + \left[(\square+\square)\left(\begin{matrix}\frac{(\square-\square)-\square^{2}(\square-\square)}{\square} \\ +(\square+\square+1)(\square-\square)(\square-\square)\end{matrix}\right)\right]\right\}^{\square}. \quad (1.11)$$

From (1.10) it follows that $\square^{\square}_{\square\square\square\square\square}\,\square(\square)$ can be written in terms of convolution as $\square^{\square}_{\square\square\square\,\square\square}\,\square(\square) = (\square * \square)(\square)$

(1.12) where

$$\square(\square) = \square^{-\square} + \sum_{\square=1-\square}^{\infty} \square_{\square}(\square,\square,\square,\square,\square,\square)\square^{\square}.$$

(1.13) Note that, the case $\square = \frac{1}{2}$ and $\square = \square$ of the differential operator $\square^{\square}_{\square\square\square\square,\square}$ was introduced by Srivastava and Patel [18]. Making use of the differential operator $\square^{\square}_{\square\square\square\square\square}\,\square(\square)$ for $p = 1$ was considered in [16]. For differential operator $\square^{\square}_{\square\square\square\square\square}\,\square(\square)$, we define a new subclass of the function class $\sum_{\square}$ as follows.

Definition 1.1 A function $\square \in \sum_{\square}$ is said to be in the class $\sum_{\square\square\square\square\square\square}(\square,\square,\square)$ if it satisfies the condition

$$\left|\frac{\frac{\square[\square\square^{\square}_{\square\square\square\square}\square(\square)]'}{\square\square^{\square}_{\square\square\square\square}\square(\square)} + 2\square(1-\square)}{\frac{\square[\square\square^{\square}_{\square\square\square\square}\square(\square)]'}{\square\square^{\square}_{\square\square\square\square}\square(\square)} + 8\square(1-\square)^{2}\square - 2\square(1-\square)}\right| < \square \text{ for some } z \in \mathrm{U}^{*}. \quad (1.14)$$

And

$\left(\begin{matrix} 0 < \square \le \frac{1}{2}, 0 \le \square < 1, \square \ge 1, 0 \le \square \le 1, \\ 0 \le \square < 1, 0 \le \alpha < 1, 0 < \square \le 1 \text{ and } \square \in \square \end{matrix}\right)$.

Note that a special case of the class $\sum_{\square\square\square\square\square\square}(\alpha,\square,\square)$ for $p = 1$ and $m = 0$ is the class of meromorphically starlike functions of order α and type $\square$ introduced earlier by Mogra et al. [12]. It is easy to check that for $m = 0$ and $\beta = 1$, the class $\sum_{\square\square\square\square\square\square}(\alpha,\square,\square)$ reduces to the class $\sum^{*}_{\square}(\alpha)$. We consider another subclass of $\sum_{\square}$ given by

$$\sum^{+}_{\square\square\square\square\square\square}(\alpha,\square,\square) = \sum^{+}_{\square\square\square\square\square\square}(\alpha,\square) \cap \sum_{\square\square\square\square\square\square}(\alpha,\square,\square).$$

(1.15) The main object of this paper is to present a systematic investigation of the classes $\sum_{\square\square\square\square\square\square}(\alpha,\square,\square)$ and $\sum^{*}_{\square}(\alpha,\square,\square)$.

II. MAIN RESULTS

2 Properties of the class $\sum_{\square\square\square\square\square\square}(\alpha,\square,\square)$

We begin this section with a necessary and sufficient condition, in terms of subordination, for a function to be in the class $\sum_{\square\square\square\square\square\square}(\alpha,\square,\square)$.

Theorem 2.1: A function $\square \in \sum_{\square}$ is in the class $\sum_{\square\square\square\square\square\square}(\alpha,\square,\square)$ if and only if

$$\frac{\square[\square\square^{\square}_{\square\square\square\square}\square(\square)]'}{\square\square^{\square}_{\square\square\square\square}\square(\square)} \prec \frac{\square(4(1-\square)\alpha-1)\square\square-\square}{1-\square\square} \quad (\square \in \square). \quad (2.1)$$

Proof: Let $\square \in \square_{\square\square\square\square\square\square}(\alpha,\square,\square)$. Then, from (1.6), we have

$$\left|\frac{-\frac{\square[\square\square^{\square}_{\square\square\square\square}\square(\square)]'}{\square\square^{\square}_{\square\square\square\square}\square(\square)} + 2\square(\square-1)}{-\frac{\square[\square\square^{\square}_{\square\square\square\square}\square(\square)]'}{\square\square^{\square}_{\square\square\square\square}\square(\square)} + 4\square(1-\square) + 2\square(\square-1)(4(1-\square)\alpha-1)}\right|^{2} < \square^{2}$$

or

$$\frac{(1-\square^{2})}{(-(4(1-\square)\alpha-1))^{2}-1}\left|-\frac{\square[\square\square^{\square}_{\square\square\square\square}\square(\square)]'}{\square\square^{\square}_{\square\square\square\square}\square(\square)} + 2\square(\square-1)\right|^{2}$$

$$-\frac{2\left[1+\square^{2}(-(4(1-\square)\alpha-1))\right]}{(-(4(1-\square)\alpha-1))^{2}-1}\square\square\left\{-\frac{1}{2\square(1-\square)}\cdot\frac{\square[\square\square^{\square}_{\square\square\square\square}\square(\square)]'}{\square\square^{\square}_{\square\square\square\square}\square(\square)}\right\} < \square^{2}$$

If $\beta \ne 1$, we have

$$\left|-\frac{\square[\square\square^{\square}_{\square\square\square\square}\square(\square)]'}{\square\square^{\square}_{\square\square\square\square}\square(\square)} - 2\square(1-\square)\right|^{2}$$

$$-2\frac{1+\square^{2}(4(\square-1)\alpha+1)}{1-\square^{2}}\square\square\left\{-\frac{1}{2\square(1-\square)}\cdot\frac{\square[\square\square^{\square}_{\square\square\square\square}\square(\square)]'}{\square\square^{\square}_{\square\square\square\square}\square(\square)}\right\}$$

$$+\left[\frac{1+\square^{2}(4(\square-1)\alpha+1)}{1-\square^{2}}\right]^{2}$$

$$< \frac{-1-\square^{2}(4(\square-1)\alpha+1)}{1-\square^{2}} + \left[\frac{1+\square^{2}(4(\square-1)\alpha+1)}{1-\square^{2}}\right]^{2}$$

That is

$$\frac{\left|-\frac{\square[\square\square^{\square}_{\square\square\square\square}\square(\square)]'}{\square\square^{\square}_{\square\square\square\square}\square(\square)} + 2\square(\square-1)\frac{1+\square^{2}(4(\square-1)\alpha+1)}{1-\square^{2}}\right|}{\frac{\square[1+(4(\square-1)\alpha+1)]}{1-\square^{2}}} < 1 \quad (2.2)$$

The above inequality shows that the values region of

$$\square(\square) = -\frac{1}{2\square(1-\square)}\cdot\frac{\square[\square\square^{\square}_{\square\square\square\square}\square(\square)]'}{\square\square^{\square}_{\square\square\square\square}\square(\square)}$$

is contained in the disk centered at $\frac{1+\square^{2}(4(\square-1)\alpha+1)}{1-\square^{2}}$ and radius $\frac{\square[1+(4(\square-1)\alpha+1)]}{1-\square^{2}}$.

It is easy to check that the function $\square(\square) = \frac{1-(4(1-\square)\alpha-1)\square\square}{1-\square\square}$ maps the unit disk U onto the disk $\frac{\left|w-\frac{1+\beta^{2}(4(\eta-1)\alpha+1)}{1-\beta^{2}}\right|}{\frac{\beta[1+(4(\eta-1)\alpha+1)]}{1-\beta^{2}}} < 1$.

Since G is univalent and $\square(0) = \square(0)$, $\square(\square) \subset \square(\square)$, we obtain that $\square(\square) \prec \square(\square)$, that is

$$-\frac{\square[\square\square^{\square}_{\square\square\square\square}\square(\square)]'}{\square\square^{\square}_{\square\square\square\square}\square(\square)} \prec 2\square(1-\square)\frac{1-(4(1-\square)\alpha-1)\square\square}{1-\square\square}$$

or

$$\frac{\square[\square\square^{\square}_{\square\square\square\square}\square(\square)]'}{\square\square^{\square}_{\square\square\square\square}\square(\square)} \prec \frac{-2\square(1-\square)+2\square(1-\square)(4(1-\square)\alpha-1)\square\square}{1-\square\square\square\square}.$$

Conversely, suppose that subordination

$$\frac{\square[\square\square^{\square}_{\square\square\square\square}\square(\square)]'}{\square\square^{\square}_{\square\square\square\square}\square(\square)} \prec \frac{\square(4(1-\square)\alpha-1)\square\square-\square}{1-\square\square} \text{ holds.}$$

Then

$$-\frac{\square[\square\square^{\square}_{\square\square\square\square}\square(\square)]'}{\square\square^{\square}_{\square\square\square\square}\square(\square)} = 2\square(1-\square)\frac{1-(4(1-\square)\alpha-1)\square\square(\square)\square\square}{1-\square\square(\square)\square\square}, \quad (2.3)$$

where $\square \in \square$. After simple calculations, from (2.3), we obtain

$$\left|\frac{\frac{\square[\square\square^{\square}_{\square\square\square\square}\square(\square)]'}{\square\square^{\square}_{\square\square\square\square}\square(\square)} + 2\square(1-\square)}{\frac{\square[\square\square^{\square}_{\square\square\square\square}\square(\square)]'}{\square\square^{\square}_{\square\square\square\square}\square(\square)} + 2\square(1-\square)(4(1-\square)\alpha-1)}\right| < \square$$

which proves that $\square \in \sum_{\square\square\square\square\square\square}(\alpha,\square,\square)$.

If $\beta = 1$, inequality (1.14) becomes

$$\left|\frac{-\frac{\alpha[\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)]'}{\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)}+2\alpha(\alpha-1)}{-\frac{\alpha[\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)]'}{\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)}+2\alpha(1-\alpha)(4(1-\alpha)\alpha-1)}\right|<\alpha \qquad (2.4)$$

From above we can easily obtain

$$-\frac{\alpha[\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)]'}{\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)}<2\alpha(1-\alpha)\frac{1-(4(1-\alpha)\alpha-1)\alpha\alpha}{1-\alpha}$$

or

$$\frac{\alpha[\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)]'}{\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)}<\frac{-2\alpha(1-\alpha)+2\alpha(1-\alpha)(4(1-\alpha)\alpha-1)2\alpha(1-\alpha)\alpha}{1-\alpha\alpha\alpha}.$$

(2.5) Where,

$$\left(\begin{matrix}0<\alpha\leq\frac{1}{2},0\leq\xi<1,\alpha\geq1,0\leq\varsigma\leq1,\\0\leq\eta<1,0\leq\alpha<<1,0<\alpha\leq1\text{ and }m\in N\end{matrix}\right).$$

Hence the proof.

Remark 2.1 Since $\alpha\alpha\frac{1-(4(1-\alpha)\alpha-1)\alpha\alpha}{1-\alpha\alpha}>\alpha$ it follows that

$$-\alpha\alpha\left\{\frac{1}{2\alpha(1-\alpha)}\cdot\frac{\alpha[\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)]'}{\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)}\right\}>\alpha.$$

Hence $\alpha^{\alpha}_{\alpha\alpha\alpha\alpha\alpha}\alpha(\alpha)\in\sum^{*}_{\alpha}(\alpha)$.

Structural formula for the class $\sum_{\alpha\alpha\alpha\alpha\alpha\alpha}(\propto,1,\alpha)$

Theorem 2.2 A function $\alpha\in\sum_{\alpha}$ is in the class $\sum_{\varphi\tau\varsigma\xi\alpha\rho}(\propto,1,\eta)$ if and only if there exists a probability measure $\mu(x)$ on ∂U such that

$$\alpha(\alpha)=\left[\alpha^{-\alpha}+\sum_{\alpha=1-\alpha}^{\infty}\frac{\alpha^{\alpha}}{\alpha_{\alpha}(\alpha,\alpha\alpha_{n},\alpha,\alpha)}\right]$$
$$*\left\{\alpha^{-\alpha}\cdot\alpha\alpha\alpha\int_{|\alpha|}2\alpha(1-\alpha)[1+\langle4(\alpha-1)\propto+1\rangle]\times\alpha\alpha\alpha(1-\alpha\alpha)\,\alpha\alpha(\alpha)\right\}$$

Where $(\alpha\in\alpha^{*})$. (2.6)

The correspondence between $\sum_{\alpha\alpha\alpha\alpha\alpha\alpha}(\propto,1,\alpha)$ and the probability measure $\mu(x)$ is one-to-one.

Proof: In view of the subordination condition (2.5), we have that $\alpha\in\int_{\alpha\alpha\alpha\alpha\alpha\alpha}(\propto,1,\alpha)$ if and only if

$$-\frac{1}{2\alpha(1-\alpha)}\cdot\frac{\alpha[\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)]'}{\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)}\in\alpha(\alpha).$$

From Lemma 1.2, we have

$$-\frac{\alpha[\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)]'}{\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)}=2\alpha(1-\alpha)\int_{|\alpha|}\frac{1-(4(1-\alpha)\alpha-1)\alpha\alpha}{1-\alpha\alpha}\,\alpha\alpha(\alpha)$$

Which is equivalent to

$$\frac{\alpha[\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)]'}{\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)}=\int_{|\alpha|}\frac{-2\alpha(1-\alpha)+2\alpha(1-\alpha)(4(1-\alpha)\alpha-1)\alpha\alpha}{1-\alpha\alpha}\,\alpha\alpha(\alpha)$$

Integrating this equality, we obtain

$$\alpha^{\alpha}\alpha^{\alpha}_{\alpha\alpha\alpha\alpha\alpha}\alpha(\alpha)=$$
$$\alpha\alpha\alpha\int_{|\alpha|}2\alpha(1-\alpha)[1+\langle4(\alpha-1)\propto+1\rangle]\,\alpha\alpha\alpha(1-\alpha\alpha)\,\alpha\alpha(\alpha)$$

or

$$\alpha^{\alpha}_{\alpha\alpha\alpha\alpha\alpha}\alpha(\alpha)=$$
$$\alpha^{-\alpha}\cdot\alpha\alpha\alpha\int_{|\alpha|}2\alpha(1-\alpha)[1+\langle4(\alpha-1)\propto+1\rangle]\times\alpha\alpha\alpha(1-\alpha\alpha)\,\alpha\alpha(\alpha) \qquad (2.7)$$

$$\alpha^{\alpha}_{\alpha\alpha\alpha\alpha\alpha}\alpha(\alpha)=(\alpha*\alpha)(\alpha).$$

Where

$$\alpha(\alpha)=\alpha^{-\alpha}+\sum_{\alpha=1-\alpha}^{\infty}\alpha_{\alpha}(\alpha,\alpha,\alpha,\alpha,\alpha,\alpha)\alpha^{\alpha}$$

Using (2.10) and above two equations we obtained following result

$$\alpha(\alpha)=\left[\alpha^{-\alpha}+\sum_{\alpha=1-\alpha}^{\infty}\frac{\alpha^{\alpha}}{\alpha_{\alpha}(\alpha,\alpha,\alpha,\alpha,\alpha,\alpha)}\right]$$
$$*\left\{\begin{matrix}\alpha^{-\alpha}\cdot\alpha\alpha\alpha\int_{|\alpha|}2\alpha(1-\alpha)[1+\langle4(\alpha-1)\propto+1\rangle]\\\times\alpha\alpha\alpha(1-\alpha\alpha)\,\alpha\alpha(\alpha)\end{matrix}\right\}.$$

Where $(\alpha\in\alpha^{*})$.

Theorem 2.3 Let $\mathbf{f}\in\sum_{\alpha\alpha\alpha\alpha\alpha\alpha\alpha}(\propto,1,\alpha)$ Then

$$\alpha^{\alpha}\alpha^{\alpha}_{\alpha\alpha\alpha\alpha\alpha\alpha}\alpha(\alpha)<(1-\alpha)^{2\alpha(1-\alpha)[1+(4(\alpha-1)\propto+1)]}\quad(\alpha\in\alpha).$$

Proof: Let $\alpha\in\sum_{\alpha\alpha\alpha\alpha\alpha\alpha\alpha}(\propto,1,\alpha)$ Then by (2.5) we have

$$\frac{\alpha[\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)]'}{\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)}<\frac{-2\alpha(1-\alpha)+2\alpha(1-\alpha)(4(1-\alpha)\alpha-1)\alpha}{1-\alpha}$$

Since the function

$$\frac{-2\alpha(1-\alpha)+2\alpha(1-\alpha)(2(1-\alpha)\alpha-1)\alpha}{1-\alpha}$$

is univalent and convex in U, in view of Goluzin's result, we obtain

$$\int_{0}^{\alpha}\frac{\alpha[\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)]'}{\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)}\,\alpha\alpha<\int_{0}^{\alpha}\frac{-2\alpha(1-\alpha)+2\alpha(1-\alpha)(4(1-\alpha)\alpha-1)\alpha}{\alpha(1-\alpha)}\,\alpha\alpha$$

Or

$$\alpha\alpha\alpha\,\alpha^{\alpha}_{\alpha\alpha\alpha\alpha\alpha\alpha}\alpha(\alpha)<\alpha\alpha\alpha\,\frac{[1-\alpha]^{2\alpha(1-\alpha)[1+(4(\alpha-1)\alpha+1)]}}{(\alpha)^{\alpha}}$$

Thus, there exists a function $\alpha\in\alpha$ such that

$$\alpha\alpha\alpha\,\alpha^{\alpha}_{\alpha\alpha\alpha\alpha\alpha\alpha}\alpha(\alpha)<\alpha\alpha\alpha\,\frac{[1-\alpha(\alpha)]^{2\alpha(1-\alpha)[1+(4(\alpha-1)\alpha+1)]}}{\alpha(\alpha)^{\alpha}},$$

which is equivalent to

$$\alpha^{\alpha}\alpha^{\alpha}_{\alpha\alpha\alpha\alpha\alpha\alpha}\alpha(\alpha)<[1-\alpha]^{2\alpha(1-\alpha)[1+(4(\alpha-1)\alpha+1)]}.$$

Structural formula for the class $\sum_{\alpha\alpha\alpha\alpha\alpha\alpha}(\propto,\alpha,\alpha)$

Theorem 2.4 Let $\alpha\in\sum_{\alpha\alpha\alpha\alpha\alpha\alpha}(\propto,\alpha,\alpha)$. Then

$$\alpha(\alpha)=\left[\alpha^{-\alpha}+\sum_{\alpha=1-\alpha}^{\infty}\frac{\alpha^{\alpha}}{\alpha_{\alpha}(\alpha,\alpha,\alpha,\alpha,\alpha,\alpha)}\right]*$$
$$\left[\alpha^{-\alpha}\alpha\alpha\alpha\left(2\alpha(1-\alpha)[\alpha+\langle4(\alpha-1)\alpha\propto+\alpha\rangle]\int_{0}^{\alpha}\frac{\alpha(\alpha)}{[1-\alpha\alpha(\alpha)]}\,\alpha\alpha\right)\right]. \qquad (2.8)$$

Where $(\alpha\in\alpha^{*})$ and $(\alpha\in\alpha)$.

Proof: Let $\alpha\in\sum_{\alpha\alpha\alpha\alpha\alpha\alpha\alpha}(\propto,\alpha,\alpha)$ and since we have obtained

$$\frac{\alpha[\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)]'}{\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)}<\frac{\alpha(4(1-\alpha)\alpha-1)\alpha\alpha-\alpha}{1-\alpha\alpha}\quad(\alpha\in\alpha),$$

$$\therefore\frac{\alpha[\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)]'}{\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)}=\frac{\alpha(4(1-\alpha)\alpha-1)\alpha\alpha(\alpha)-2\alpha(1-\alpha)}{1-\alpha\alpha(\alpha)}\quad(\alpha\in\alpha)$$

(2.9) From above equation, we have

$$\frac{\alpha[\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)]'}{\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)}+\frac{2\alpha(1-\alpha)}{\alpha}=\frac{2\alpha(1-\alpha)(4(1-\alpha)\alpha-1)\alpha\alpha(\alpha)-2\alpha(1-\alpha)}{1-\alpha\alpha(\alpha)}$$

$(z\in U^{*})$.

Integrating above implies that

$$\frac{\alpha\alpha\alpha\left[\alpha^{-\alpha}\alpha\bar{\xi}_{\alpha\alpha\alpha\alpha}\alpha(\alpha)\right]}{2\alpha(1-\alpha)[1+(4(\alpha-1)\alpha+1)]}=\alpha\int_{0}^{\alpha}\frac{\alpha(\alpha)}{[1-\alpha\alpha(\alpha)]}\,\alpha\alpha. \qquad (2.10)$$

$$\square^{\square}_{\square\square\square\ \square\square}\square(\square) = (\square * \square)(\square),$$
where
$$\square(\square) = \square^{-\square} + \sum_{\square=1-\square}^{\infty}\square_{\square}(\square,\square,\square,\square,\square,\square)\square^{\square}.$$
Using (2.10) and above two equations we obtained following result
$$\square(\square) = \left[\square^{-\square} + \sum_{\square=1-\square}^{\infty}\frac{\square^{\square}}{\square_{\square}(\square,\square,\square,\square,\square,\square)}\right] * \left[\square^{-\square}\square\square\left(2\square(1-\square)[1+\langle 4(\square-1)\propto+1\rangle]\square\int_0^{\square}\frac{\square(\square)}{[1-\square\square(\square)]}\square\square\right)\right].$$

Theorem 2.5 If $\square \in \square_{\square}$ belongs *to* $\square_{\square\square\square\square\square\square}(\propto,\square,\square)$, then
$$\square^{\square}_{\square\square\square\square\square}\square(\square) * \left\{\begin{array}{c}\frac{-2\square(1-\square)\square^{-\square}+(\square+1)\square^{1-\square}}{(1-\square)^2}(1-\square\square^{\square\square}) \\ +\frac{\square^{-\square}}{(1-\square)}[2\square(1-\square)-2\square(1-\square)\langle 4(1-\square)\propto-1\rangle\square\square^{\square\square}]\end{array}\right\} \neq 0 \quad (2.11)$$
for $(\square \in \square^{*})$ and $\square \in (0,2\square)$.
Proof: Let $\square \in \sum_{\square\square\square\square\square\square}(\propto,\square,\square)$. Then, from (2.1) it follows
$$-\frac{\square[\square\bar{\square}_{\square\square\square\square}\square(\square)]'}{\square\bar{\square}_{\square\square\square\square}\square(\square)} \neq \frac{-2\square(1-\square)\langle 4(1-\square)\propto-1\rangle\square\square^{\square\square}+\square}{1-\square\square^{\square\square}}$$
$(\square \in \square),\ \square \in (0,2\square)$ (2.12)
It is easy to see that the condition (2.12) can be written as follows $(1-\square\square^{\square\square})\square[\square^{\square}_{\square\square\square\square\square}\square(\square)]^{\square\square^{\square\square}}\square^{\square}_{\square\square\square\square\square}\square(\square) \neq 0$ (2.13) Note that
$$\square^{\square}_{\square\square\square\square\square}\square(\square) = \square^{\square}_{\square\square\square\ \square\square}\square(\square) * \left(\square^{-\square} + \square^{1-\square} + \cdots + \frac{1}{\square} + 1 + \frac{\square}{1-\square}\right) = \square^{\square}_{\square\square\square\square\square}\square(\square) * \frac{\square^{-\square}}{1-\square}$$
(2.14) And
$$\square[\square^{\square}_{\square\square\square\square\square}\square(\square)]' = \square^{\square}_{\square\square\square\square\square}\square(\square) * \left(-\square\square^{-\square} + (1-\square)\square^{1-\square} - \cdots - \frac{1}{\square} + \frac{\square}{(1-\square)^2}\right)$$
$$= \square^{\square}_{\square\square\square\square\square}\square(\square) * \frac{-\square\square^{-\square}+(1+\square)\square^{1-\square}}{(1-\square)^2}$$
(2.15) Where,
$$\left(\begin{array}{c}0<\square\leq\frac{1}{2}, 0\leq\square<1, \square\geq 1, 0\leq\square\leq 1, \\ 0\leq\square<1, 0\leq\propto<1, 0<\square\leq 1\ \square\square\square\ \square\in\square\end{array}\right).$$
By using (2.13), (2.14) and (2.15), we obtained
$$-\frac{\square[\square\bar{\square}_{\square\square\square\square}\square(\square)]'}{\square\bar{\square}_{\square\square\square\square}\square(\square)} \neq \frac{-2\square(1-\square)\langle 4(1-\square)\propto-1\rangle\square\square^{\square\square}+\square}{1-\square\square^{\square\square}}$$
$(\square \in \square),\ \square \in (0,2\square)$
Coefficient estimates:

Theorem 2.6 Let f of the form (1.6) is in the class $\sum_{\square\square\square\square\square\square}(\propto,\square,\square)$.Then, for $\square \geq 3-2\square(1-\square)$
$$\frac{|\square_{\square}|(\square+\square)\square_{\square}(\square,\square,\square,\square,\square,\square)}{2\square(1-\square)[1+\langle 4(\square-1)\propto+1\rangle]} \leq \square \quad (2.16)$$
Where $\Phi_{\mathrm{n}}(\varrho,\tau,\xi,\mathbf{m},\mathbf{p})$ is given by (1.11)

Proof: To prove the coefficient estimates (2.16) we use the method of Clunie and Koegh [4].
Let $\square \in \sum_{\square\square\square\square\square\square}(\propto,\square,\square)$.
We have
$$\frac{\frac{\square[\square\bar{\square}_{\square\square\square\square}\square(\square)]'}{\square\bar{\square}_{\square\square\square\square}\square(\square)}+2\square(1-\square)}{\frac{\square[\square\bar{\square}_{\square\square\square\square}\square(\square)]'}{\square\bar{\square}_{\square\square\square\square}\square(\square)}+2\square(1-\square)\langle 4(1-\square)\propto-1\rangle} = \square\square(\square)$$
where w is analytic in U and $|\square(\square)| \leq \square$ for $z \in U$. Then
$$\square[\square^{\square}_{\square\square\square\square\square}\square(\square)]' + \square\square^{\square}_{\square\square\square\square\square}\square(\square) = \square\square(\square)\left[\begin{array}{c}\square\left(\square^{\square}_{\square\square\square\square\square}\square(\square)\right)' \\ +2\square(1-\square)\langle 4(1-\square)\propto-1\rangle\square^{\square}_{\square\square\square\square\square}\square(\square)\end{array}\right]. \quad (2.17)$$
$$\square\square(\square) = \sum_{\square=1}^{\infty}\square_{\square}\square^{\square},$$
making use of (1.10) and (2.17), we obtain
$$\sum_{\square=1-\square}^{\infty}(\square+\square)\square_{\square}(\square,\square,\square,\square,\square,\square)\square_{\square}\square^{\square+\square} = \left\{\frac{-\square[1+\langle 4(\square-1)\propto+1\rangle]}{1}\right\}\sum_{\square=1}^{\infty}\square_{\square}\square^{\square}. \quad (2.18)$$
Equating the coefficients in (2.18), we have
$$\square\square_{\square-\square}(\square,\square,\square,\square,\square,\square)\square_{\square-\square} = -2\square(1-\square)[1+\langle 4(\square-1)\propto+1\rangle]\square_{\square},$$
for $\mathbf{n} = 1,2$
And
$$\square\square_{\square-\square}(\square,\square,\square,\square,\square,\square)\square_{\square-\square} = -2\square(1-\square)[1+\langle 4(\square-1)\propto+1\rangle]\square_{\square} + \sum_{\square=1-\square}^{\square-1-\square}[\square+2\square(1-\square)\langle 2(1-\square)\propto-1\rangle]\square_{\square}\square_{\square}\square_{\square-\square-\square} \quad (2.19)$$
for $n \geq 3$.
From (2.19), we obtain
$$\left\{\begin{array}{c}-4\square(1-\square)(1-\square)+ \\ \sum_{\square=1-\square}^{\square-1-\square}[\square+2\square(1-\square)\langle 4(1-\square)\propto-1\rangle\square_{\square}]\square_{\square}\square^{\square+\square}\end{array}\right\} \times \sum_{\square=1}^{\infty}\square_{\square}\square^{\square} = \sum_{\square=1-\square}^{\square-\square}[(\square+\square)\square_{\square}(\square,\square,\square,\square,\square,\square)]\square_{\square}\square^{\square+\square} + \sum_{\square=1-\square}^{\square-1-\square}\square_{\square}\square^{\square+\square} \quad (2.20)$$
It is known that, if $\square(\square) = \sum_{\square=1-\square}^{\square-1-\square}\square_{\square}\square^{\square}$ is analytic in $\mathbf{U}$, then for $0<\square<1$
$$\sum_{\square=0}^{\infty}|\square_{\square}|^2\square^{2\square} = \frac{1}{2\square}\int_0^{2\square}|\square(\square\square^{\square\square\square})|^2\square\square.$$
(2.21) Since $|\sum_{\square=1}^{\infty}\square_{\square}\square^{\square}| < \square|\square| < \square$,
Making use of (2.20) and (2.21), we have
$$\sum_{\square=1-\square}^{\square-1-\square}\left[(\square+2\square(1-\square))^2[\square_{\square}(\square,\square,\square,\square,\square,\square)]^2\right]|\square_{\square}|^2\square^{2(\square+\square)} + \sum_{\square=\square+1-\square}^{\infty}|\square_{\square}|^2\square^{2(\square+\square)} \leq \square^2 \times \left\{\begin{array}{c}\square^2[1+\langle 4(\square-1)\propto+1\rangle]^2 + \\ [\square+2\square(1-\square)\langle 4(1-\square)\propto-1\rangle]^2 \\ \times\left(\square_{\square}(\square,\square,\square,\square,\square,\square)\right)^2|\square_{\square}|^2\square^{2(\square+\square)}\end{array}\right\}.$$
Letting $\square \to 1$, we obtain
$$\sum_{\square=1-\square}^{\square-1-\square}[\square+2\square(1-\square)]^2\left(\square_{\square}(\square,\square,\square,\square,\square,\square)\right)^2|\square_{\square}|^2 \leq 4(1-\square)^2\square^2\square^2[1+\langle 4(\square-1)\propto+1\rangle]^2 + \sum_{\square=1-\square}^{\square-1-\square}[\square+2\square(1-\square)\langle 4(1-\square)\propto-1\rangle]^2\left(\square_{\square}(\square,\square,\square,\square,\square,\square)\right)^2|\square_{\square}|^2.$$
The above inequality implies
$$\square^2\square_{\square-\square}(\square,\square,\square,\square,\square,\square)^2|\square_{\square-\square}|^2 \leq 4(1-\square)^2\square^2\square^2[1+\langle 4(\square-1)\propto+1\rangle]^2 + \sum_{\square=1-\square}^{\square-1-\square}[\square+2\square(1-\square)]^2\left(\square_{\square}(\square,\square,\square,\square,\square,\square)\right)^2|\square_{\square}|^2$$

Finally, replacing n − p by n, we have

$$|a_n| \le \frac{4(1-\alpha)\Gamma\left[\frac{1+(4(\beta-1)\alpha+1)}{2}\right]}{(\beta+2\beta(1-\alpha))\Gamma_2(\alpha,\beta,\alpha,\beta,\alpha,\beta)},$$

Where,

$$\left(\begin{array}{c} 0 < \alpha \le \frac{1}{2}, 0 \le \beta < 1, \beta \ge 1, 0 \le \gamma \le 1, \\ 0 \le \beta < 1, 0 \le \alpha < 1, 0 < \gamma \le 1 \ \beta\beta\beta \ \ \gamma \in \gamma \end{array}\right).$$

Thus, the proof of our theorem is completed. Theorem 2.6 enables us to obtain a distortion result for the class $\sum_{\alpha\beta\gamma\alpha\beta\gamma}(\alpha, \beta, \gamma)$.

Corollary 2.1: If $f \in \sum_{\alpha\beta\gamma\alpha\beta\gamma}(\alpha, \beta, \gamma)$ is given by (1.6), then for $0 < |z| = r < 1$

$$|f(z)| \le \frac{1}{r^2} + \beta\gamma[1 + \langle 4(\gamma - 1)\alpha + 1\rangle] \times r^{1-\beta}\sum_{n=1-\beta}^{\infty}\frac{1}{(\beta+\gamma)\Gamma_2(\alpha,\beta,\alpha,\beta,\alpha,\beta)}.$$

$$|f(z)| \ge \frac{1}{r^2} + 2\beta(1-\gamma)\beta[1 + \langle 4(\gamma - 1)\alpha + 1\rangle] \times r^{1-\beta}\sum_{n=1-\beta}^{\infty}\frac{1}{(\beta+2\beta(1-\alpha))\Gamma_2(\alpha,\beta,\alpha,\beta,\alpha,\beta)}.$$

And

$$|f'(z)| \ge \frac{2\beta(1-\alpha)}{r^{2+1}} - 2\beta(1-\gamma)\beta[1 + \langle 4(\gamma - 1)\alpha + 1\rangle] \times r^{2-\beta}\sum_{n=1-\beta}^{\infty}\frac{1}{(\beta+2\beta(1-\alpha))\Gamma_2(\alpha,\beta,\alpha,\beta,\alpha,\beta)}.$$

$$|f'(z)| \le \frac{2\beta(1-\alpha)}{r^{2+1}} + 4\beta(1-\gamma)^2\beta[1 + \langle 4(\gamma - 1)\alpha + 1\rangle] \times r^{2-\beta}\sum_{n=1-\beta}^{\infty}\frac{1}{(\beta+2\beta(1-\alpha))\Gamma_2(\alpha,\beta,\alpha,\beta,\alpha,\beta)}.$$

In the sequence we give a sufficient condition for a function to belong to the class $\sum_{\alpha\beta\gamma\alpha\beta\gamma}(\alpha, \beta, \gamma)$.

Theorem 2.7 Let $f \in \sum_{\Gamma}$ be given by (1.6). If for $0 \le \alpha < 1$ and $0 < \beta \le 1$

$$\sum_{n=1-\beta}^{\infty}\{\beta(\beta+1) + 2\beta(1-\gamma)[1 + \beta\langle 4(1-\gamma)\alpha - 1\rangle]\} \times \Gamma_n(\alpha,\beta,\gamma,\alpha,\beta,\gamma)|a_n| \le \beta(1-\gamma)\beta[1 + \langle 4(\gamma-1)\alpha + 1\rangle] \quad (2.22)$$

$f \in \sum_{\alpha\beta\gamma\alpha\beta\gamma}(\alpha, \beta, \gamma)$.

Proof: Assume that $f(z) = z^{-\beta} + \sum_{n=1-\beta}^{\infty} a_n z^n$.

We have

$$\begin{aligned}\beta &= \left|\beta[z^{\beta}_{\alpha\beta\gamma\alpha\beta} f(z)]' + 2\beta(1-\gamma)z^{\beta}_{\alpha\beta\gamma\alpha\beta} f(z)\right| \\ &\quad -\beta\left|\beta[z^{\beta}_{\alpha\beta\gamma\alpha\beta} f(z)]' + 2\beta(1-\gamma)\langle 4(1-\gamma)\alpha - 1\rangle z^{\beta}_{\alpha\beta\gamma\alpha\beta} f(z)\right| \\ &= \left|\sum_{n=1-\beta}^{\infty}\big(\beta + 2\beta(1-\gamma)\big)\Gamma_n(\beta,\gamma,\beta,\gamma,\beta,\gamma)a_n z^{\beta}\right| \\ &\quad -\beta\left|\frac{-2\beta(1-\gamma)[1+(4(\beta-1)\alpha+1)]}{r^2}\right| \\ &\quad +\sum_{n=1-\beta}^{\infty}[\beta + 2\beta(1-\gamma)\langle 4(1-\gamma)\alpha - 1\rangle]\Gamma_n(\beta,\gamma,\beta,\gamma,\beta,\gamma)a_n z^{\beta}\end{aligned}$$

For $0 < |z| = r < 1$, we obtain

$$r^{\beta}\beta \le \sum_{n=1-\beta}^{\infty}(\beta + \beta)\Gamma_n(\beta,\gamma,\beta,\gamma,\beta,\gamma)a_n r^{\beta+\beta} - \beta\left[\begin{array}{c} 2\beta(\beta-1)(\langle 4(\beta-1)\alpha+1\rangle + 1) \\ -\sum_{n=1-\beta}^{\infty}|\beta + 2\beta(1-\gamma)\langle 4(1-\gamma)\alpha - 1\rangle|\Gamma_n a_n r^{\beta+\beta}\end{array}\right].$$

Or

$$r^{\beta}\beta \le \sum_{k=1-p}^{\infty}\{k(\beta+1) + 2p(1-\eta)[1 + \beta\langle 4(1-\eta)\alpha - 1\rangle]\} \times \Phi_k|a_k|r^{k+p} - 2p(1-\eta)\beta[1 + \langle 4(\eta-1)\alpha + 1\rangle].$$

Since the above inequality holds for all r $(0 < r < 1)$, letting $r \to 1$, we have

$$\beta \le \sum_{n=1-\beta}^{\infty}\left\{\begin{array}{c}\beta(\beta+1) + 2\beta(1-\gamma) \\ \times [1 + \beta\langle 4(1-\gamma)\alpha - 1\rangle]\end{array}\right\} \times \Gamma_n|a_n| - 2\beta(1-\gamma)\beta[1 + \langle 4(\beta-1)\alpha + 1\rangle].$$

Making use of (2.22), we obtain $M \le 0$, that is

$$\left|\frac{\beta[z^{\beta}_{\alpha\beta\gamma\alpha\beta} f(z)]'}{z^{\beta}_{\alpha\beta\gamma\alpha\beta} f(z)} + 2\beta(1-\gamma)\right| < \beta\left|\frac{\beta[z^{\beta}_{\alpha\beta\gamma\alpha\beta} f(z)]'}{z^{\beta}_{\alpha\beta\gamma\alpha\beta} f(z)} + 2\beta(1-\gamma)\langle 4(1-\gamma)\alpha - 1\rangle\right|.$$

Consequently, $f \in \sum_{\alpha\beta\gamma\alpha\beta\gamma}(\alpha, \beta, \gamma)$.

Where,

$$\left(\begin{array}{c} 0 < \alpha \le \frac{1}{2}, 0 \le \beta < 1, \beta \ge 1, 0 \le \gamma \le 1, \\ 0 \le \beta < 1, 0 \le \alpha < 1, 0 < \gamma \le 1 \ \beta\beta\beta \ \ \gamma \in \gamma \end{array}\right).$$

3 *Properties of the class* $\sum^{+}_{\alpha\beta\gamma\alpha\beta\gamma}(\alpha, \beta, \gamma)$

We begin this section by proving that the condition (2.22) is both necessary and sufficient for a function to be in the class $\sum^{+}_{\alpha\beta\gamma\alpha\beta\gamma}(\alpha, \beta, \gamma)$.

Theorem 3.1 Let $f \in \sum^{+}_{\Gamma}$. Then f belongs to the class $\sum^{+}_{\alpha\beta\gamma\alpha\beta\gamma}(\alpha, \beta, \gamma)$ if and only if

$$\sum_{n=1-\beta}^{\infty}\{\beta(\beta+1) + 2\beta(1-\gamma)[1 + \beta\langle 2(1-\gamma)\alpha - 1\rangle]\} \times \Gamma_n(\beta,\gamma,\beta,\gamma,\beta,\gamma)\Gamma_n \le 2\beta(1-\gamma)\beta[1 + \langle 4(\beta-1)\alpha + 1\rangle].$$

Proof: In view of Theorem 2.7, we have to prove "only if" part. Assume that

$$f(z) = z^{-\beta} + \sum_{n=1-\beta}^{\infty}\Gamma_n z^{\beta} \qquad (\Gamma_n \ge 0,\ \beta \in \gamma)$$

is in the class $\sum^{+}_{\alpha\beta\gamma\alpha\beta\gamma}(\alpha, \beta, \gamma)$.

Then

$$\left|\frac{\frac{\beta[z^{\beta}_{\alpha\beta\gamma\alpha\beta} f(z)]'}{z^{\beta}_{\alpha\beta\gamma\alpha\beta} f(z)} + 2\beta(1-\gamma)}{\frac{\beta[z^{\beta}_{\alpha\beta\gamma\alpha\beta} f(z)]'}{z^{\beta}_{\alpha\beta\gamma\alpha\beta} f(z)} + 2\beta(1-\gamma)\langle 4(1-\gamma)\alpha - 1\rangle}\right| = \left|\frac{\sum_{n=1-\beta}^{\infty}(\beta + 2\beta(1-\gamma))\Gamma_n(\beta,\gamma,\beta,\gamma,\beta,\gamma)a_n z^{\beta+\beta}}{\frac{2\beta(1-\gamma)[1+(4(\beta-1)\alpha+1)]}{r^{\beta}} - \sum_{n=1-\beta}^{\infty}[\beta + 2\beta(1-\gamma)\langle 4(1-\gamma)\alpha - 1\rangle]\Gamma_n a_n r^{\beta}}\right| < \beta$$

for all $z \in U$. Since $\mathrm{Re}\, z \le |z|$ for all z, it follows that

$$\mathrm{Re}\left\{\frac{\sum_{k=1-p}^{\infty}(k+2p(1-\eta))\Phi_k(\varrho,\tau,\varsigma,\xi,m,p)a_k z^{k+p}}{\frac{2p(1-\eta)[1+(4(\eta-1)\alpha+1)]}{z^{p}} - \sum_{k=1-p}^{\infty}[k+2p(1-\eta)\langle 4(1-\eta)\alpha - 1\rangle]\Phi_k a_k z^{k}}\right\} < \beta. \quad (3.1)$$

We choose the values of z on the real axis such that $\frac{1}{2\beta(1-\gamma)} \cdot \frac{\beta[z^{\beta}_{\alpha\beta\gamma\alpha\beta} f(z)]'}{z^{\beta}_{\alpha\beta\gamma\alpha\beta} f(z)}$ is real. Upon clearing the denominator in (3.1) and letting $z \to 1$ through positive values, we obtain

$$\sum_{n=1-\beta}^{\infty}\big(\beta + 2\beta(1-\gamma)\big)\Gamma_n(\beta,\gamma,\beta,\gamma,\beta,\gamma)\Gamma_n \le 2\beta(1-\gamma)\beta[1 + \langle 4(\beta-1)\alpha + 1\rangle] - \sum_{n=1-\beta}^{\infty}\beta\,[\beta + 2\beta(1-\gamma)\langle 4(1-\gamma)\alpha - 1\rangle]\Gamma_n(\beta,\gamma,\beta,\gamma,\beta,\gamma)\,\Gamma_n$$

Or

$$\sum_{n=1-\beta}^{\infty}\{\beta(\beta+1) + 2\beta(1-\gamma)[1 + \beta\langle 4(1-\gamma)\alpha - 1\rangle]\}\Gamma_n(\beta,\gamma,\beta,\gamma,\beta,\gamma)\Gamma_n \le \beta\beta[1 + \langle 4(\beta-1)\alpha + 1\rangle].$$

Where,
$\left(\begin{matrix} 0 < \square \le \frac{1}{2}, 0 \le \square < 1, \square \ge 1, 0 \le \square \le 1, \\ 0 \le \square < 1, 0 \le \propto < 1, 0 < \square \le 1 \ \square\square\square \ \ \square \in \square \end{matrix} \right).$
Hence, the result follows.

Corollary 3.1: If $\mathbf{f} \in \sum_{\mathrm{p}}^{+}$ given by (1.7) is in the class $\sum_{\square\square\square\square\square\square}^{+}(\propto, \square, \square)$
then
$\mathbf{a}_{\mathrm{n}} \le \frac{2\mathrm{p}(1-\eta)\beta[1+(4(\eta-1)\propto+1)]}{\{\mathrm{n}(\beta+1)+2\mathrm{p}(1-\eta)[1+\beta(4(1-\eta)\propto-1)]\}\Phi_{\mathrm{k}}(\varrho,\tau,\varsigma,\xi,\mathrm{m},\mathrm{p})}$
$\mathbf{n} \ge 1 - 2\mathbf{p}(1-\eta).$ (3.2)
with equality for the functions of the form
$\square_{\square}(\square) = \frac{1}{\square^{\square}} - \frac{2\square(1-\square)\square[1+(4(\square-1)\propto+1)]}{\{\square(\square+1)+2\square(1-\square)[1+\square(4(1-\square)\propto-1)]\}\square_{-\square}(\square,\square,\square,\square,\square,\square)} \square^{\square}$
Coefficient estimates obtained in Corollary 3.1 enables us to give a distortion result for the class $\sum_{\square\square\square\square\square\square}^{+}(\propto, \square, \square)$.

Theorem 3.2 If $\square \in \sum_{\square\square\square\square\square\square}^{+}(\propto, \square, \square)$,
then for $0 < |\square| = \square < 1$
$|\square(\square)| \ge \frac{1}{\square^{\square}} - \frac{2\square(1-\square)\square[1+(4(\square-1)\propto+1)]}{\{\square[1-2\square(1-\square)+(4(1-\square)\propto-1)2\square(1-\square)]+1\}\square_{1-\square}(\square,\square,\square,\square,\square,\square)} \square^{1-\square}$
And
$|\square(\square)| \ge \frac{1}{\square^{\square}} + \frac{2\square(1-\square)\square[1+(4(\square-1)\propto+1)]}{\{\square[1-2\square(1-\square)+(4(1-\square)\propto-1)2\square(1-\square)]+1\}\square_{1-\square}(\square,\square,\square,\square,\square,\square)} \square^{1-\square}$

where equality holds for the function
$\square_{\square}(\square) = \frac{1}{\square^{\square}} + \frac{2\square(1-\square)\square[1+(4(\square-1)\propto+1)]}{\{\square[1-2\square(1-\square)+(4(1-\square)\propto-1)2\square(1-\square)]+1\}\square_{1-\square}(\square,\square,\square,\square,\square,\square)} \square^{1-\square}$

at $\square = \square\square, \square.$

Proof: Suppose $\square \in \sum_{\square\square\square\square\square\square}^{+}(\propto, \square, \square)$, Making use of inequality
$\sum_{\square=1-\square}^{\infty} \square_{\square} \le \frac{2\square(1-\square)\square[1+(4(\square-1)\propto+1)]}{\{\square[1-2\square(1-\square)+(4(1-\square)\propto-1)2\square(1-\square)]+1\}\square_{1-\square}(\square,\square,\square,\square,\square,\square)}$ (3.3)
which follows easily from Theorem 3.1, the proof is trivial. Now, we prove that the class $\sum_{\varrho\tau\varsigma\xi\mathrm{mp}}^{+}(\propto, \beta, \eta)$ is closed under convolution.

Theorem 3.3 Let $\square(\square) = \square^{-\square} + \sum_{\square=1-\square}^{\infty} \square_{\square} \square^{\square}$ be analytic in $\square^{*}$ and $0 \le \square_{\square} \le 1$. If f given by (1.7) is in the class $\sum_{\varrho\tau\varsigma\xi\mathrm{mp}}^{+}(\propto, \beta, \eta)$ then f $*$ h
is also in the class $\sum_{\square\square\square\square\square\square}^{+}(\propto, \square, \square)$.

Proof: Since $\mathbf{f} \in \sum_{\varrho\tau\xi\mathrm{mp}}^{+}(\propto, \beta, \eta)$, then by Theorem 3.1, we have
$\sum_{\square=1-\square}^{\infty}\left\{\square(\square+1) + 2\square(1-\square)\left[1+\square\left\langle\frac{4(1-\square)}{\propto-1}\right\rangle\right]\right\}\square_{\square}(\square,\square,\square,\square,\square,\square)\square_{\square}$
$\le 2\square(1-\square)\square[1+\langle 4(\square-1)\propto+1\rangle].$
In view of the above inequality and the fact that
$(\square * \square)(\square) = \square^{-\square} + \sum_{\square=1-\square}^{\infty} \square_{\square}\square_{\square} \square^{\square}$
we obtain

$\sum_{\square=1-\square}^{\infty}\{\square(\square+1) + 2\square(1-\square)[1+\square\langle 4(1-\square)\propto - 1\rangle]\}\square_{\square}(\square,\square,\square,\square,\square,\square)\square_{\square}\square_{\square}$
$\le \sum_{\square=1-\square}^{\infty}\{\square(\square+1) + 2\square(1-\square)[1+\square\langle 4(1-\square)\propto - 1\rangle]\}\square_{\square}(\square,\square,\square,\square,\square,\square)\square_{\square}$
$\le 2\square(1-\square)\square[1+\langle 4(\square-1)\propto+1\rangle].$
Therefore, by Theorem 3.1, the result follows. The next result involves an integral operator which was investigated in many papers [2], [6], [20].

Theorem 3.4 If $\square \in \sum_{\square\square\square\square\square\square}^{+}(\propto, \square, \square)$, then the integral operator
$\square_{\square,\square}(\square) = \frac{\square}{\square^{\square+\square}}\int_{0}^{\square} \square^{\square+2\square(1-\square)+1}\square(\square)\square\square, \qquad \square > 0.$
It is also in the class $\sum_{\square\square\square\square\square\square}^{+}(\propto, \square, \square)$.
Proof: It is easy to check that
$\square_{\square,\square}(\square) = \square(\square) * \left(\square^{-\square} + \sum_{\square=1-\square}^{\infty} \frac{\square}{\square+2\square(1-\square)+\square}\square^{\square}\right).$
Since $0 < \frac{\square}{\square+2\square(1-\square)+\square} \le 1,$ by Theorem 3.3, the proof is trivial.
Where,
$\left(\begin{matrix} 0 < \square \le \frac{1}{2}, 0 \le \square < 1, 0 \le \square \le 1, \square \ge 1, \\ 0 \le \square < 1, 0 \le \propto < 1, 0 < \square \le 1 \ \square\square\square \ \ \square \in \square \end{matrix} \right).$

4 Neighborhoods and partial sums

Following earlier investigations on the familiar concept of neighborhoods of analytic functions by Goodman [7], Ruschweyh [17] and more recently by Liu and Srivastava [9], [10], Liu [11], Altinta¸s et al. [1], Orhan and Kamali [14], Srivastava and Orhan [19], Orhan [15], Deniz and Orhan [5] and Aouf [3], we define the *(n, δ)*− neighborhood of a function $\square \in \sum_{\square}$ of the form (1.6) as follows.
Definition 4.1 For $\square = \frac{2(\square-\square)^2+\frac{1}{\square}(\square-\square)(1-\square^2)}{1+2(\square-\square)^2+\frac{1}{\square}(\square-\square)(1-\square^2)} > 0$ and nonegative sequence
$\mathbf{S} = \{\mathrm{s_k}\}_{\mathrm{k=1-p}}^{\infty}$
Where
$\square_{\square} = \frac{\{\square[1-2\square(1-\square)+|(4(1-\square)\propto-1)|2\square(1-\square)]+1\}\square_{1-\square}(\square,\square,\square,\square,\square,\square)}{2\square(1-\square)\square[1+(4(\square-1)\propto+1)]}$ (4.1)
$(\square \ge 1-\square, \square \in \square, 0 \le \square < 1, 0 < \square \le 1).$
The $(\square, \square)$ − neighborhood of a function $\square \in \sum_{\square}$ of the form (1.6) is defined by
$\square_{\square}(\square) = \left\{ \square \in \sum_{\square} : \square(\square) = \square^{-\square} + \sum_{\square=1-\square}^{\infty} \square_{\square}\square^{\square} \ \square\square\square \ \ \sum_{\square=1-\square}^{\infty} \square_{\square}|\square_{\square} - \square_{\square}| \le \square \right\}.$ (4.2)
For $\square_{\square} = \square,$ Definition 1.4 corresponds to the $(\square, \square)$ − neighborhood considered by Ruscheweyh [17]. Making use of Definition 4.1, we prove the first result on $(\square, \square)$-neighborhood of the class $\sum_{\square\square\square\square\square\square}(\propto, \square, \square)$.

Theorem 4.1 Let $\square \in \sum_{\square\square\square\square\square\square}(\propto, \square, \square)$ be given by (1.6). If f satisfies
$[\square(\square) + \square\square^{\square}](1+\square)^{-1} \in \sum_{\square\square\square\square\square\square}(\propto, \square, \square)$
(4.3) $(\square \in \square, |\square| < \square, \ \square \ge 0)$
then

$$\square_{\square}(\square) \subset \sum_{\square\square\square\square\square\square}(\propto, \square, \square) \qquad (4.4)$$

Proof: It is not difficult to see that a function $\square \in \sum_{\square\square\square\square\square\square}(\propto, \square, \square)$ if and only if

$$\frac{\square[\square\bar{\square}_{\square\square\square\square}\square(\square)]' + 2\square(1-\square)\square\bar{\square}_{\square\square\square\square}\square(\square)}{\square\square[\square\bar{\square}_{\square\square\square\square}\square(\square)]' + \square(4(1-\square)\propto-1)2\square(1-\square)\square\bar{\square}_{\square\square\square\square}\square(\square)} \neq \square$$

(4.5) $(\square \in \square,\ \sigma \in C,\ |\sigma| = 1)$

which is equivalent to

$$\frac{(\square*\square)(\square)}{\square^{-\square}} \neq 0 \qquad (\square \in \square).$$

(4.6) Where for convenience,

$$\square(\square) = \square^{-\square} + \sum_{\square=1-\square}^{\infty} \square_{\square}\square^{\square}$$

$$= \square^{-\square} + \sum_{\square=1-\square}^{\infty} \frac{\{\square\square[\square+(4(1-\square)\propto-1)2\square(1-\square)]-(\square+2\square(1-\square))\}\square_{\square}(\square,\square,\square,\square,\square,\square)}{2\square(1-\square)\square[1+(4(\square-1)\propto+1)]\square}\square^{\square}$$

(4.7) From (4.7) it follows that

$$|\square_{\square}| = \left|\frac{\{\square\square[\square+(4(1-\square)\propto-1)2\square(1-\square)]-(\square+2\square(1-\square))\}\square_{\square}(\square,\square,\square,\square,\square,\square)}{2\square(1-\square)\square[1+(4(\square-1)\propto+1)]\square}\right|$$

$$\leq \frac{\{\square\square[\square+|(4(1-\square)\propto-1)|2\square(1-\square)]+(\square+2\square(1-\square))\}\square_{\square}(\square,\square,\square,\square,\square,\square)}{2\square(1-\square)\square[1+(4(\square-1)\propto+1)]\square}$$

$(\square \geq 1-\square,\ \square \in \square).$ (4.8)

Furthermore, under the hypotheses (4.3), using (4.6) we obtain the following assertions:

$$\frac{[\square(\square)+\square\square^{\square}](1+\square)^{-1}*\square(\square)}{\square^{-\square}} \neq 0.$$

Or

$$\frac{(\square*\square)(\square)}{\square^{-\square}} \neq \square \qquad (\square \in \square).$$

which is equivalent $\left|\frac{(\square*\square)(\square)}{\square^{-\square}}\right| \geq \square$ (4.9)

Now, if we let

$$\square(\square) = \square^{-\square} + \sum_{\square=1-\square}^{\infty} \square_{\square}\square^{\square} \in \square_{\square}(\square),$$

then we have

$$\left|\frac{\square(\square)-\square(\square)1*(\square)}{\square^{-\square}}\right| = \left|\sum_{\square=1-\square}^{\infty} (\square_{\square}-\square_{\square})\square^{\square+\square}\right|$$

$$\leq \sum_{\square=1-\square}^{\infty} \frac{\{\square[\square+|(4(1-\square)\propto-1)|2\square(1-\square)]+(\square+2\square(1-\square))\}\square_{\square}(\square,\square,\square,\square,\square,\square)}{2\square(1-\square)\square[1+(4(\square-1)\propto+1)]}$$

$$|\square_{\square}-\square_{\square}||\square|^{\square+\square} < \square$$

$(\square \in \square, \square \geq 1-\square, \square \in \square, \square > 0).$

Where,

$$\left(\begin{matrix} 0 < \square \leq \frac{1}{2}, 0 \leq \square < 1, 0 \leq \square \leq 1 \\ , \square \geq 1, 0 \leq \square < 1, 0 \leq \propto < 1, 0 < \square \leq 1\ \square\square\square\ \square \in \square \end{matrix}\right).$$

Thus, for any complex number σ with $|\sigma| = 1$, we have

$$\frac{(\square*\square)(\square)}{\square^{-\square}} \neq 0 \qquad (\square \in \square)$$

which implies $\square \in \sum_{\square\square\square\square\square\square}(\propto, \square, \square)$. The proof of the theorem is completed. In the sequence we give the definition of $(\square, \square)-$ neighborhood of a function $\square \in \sum_{\square}^{+}$ of the form (1.7).

Definition 4.2 For $\delta > 0$ and a non-negative sequence $\square = \{\square_{\square}\}_{\square=1-\square}^{\infty}$ where

$$\square_{\square} = \frac{\{\square(\square+1)+2\square(1-\square)[1+\square(4(1-\square)\propto-1)]\}\square_{\square}(\square,\square,\square,\square,\square,\square)}{2\square(1-\square)\square[1+(4(\square-1)\propto+1)]}$$

$(\square \geq 1-\square, \square \in \square, 0 \leq \propto < 1, 0 < \square \leq 1)$

The $(\square, \square)-$neighborhood of a function $\mathbf{f} \in \sum_{\mathbf{p}}^{+}$ of the form (1.7) is defined by

$$\square_{\square}(\square) = \left\{\begin{matrix} \square \in \sum_{\square}^{+}: \square(\square) = \square^{-\square} + \sum_{\square=1-\square}^{\infty} \square_{\square}\square^{\square} \\ \square\square\square\ \sum_{\square=1-\square}^{\infty} \square_{\square}|\square_{\square}-\square_{\square}| \leq \square \end{matrix}\right\}$$

(4.10) We have the following result on $(\square, \square)-$neighborhood of the class $\sum_{\varrho\tau\zeta\xi\mathrm{mp}}^{+}(\alpha, \beta, \eta)$.

Theorem 4.2 If the function f given by (1.7) is in the class $\sum_{\square\square\square\square\square\square}^{+}(\square, \square, \square)$, Then $\square_{\square}(\square) \subset \sum_{\square\square\square\square\square\square}^{+}(\square, \square, \square)$,

(4.11) Where

$$\square = \frac{2(\square-\square)(\square-\square)+\frac{(\square-\square)-\square^{2}(\square-\square)}{\square}}{1+2(\square-\square)(\square-\square)+\frac{(\square-\square)-\square^{2}(\square-\square)}{\square}}.$$

The result is the best possible in the sense that δ cannot be increased.

Proof: For a function $\mathbf{f} \in \sum_{\varrho\tau\zeta\xi\mathrm{mp}}^{+}(\alpha, \beta, \eta)$ of the form (1.7), Theorem 3.1 immediately yields

$$\sum_{\square=1-\square}^{\infty} \frac{\{\square(\square+1)+2\square(1-\square)[1+\square(4(1-\square)\propto-1)]\}\square_{\square}(\square,\square,\square,\square,\square,\square)}{2\square(1-\square)\square[1+(4(\square-1)\propto+1)]}\ \square_{\square}$$

$$\leq \frac{1}{\square_{1-\square}(\square,\square,\square,\square,1,\square)} \qquad (4.12)$$

Let

$$\square(\square) = \square^{-\square} + \sum_{\square=1-\square}^{\infty} \square_{\square}\square^{\square} \in \square_{\square}(\square)$$

with

$$\square = \frac{2(\square-\square)(\square-\square)+\frac{(\square-\square)-\square^{2}(\square-\square)}{\square}}{1+2(\square-\square)(\square-\square)+\frac{(\square-\square)-\square^{2}(\square-\square)}{\square}} > 0.$$

From the condition (4.10) we find that

$$\sum_{\square=1-\square}^{\infty} \square_{\square}|\square_{\square}-\square_{\square}| \leq \square$$

(4.13) Using (4.12) and (4.13), we obtain

$$\sum_{\square=1-\square}^{\infty} \square_{\square}\ \square_{\square} \leq \sum_{\square=1-\square}^{\infty} \square_{\square}\ \square_{\square} + \sum_{\square=1-\square}^{\infty} \square_{\square}|\square_{\square}-\square_{\square}|$$

$$\leq \frac{1}{\square_{1-\square}(\square,\square,\square,\square,1,\square)} + \square = 1$$

Thus, in view of Theorem 3.1, we get $\square \in \sum_{\square\square\square\square\square\square}^{+}(\square, \square, \square)$. To prove the sharpness of the assertion of the theorem, we consider the functions $\mathbf{f} \in \sum_{\varrho\tau\zeta\xi\mathrm{mp}}^{+}(\alpha, \beta, \eta)$ and $\mathbf{f} \in \sum_{\mathbf{p}}^{+}$ given by

$$\square(\square) = \square^{-\square} + \frac{2\square(1-\square)\square[1+(4(\square-1)\propto+1)]}{\{\square[1-2\square(1-\square)+(4(1-\square)\propto-1)2\square(1-\square)]+1\}\square_{1-\square}(\square,\square,\square,\square,\square,\square)}\square^{1-\square}$$

(4.14) And

$$\square(\square) = \square^{-\square} + \left[\begin{matrix} \frac{2\square(1-\square)\square[1+(4(\square-1)\propto+1)]}{\{\square[1-2\square(1-\square)+(4(1-\square)\propto-1)2\square(1-\square)]+1\}\square_{1-\square}(\square,\square,\square,\square,\square,\square)} \\ + \frac{2\square(1-\square)\square[1+(4(\square-1)\propto+1)]}{\{\square[1-2\square(1-\square)+(4(1-\square)\propto-1)2\square(1-\square)]+1\}\square_{1-\square}(\square,\square,\square,\square,\square,\square)} \end{matrix}\right]\square^{1-\square}$$

(4.15) Where,

$$\left(\begin{matrix} 0 < \square \leq \frac{1}{2}, 0 \leq \square \leq 1, 0 \leq \square < 1, \square \geq 1, \\ 0 \leq \square < 1, 0 \leq \propto < 1, 0 < \square \leq 1\ \square\square\square\ \square \in \square \end{matrix}\right).$$

Hence the function $\square \in \square_{\square}(\square)$ but according to Theorem 3.1, $\square \in \sum_{\square\square\square\square\square\square}^{+}(\square, \square, \square)$. Consequently, the proof of our theorem is completed. Next, we investigate the ratio of real parts of functions of the form (1.6) and their sequences of partial sums defined by

$$\square_{\square}(\square) = \begin{cases} \square^{-\square}, & \square = 1, 2, \cdots, -\square \\ \square^{-\square} + \sum_{\square=1-\square}^{\square-1} \square_{\square}\square^{\square} & \square = 1-\square, 2-\square, \cdots \end{cases}$$

(4.16) We also determine sharp lower bounds for $\square\square\left\{\frac{\square(\square)}{\square_{\square}(\square)}\right\}$ and $\square\square\left\{\frac{\square_{\square}(\square)}{\square(\square)}\right\}$.

Theorem 4.3 Let $\square \in \sum_{\square}$ be given by (1.6) and let $\square_{\square}(\square)$ be given by (4.16). Suppose that

$$\sum_{\square=1-\square}^{\infty} \square_{\square}|\square_{\square}| \leq 1 \qquad (4.17)$$

Where

$$\square_{\square} = \frac{\{\square(\square+1)+2\square(1-\square)[1+\square(4(1-\square)\propto-1)]\}\square_{\square}(\square,\square,\square,\square,\square)}{2\square(1-\square)\square[1+(4(\square-1)\propto+1)]}$$

Then, for $\square \geq 1-\square$, we have

$$\square\square\left\{\frac{\square(\square)}{\square_{\square}(\square)}\right\} > 1-\frac{1}{\square_{\square}} \qquad (4.18)$$

And

$$\square\square\left\{\frac{\square_{\square}(\square)}{\square(\square)}\right\} > \frac{\square_{\square}}{1+\square_{\square}} \qquad (4.19)$$

The results are sharp for each $\mathbf{m} \geq 1-\mathbf{p}$ with the extremal function given by

$$\square(\square) = \square^{-\square} - \frac{1}{\square_{\square}}\square^{\square} \qquad (4.20)$$

Proof: Under the hypotheses of the theorem, we can see from (4.17) that

$$\square_{\square+1} > \square_{\square} > 1 \qquad (\square > 1-\square).$$

Therefore, by using hypotheses (4.17) again, we have

$$\sum_{\square=1-\square}^{\square-1} |\square_{\square}| + \square_{\square}\sum_{\square=\square}^{\infty} |\square_{\square}| \leq \sum_{\square=1-\square}^{\infty} \square_{\square}|\square_{\square}| \leq 1. \qquad (4.21)$$

Let

$$\square(\square) = \square_{\square}\left[\frac{\square(\square)}{\square_{\square}(\square)} - \left(1-\frac{1}{\square_{\square}}\right)\right] = 1 + \frac{\square_{\square}\sum_{\square=\square}^{\infty}\square_{\square}\square^{\square+\square}}{1+\sum_{\square=1-\square}^{\square-1}\square_{\square}\square^{\square+\square}} \qquad (4.22)$$

Applying (4.21) and (4.22), we find

$$\left|\frac{\square(\square)-1}{\square(\square)+1}\right| = \left|\frac{\square_{\square}\sum_{\square=\square}^{\infty}\square_{\square}\square^{\square+\square}}{2+2\sum_{\square=1-\square}^{\square-1}\square_{\square}\square^{\square+\square}+\square_{\square}\sum_{\square=\square}^{\infty}\square_{\square}\square^{\square+\square}}\right|$$

$$\leq \frac{\square_{\square}\sum_{\square=1-\square}^{\square-1}|\square_{\square}|}{2-2\sum_{\square=1-\square}^{\square-1}|\square_{\square}|-\square_{\square}\sum_{\square=1-\square}^{\square-1}|\square_{\square}|} \leq 1 \quad (\square) \in \square \qquad (4.23)$$

From above it is clear that $\mathrm{Re}\, w(z) > 0$, $(z \in U)$. From (4.22), we immediately obtain (4.18). To prove that the function *f* defined by (4.20) gives sharp result,

we can see that for $\square \to 1-$

$$\frac{\square(\square)}{\square_{\square}(\square)} = 1-\frac{1}{\square_{\square}}\square^{\square} \to 1-\frac{1}{\square_{\square}}$$

which shows that the bound in (4.18) is the best possible.

Similarly, if we let

$$\square_{\square} = (1+\square_{\square})\left[\frac{\square_{\square}(\square)}{\square(\square)} - \frac{\square_{\square}}{1+\square_{\square}}\right]$$

$$= 1 - \frac{(1+\square_{\square})\sum_{\square=\square}^{\infty}\square_{\square}\square^{\square+\square}}{1+\sum_{\square=1-\square}^{\infty}\square_{\square}\square^{\square+\square}} \qquad (4.24)$$

and making use of (4.21), we find that

$$\left|\frac{\square(\square)-1}{\square(\square)+1}\right| = \left|\frac{-(1+\square_{\square})\sum_{\square=\square}^{\infty}\square_{\square}\square^{\square+\square}}{2+2\sum_{\square=1-\square}^{\infty}\square_{\square}\square^{\square+\square}-(1+\square_{\square})\sum_{\square=\square}^{\infty}\square_{\square}\square^{\square+\square}}\right|$$

$$\leq \frac{(1+\square_{\square})\sum_{\square=\square}^{\infty}|\square_{\square}|}{2-2\sum_{\square=1-\square}^{\infty}|\square_{\square}|+(1+\square_{\square})\sum_{\square=\square}^{\infty}|\square_{\square}|} \leq 1 \qquad (4.25)$$

Where,

$$\left(\begin{array}{c} 0 < \square \leq \frac{1}{2}, 0 \leq \square \leq 1, 0 \leq \square < 1, \square \geq 1, \\ 0 \leq \square < 1, 0 \leq \propto < 1, 0 < \square \leq 1 \ \square\square\square \ \square \in \square \end{array}\right).$$

which leads immediately to the assertion (4.19) of the theorem. The bound in (4.19) is sharp for each m ≥ 1−p, with the extremal function f given by (4.20). The proof of the theorem is now completed.

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Job Satisfaction in Organizational Executives

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Abstract- This paper is aimed at examining the level of job satisfaction among the executives serving in various organizations. For the purpose of conducting this study, a sample comprising 69 executives was selected with a deliberate effort to ensure that the sample comprises executives serving at junior, middle and senior levels. All subjects were administered MAO-R (Motivational Analysis of Organization – Roles), a questionnaire to assess level of job satisfaction. It was found that the job satisfaction level was high among executives. The need for achievement and the need for influence were more satisfied as compared to other needs. Detailed implications of this study will be discussed in the paper.

Index Terms- Job Satisfaction, Need for Achievement, Need for Affiliation, Need for Control Need for Extension, Need for Influence.

I. INTRODUCTION

In the present day scenario of stiff competition and stresses, an executives faces great challenges in order to keep himself motivated, an executive needs to be highly self inspired and resilient. Serving an organization where one is neck deep in work due to shortage and then be denied a promotion for want of vacancies brings us to the question of 'Job Satisfaction' in the organizations.

1.1 Job Satisfaction

Job satisfaction can be viewed as the overall attitude, or it can apply to the parts of an individual's job. It is like an attitude, generally acquired over a period of time as an employee gains more and more information about the workplace. Nevertheless, job satisfaction is dynamic, for it can decline even more quickly than it developed. Managers can not establish the conditions leading to high satisfaction now and then neglect it, for employee needs may change suddenly. Managers need to pay attention to employees' attitudes week after week, month after month, year after year.

Job satisfaction is one part of life satisfaction. The nature of one's environment of the job influences one's feelings on the job. Similarly, since a job is an important part of life, job satisfaction influences one's general life satisfaction. Managers may need to monitor not only the job and immediate work environment but also their employees' attitudes towards other part of life.

According to many studies, occupational stress, fatigue and job satisfaction are widespread problems in western societies and lately in the Indian society too. Since new management approaches seek constant development by placing the human factor in the foreground, and since institutions can only progress based on the views, attitudes, and perceptions of their human resources, the number of studies related to employee satisfaction has increased very rapidly (Witt and Beokermen, 1991; Jenkins, 1993; Judge and Watanabe, 1993). We know that satisfaction of the employees is crucial for achieving organizational excellence.

Job satisfaction describes how content an individual is with his or her job. The happier people are within their job, the more satisfied they are said to be. Job satisfaction is not the same as motivation, although it is clearly linked. Job design aims to enhance job satisfaction and performance; methods include job rotation, job enlargement and job enrichment. Other influences on satisfaction include the management style and culture, employee involvement, empowerment and autonomous work position. Job satisfaction is a very important attribute which is frequently measured by organizations.

1.2 Models of Job Satisfaction

Affect Theory

Edwin A. Locke's Range of Affect Theory (1976) is arguably the most famous job satisfaction model. The main premise of this theory is that satisfaction is determined by a discrepancy between what one wants in a job and what one has in a job. Further, the theory states that how much one values a given facet of work (e.g. the degree of autonomy in a position) moderates, how satisfied / dissatisfied one becomes when expectations are / aren't met. When a person values a particular facet of a job, his satisfaction is more greatly impacted both positively (when expectations are met) and negatively (when expectations are not met), compared to one who doesn't value that facet. To illustrate, if Employee 'A' values autonomy in the workplace and Employee 'B' is indifferent to autonomy, then Employee 'A' would be more satisfied in a position that offers a high degree of autonomy and less satisfied in a position with little or no autonomy compared to Employee 'B'.

Two-Factor Theory (Motivator-Hygiene Theory)

Fredrick Herzberg's Two Factor Theory (1966) (also known as Motivator Hygiene Theory) attempts to explain satisfaction and motivation in the work place. This theory states that satisfaction and dissatisfaction are driven by different factors – motivation and hygiene factors, respectively. An employee's motivation to work is continually related to job satisfaction of a subordinate. Motivating factors are those aspects of the job that make people want to perform, and provide people with satisfaction, for example achievement in work, recognition and promotion opportunities. These motivating factors are considered to be intrinsic to the job, or the work carried out. Hygiene factors include aspects of the working environment such as pay, company policies, supervisory practices, and other working conditions.

1.3 Factors Responsible for Job Satisfaction

Personal Factors

These factors include the individual employee personality, age, gender, number of dependents, education, intelligence, time on the job etc. Certain personality traits have been viewed as an important cause of job dissatisfaction.

Factors Inherent in the Job

These factors include the type of work to be performed, skills required for work performance, occupational status involved in the job, difference in work situations etc. The type of work inherent in the job is very important. Varied work generally brings about more job satisfaction than does routine work.

As regards to relation of occupational status to job satisfaction, it has been pointed out that occupational status is related to job satisfaction but it is not identical with it. Such status does not depend entirely on the job itself and it depends not only on way the employee regards the status of his job but also on how it is considered by others whose opinions or views he regards to be important. Generally the employees in those jobs are relatively more dissatisfied which have less social status or prestige. Job satisfaction is also related to the place of work situation. It is believed that employees at the work place in large cities are generally less satisfied with their jobs than work in small towns or cities.

Factors Under the Control of the Management

These factors include security (economic as well as social), wages and salaries or pay, fringe benefits, opportunities for advancement, working conditions, type and quality of supervision etc. It is believed that industrial employees mostly want steady work and security for life. The importance of economic as well as security as a factor in job satisfaction varies with the marital status and the number of dependents.

1.4 Needs Influencing Job Satisfaction

Need Achievement

Some people have a compelling drive to succeed and they strive for personal achievement rather than the rewards of success that accompany it. They have a desire to do something better or more efficiently than it has been done before. This drive is the achievement need. From research into the achievement need, McClelland (1958) found that high achievers differentiate themselves from others by their desire to do things better. They seek situations where they can attain personal responsibility for finding solutions to problems, where they can receive rapid feedback on their performance so they can set moderately challenging goals. High achievers are not gamblers: they dislike succeeding by chance. They prefer the challenge of working at a problem and accepting the personal responsibility for success or failure, rather than leaving the outcomes to chance or the actions of others. People with achievement need are enterprising in nature as they take calculated risks in achieving such targets for which they have to apply their imagination and stretch their limbs. They give single-minded attention to accomplishment of their tasks. Further, high achievers find goal achievement intrinsically satisfying and they do not necessarily crave for material rewards.

Achievement motive is not necessarily present in all individuals. It also differs in different cultures. For instance, McClelland (1961) in his research on the achievement motive in various cultures found that under-developed societies are characterized by low levels of achievement motive and even in the developed countries like the USA, only 10 percent of the population has high levels of achievement need. The achievement motive among people has been found to depend on child –rearing practices, maternal warmth, personal and occupational experiences and the type of organization for which they work. Apart from gaining understanding of human behaviour, the concept of 'need for achievement' is also important for understanding how people respond to the work environment. This helps in designing jobs for achievement – oriented people.

Need Affiliation

The need for affiliation can be viewed as the desire to be liked and accepted by others. It is the need for human companionship. Individuals with a high affiliation motive strive for friendship, prefer cooperative situations rather than competitive ones, and desire relationships involving in high degree of mutual understanding. Research has shown that people with a high affiliation need generally display the following characteristics:

(a) A strong desire for approval and reassurance from others.

(b) A tendency to conform to the wishes and norms of those whom they value.

(c) A sincere interest in the feelings of others.

Such people are motivated to express their emotions and feelings to others and expect other people to do the same in return. High need affiliation individuals tend to take job characterized by much of inter-personal contacts. They receive an inner satisfaction from socializing with people with little regard for personal feelings about them. They play important integrative role in group or inter group activities. But their overemphasis on social relationships may interfere with the process of accomplishing the task.

People possess the above needs in varying degrees. However, one of the needs will tend to be more characteristic of the individual rather than the other two. Individuals with a high need for achievement thrive on jobs and projects that tax their skills and abilities. Such individuals are goal oriented in their activities, seek a challenge and want task relevant feedback. Individuals with high affiliation needs value interpersonal relationships and exhibit sensitivity towards other people's feelings.

Need Influence

It is the need to have impact, to be influential, and to control others. Individuals high in influence need enjoy being "in charge", strive for influence over others, prefer to be placed into competitive and status oriented situations. People driven by influence need wish to create an impact on their organizations and are willing to take risks to do so. Such people are able to influence others to achieve organizational goals.

Need Extension

For the overall betterment of an individual as also the organisation, continuous enhancement of skills and resources is required. Continuous attention is required to be paid to motivational aspects, self system and group belonging. Need for extension is thus development of people, their competencies and the development of the organization as a whole. Constant endeavour on the part of all executives is required for enhancement in order to attain efficiency. This endeavour is directed at improvement and enhancement of self, subordinates, organizational infrastructure and procedures. While carrying out constant improvement, optimum utilization of resources is required to be made. At the same time creation of new resources is also desirable through judicious utilization of funds and grants. Efficiency is also required to be enhanced by training, monitoring and supervision. Successful attainment of personal and organizational enhancement leads to job satisfaction.

Need Control

Executives in the organizations have subordinates who are required to execute roles assigned to them over divergent conditions. Executives need to exercise control over their subordinates as also resources. The resources required to be managed are time, personnel, material and information.

The control required to be exercised is both, external and internal. By external control it is meant that control is required to be administered on external agencies which contribute to the final outcome. Internally, executives are also required to control themselves while working towards a goal in which other subordinates are also involved.

1.5 Job Satisfaction and Withdrawal Behaviors

Based on the research that shows job satisfaction predicts withdrawal behaviors like turnover and absenteeism, researchers have been able to statistically measure the financial impact of employee attitudes on organizations (e.g., Cascio, 1986; Mirvis & Lawler, 1977). Using these methods can be a powerful way for practitioners to reveal the costs of low job satisfaction and the value of improved employee attitudes on such outcomes as absenteeism and retention.

II. Methodology

2.1 Hypothesis

It was hypothesized that

(a) The job satisfaction will be high among the executives.

(b) There will be different levels of needs influencing overall job satisfaction.

2.2 Participants

69 executives from senior, middle and junior levels participated in the study. The average age of participants was 37 years, with the youngest being 22 years of age and the oldest aged 58 years. All participants were graduates.

2.3 Procedure

Subjects were administered "Motivational Analysis of Organization – Roles (MAO-R)", developed by Dr Udai Pareek, (1988).

2.4 Measures

MAO-R is developed by Dr Udai Pareek (1988). This self administered 25 item scale is intended to assess five needs: achievement, affiliation, influence, control and extension. MAO-R is a 5-point scale (1-5), containing five items for each need and a total of 25 items. The difference between perceived satisfaction and desired satisfaction gives the score. The score on each need will range from -20 to +20. The higher the score, the higher the dissatisfaction. The retest reliability of this scale ranged from .40 -.70. For validity: Item-analysis found all correlations significant at 0.001 levels. Factor analysis revealed that all factors together explain about 100% variance.

III. Result & Discussion

Table 1
Level of Job Satisfaction Among Organizational Executives

Job satisfaction variables	Mean
Achievement	6.09
Influence	6.68
Control	6.81
Extension	6.94
Affiliation	7.17
Total	33.70

The overall score of 33.70 is indicative of high level of job satisfaction as lower the value, higher is the satisfaction and vice-versa. The high level of job satisfaction amongst the executives arises from the nature of work, organizational culture and organizational climate. The need for achievement and the need for influence are more satisfied as compared to other needs.

Table 1 shows that need for achievement is 6.09. It means that the organizational executives at all levels are self motivated and strive to achieve assigned tasks. Above finding is supported by earlier studies also (Dayal and Saiyadain, 1970 ; Singh and Srivastava, 1983 and Padaki and Dolke, 1970). Good work is always appreciated and rewarded. Table 1 also shows that need for affiliation is 7.17. Need affiliation is not being fulfilled by the executives to the level desired by them because of formal nature of the organization and frequent transfers.

Need for influence is 6.68. (shown in Table1) It means an executive is able to exercise influence commensurate to his age and service and can achieve results relevant to the influence he exercises. Need for control is 6.81 which is average as compare to other needs (shown in Table 1). It means absolute control is not a reality, being relative in nature and control is associated with the position an executive holds in the organizational hierarchy. Need for extension is 6.94. (shown in Table1). Need extension is desired more in organization by executives, being low on account of rapid changes in technology, competitive environment, organizational restrictions and short tenures and frequent transfers.

IV. Conclusion

It can be concluded that the organizational executives are satisfied with their job and their needs are fulfilled in varying degrees in the organization. The overall need for achievement is dominant among the organizational executives in comparison to other needs related to job satisfaction.

V. Questionnaire

5.1 Challenging Work
(P) What level of challenging work you get to do?
(D) What level do you desire?

5.2 Influence or Impact
(P) What influence or impact are you allowed to make on others in your organization?
(D) What influence or impact do you desire to make?

5.3Authority to Admonish (Punish)
(P) What authority do you have to admonish (punish) individuals who do not conform?
(D) What level of such authority you desire to have?

5.4Working with Friendly People
(P) What level of opportunity is available to you to do something useful for others in the organization?
(D) What level of opportunity would you like to have?

5.5Doing Something Useful
(P) What level of opportunity is available to you to do something useful for others in the organization?
(D) What level of opportunity do you desire?

5.6Feedback on Performance
(P) How expeditiously are you apprised of your performance in the organization?
(D) What level of promptness in feedback on your performance you desire?

5.7Autonomy and Independence
(P) What level of autonomy / freedom you enjoy to work independently in your organization?
(D) What level of autonomy you desire?

5.8Freedom to Direct / Instruct Subordinates
(P) What level of autonomy you have to direct / instruct subordinates in your organization?
(D) What level of autonomy do you desire?

5.9Development of Close Personal Relations
(P) What opportunities do you get to develop close personal relations with colleagues in the organization?
(D) What level of such opportunity do you desire?

5.10 Development of Subordinates
(P)What the level of opportunity is available to you for developing your subordinates?
(D) What level of such opportunity do you desire?

5.11Setting Standards of Excellence
(P) What level of opportunity is provided to you for setting standards of excellence in your organization?
(D) What level would you like to have?

5.12Providing Suggestions to Superiors
(P) What level of freedom do you enjoy to provide suggestions to your superiors?
(D) What level do you desire?

5.13Control over People
(P) How far can you control people in your organization?
(D) What level do you desire?

5.14Sharing of Feelings and Emotions
(P) To what level are you allowed to share your feelings and emotions with others in the organization?
(D) What level would you like to enjoy in this regard?

5.1Helping Others
(P) What do you rate your possibilities for helping others in your organization?
(D) What is the level of opportunities you desire in order to help others?

5.16Rewarding of Efficiency
(P) How good are the avenues for rewarding efficient people in your organization?
desire?

5.17Contribution to Significant Decisions
(P) What level of opportunity is provided in your organization to contribute in making significant decisions?
(D) What level of such opportunity do you desire?

5.18Admonishment of Non - Performers
(P) How much authority you enjoy to admonish / punish non performers in your organization?
(D) What level of such authority do you desire?

5.19Interaction with Colleagues
(P) What level of freedom do you enjoy to interact with your colleagues?
(D) What level of such freedom do you desire?

5.20Cooperation with Others in a Common Task
(P) How far can you cooperate with others in a common task?
(D) What level do you desire for such cooperation?

5.21Stretching of Abilities / Skills
(P) What opportunities does the organization provide you to stretch your abilities / skills?
(D) What level do you desire to have?

5.22Recognition for Work

(P) What are the chances of getting recognition for work produced by you in your organization?
(D) What level do you desire in this regard?

5.23Reporting by Subordinates / Other Sections
(P) How good is reporting by your subordinates to you?
(D) How good do you want it to be?

5.24 Interaction with Others on Matters Other Than Tasks
(P) What level of freedom do you enjoy to interact with your colleagues?
(D) What level of such freedom do you desire?

5.25 Team Work
(P) How much freedom do you have to work in teams?
(D) What level of freedom do you desire?

VI. Instruction for Answering the Questionnaire

6.1 Personal Information

Designation ……. with effect from ……

Date of birth ……Age (in years)……. Gender (male / female)

Date of enrolment ……Seniority ……

Service (in years) ……..

6.2 Instructions

Your job in your organization, including your current role / appointment, provides you with some degree of freedom of action / opportunities / facilities to achieve assigned goals. Some of the functional aspects of your job are listed below vide statements 1-25. You are requested to read these statements carefully and then answer each under headings 'P' or 'D'; with 'P' indicating the present level of freedom of action / opportunities / facilities you actually enjoy and 'D' implying the level you would desire to have.

Please accord level to your answers on a scale of 1 to 5 with each numerical value meaning as under :-

(a) 1 = No freedom of action / opportunity / facility at all.
(b) 2 = Very little freedom of action / opportunity / facility .
(c) 3 = Some freedom of action / opportunity / facility.
(d) 4 = Good freedom of action / opportunity / facility.
(e) 5 = Great deal freedom of action / opportunity / facility.

Please fill in the personal information details before proceeding with the actual task.

6.3 Answers

Item	P	D
1		
2		
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Word Sense Disambiguation Using Selectional Restriction

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Abstract- Word sense disambiguation (WSD) is still an open research area in natural language processing and computational linguistics. It is from both theoretical and practical point of view. Here, the problem is to find the sense for word in given a context, It is a technique of natural language processing(NLP) ,which requires queries and documents in NLP or texts from Machine Translation (MT). MT is an automated system which involves Marathi, Urdu, Bengali, Punjabi, Hindi, and English etc. Most of the work has been completed in English, and now the focus has been shifted to other languages. The applications of WSD are disambiguation of content in information retrieval (IR), machine translation (MT), speech processing, lexicography, and text processing. In this paper, we have used knowledge based approach along with selectional restriction. It is used to block the formation of component word meanings representation that contains selectional restriction. We have developed a WSD tool using Hindi wordnet. Wordnet is built from co-occurrence, and collocation and it includes synset or synonyms which belong to either noun, verb, adjective, or adverb. In this paper we shall discuss the implementation of our tool and its evaluation.

Index Terms- Word sense disambiguation, knowledge based approach using Selectional Restriction, Hindi wordnet, etc.

I. INTRODUCTION

A.Word Sense Disambiguation

Word sense disambiguation is the problem to find the sense of a word in natural language context, where the word have multiple meanings. The sense of a word in a text depends on the context in which it is used; the context of the ambiguous word is determined by other neighboring words. This is called as local context or sentential context. This task needs lot of words as well as world knowledge. A WSD is the process of identifying the sense of the word.

Example: गौतम बुद्ध ने गया में ज्ञान प्राप्त किया था

or राम स्कूल गया

Here, we can easily see that ,In first sentence the word' गया' refers to 'a place name' in the former sentences, whereas the word 'गया' refers to a 'verb'

Words do not have well-defined boundaries between their senses, and our task is to determine, which meaning of word is indented in a given text. This is one of the first problems that is encountered by any natural language processing system which is refered to as lexical ambiguity. WSD is a research area in NLP, which is very useful nowdays. It is the technique of natural language processing (NLP).It can be represented by tasks, performance, knowledge source, computational complexity, assumptions and application for WSD algorithms.WSD involve more word knowledge or common sense, which identifies Dictionary or thesauri. It is also helpful in many application such as information extraction (IE), information retrieval (IR), and speech recognition (SR). Word sense disambiguation is important for lexical knowledge and word knowledge.

II. HISTORY

The problem of WSD was first formulated as a computational task during the early days of machine translation in the 1940s. In 1949 Warren Weaver, in his famous 1949 memorandum on machine translation, first introduced the problem in a computational context. Early researchers understood well the significance and difficulty of WSD in NLP. In 1960, Bar-Hillel, used the example to argue that WSD could not be solved by "electronic computer". In 1970s WSD was a subtask of semantic interpretation systems or model developed within the field of artificial intelligence, but since WSD systems were largely rule-based and hand-coded. By the 1980s large lexical resources, such as the Oxford Advanced Learner's Dictionary of Current English (OALD), became available: hand-coding was replaced with knowledge based approach automatically extracted from these resources, but word sense disambiguation remained still knowledge-based or dictionary-based. In the 1990s the statistical linguistics revolution swept through computational linguistics, and WSD became a paradigm problem on which to apply supervised machine learning techniques. The 2000s saw supervised techniques to reach a plateau in accuracy, and for this reason attention has been shifted to coarser-grained senses, domain adaptation, semi-supervised and unsupervised corpus-based systems, combinations of different methods or approachs, and the return of knowledge-based systems via graph-based methods. Still, supervised method continue to perform best hybrid Systems, minimizing or eliminating use of sense tagged text by taking the advantage of the Websystem.

III. WORDNET CONCEPT

A.Wordnet

Wordnet is a network of words linked by lexical and semantic relationship. Wordnets for Hindi and Marathi being built at IIT Bombay are amongst the first IL wordnets. Wordnet is an electronic large lexical database of English, and it is a combination of dictionary and thesaurus which is being created

and maintained by cognitive science lab of Princeton university. In this way, the Hindi wordnet is inspired by the English wordnet. The wordnet refers the lexical information in senses and set of words.in which describes the meaning of the word in a specific text. Wordnet is the existence of various relations between the word forms (e.g. lexical relations, such as synonymy and antonymy) and the synsets (meaning to meaning or semantic relations e.g. hyponymy/hypernymy relation, meronymy relation). Wordnet has four types of part of speech (POS), such as noun, verb, adverb, adjective. POS tagging is the process of identifying lexical category of a word in a sentence on the basis of its context.

B. Synset

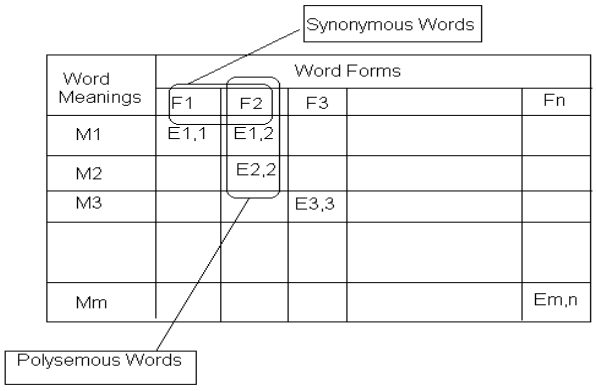
- The smallest unit in wordnet.
- A synonym set.
- Represent a specific meaning of a word.

Synsets are connected to one another through semantic and lexical relations. Each word meaning can be represented by a set of word-forms.It is called as synonym set or synset. Synsets are made by content words such as noun, verb, adjective, and adverb.

C. Lexical Matrix

The lexical matrix is a part of the language system.It refers the link between word form and the word meanings.The following table is representing the lexical matrix.It is called as lexical matrix.It shown of the lexical information by an organization.Word forms are imagined as headings for the columns and word meanings for the rows.Rows represent only synonymy while columns represent polysemy.

Table 1. Lexical Matrix



For example the word 'कमल' of synset like {कमल,पंकज,सरोज,नीरज}gives the meanings'फूल' (पानी में होने वालेएक पौधे का फूल जो बहुत सुन्दर होता है) belongs to a synset, whose members from a row in the matrix, and the row numbers gives a ID to the synset.'कमल' has different meanings,(रक्त में पित्त वर्दक के जमा हो जाने से उत्पन्न एक रोग जिससे शरीर व आँखे पीले पड़ जाते है) which comes in the column by the word.

IV. SEMANTIC RELATIONS IN WORDNET

The Hindi wordnet is inspired by the English wordnet, semantic relation use in structuring lexical data. They have been extensively used in wordnet and evaluated also, and they are mainly used to structure the lexicon such as, and the semantic relations is following below.

Types of semantic relations (It is based on POS):

- Hypernymy (kind-of):'घर ' is hypernym of 'गृह '
- Hyponymy (kind-of):'गृह ' is a hyponym of 'घर'
- Holonymy (part-of):'आवास' is a holonym of 'निवास'
- Meronymy (part-of):'बरामदा' is a meronym of 'आगंन'

For example, we have the synset {घर, गृह}. The hypernymy relation (Is-A) of it links to {आवास,निवास}. Its meronymy relation (Has-A) links to {आँगन, बरामदा} and hyponymy relation to {सराय} and {झोपड़ा}.

V. LITERATURE SURVEY

In the last 15 yaers, the NLP community has an increasing interest in machine learning based approaches for automated classification of word senses. This is evident from the number of supervised WSD approaches that have spawned. Today, the supervised approaches for WSD are the largest number of algorithms, used for disambiguation, Robert R.Korfhage [11]. Supervised WSD uses machine learning techniques on a sense data set to classify the senses of the words Fellbaum Christiane [3], there are a number of classifiers also called word senses that assign an appropriate sense to an instance of a single word. The supervised algorithms thus perform target-word WSD. Any algorithm uses certain features associated with a sense for training.Supervised algorithms trained a model based on the corpus provided to it. This corpus needs to be manually annotated, and the size of the corpus needs to be large enough in order to train a generalized model. Semi-supervised, also known as minimally supervised algorithms and it make some assumptions about the language and discourse in order to minimize these restrictions, Kieinberg.M.Jon [4]. The common threads of operation of these algorithms are these assumptions and the seeds used by them for word sense disambiguation purposes. However, these are fundamental overlap based algorithms which suffer from overlap sparsity, dictionary, thesauri, definitions being generally small in length, Yarowsky [8]. Supervised learning algorithms for WSD are mostly word specific classifiers.The requirement of a large training corpus renders these algorithms unsuitable for resource scarce languages, Fellbaum Christiane [3].

There are three approaches for WSD:

A.Supervised Approach

Supervised is based on a labeled training set and corpus, and it is a learning system, has a training set of featured-encoded inputs and their, appropriate sense label or category, because they can cope with the high dimensionally of the feature spaces, Mark Stevenson[7], basically, supervised approaches applied to

the problem of WSD, and used to machine learning techniques for classifier from senses,Yarowsky[8].

B.Unsupervised Learning Approach

In approach, Stevenson, Mark Stevenson [7], unsupervised approach based on unlabeled corpora, and it is learning system, has a training set of feature encoded inputs but not their appropriate sense label or category. it only use the information available in raw text, do not use outside knowledge sources or manual annotations, unsupervised reduces WSD to the problem of finding the targeted words that occurs in the most similar contexts and placing them in the cluster, Agirre Eneko,and Rigau[14].

C.Knowledge Based Approach

This approach based on wordnet, and knowledge based approach to word sense disambiguation taken place, when experimental are conducted on extremely limited domains, Robert R.Korfhage[11], Here the knowledge resources are dictionary, thesauri, collocation, and ontology etc.In knowledge based approach to disambiguate ,we will determines to the target word, along with a context, Kleinberg.M.Jon[7].

VI. PROPOSED WORK

In our thesis, we are using knowledge based approach using selection restriction and developed a WSD tool with Hindi wordnet. Our system currently deals with POS tagger such as noun, verb, adjective, adverb.To given corpora to assign correct sense to the words.This is sense tagging, needs word sense disambiguation (WSD). It is highly important for Question Answering, Machine Translation, Text Mining tasks. Work is depend on to including words of other part of speech. We have taken the database of text files saved from Hindi Wordnet. It prepared by IIT, Bombay but in future, In directly, the database for Hindi language's WSD can use the database prepared for Hindi Wordnet.

Selection restriction s'Algorithim

Selectional preference used here is that proposed by Resnik, combining statistical linguistics and knowledge-based approaches. The basis of the approach is a probabilistic model capturing the co-occurrence and collocation behavior of predicates and conceptual classes in the taxonomy.

The prior distribution PrR(c) captures the probability of a word-class occurring as the argument in predicate-argument relation R, regardless of the identity of the predicate. For example, given the verb relationship, the prior probability for tends to be significantly higher than the prior probability for (insect). However, once the identity of the predicate is taken into account, the probabilities can change, if the verb is buzz, then the probability for (insect),it can be expected to be higher than its prior, and (person) will likely be lower. In probabilistic terms, it is the difference between this conditional or posterior distribution and the prior distribution that determines selectional restriction. Information theory provides an appropriate way to quantify the difference between the prior and posterior distributions, in the way of relative entropy. The model defines the selectional preference strength of a predicate as:

$$Sr(p) = D(pr(c \mid p) \| Pr(c))$$
$$= \sum Pr(c \mid p) \log Pr(c \mid p) / Pr(c \mid p) \quad (1)$$

Intuitively, Sr(p) measures the information, in bits, p predicate provides about the conceptual class of its argument. The better Pr(c) approximates Pr(c | p), the leas influence p is having on its argument, and therefore the less strong its selectional restriction or preference. Given this definition, a natural way to characterize the "semantic fit of a particular class as the argument to a predicate is by its relative contribution to the overall selectional restriction strength. In particular, word to classes that fit very well can be expected to have higher posterior probabilities, compared to their priors, as is the case for (insect) in. Formally, selection association is defined as given below:-

$$Ar(p, c) = 1 / Sr(p).Pr(c \mid p) \log Pr(c \mid p) / Pr(c) \quad (1)$$

This model of selectional restriction or preferences has turned to make reasonable predictions about human judgments of argument plausibility obtained by psycholinguistic methods (Resnik, 1993a). The selectional association has also been used recently to explore apparent cases of syntactic or lexical system optionality (Paola Merlo, personal communication).

Selectional restriction depends on knowledge based algorithm. A knowledge based algorithm is one which depends upon the selectional restriction to restrict the number of meaning of a target word in given a context, and it is to determine word to word relation. The problem of lexical and syntactic ambiguity encountered in NLP. It is also called as selectional preferences. A selectional preferences or restriction are constraints on semantic type that a sense imposes on the words, which it combines usually through grammatical relation in sentences.

Now consider the following example:

Here, we have taken the word in Hindi as '**खाना**', which consists of two senses, as the first '**खाना**' and other '**भोजन**'.In the given context, such as "**मुझे आम खाना है**".so, here the word '**खाना**' is the same sense of the word '**खाना**'. Thus, this technique shows any sense of the word depends on its context.What's that word makes sense. Namely the 'ordinary' word has been seen as a reference to the selection of what is the sense of the word.

The selection restriction approach to disambiguation has many requirements to be very useful in large scale practical application using with wordnet and it have been developed part of speech (POS) tagger. These systems are designed to make minimal assumption about what information will be available from the processes. The knowledge based approach uses of external lexical resources like dictionary or thesauri. In knowledge based approach, system are trained to perform the task of word sense disambiguation.

In this approach, what is learned to classifier that is used to assign examples to one of a fixed senses.WSD is refers to the knowledge resources to the senses of word in given a sentences. These are some knowledge resources such as dictionary, thesauri, ontology, glossaries, collocations, etc.

Knowledge based approach have a knowledge resources of machine readable dictionary(MRD), and selectional restriction(SR) in front of corpus, for example Wordnet. They may use grammar rules for word sense disambiguation. It is a fundamental component of WSD, and can be provide the data, in which associate sense with words.This is one of the most important knowledge based in natural language processing and has been used for syntactic lexical and word sense disambiguation, and the degree of the selectional restriction for a word combinations from a tagged corpus, based on the multiple regression model.

VII. RESULT

Synset format: The word 'प्रेम'

ID: 121(a unique number identifying a synset)
CATEGORY: NOUN (POS category of the words)
CONCEPT: अपने से छोटों के लिए प्रेम होना चाहिए (The part of the gloss that gives a brief summary of what the synset represents)
EXAMPLE: "चाचा नेहरू को बच्चों से बहुत ही स्नेह हुआ करता था" (one or more example of the word in the synset used in context)
SYNSET: स्नेह, नेह, लगाव, ममता (The set of synonymous word)

VIII. EVALUATION

In our paper, the evaluation of WSD, We developed the WSD tool followed by Hindi wordnet. We used a small corpus with word occurrences and collocations; firstly the corpus was tagged by part of speech (POS) tagger. A parsed, sense-tagged corpus determined by Hindi wordnet sense-tagged corpus.

The test for the verb-object relationship was constructed by a selectional restriction model on the corpus. The 100 verbs that select most strongly for their objects were identified, including verbs appearing only once in the corpus, test instances of the form (verb, object, correct sense),were the extracted from the merged test corpus, including all triples where verb was one of the 100 test verbs.

Here, we contained some words like {सोना, आम, गया, चैन, बाली}.These files contained 5463 words, out of which we could disambiguate 5236. The accuracy of our approach was 66.92%, which means that our system disambiguated correctly 3492 out of 5463 words.

IX. CONCLUSION

In this paper, we describe a knowledge based approach using selectional restriction for Hindi language. Our methods can be improved by parts of speech (POS) and Hindi wordnet .Manually, we have added these links in the Wordnet database available in MySQL format for some words.This method gives a multiple occurrences for the word in given a context, if a word occurs multiple times in different senses in the same text, it is high likely that our methods would assign the same synset or synonyms to all its occurrences, for example the word ''occurs in the text with the meaning 'धन'as well as 'दौलत' but the synonyms assigned to all occurrences of 'धन'is{रुपया,पैसा,मूल्धन},since The wordnet relation system applied at training time and frequency data (no. of tagged senses)applied at runtime. The Hindi wordnet depend only the lexical data files distributed with wordnet not on any code. The accuracy of WSD highly depends on the part of speech (POS) tagger module. The efficiency of our work is limited due to the fact that, it can't tag some words correctly with POS tagger.
Here, Since the POS tagger plays an important role in the WSD we need to improve the accuracy of the POS tagger in order to disambiguate a word correctly.

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Development of A Grid Connected Inverter for Solar PV Systems with Energy Capture Improvement Based On Current Control Strategy

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Abstract- In this paper a single stage inverter for solar PV system with energy capture improvement based on voltage control to solve fast changing irradiation problem is proposed. A cascaded control structure with a dc link voltage control loop and a current control loop is used. The maximum power capturing controller is applied to the reference of the outer loop control dc voltage photovoltaic,. without PV array power measurement. In order to generate the correct maximum power point reference voltage under rapidly changing irradiation, a robust maximum power point capturing controller has been proposed. In this controller, the d-axis grid current component reflecting the power grid side and the signal error of a proportional–integral outer voltage regulator is designed to reflect the change in power caused by the irradiation variation. Hence, with this information, the proposed algorithm can greatly reduce the power losses caused by the dynamic tracking errors under rapid weather changing conditions. The superiority of the newly proposed method is supported by simulation and experimental results. The robust tracking capability under rapidly increasing and decreasing irradiance is verified experimentally with a PV array emulator. The performance of the power flow depends largely on the quality of the applied current control strategy.

Index Terms- Fast changing irradiation, maximum power capturing,gridconected PV systems, fast transient response, proportional-integral outer voltage regulator,PV array model.

I. INTRODUCTION

Because of the changes caused by the atmospheric conditions the voltage-power characteristic of a photovoltaic array is nonlinear and time varying. The task of a maximum power capturing in a PV power system is to continuously tune the system so that it draws maximum power from the PV array. The grid connected PV systems have become more popular because they do not need battery backups to ensure maximum power point capturing[1] . The number of power stages undermines the overall efficiency, reliability, and compactness of the system besides increasing the cost[2]. The typical configurations of a grid-connected PV system are single or two stages. In two stages, the first is used to boost the PV array voltage and track the maximum power; the second allows the conversion of this power into high-quality ac voltage[3]. The single stage has numerous advantages, such as simple topology, high efficiency, etc.

The control strategy has to be designed in order to track the maximum available power and to properly transfer it from the PV array to the grid simultaneously. The main component of the single-stage grid connected PV system is the three-phase voltage source inverter . A simple inductors *L* are used as a filter interfacing inverter and mains. This project is based on proposed maximum power point and the performance of the power flow depends largely on the quality of the applied current control strategy. In this paper, the current control has been implemented in a rotating synchronous reference frame *d*, *q* because the controller can eliminate a steady-state error and has fast transient response by decoupling control. Energy-balance modeling and discrete control for single-phase grid-connected PV central inverters[4].is proposed.

In this paper, in order to generate the correct maximum power point reference voltage under rapidly changing irradiation, a robust maximum power point extraction controller has been proposed. In this method the *d*-axis grid current component reflecting the power grid side and the signal error of a proportional–integral outer voltage regulator is designed to reflect the change in power caused by the irradiation variation. Hence, with this information, the proposed a method can greatly reduce the power losses caused by the dynamic tracking errors under rapid weather changing conditions[5-8]. The superiority of the newly proposed method is supported by simulation and experimental results.

II. PHOTOVOLTAIC INVERTER

The inverter is the heart of the PV system and is the focus of all utility-interconnection codes and standards. A Solar inverter or PV inverter is a type of electrical inverter that is made to change the direct current (DC) electricity from a photovoltaic array into alternating current (AC) for use with home appliances and possibly a utility grid. Since the PV array is a dc source, an inverter is required to convert the dc power to normal ac power that is used in our homes and offices. To save energy they run only when the sun is up and should be located in cool locations away from direct sunlight. The PCU is a general term for all the equipment involved including the inverter and the interface with the PV (and battery system if used) and the utility grid. It is very important to point out that inverters are by design much safer than rotating generators.

Maximum power point capturing
Main article: Maximum power point tracker

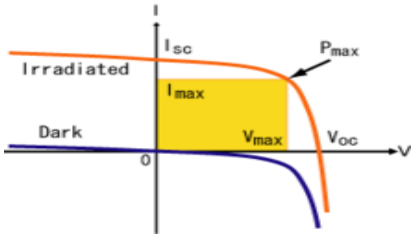

Fig: 1, V-I curve for a solar cell, showing the maximum power point P_{max}.

Maximum power point tracking is a technique that solar inverters use to get the most possible power from the PV array. Any given PV module or string of modules will have a maximum power point: essentially, this defines current that the inverter should draw from the PV in order to get the most possible power .A maximum power point tracker is a high efficiency DC to DC converter that presents an optimal electrical load to a solar panel or array and produces a voltage suitable for the load.

A PV generator is a combination of solar cells, connections, protective parts, supports, etc. In the present modeling, the focus is only on cells. Solar cells consist of a p-n junction; various mode lings of solar cells have been proposed in the literature.
Thus, the simplest equivalent circuit of a solar cell is a current source in parallel with a diode. The output of the current source is directly proportional to the light falling on the cell (photocurrent). During darkness, the solar cell is not an active device; it works as a diode, i.e., a p-n junction. It produces neither a current nor a voltage.

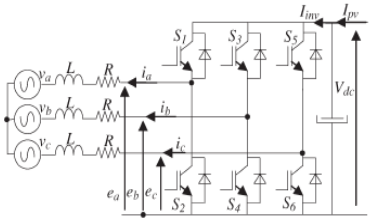

Fig: 2 Three-phase VSI

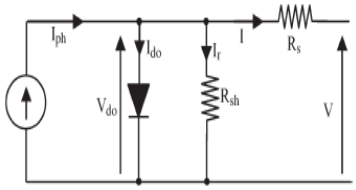

Fig:3 Solar cell electrically equivalent circuit

In general, the output current of a solar cell is expressed by

$$I = I_{ph} - I_o \left(\exp\left(\frac{q}{n.k.T}(V + RsI)\right) - 1\right) - \left(\frac{V+IR_S}{R_{Sh}}\right) \quad (1)$$

In (3), the resistances can be generally neglected, and thus, it can be simplified to

$$I = I_{ph} - I_o \left(\exp\left(\frac{q}{(n.k.T}V\right) - 1\right) \quad (2)$$

If the circuit is opened, the output current $I = 0$, and the open-circuit voltage V_{oc} is expressed by

$$V_{oc} = \left(\frac{n.k.T}{q}\right) \text{In}\left(\frac{I_{ph}}{I_o} + 1\right) \approx \left(\frac{n.k.T}{q}\right) \text{In}\left(\frac{I_{ph}}{I_o}\right) \quad (3)$$

If the circuit is shorted, the output voltage $V = 0$, the average current through the diode is generally neglected, and the short circuit current I_{sc} is expressed by using

$$I_{sc} = I = \frac{I_{ph}}{\left(1+\frac{R_S}{R_{Sh}}\right)} \quad (4)$$

Finally, the output power P is expressed by

$$P = VI = \left(I_{ph} - I_{do} - \frac{V_{do}}{R_{sh}}\right) V. \quad (5)$$

III. CURRENT CONTROLLER

According to(3), V_{oc} strategy guarantees fast transient response and high static performance via internal current control loops.

Current Control

It can be seen that there is cross-coupling between the *d* and *q* components. However, cross-coupling can affect the dynamic performance of the regulator. Therefore, it is very important to decouple the two axes for better performance. This effect can be accomplished with the feed forward decoupling control method. Assuming that

$$V_{rd} = -V_d + d_d V_{dc} + \omega L i_q$$
$$V_{rd} = -V_q + d_d V_{dc} - \omega L i_d \quad (6)$$

where ω is the angular frequency of the utility. Then, the system model is transformed to

$$\begin{cases} \dfrac{di_d}{dt} = -\dfrac{R}{L} i_d + \dfrac{1}{L} v_{rd} \\ \dfrac{di_q}{dt} = -\dfrac{R}{L} i_q + \dfrac{1}{L} v_{rq} \\ \dfrac{dV_{dc}}{dt} = \dfrac{I_{pv}}{C} - \dfrac{V_d + v_{rd}}{CV_{dc}} i_d - \dfrac{V_q + v_{rq}}{CV_{dc}} i_q \end{cases} \quad (7)$$

The cross-coupling variables are eliminated in the aforementioned model. Hence, the currents i_d and i_q can be controlled independently by acting upon inputs V_d and V_q, respectively. Furthermore, by using PI-type regulators, a fast dynamic response and zero steady-state errors can be achieved. Since the switching frequency is much higher than the line frequency, the sampling and hold delay is neglected.

The diagram is suitable for both i_d and i_q loops. From the diagram, the closed-loop transfer function of the *d*, *q* current loops is

$$\frac{i_q(s)}{I_q^*(s)} = \frac{i_d(s)}{i_d^*(s)} = \frac{k_{ip}}{L} \frac{S+\frac{k_{ii}}{k_{ip}}}{S^2+\frac{(k_{ip}+R)}{L}S+\frac{k_{ii}}{L}} \quad (8)$$

PV Power Calculation

In the synchronous rotating frame *d*, *q*, the active and reactive powers of a three-phase grid-connected VSI are given by

$$\begin{cases} P = \frac{3}{2}(v_d i_{d+v_{di_q}}) \\ Q = \frac{3}{2}(v_d i_{d-v_q i_d}) \end{cases} \quad (9)$$

If the three-phase grid voltage is ideally sinusoidal without any harmonics, then in the *d*, *q* frame, the grid voltage vector is given by (24)

In practice, the grid voltage is non sinusoidal due to harmonics. Therefore, both V_d and V_q will not be constant but have slight ripples whose frequencies and magnitudes depend on the harmonic components. However, in steady state, the average value of V_q is still equal to zero. Consequently, (23) can be rewritten as (25).

Its active power depends on the *d*-axis current, and the reactive power depends on the *q*-axis current. Furthermore, in order to achieve unity power factor fundamental current flow, the *q* component of the command current vector is set to zero

$$\begin{cases} P = \frac{3}{2}(v_d i_d) \\ Q = \frac{3}{2}(v_d i_q) \end{cases} \quad (10)$$

Thus, the dc bus-voltage control loop under changing irradiation can be modeled with the block diagram, where the current of PV array is an input disturbance.

Between voltage reference V_{dc}^* and voltage measurement V_{dc} is the following:

$$\varepsilon(s) = A(s)\, V_{dc}^*(s) + B(s)\, i_G(s) \quad (11)$$

$$A(s) = \frac{CS^2}{CS^2 + d_d k_{up} S + d_d k_{ui}}$$

$$B(s) = \frac{d_d}{C} \frac{S}{S^2+\frac{d_d k_{up}}{c}S+\frac{d_d k_{ui}}{c}}$$

If we consider only the impact of perturbation i_G, we can write

$$\varepsilon(s) = \frac{d_d}{C} \frac{S}{S^2+\frac{d_d k_{up}}{c}S+\frac{d_d k_{ui}}{c}} i_G(s) \quad (12)$$

IV. SIMULATION RESULTS

This section presents the simulation results of the classical and the proposed methods in order to validate the performance of the control scheme.

Computer simulation has been done using MATLAB/SIMULINK simulation package

The full diagram of the control methodology and the modulation is shown.

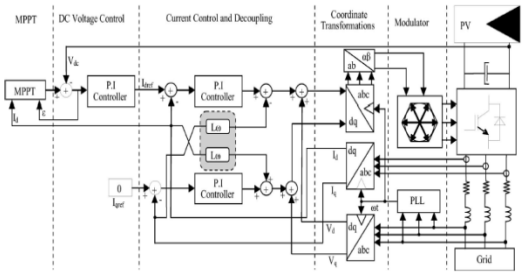

Fig: 4 Grid-connected PV system with the proposed tracker.

PROPOSED MPPT WITH PERTURBATION AND OBSERVATION METHOD.

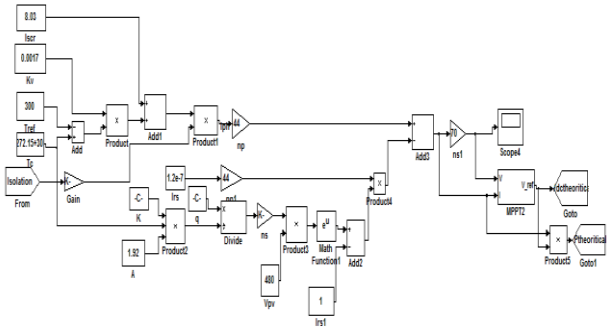

Fig: 5.Theoretical calculations of classical method.

In the conventional method, the MPP is obtained from the PV array power by multiplying the voltage and current of PV arrays and comparing it with the previously measured power.

In the case of a sudden increase in irradiance, the P&O algorithm reacts as if the increase occurred as a result of the previous perturbation of the array operating voltage. The next perturbation, therefore, will be in the same direction as the previous one.

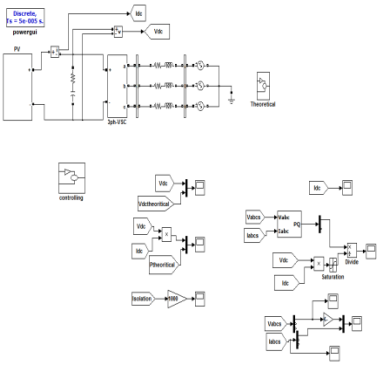

Fig:6 Simulation circuit of classical method

The characteristics of Solar PV module are used for the PV array model in the simulation and experiment. The module

provides 60 W of nominal maximum power and a 21.1-V open-circuit voltage at an irradiation of 1 kW/m2 and an ambient temperature of 25◦C.

To compare the performance of the proposed MPPC method with that of the classical method, the simulations are configured under exactly the same conditions to compare the performances. The PV array in simulation is composed of ten series connected Modules.

In order to verify the effect of rapidly changing irradiation, an irradiation ramp change was used. A 25-s period for the increasing and decreasing ramps was selected. This irradiation change starts from 225 W/m2, stops at 1000 W/m2, waits at this level for 25s, and decreases again back to 225 W/m2 with a constant slope. The temperature is considered constant during the simulation.

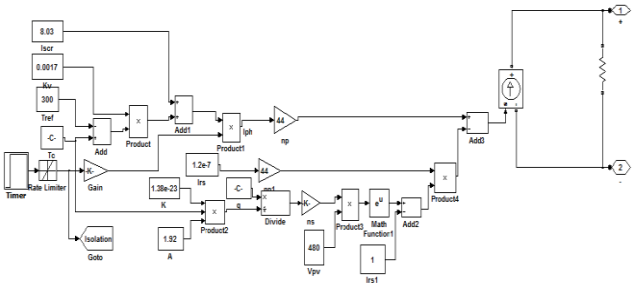

Fig: 7.Designed circuit of photo voltaic system

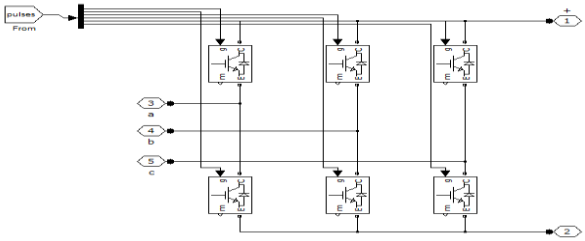

Fig: 8.. Circuit diagram of 3-Ph voltage source converter

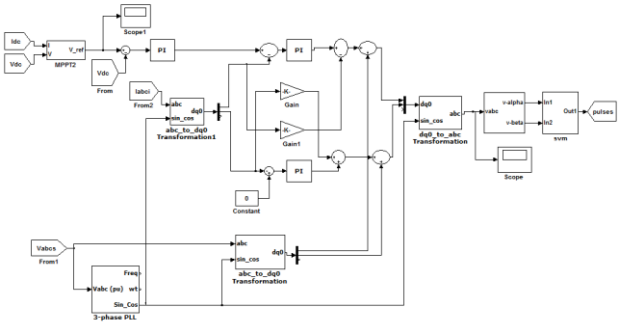

Fig: 9 Controlling circuit of voltage source converter in classical method

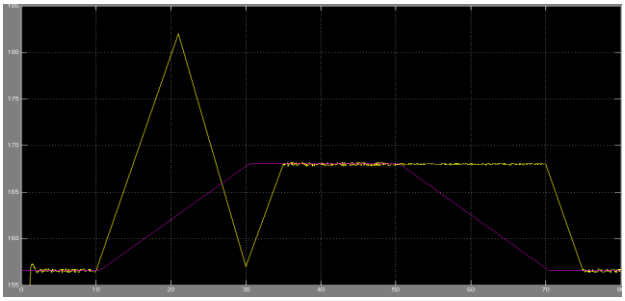

Fig: 10 PV array voltage with classical and theoretical MPP voltage during a trapezoidal irradiation profile.

In above fig. under a decrease of irradiation (60–80 s), we can see that the voltage of PV array varies between *V* and *V* +Δ*v* since it decreases the PV array power in the two directions of perturbation. This is because the power change caused by irradiation decrease in sunshine is greater than the variation caused by the voltage perturbation.

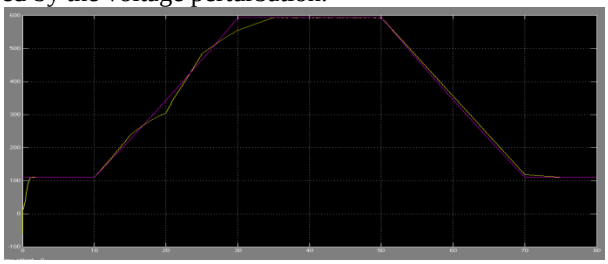

Fig: 11. Simulation measurement of the PV array power during a trapezoidal irradiation profile, using the classical MPPC method, compared to the theoretical MPP power.

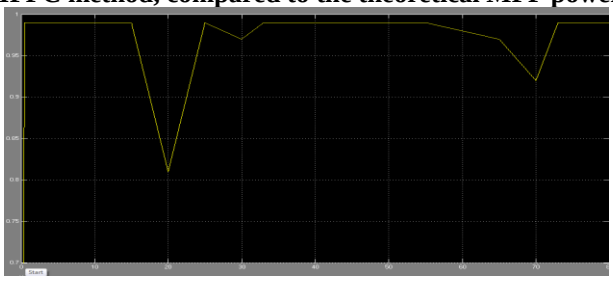

Fig: 12. Simulation measurement of the instantaneous efficiency with classical method

PROPOSED MPPT WITH VOLTAGE-ORIENTED CONTROL METHOD.

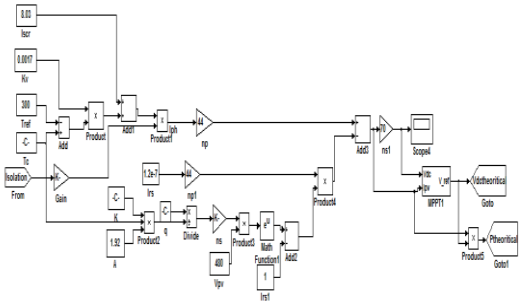

Fig: 13.Theoretical calculations of voltage-oriented control method.

To overcome the limitations of the classical method, the proposed MPPT enables us to decouple the change in power caused by the simultaneous increment perturbation and irradiation variation.

The irradiation variation is estimated by using the signal error of the PI controller of the dc voltage control. The PI regulator is designed to assure zero signal error if the atmospheric conditions are constant and a constant signal error in the opposite case. Hence, the signal error reflects only the change in power caused by the irradiation variation.

After that, in order to calculate the total change in the PV array power, the *d*-axis grid current component is used. Finally, the change in power caused by the previous perturbation is obtained by a simple subtraction; therefore, the correct direction of the MPP can be identified.

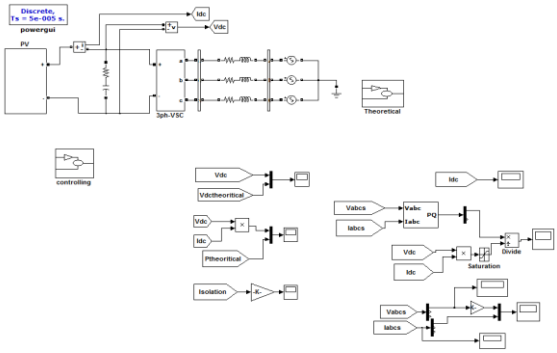

Fig: 14.Simulation circuit of voltage-oriented control method

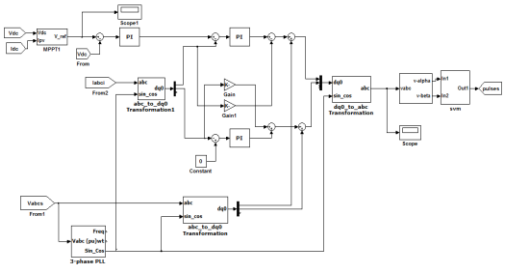

Fig: 15 Controlling circuit of voltage source converter in voltage-oriented control method

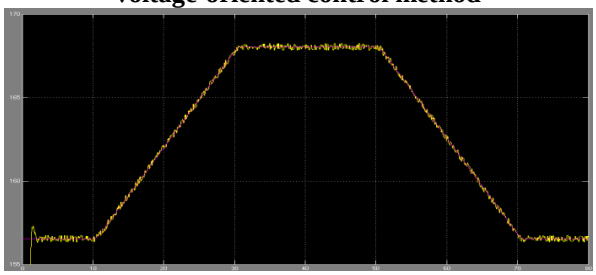

Fig: 16. PV system voltage with the proposed MPPC and theoretical MPP voltage during a trapezoidal irradiation profile.

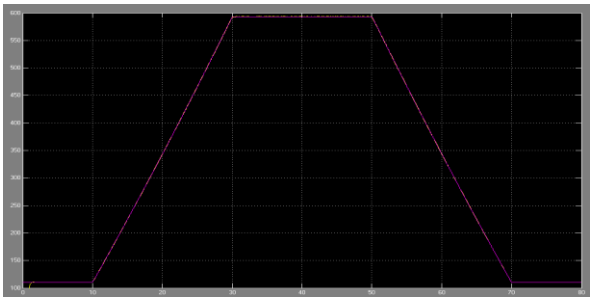

Fig: 17 Simulation measurement of the PV array power during a trapezoidal irradiation profile, using the proposed MPPC method, compared to the theoretical MPP power.

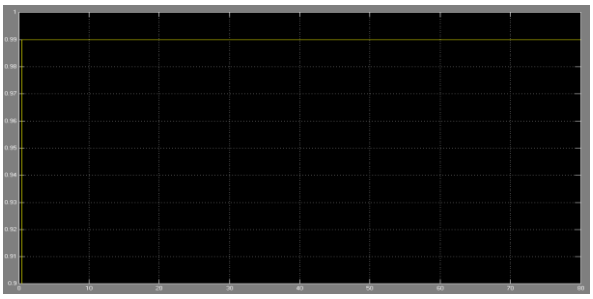

Fig: 18. Simulation measurement of the instantaneous efficiency with proposed MPPC

V. CONCLUSION

1. In order to avoid mistakes in the classical method due to the fast-changing irradiation, this paper has proposed an improved MPPC controller without PV array power measurement. This control scheme uses the *d*-axis grid current component and the signal error of the outer voltage regulator. The robust tracking capability under rapidly increasing and decreasing irradiance is verified experimentally with a PV array emulator.

This MPPC method permits one to differentiate the contribution of increment perturbation and irradiation change in power variation, hence identifying the correct direction of the MPP. In the simulation and experimental results, the robust tracking capability under rapidly increasing and decreasing irradiance has been proved. And overcome the limitations of the classical method. the proposed MPPC enables us to decouple the change in power caused by the simultaneous increment perturbation and irradiation variation. The steady-state and dynamic responses illustrated the perfect desired reference tracking controller. The output power losses caused by the dynamic tracking errors are significantly reduced, particularly under fast changing irradiation. The irradiation variation is estimated by using the signal error of the PI controller of the dc voltage control. The PI regulator is designed to assure zero signal error if the atmospheric conditions are constant and a constant signal error in the opposite case. Hence, the signal error reflects only the change in power caused by the irradiation variation. After that, in order to calculate the total change in the PV array power, the *d*-axis grid current component is used.

Finally, the change in power caused by the previous perturbation is obtained by a simple subtraction; therefore, the correct direction of the MPP can be identified .MPPC method

permits one to differentiate the contribution of increment perturbation and irradiation change in power variation, hence identifying the correct direction of the MPP The robust tracking capability under rapidly increasing and decreasing irradiance has been proved.

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Low Power and Area Efficient Design of VLSI Circuits

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***Abstract*—** In deep submicron technologies, leakage power becomes a key for a low power design due to its ever increasing proportion in chip's total power consumption. Power dissipation is an important consideration in the design of CMOS VLSI circuits. High power consumption leads to reduction in battery life in case of battery powered applications and affects reliability packaging and cooling costs. We propose a technique called LCPMOS for designing CMOS gates which significantly cuts down the leakage current without increasing the dynamic power dissipation. LCPMOS, a technique to tackle the leakage problem in CMOS circuits, uses single additional leakage control transistor, driven by the output from the pull up and pull down networks,which is placed in a path from pull down network to ground which provides the additional resistance thereby reducing the leakage current in the path from supply to ground. The main advantage as compared to other techniques is that LCPMOS technique does not require any additional control and monitoring circuitry, thereby limits the area and also decreases the power dissipation in active state. Along with this, the other advantage with LCPMOS technique is that it reduces the leakage power to an extent of 91.54%,which is more efficient in aspects of area and power dissipation compared to other leakage power reduction techniques.

Index Terms- sub threshold leakage current; LCPMOS; voltage scaling; LCT; self-controlled LCT; deep-submicron.

I. INTRODUCTION

The main sources for power dissipation are: 1) capacitivepower dissipation due to the charging and discharging of theload capacitance; 2) short-circuit currents due to the existenceof a conducting path between the voltage supply and ground for the brief period during which a logic gate makes a transition; and 3) leakage current. The leakage current consists of reverse-bias diode currents and subthreshold currents. The former is due to the stored charge between the drain and bulk of active transistors while the latter is due to the carrierdiffusion between the source and drain of the OFF transistors as shown in fig1.

Digital integrated circuits are found everywhere in modern life and many of them are embedded in mobile devices where limited power resource is available (e.g. mobile phones, watches, mobile computers…). To permit a usable battery runtime, such devices must be designed to consume the lowest possible power. Furthermore, low power is also very important for non-portable devices, too. Indeed reduced power consumption can highly decrease the packaging costs and highly increase the circuit reliability, which is tightly related to the circuit working temperature. Hence, low power consumption is a zero-order constraint for most ICs manufactured today. In fact, higher performance-per-watt is the new mantra for micro-processor chip manufacturers today.In order to achieve high density and high performance, CMOS technology feature size and threshold voltage have been scaling down for decades. Because of this trend, transistor leakage power has increased exponentially. The reduction of the supply voltage is dictated by the need to maintain the electric field constant on the ever shrinking gate oxide. Unfortunately, to keep transistor speed (proportional to the transistor "on" current) acceptable, the threshold voltage must be reduced too, which results in an exponential increase of the "off" transistor current, i.e. the current constantly flowing through the transistor even when it should be "non-conducting".

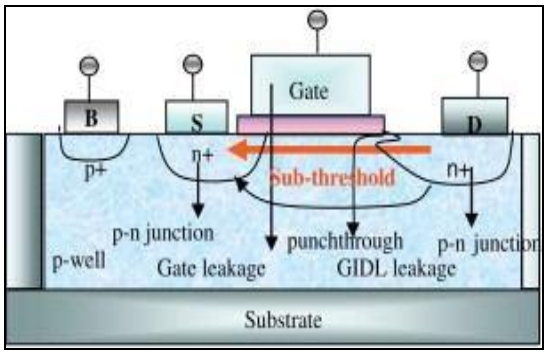


Fig.1:StaticCMOSleakagesources.

As the feature size becomes smaller, shorter channel lengths result in increased sub-threshold leakage current through a transistor increases when it is off as shown in fig2. Low threshold voltage also results in increased sub-threshold leakage current because transistors cannot be turned off completely. For these reasons, static power consumption, i.e. leakage power dissipation has become a significant portion of total power consumption for current and future silicon technologies.To solve the power dissipation problem, many researchers have proposed different ideas from the device level to the architectural level and above.

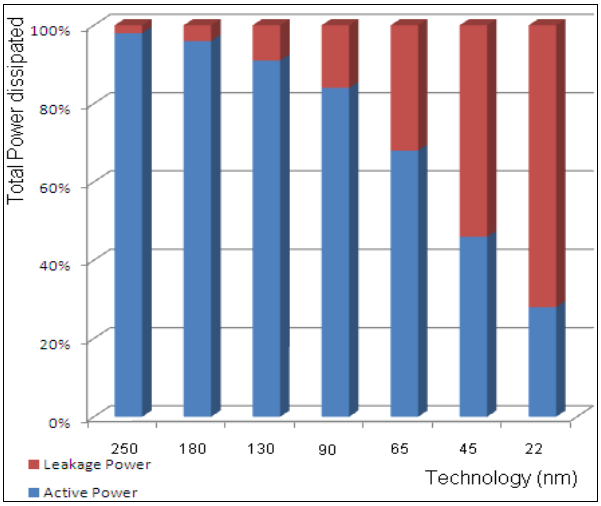


Fig.2:TechnologyVsLeakagePower

In this paper, we describe a new leakage power reduction technique called LCPMOS (Leakage Control PMOS) for designing CMOS circuits. The rest of the paper is organized as follows. Section II describes briefly the prior works on leakage power reduction and their limitations. Section III introduces the transistor models used for estimating the leakage power. Our design strategy and an approach for minimizing the area overhead are described in Sections IV. Results are presented in Section V, followed by conclusions in Section VI respectively.

I. LIMITATIONSWITH RELATEDWORK

A. MTCMOS

A high-threshold NMOS gating transistoris connected between the pull-down network and the ground, and low-threshold voltage transistors are used in the gate.The reverse conduction paths exist, which tends the noisemargin to reduceormay result in completef ailure of the gate.There also exists a performance penalty due to the high-threshold transistors in series with all the switching current paths.

Dual V_T technique is a variation in MTCMOS,in which the gates in the criticalpath use low-threshold transistors and high-thresholdt ransistors for gates in non-criticalpath[3],[7].Both the methods requires additional mask layers for each value of V_T in fabrication,which is a complicated task depositing two different oxides thickness, hence making the fabrication process complex. The techniques also suffer from turning-on latency i.e., the idle of circuit cannot be used immediately after reactivated since sometime is needed to return to normal operating condition. The latency is typically a fewcycles for former method, and for Dual technology, is much higher.When the circuitis active, these techniques are not effective in controlling the leakagepower.

B. SLEEP TransistorTechnique

This is a State-destructive technique which cuts off either pull-up or pull-down or both the networks from supply voltage or ground or both using sleep transistors. This technique is MTCMOS, which adds high-V_{th} sleep transistors between pull-up networks and V_{dd} and pull-downnetworks and gnd while for fast switching speeds,low-V_{th} transistors are used in logic circuits[8]. this technique dramatically reduces leakagepower during sleep mode. However, the area and delay are increased due to additional sleep transistors. During the sleepmode, the state will be lost as the pull-up and pull-down networks will have floating values. These values impact the wakeup time and energy significantly due to therequirement to recharge transistors which lost state during sleep.

C. ForcedStack

In this technique, every transistor in the network is duplicated with both the transistors bearing half the original transistor width[6].Duplicated transistors cause a slight reverse bias between the gate and source when both transistors are turnedoff. Because sub-threshold current is exponentially dependent on gate bias, it obtains substantial current reduction. It overcomes the limitation with sleep technique by retaining state but it takes more wakeup time

D. ZIGZAGTechnique

Wake-up cost can be reduced in zigzag technique but still state losing is a limitation. Thus, any particular state which is needed upon wakeup must be regenerated somehow. For this, the technique may need extra circuitry to generate a specific input vector.

E. SLEEPYSTACK Technique

This technique combines the structure of the Forcedstack technique and the sleeptransistor technique. In the sleepy stack technique, one sleep transistor and two half sized transistors replaces each existing transistor[10]. Although using of W0/2 for the width of the sleeptransistor, changing the sleep transistor width may provide additional tradeoffs between delay,power and area.It also requires additional control and monitory circuit,for the sleep transistors.

F. LEAKAGE FEEDBACKTechnique

This technique is based on the sleep approach. To maintain logic during sleep mode, the leakage feedback technique uses two additional transistors and the two transistors are driven by the output of an inverter which is drivenby output of the circuit implemented utilizing leakage feedback. Performance degradation and increase in area are the limitations along with the limitation of sleep technique.

G. SLEEPYKEEPER Technique

This technique consists of sleep transistors connected to the circuit with NMOS connected to V_{dd} and PMOS to Gnd. This creates virtual power and ground rails in the circuit, which affects the switching speed when the circuitis active [9]. The identification of the idle regions of the circuit and the generation of the sleep signal need additional hardware capable of predicting the circuit states accurately, increasing the area requirement of the circuit. This additional circuit consumes power throughout the circuit operation to continuously monitor the circuit state and control the sleep transistors even though the circuit is in an idle state.

H. LECTOR Technique

This technique consists of two self controlled transistors which increases the resistance in the path from source to ground, which increases the area of the circuit, one of the most important constraint in the design of vlsi circuits.

II. LCPMOS

In this proposed technique, we introduce a single leakage control transistor within the logic gate for which the gate terminal of leakage control transistor (LCT) is controlled by the output of the circuit itself. Which increases the resistance of the path from pull down network to ground thereby increasing the resistance from $V_{dd}$to ground, leading to significant decrease in leakage currents. The main advantage as compared to other techniques is that LCPMOS technique does not require any additional control and monitoring circuitry, thereby limits the area and also the power dissipation in active state.

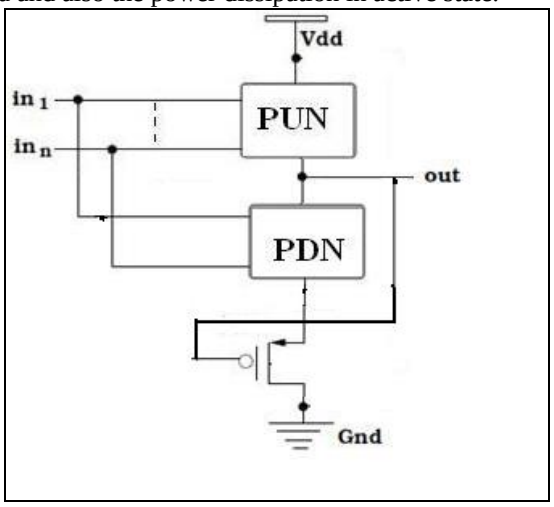


Fig.3. LCPMOSCMOSGate

The topology of a LCPMOS CMOS gate is shown in Figure 5. OneLCTs are introduced betweennodes N1 and Gnd. The gate terminal of LCT is controlled by the output of the circuit itself. As LCT is controlled by output , no external circuit is needed; thereby the limitation with the sleep transistor technique has been overcome. The introduction of LCT increases the resistance of the path from V_{dd} toGnd, thus reducing the leakage current.

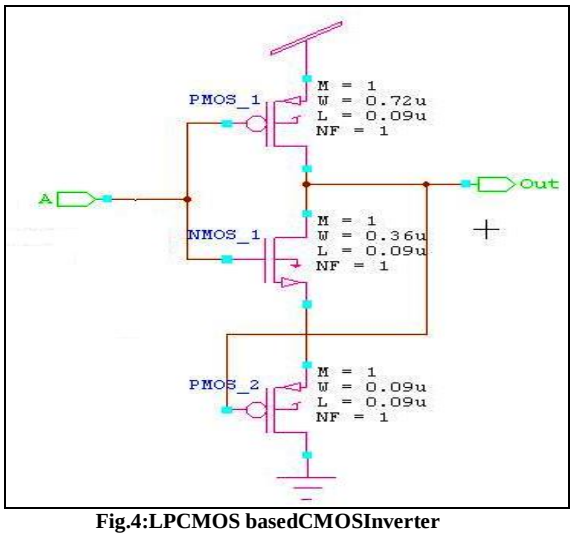


Fig.4:LPCMOS basedCMOSInverter

Leakage ControlPMOS(LCPMOS) technique is illustrated in detail with the case of an inverter. A LCPMOS INVERTER is shown in Figure6. A PMOS is introduced as LCT between N1 and Gnd nodes of inverter.

WhenV_{dd}=1V, input A=0, the output is high. As the output drives the LCT the LCT goes to OFF state hence provides high resistance path between V_{dd} and Gnd. When A=1,theoutput is low; hence LCT will be in ON state hence output is low. LCPMOS inverter for all possible inputs are tabulated in TableI.

TABLEI. STATEMATRIXOFLCPMOSINVERTER

Transistor Reference	**InputVector(A)**	
	0	**1**
M1	ONState	OFFState
M2	OFFState	ONState
LCT	Near Cut-OFF State	ONState

In the sleep related technique, the sleep transistors have to be able to isolate the power supply and/or groundf rom the rest of the transistors of the gate. Hence, they need to be made bulkier dissipating more dynamic power. This offsets the savings yielded when thecircuitisidle. Sleep transistor technique depends on input vector and it needs additional circuitry to monitor and control the switch in sleep transistors, consuming power in both active and idle states. In comparison, LCPMOS generates the required control signals with in the gate and is also vector independent.

Single transistor is added in LCPMOS technique in every path from V_{dd} to Gnd irrespective of number of transistors in pull-up and pull-down network. Where as, forced stack save 100% area overhead. The loading requirement with LCT is a constant which is much lower.

IV.APPLYINGLCPMOSTO CMOSCIRCUITS

Various circuit applications of the LCPMOS technique are explored in this section. The LCPMOS technique is applied to the following CMOS circuits and also the irrespective basecase are implemented to calculate the amount of leakage power

reduced in LCPMOS technique.

A. *LCPMOS basedNOT gate*

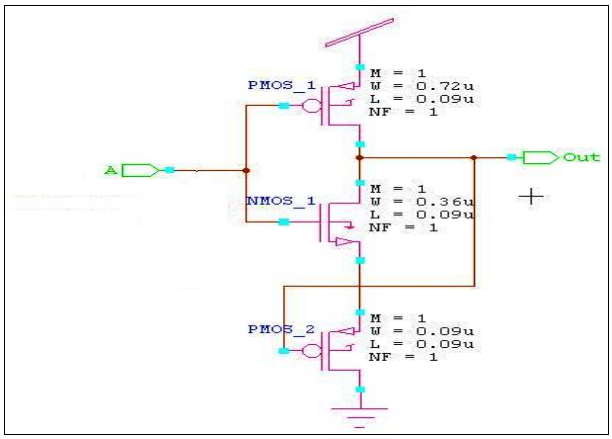


Fig.5:2-inputLCPMOSNAND

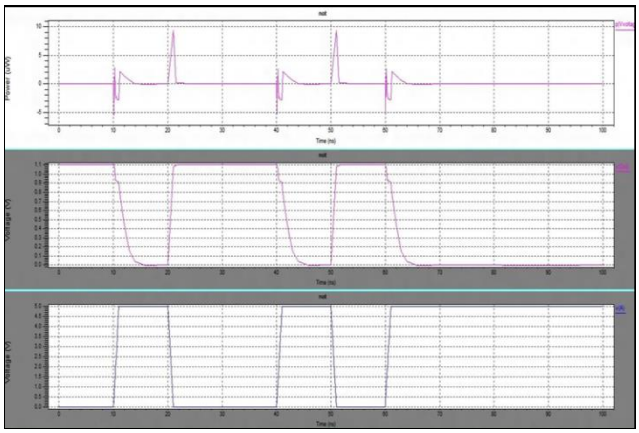


Fig.6:SimulationwaveformsofLCPMOS NOT

The CMOS INVERTER is shown in Figure5 with the one LCT added between pull-down network and gnd. The simulation waveforms of LCPMOS NOT from Figure8 show that the basic characteristics of NOT are retained by LCPMOS NOT.

B. *LCPMOS basedNANDgate*

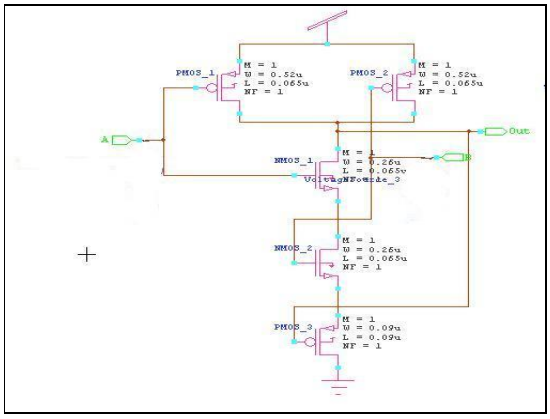


Fig.7:2-inputLCPMOSNAND

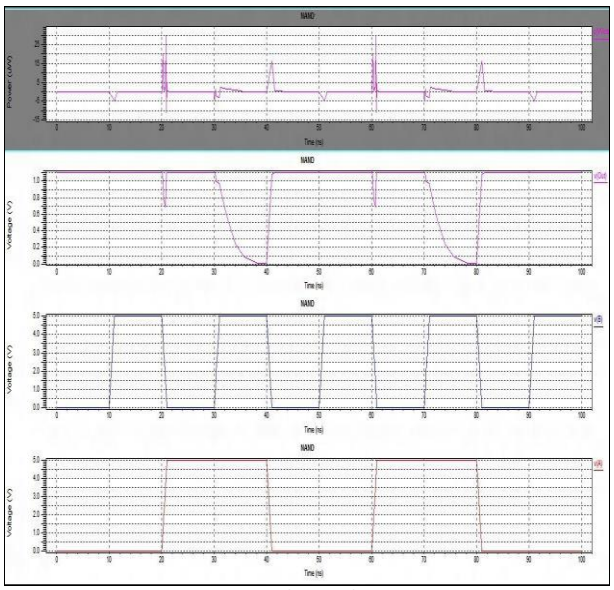


Fig.8:SimulationwaveformsofLCPMOS NAND

The 2-input CMOS NAND gate is shown in Figure7with the one LCT added between pull-down network and gnd. The simulation wave forms of LCPMOS NAND from Figure8 show that the basic characteristics of NAND are retained by LCPMOS NAND.

C . LCPMOS basedNORgate

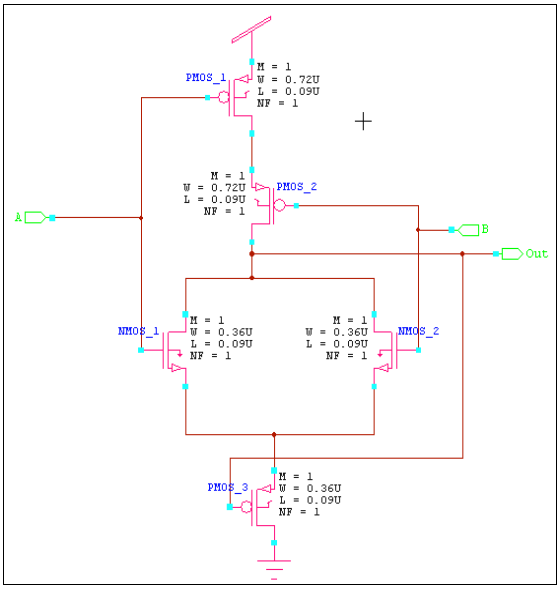
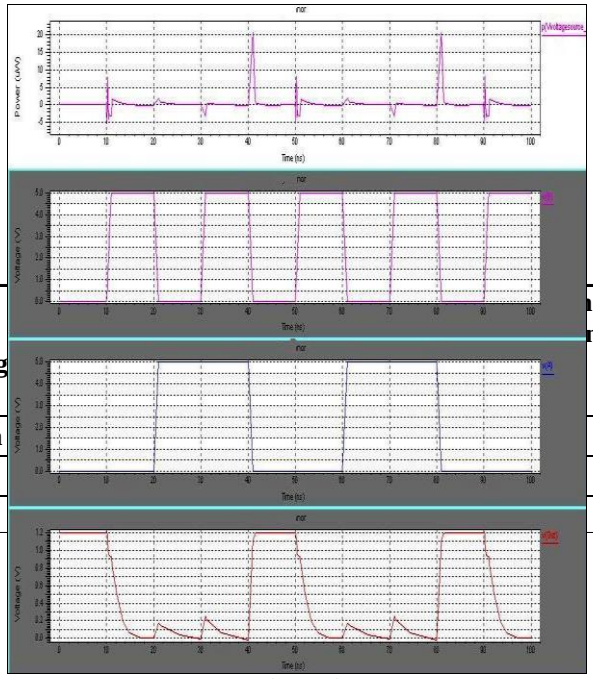


Fig.9:2-inputLCPMOSNOR



Technolog			n
180nm			
90nm			
65nm			

Fig.10:SimulationwaveformsofLCPMOS NOR

The 2-input CMOS NOR gate is shown in Figure8 with the one LCT added between pull-down network and gnd. The simulation waveforms of LCPMOS NOR from Figure8 show that the basic characteristics of NOR are retained by LCPMOS NOR.

V. EXPERIMENTALRESULTS

The leakage power is measured using the S-EDIT simulator. The results obtained through the technique for NOT gate is shown in TableIII. Simulation for the NOT is performed by taking three different process parameters Viz.180nm, 90nm, 65nm.

TABLEII. NOT RESULTSFOR VARIOUSTECHNOLOGIES

Technology	Leakagepower (uW)			%age decreasein power dissipation (LCPMOS)
	BASE CASE	LECTOR	LCP MOS	
180nm	130	78	39	70
90nm	110	31	9.3	91.54
65nm	98	5	3.8	95.4

Table III gives the results for 2-input NAND for,180nm, 90nm and 65nm technologies. Table IV gives the results for 2-input NOR for ,180nm, 90nm and 65nm technologies.

TABLEIII. NAND RESULTSFOR VARIOUSTECHNOLOGIES

Technology	Leakagepower (uW)			%age decreasein power dissipation (LCPMOS)
	BASE CASE	LECTOR	LCP MOS	
180nm	140	90	70	50
90nm	125	37	30	76
65nm	115	75	12.5	89.13

TABLEIV. NORRESULTSFOR VARIOUSTECHNOLOGIES

Leakage power dissipation is taken as the average of power dissipations obtained at all the possible input vectors of the CMOS circuit. There are 4 possible combinations for 2-input NAND, hence the averageof the four powerdissipations gives the leakage power. In each case, the leakage power is measured by exciting the circuits for 3cases (Conventional and LECTOR and LCPMOS) with same set of input vectors.

VI.CONCLUSION

The increase in leakage power because of the scaling down of device dimensions, supply and threshold voltages in order to achieve high performance and low dynamic power dissipation, becomes more with the deep-submicron and nano meter technologies and thus it becomes a great challenge to tackle the problem of leakage power. LCPMOS uses one LCT which is controlled by the output of circuit itself. LCPMOS achieves the reduction in leakage power compared to other leakage reduction techniques,such as LECTOR, sleepy stack, sleepykeeper, etc, along with the advantage of not affecting the dynamic power, since this technique does not require any additional control and monitor circuitry and also in this technique, the exact logic state is maintained. The LCPMOS technique whenapplied to generic logic circuits achieves up to 80-92% leakage reduction over the respective conventional circuits without affecting the dynamic power. A tradeoff between Propagation delay and area overhead exists here.

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A Verilog Model of Universal Scalable Binary Sequence Detector

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Abstract- This paper presents a Verilog based Universal Sequence Detector, which will be able to detect a binary sequence, from a sequence of inputs. The Sequence Detector looks for some specified sequence of inputs and outputs 1, whenever the desired sequence has found. The sequence detector is like a lock which unlocks (outputs 1), only when a combination appears. Coding of design is done in Verilog HDL and the design is tested and simulated in ModelSim Simulator and is implemented on Xilinx Virtex 4 XC4VFX12 FPGA device.

Index Terms- FPGA, Xilinx, ModelSim, Sequence Detector, HDL.

I. INTRODUCTION

In a computer network like Ethernet, digital data is sent one bit at a time, at a very high rate. Such a movement of data is commonly called a bit-stream. One characteristic is unfortunate, particularly that any one bit in bit stream looks identical to many other bits. Clearly it is important that a receiver can identify important features in a bit stream. As an example, it is important to identify the beginning and ending of a message. This is the job of special bit sequences called flags. A slag is simply a bit sequence that serves as a marker in the bit stream. To detect a flag in bit stream a sequence detector is used.

In this paper we have developed a universal sequence detector. The design implemented in Verilog HDL Hardware Description Language. It was simulated using ModelSim simulator and then is tested for the validation of the design on Virtex 4 XC4VFX12 FPGA.

A) Field Programmable Gate Array

Field Programmable Gate Array (FPGA) is a reconfigurable hardware platform useful for the implementation of high digital functions. Using fixed point, parallel computational structures, FPGA provides computational speeds as much as 100 times greater than those possible with Digital Signal Processors (DSP). The extremely fast computational capability of FPGAs allows a few microseconds for real-time computation of algorithms in spite of their complexities. Furthermore, as DSPs, FPGAs are very low cost components. Virtex 4 XC4VFX12 FPGA's are ideal for low-cost, high-volume applications and are targeted as replacements for fixed-logic gate arrays. The Virtex 4 XC4VFX12 FPGA is not only available for a very low cost, but it integrates many architectural features associated with high-end programmable logic. This combination of low cost and integrated features has made it an ideal replacement for ASICs (Application Specific Integrated Circuits).

B) ModelSim Simulator

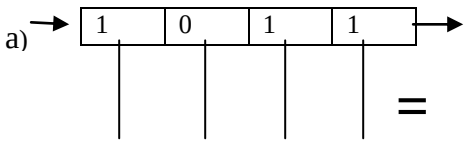
ModelSim simulator is a Mentor Graphics product which is capable to simulate various programs of Hardware Description Languages. It combines single kernel simulator (SKS) technology with a unified debug environment for Verilog, VHDL, and SystemC. The combination of industry-leading, native SKS performance with the best integrated debug and analysis environment make ModelSim the simulator of choice for both ASIC and FPGA designs.

II. SYSTEM IMPLEMENTATION

Binary sequences are inserted at the beginning (or end) of the data frame or sub frame emanating from a digital data processor of a space craft. Sequence detectors are used in the detecting equipment on the ground to provide flags which indicate the beginning (or end) of a data block. The sequence detector is in essence an electronic combination lock which is opened for one digit period only when the proper sequence of binary digits is entered. In general sequence detectors can be designed using the state machines. Those detectors which were designed using state machines are limited to detect a particular sequence. As the state machine varies from one required sequence to other, it will be difficult to design a universal detector which will able to detect any sequence of any length.

Universal Scalable Sequence detector design can be achieved if we use a scalable shift register. So we have designed Detector using scalable shift register, which stores the input sequence and compares it with the required sequence to be detected. Scalable register in the sense that, the register facilitates us in storing any number of bits. In general, the required sequence that should be detected is stored in internal register.

Whenever the input binary sequence is applied to the shift register, it takes the sequence bit by bit for every clock cycle and shifts to the right in the next clock cycle. If the sequence resting in the shift register matches with the required sequence, present in the internal register, an output pulse will be generated, acknowledging the user about the detection of the sequence.



outpu

b	1	0	0	1

Figure1: Representation of Detection of Required Sequence

The representation of detection of required sequence is shown in the above figure. The first block a) containing four memory spaces is the scalable shift register and the second block b) containing four memory spaces represent an internal register. Here in this case the sequence to be detected is 1001, which will be stored in internal register. The Scalable register takes the size of internal register. The major sequence from which the require sequence is to be detected will be carried from the LSB of the shift register to the MSB of it. Whenever the binary contents of the shift register matches with that in the Internal register, an output pulse will be generated. Similarly we can be able to detect any sequence of any length. The sequence to be detected will be stored in the Internal register, shift register of same memory will be generated and then compared, finally detecting the required sequence.

III. SIMULATION RESULTS

The design is simulated in ModelSim PE student Edition Figure 3 shows the timing waveform of the design obtained with ModelSim PE student Edition Simulator for a 8 bit sequence. Design is implemented on Xilinx Virtex 4 XC4VFX12 FPGA device. The top level of RTL schematic obtained by synthesizing the design in Xilinx is shown in Figure 2 and the pins are described in Table 1.

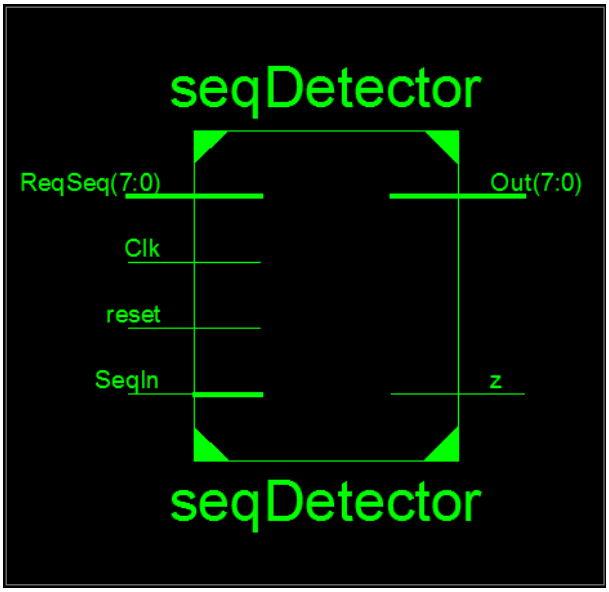


Figure 2: Top level block diagram of universal Scalable binary sequence detector

Table 1: Pin Description

Input Signal	**Description**
ReqSeq(7:0)	Required Sequence to be detected here in this example it is of 8 bit
Clk	System Clock signal input
reset	Reset signal input
Seqin	Input binary sequence from which required sequence to be detected
Output signal	**Description**
Out(7:0)	Shift register output at every clock
z	The Output if sequence is detected

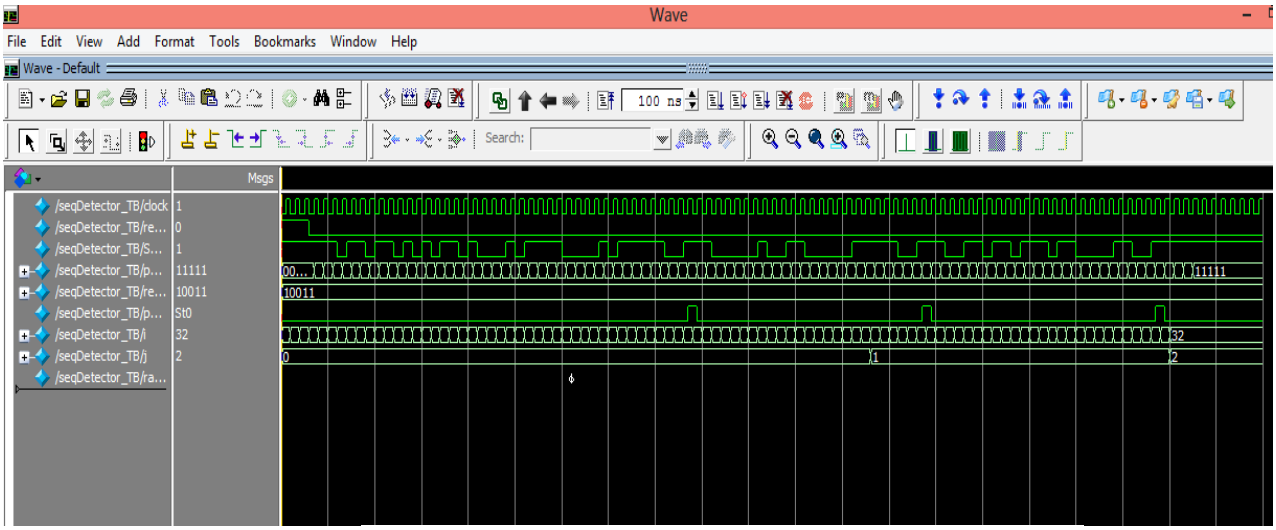


Figure 3: Simulation results obtained from Modelsim simulator

The percentage utilization of FPGA in terms of number of slices, number of flip-flops and number of Look up tables (LUTs) are evaluated .The Device utilization summary obtained from Xilinx is shown in Figure 4

Device Utilization Summary			
Logic Utilization	**Used**	**Available**	**Utilization**
Number of Slice Flip Flops	41	10,944	1%
Number of 4 input LUTs	62	10,944	1%
Number of occupied Slices	52	5,472	1%
Number of Slices containing only related logic	52	52	100%
Number of Slices containing unrelated logic	0	52	0%
Total Number of 4 input LUTs	93	10,944	1%
Number used as logic	62		
Number used as a route-thru	31		
Number of bonded IOBs	20	320	6%
IOB Flip Flops	1		
Number of BUFG/BUFGCTRLs	1	32	3%
Number used as BUFGs	1		
Average Fanout of Non-Clock Nets	2.46		

Figure 4: Device utilization summary

The complete schematic of the design showing all the inputs and outputs at the inside level is shown in Figure 5. In this example it detects the binary sequence of 8 bits. It includes AND gates, Magnitude comparators, D flipflops , NOT gates.

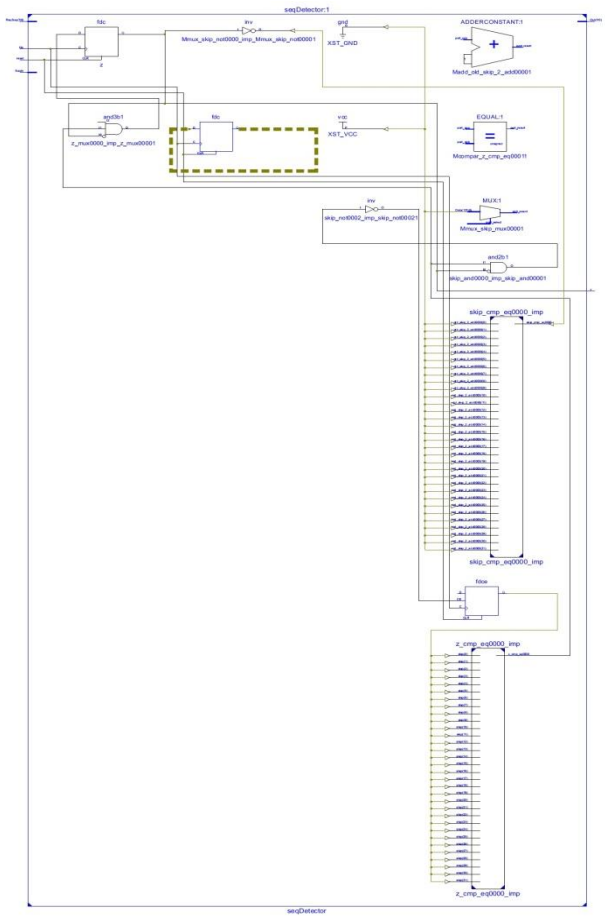


Figure 5: Internal view block diagram of universal scalable binary sequence detector

IV. CONCLUSION

In this paper we have developed a universal scalable binary sequence detector. The design implemented in Verilog HDL Hardware Description Language. It was simulated using ModelSim simulator and then is tested for the validation of the design on Virtex 4 XC4VFX12 FPGA board and the design is proved to be efficient.As the name suggests it is sacalabe and can be used to detect the sequence of any length within the storage capacity of register.

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Enhanced Antibiofilm Activity of Chitosan Stabilized Chemogenic Silver Nanoparticles Against Escherichia Coli

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Abstract- Microbial biofilms, which are formed when bacterial and/or yeast cells adhere to abiotic and biotic surfaces represent the most prevalent type of virulent factor involved in the crucialdevelopment of clinical infection exhibiting resistance to antimicrobial agents. Metalnanotechnology chemistry has the potential to prevent the formation of these life-threatening biofilms on life supporting devices. The present study is undertaken to evaluate the antibiofilm effect of metallic silver nanoparticles (AgNP) stabilized with biocompatible chitosan polymer against biofilm of clinical isolate of Escherichia coli and their effect on biochemical composition of biofilm matrix in terms of total carbohydrate and total protein under in vitro condition. Free and chitosan stabilized AgNPs were characterized by Fourier Transform Infrared Spectroscopy(FTIR), Transmission Electeron Microscopy (TEM), Scanning Electron Microscopy (SEM) and EDS (Energy Dispersive Spectroscopy). Both free as well as the stabilized nanoparticles were found to be effective in bringing about the inhibition of biofilm formation. But enhanced biofilm inhibitory effect inhibition was observed in case of polymer stabilized nanoparticles. Biochemical composition of biofilm matrix in terms of total carbohydrates and total proteins was appreciably reduced in all tested concentration of polymer stabilized nanoparticles compared to free AgNPs.

Index Terms- Silver Nanoparticles; Chitosan; *Escherichia coli*; Biofilm; Matrix,Inhibition

I. INTRODUCTION

Nanobiotechnology, the convergence of nanotechnology and biotechnology and in particular its applications in the medical sector are considered as one of the most promising and most advanced areas of nanotechnology [1] The application of nanotechnology in the field of healthcare has come under great attention in recent times. There are many treatments today that take a lot of time and are also very expensive. Using nanotechnology, quicker and much cheaper treatments can be developed. By performing further research on this technology, cures can be found for diseases that have no cure today. The application of such a technology can be used for the inhibition of biofilm formation on the surgical and medical devices which are of higher threat in the process of treatments. Microbial biofilms develop when microorganisms irreversibly adhere to a submerged surface and produce extracellular polymers that facilitate adhesion and provide a structural matrix [2, 3]. This surface may be inert, nonliving material or living tissue.Biofilm-associated microorganisms behave differently from freely suspended organisms with respect to growth rates and ability to resist antimicrobial treatments and therefore pose a public health problem [4, 5]. Due to increasing tolerance of the biofilm community to antibiotics,biocides and mechanical stress, it has become difficult to completely eradicate mature biofilms.Common treatments to prevent or remove biofilm include disinfection, minimizing nutrients in the feed or altering surface materials to prevent bacterial attachment or clean-in-place (CIP) to remove mature biofilm by chemical or mechanical shear.Nevertheless, nanoscale materials have recently appeared as one of the most promising strategies to control biofilm infections related to indwelling medical devices, especially due to their high surface area to volume ratio and unique chemical and physical properties [6]. A nanomaterial has a diameter ranging from1 and 100 nm, and they can be made from different materials like copper, zinc, titanium, magnesium, gold, alginate and silver. The use of silver nanoparticles (NPs) is now considered as one of the most promising strategies to combat biofilm infections related to indwelling medical devices [7].Another important factor is the protection provided by the encapsulation of the drug in the biological milieu, decreasing toxicity and allowing the drug to reach the specific site [8].Chitosan a natural polymer has been reported as a polymer-based protective agent to stabilize the metal nanoparticles [9]. Because of the biocompatibility, biodegradability, nontoxicity and adsorption properties of chitosan, it was used as a stabilizing agent to prepare Ag, Au and Pt nanoparticles. These chitosan-protected nanoparticles can be easily integrated into systems relevant for pharmaceutical, biomedical, and biosensor applications. Therefore, it has attracted considerable interest due to its medicinal properties, such as antifungal, antibacterial, antiprotozoan, anticancer, antiplaque, antitartar, hemostatic, wound healing and potentiates anti-inflammatory response, immunopotentiation, antihypertensive, serum cholesterol lowering,immune enhancer, increases salivary secretion (anti-xerostomial) and helps in the formation of bone substitute materials [10]. Hence an attempt has been made in this study to evaluate anti biofilm activity of biocompatible chitosan stabilized silver nanoparticles against clinical isolate of *Escherichia coli* under *in vitro* conditions.

II. MATERIALS AND METHODS

Synthesis of silver nanoparticles

Silver nanoparticles were synthesized by chemical reduction of 0.1 M silver nitrate with 0.1 M sodium borohydride as reducing agent[11].Synthesis of silver nanoparticles (AgNPs) was confirmed by the conversion of the reaction mixture into brown colour.Further characterization of the synthesized AgNPs was carried out with the determination of plasmon absorption maxima using UV-Visible spectroscopy (Thermoscientific Spectrascan UV 2700 spectrophotometer) operating in the absorbance mode. The particle morphology was recorded by Scanning Electron Microscopy (SEM) by using Carl Zeiss subra Germany equipped with an energy dispersive spectroscopy (EDS) capability. Synthesized nanoparticles were purified by successive centrifugation at 10,000 rpm and the collected pellets were washed thrice with deionized water, the washed suspension thus obtained was freeze dried.

Synthesis and characterization of chitosan coated AgNPs nanoparticles

For the synthesis of chitosan stabilized silver nanoparticles (CS-AgNPs), 5ml of 0.1 M silver nitrate, 1ml of 0.1 M sodium borohydride and 10 ml of a solution containing chitosan (6.92 mg mL-1) were mixed and stirred for 3 hours to obtain a homogeneous solution. The homogenous solution thus obtained was transferred to the screw cap vial and incubated for 12 h at 95°C [12].The colour of the solution changed from colourless toil ght yellow and finally to yellowish brown which primarily confirmed the coating of chitosan onto silver nanoparticles.

Bacterial strain and growth condition

Clinical isolate of *Escherichia coli* was obtained from Madurai Medical college hospital,Madurai, Tamil Nadu, India. The strain was isolated from patient with severe urinary tract infection and maintained on the slope of tryptic soy agar slant. *E. coli* from the slant was inoculated in to tryptic soy broth (Hi media) and incubated in an orbital shaker at 150 rpm at 37□C for 24 hours. Cells were collected by centrifugation and the collected cell pellets were washed twice in Phosphate buffered saline (PBS) and resuspended in PBS to give a suspension with an OD value of 0.8 (Standard MC value) at 520 nm.

Biofilm inhibition assay in microtitre plates

Freeze dried free and the polymer stabilized AgNPs were dissolved in deionised water at different concentrations (10, 25, 50, 75 and 100 µg /ml) in sterile screw cap vials, shaken well to obtain complete homogenous mixture and were used for biofilm inhibition assay. Biofilm inhibition was carried out in 96 well plate and on nitrocellulose membrane adopting modified spectrophotometric assay described below [12]. 100µl of *E coli* cell suspension was prepared and added into 96 well titre plate and different concentrations of nanoparticles (25, 50, 75 and 100 µg/ml) were added and the plates were at incubated 37º C for 3 days. After the incubation, the liquid suspension was removed and 100 µl of 1% w/v aqueous solution of crystal violet was added. Following staining at room temperature for 30 minutes, the dye was removed and the wells were washed thoroughly and 95% ethanol was added and incubated for 15 minutes. The reaction mixture was read spectrophotometrically at 570 nm. Inhibition of biofilm formation was calculated by using the following formula.

% inhibition = (OD Control –OD Treated) / OD Control × 100

Biofilm inhibition assay on nitrocellulose membrane

Sterile nitrocellulose membrane filter (Rankem, New Delhi, India) with 47 µm and 0.45 µm dia was placed in the 6 well tissue culture plate to which 100 µl of cell suspension of *E. coli*,free and coated nanoparticles were added separately in the above mentioned concentrations and incubated for 3 days at 37□C. Five replicates including a control were maintained. After the incubation period, the inoculated filter was taken and the biofilm inhibition assay was carried out by the modified spectrophotometric inhibition assay described earlier. The filter was stained with 1.0% crystal violet and incubated for 1 hour. After staining, the filter was washed thoroughly with 1 % ethanol and the washed solution was collected in sterile screw cap vial and the reaction mixture was read at 570 nm. Biofilms were examined by SEM after processing of samples by a freeze-drying technique [13] .Biofilms formed on membrane were fixed with glutaraldehyde (2.5%, v/v, in 0.1 M cacodylate buffer, pH 7.0), washed thrice in distilled water and then plunged into a liquid propane/isopentane mixture (2: 1, v/v) at 2196 µC before freeze-drying under vacuum (1026 torr, 1.361024 Pa). Samples were finally coated with gold and palladium and viewed under Carl zeiss subra (Germany) scanning electron microscope.

Isolation and biochemical analysis of biofilm matrix

Biofilm matrix material was isolated from the microtitre plate and nitrocellulose membrane as previously described [14] .Adherent biofilms were transferred to screw cap bottles containing 10 ml distilled water. The bottles were sonicated for 5 min in an ultrasonic water bath and vortexed vigorously for 1 min to disrupt the biofilms. Cell suspensions were then pooled and centrifuged. The collected supernatant was used as source for studying biochemical composition in terms of total protein determined by Lowry et al. and total carbohydrate by Dubois et al [15]

III. RESULTS AND DISCUSSION

The harsh and potentially fatal consequences of microbial biofilm infections generated efforts to prevent their formation, particularly on indwelling medical devices using chemical and mechanical approaches [16]. Biofilms, the predominant mode of device-related microbial infection, exhibit resistance to antimicrobial agents. They can serve as hides for disease and are often associated with high level antimicrobial resistance of the associated organisms. The metallic nanoparticles are thoroughly being explored and extensively investigated as potential antimicrobials [16]..Biofilm inhibitory effect of metallic nanoparticles against pathogenic bacteria specifically against the biofilms formed by them has been gaining importance recently as the indwelling device-related infections constitute a major cause of morbidity and mortality in hospitalized patients [17]. While studies on the biofilm inhibitory effect of silver nanoparticles are going at an accelerated rate [18, 19, 20] their stabilization using biocompatible polymer for increased antibiofilm activity still needs more attention.

Synthesis and characterization of free AgNPs

Silver nanoparticles synthesis adopting chemical reduction was primarily confirmed by colour change of the reaction mixture from pale yellow to brown which clearly indicates the formation of nanoparticles (Fig. 1a Inset). The characteristic brown colour due to the excitation of plasmon vibrations in the nanoparticles provides a convenient signature of their formation.Synthesized AgNPs were characterized by UV-Vis spectroscopy which reveals a strong broad surface plasmon peak located at 420 nm (Fig. 1). Particle morphology and size recorded by SEM analysis reveals smooth, spherical particles with the size in the range of 19-44 nm (Fig. 2a). The elemental composition of the sample was disclosed by EDS analysis in which strong signals of silver were observed (42.44% in mass) at 3 keV, while weaker signals from C, O, Al and S were also recorded confirming the presence of silver nanoparticles (Fig. 2b).

Synthesis and characterization of chitosan stabilized AgNPs

The AgNPs synthesized were coated with the biopolymer chitosan which acts as a stabilizing agent. The rationale behind selecting chitosan as stabilizing agent is that it shows unique polycationic, chelating and film forming properties as it is an oxygen rich linear polysaccharide having active amino and hydroxyl groups [21]. Therefore, chitosan exhibits a number of interesting biological activities such as biocompatibility, biodegradability, non-toxicity,non-antigenicity and adsorption properties.[22.] These chitosan stabilized nanoparticles were characterised by FTIR, TEM and EDS.The IR spectra of the free as well as chitosan capped silver nanoparticles (Fig. 3) show two prominent peaks, one at 3438 cm-1 corresponding to O – H stretching and a second peak at 1642 cm-1 corresponding to C=C stretching vibrations respectively. Another peak at 794 cm-1 corresponding to -C-H bending vibrations was found only in stabilized silver nanoparticles indicating that chitosan is involved in the process of stabilisation as this peak was also found in the IR spectrum of chitosan.The TEM and EDS analyses were carried out for quantitative deduction of size, morphology and localization of elements in the nano specimens. The TEM micrograph of CSAgNPs reveals the uniform spherical smooth morphology within the size range of 101.78 nanometers and electron dense thin chitosan coated shell of the diameter in the range of 3-5 nanometers (Fig. 4a). Such size distribution analysis primarily confirms that the particles are well dispersed and less aggregated. The EDS graph recorded to unearth the presence of various elements is given in Fig. 4b. Peaks of silver were observed along with significant peaks of C, O and N reflecting the presence of elements constituting chitosan.

Biofilm inhibition

The biofilm inhibition studies carried out using both the free and chitosan stabilized AgNPs at all the tested concentration have successfully inhibited the biofilm formation of *E. coli*.The results were represented as inhibition percentage of biofilm development (Table 1). Both the free and stabilized nanoparticles have been found to show distinct effect on biofilm formation in dose dependent manner. But significant effect (P>0.05) was recorded in case of chitosan stabilized nanoparticles. The results in Table 1 clearly indicate the enhanced antibiofilm effect of coated nanoparticles in microtitre plate assay. While 100 µg/ml of free AgNPs resulted in 80 % inhibition, the same concentration of CS-AgNPs brought about complete inhibition of the biofilm. Namasivayam et al [23] have reported the synergistic effect of biogenic silver nanoparticles and plant products and also with antibiotics on the biofilm of clinical isolates of *Staphylococcus aureus* and *Candida tropicalis*. His group has further studied the biofilm inhibitory effect of chemogenic silver nanoparticles and antibacterial antibiotics coated catheters against *Staphylococcus aureus* biofilm [24]. nAg immobilized on glass was found to inhibit the biofilm formation of *Streptococcus oralis* which is of great importance for materials that require durable antibacterial effect on their surfaces, as it is the case of dental implants [17]. The increased antibiofilm effect of chitosan stabilized metallic nanoparticles may be due to the inhibition of exopolysaccharide synthesis limiting the formation of biofilm [25] or due to diffusion of CS-AgNPs through the channels present in the biofilms followed by the sustained release of metal nanoparticles which may then impart antimicrobial functions [26, 27]. A study carried out by Mohanty et al. [28] revealed that starch stabilised silver nanoparticles was showing considerable antimicrobial and antibiofilm activity. Further he has reported that these AgNPs were found to disrupt biofilm formation and exhibit better antibacterial activity compared tohuman cationic antimicrobial peptide LL-37.Nitrocellulose membrane consists of a series of repeating structural units called nitrocellulose polymer. Nitrocellulose membrane is suitable to catch bacteria in a filtration processbecause its tiny pores can capture microorganisms like bacteria [29]. So these nitrocellulose membranes were selected for further studies on the biofilm inhibitory effect of chitosan coated metallic nanoparticles. The results of biofilm inhibition studies on nitrocellulose membrane using free and coated AgNPs were given in Table 1. It has been identified that 81 % biofilminhibition was brought about by 100 µg/ml of free AgNPs while complete inhibition (100%) was successfully achieved with 75 µg/ml of CS-AgNPs. Scanning electron microscopy of the biofilm derived from nitrocellulose membranes treated with CS-AgNPs reveals degeneration of biofilm with weakened cell masses (Fig. 5b) while the control exposed compact tightly packed cell aggregates (Fig. 5a). ***3.4.***

Effect on biochemical composition of biofilm matrix

The matrix is one of the most distinctive features of a microbial biofilm. It forms a three dimensional, gel- like, highly hydrated and locally charged environment in which the microorganisms are largely immobilized. Matrix-enclosed micro colonies, sometimes described as stacks or towers, are separated by water channels which provide a mechanism for nutrient circulation within the biofilm. The composition of the matrix varies according to the nature of the organism and reduction in the biochemical composition of the biofilm matrix leads to weakening of the biofilm thus facilitating the entry of the drugs [13].Biochemical composition of biofilm matrix in terms of total carbohydrate and total protein were used as an index of inhibition. Table 2 shows a gradual reduction in the concentrations of carbohydrates and proteins with increasing concentrations of free and stabilize dnanoparticles. Another notable feature is the significant reduction in the biochemical compositionbrought about by the stabilized nanoparticles

compared to the free counterparts. Similar findings were reported earlier by Namasivayam et al.[24] and Ganesh and Namasivayam [30] in which drastic reduction in carbohydrates and proteins of biofilm matrix derived from *Staphylococcus aureus* treated with silver nanoparticles and nano zero valent iron respectively. These results clearly indicate that these nanoparticles brought about antibiofilm effect mainly by inhibiting the formation of biofilm rather than disrupting the preformed biofilms.

IV. CONCLUSION

Particular attention is oriented nowadays towards the need for antimicrobial textiles and polymers that are able to reduce or eliminate infections completely especially those caused by antibiotic-resistant bacterial strains. Therefore, the development of nanoparticles possessing antimicrobial properties has recently received growing interest from both academic and industrial sectors. The present study demonstrated the synthesis of metallic silver nanoparticles by chemical reduction method followed by stabilization using biocompatible polymer chitosan which was found to exhibit enhanced antibiofilm activity. This study can further be used to prevent or minimize bacterial infections leading to the development of new generation of antimicrobial agents.

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Table 1. Effect of free and CS-AgNPs on biofilm inhibition of *E.coli*

S.No	Concentration (µg/ml)	Biofilm inhibition (%)			
		Microtitre plate		Nitrocellulose membrane	
		Free AgNps	CS-AgNps	Free AgNps	CS-AgNps
1	10	32.1	71.6	40.4	80.1
2	25	54.4	80.0 a	54.3	89.0
3	50	65.0	89.7 a	62.1	94.3
4	75	70.5	95.0 a	73.5	100.0
5	100	80.0	100.0 a	81.0	100.0

Table 2. Total carbohydrates and total protein of biofilm matrix of *E.coli* grown in microtitre plate and nitrocellulose membrane treated with free and stabilized nanoparticles

S.No	Concentration **(µg/ml)**	Assay	Total carbohydrate **(mg)**		Total protein **(mg)**	
			Free AgNp	**CS-AgNp**	**Free AgNp**	**CS-AgNp**
1	10	MP	92.4	76.1	91.0	75.4
		NA	87.4	71.1	87.3	70.2
2	25	MP	81.3	68.1	84.1	64.1
		NA	76.2	64.2	76.5	60.2
3	50	MP	74.3	55.2	72.0	58.2
		NA	70.2	51.2	68.4	52.0
4	75	MP	63.2	34.1	67.4	49.2
		NA	59.2	31.2	62.1	42.1
5	100	MP	55.1	21.2	55.1	35.2
		NA	50.2	18.5	50.1	28.5
	Control	125.0			102.4	

Figure 1. UV-Vis Spectra of synthesized AgNPs and inset shows the colour change during synthesis,

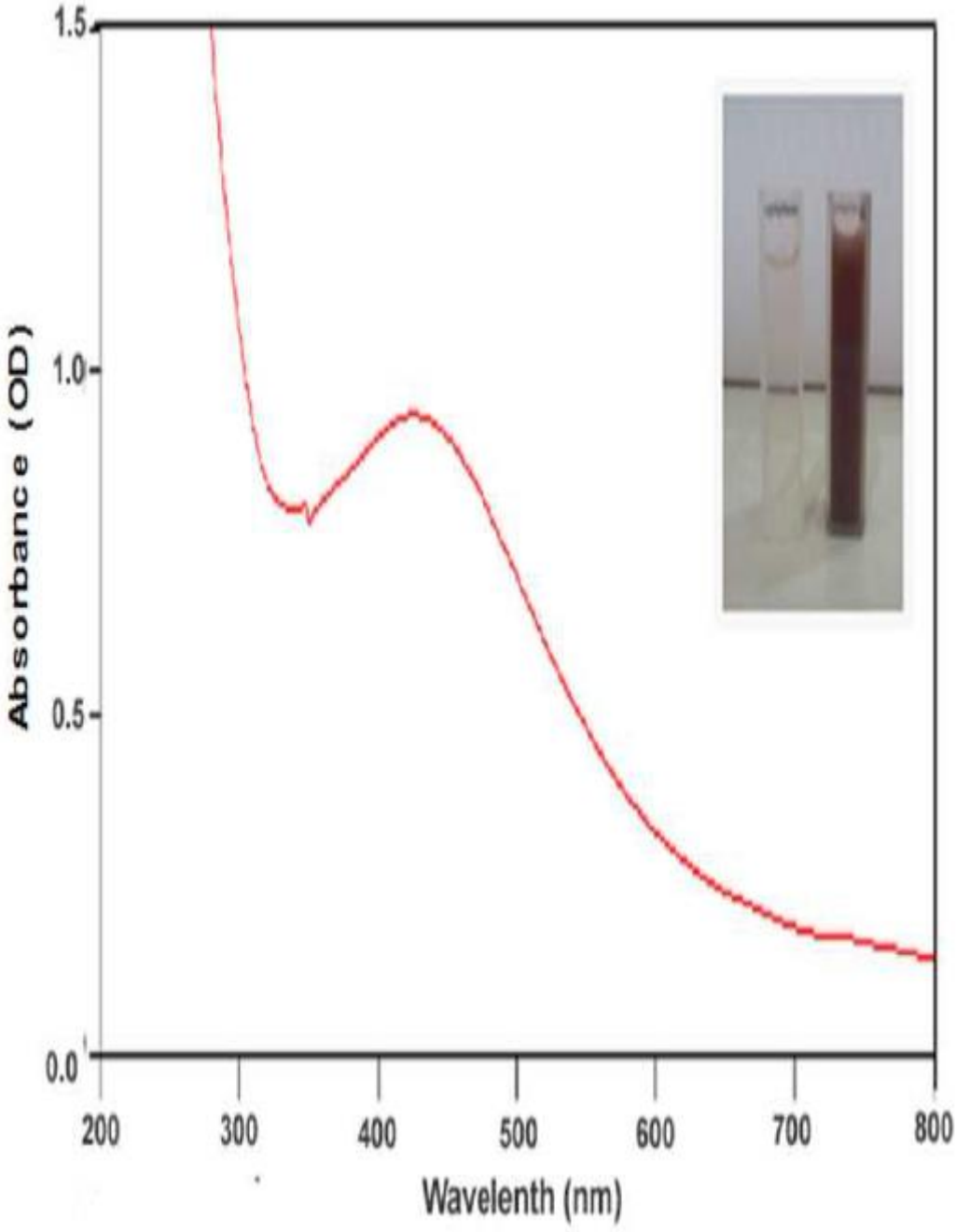


Figure 2 a) SEM and b) EDS images showing the presence of AgNPs

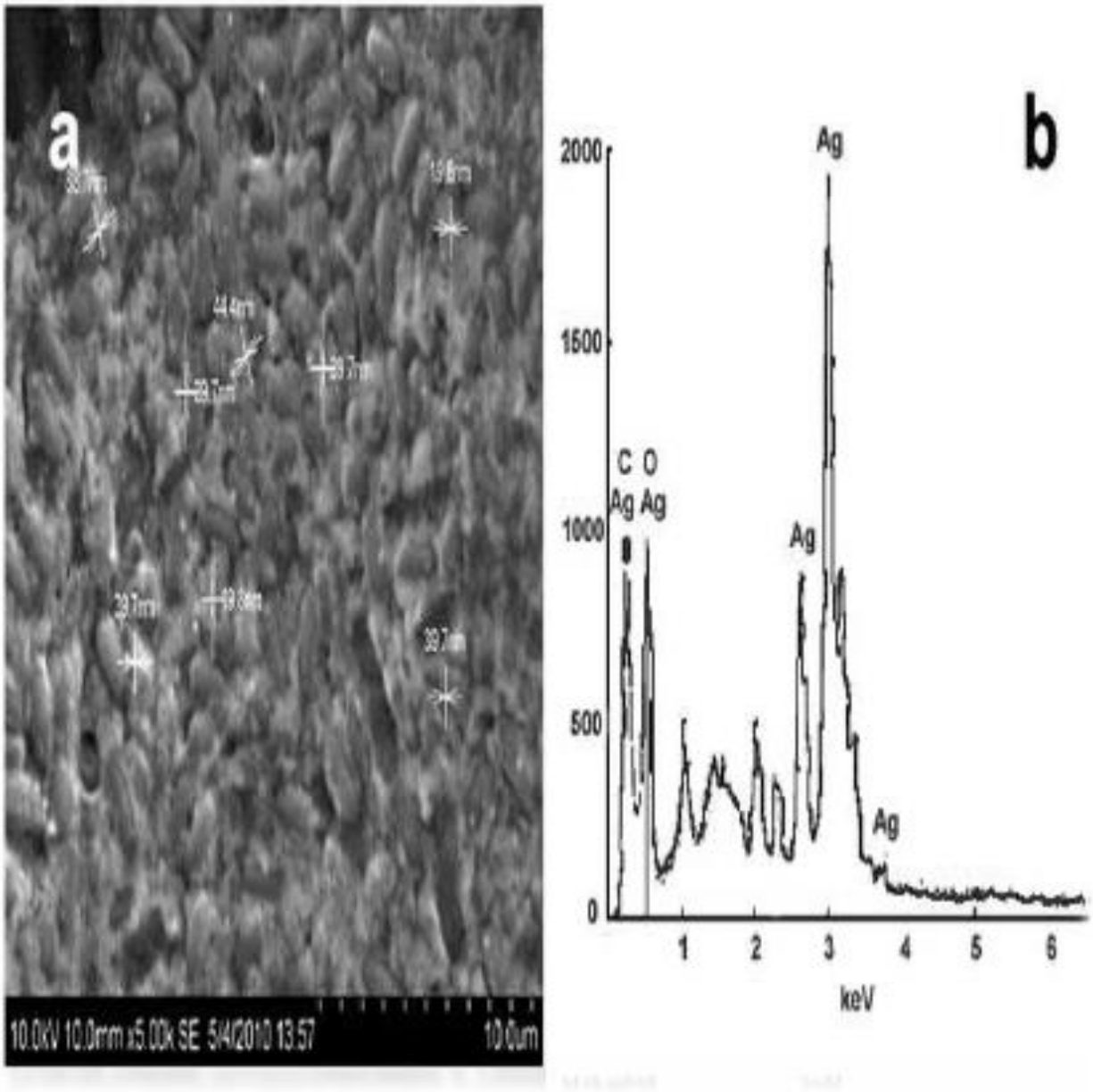


Figure 3. a) FTIR Spectra of CS-AgNPs

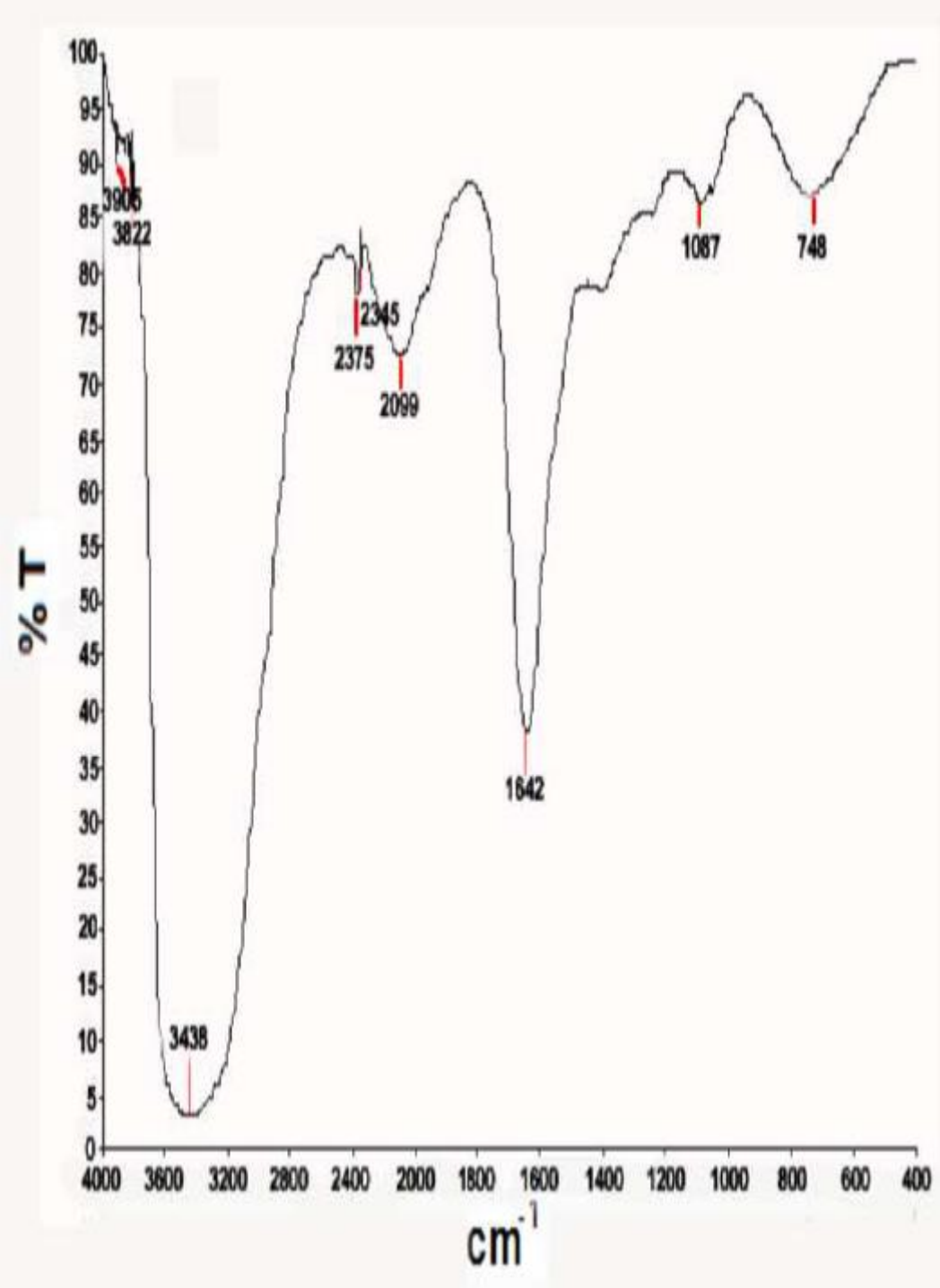


Figure 4. a) TEM and b) EDS images showing the presence of CS-AgNPs.

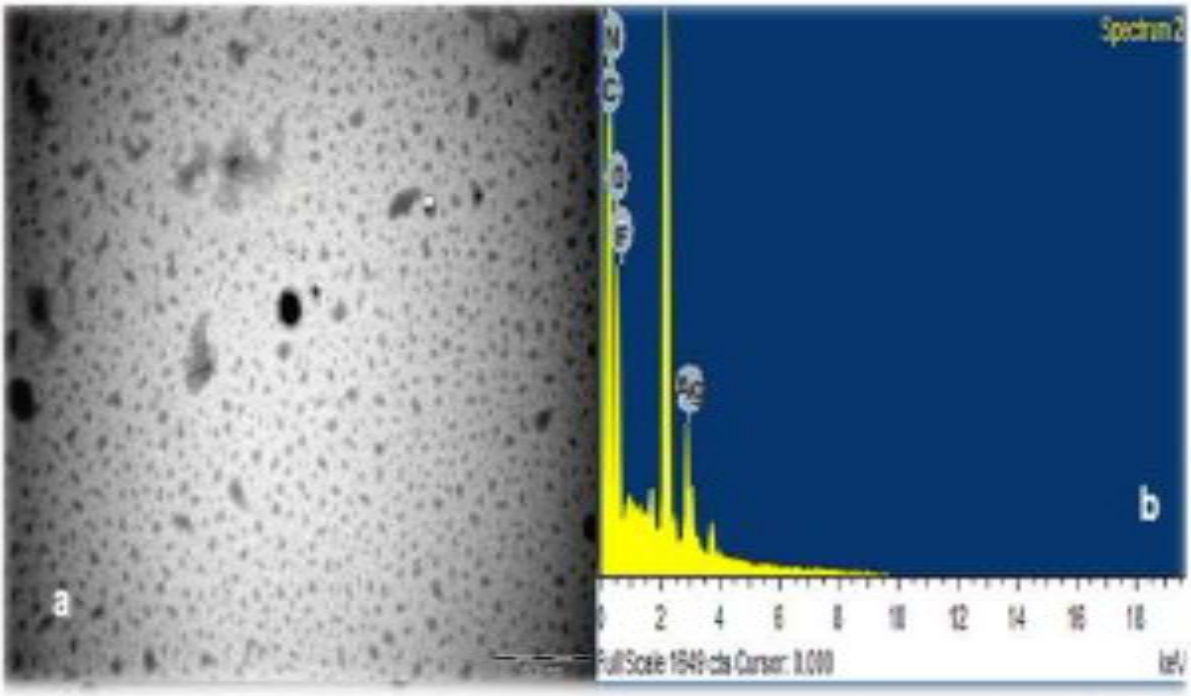
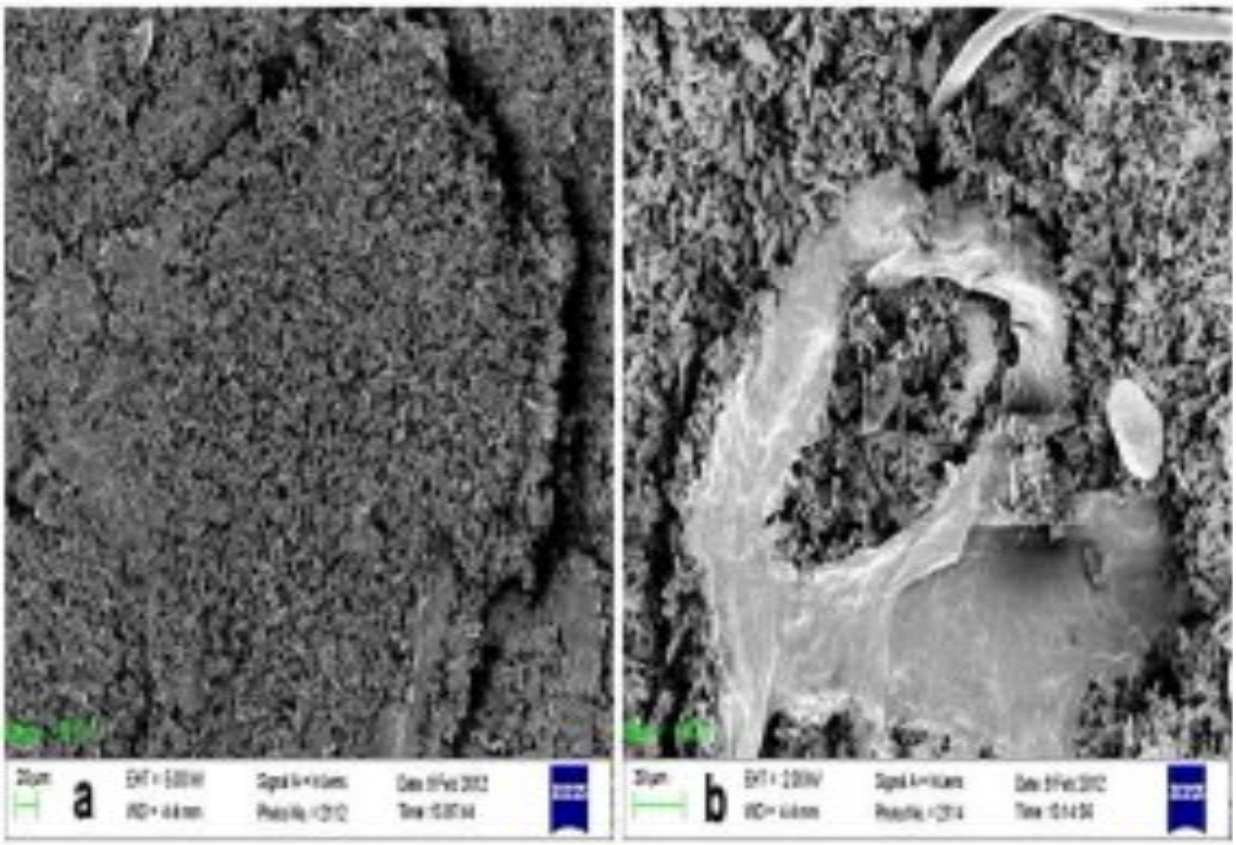


Figure 5. SEM images showing (b) the effect of CS-AgNPs on *Escherichia coli* biofilm derived from nitrocellulose membrane along with (a) being the control



Compiler Construction

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Abstract- Compiler construction is a widely used software engineering exercise, but because most students will not be compiler writers, care must be taken to make it relevant in a core curriculum. The course is suitable for advanced undergraduate and beginning graduate students. Auxiliary tools, such as generators and interpreters, often hinder the learning: students have to fight tool idiosyncrasies, mysterious errors, and other poorly educative issues. We introduce a set of tools especially designed or improved for compiler construction educative projects in C.We also provide suggestions about new approaches to compiler Construction. We draw guidelines from our experience to make tools suitable for education purposes. The final result of this paper is to provide a general knowledge about compiler design and implementation and to serve as a springboard to more advanced courses. Although this paper concentrates on the implementation of a compiler, an outline for an advanced topics course that builds upon the compiler is also presented by us.

Index Terms- Lex,Yacc Parser, Parser-Lexer, Symptoms & Anomalies.

I. INTRODUCTION

Computer programs are formulated in a programming language and specify classes of computing processes. Computers, however, interpret sequences of particular instructions, but not program texts. Therefore, the program text must be translated into a suitable instruction sequence before it can be processed by a computer. This translation can be automated, which implies that it can be formulated as a program itself. The translation program is called a *compiler*, and the text to be translated is called *source text* (or sometimes *source code*). Compilers and operating systems constitute the basic interfaces between a programmer and the machine. Compiler is a program which converts high level programming language into low level programming language or source code into machine code. It focuses attention on the basic relationships between languages and machines. Understanding of these relationships eases the inevitable transitions to new hardware and programming languages and improves a person's ability to make appropriate trade off in design and implementation. Many of the techniques used to construct a compiler are useful in a wide variety of applications involving symbolic data. In particular, every man-machine interface is a form of programming language and the handling of input involves these techniques. The term compilation denotes the conversion of an algorithm expressed in a human-oriented source language to an equivalent algorithm expressed in a hardware-oriented target language. We shall be concerned with the engineering of compilers their organization, algorithms, data structures and user interfaces. The term compilation denotes the conversion of an algorithm expressed in a human-oriented source language to an equivalent algorithm expressed in a hardware-oriented target language. We shall be concerned with the engineering of compilers their organization, algorithms, data structures and user interfaces. It is not difficult to see that this translation process from source text to instruction sequence requires considerable effort and follows complex rules. The construction of the first compiler for the language Fortran(formula translator) around 1956 was a daring enterprise, whose success was not at all assured. It involved about 18 man years of effort, and therefore figured among the largest programming projects of the time.

The term compilation denotes the conversion of an algorithm expressed in a human-oriented source language to an equivalent algorithm expressed in a hardware-oriented target language. We shall be concerned with the engineering of compilers (their organization, algorithms, data structures and user interfaces Programming languages are tools used to construct formal descriptions of finite computations (algorithms). Each computation consists of operations that transform a given initial state into some final state. A programming language provides essentially three components for describing such computations:

- Data types, objects and values with operations defined upon them.
- Rules fixing the chronological relationships among specified operations.
- Rules fixing the (static) structure of a program

II. LEX AND YACC

A. Availability

Lex and yacc were both developed at Bell. T. Laboratories in the 1970s. Yacc was the first of the two, developed by Stephen C. Johnson. Lex was designed by Mike Lesk and Eric Schmidt to work with yacc. Both lex and yacc have been standard UNIX utilities since 7th Edition UNIX. System V and older versions of BSD use the original AT&T versions, while the newest version of BSD uses flex (see below) and Berkeley yacc. The articles written by the developers remain the primary source of information on lex and yacc.

In programs with structured input, two tasks that occur over and over are dividing the input into meaningful units, and then discovering the relationship among the units. For a text search program, the units would probably be lines of text, with a distinction between lines that contain a match of the target string and lines that don't. For a C program, the units are variable names, constants, strings, operators, punctuation, and so forth. This division into units (which are usually called tokens) is

known as lexical analysis,or lexing for short. Lex helps you by taking a set of descriptions of possible tokens and producing a C routine, which we call a lexical analyzer, or a lexer, or a scanner for short, that can identify those tokens. The set of descriptions you give to lex is called a lex specification

Sample program. : program for lex

```
%{
#include "y.tab.h"
#include <stdlib.h>
void yyerror(char *);
%}
%%
[0-9]+
{               yylval = atoi(yytext);  return INTEGER; }
[-/*+\n]          { return *yytext; }
[ \t];
               yyerror("Unknown character");
%%
int yywrap(void) {    return 1;}
```

program for independent parser

```
%{
#include <stdio.h>
int yylex(void);
void yyerror(char *);
%}
%token INTEGER
%%
program     :   program expr '\n'        { printf("%d\n", $2); }
          | ;
expr        :   INTEGER
          | expr '+' expr          { $$ = $1 + $3; }
          | expr '*' expr          { $$ = $1 * $3; }
          | expr '-' expr              { $$ = $1 - $3; }
          | expr '/' expr              { $$ = $1 / $3; }
          ;
%%
void yyerror(char *s)
{
printf(stderr, "%s\n", s);
}
int main(void)
{
yyparse();
return 0;
}
```

B. Grammar

For some applications, the simple kind of word recognition we've already done may be more than adequate; others need to recognize specific sequences of tokens and perform appropriate actions. Traditionally, a description of such a set of actions is known as a ***grammar.***

Parser-Lexer Communication

When you use a lex scanner and a yacc parser together, the parser is the higher level routine. It calls the lexer **yylex()** whenever it needs a token from the input. The lexer then scans through the input recognizing tokens. As soon as it finds a token of interest to the parser, it returns to the parser, returning the token's code as the value of **yyfex().** Not all tokens are of interest to the parser-in most programming languages the parser doesn't want to hear about comments and whitespace, for example. For these ignored tokens, the lexer doesn't return so that it can continue on to the next token without bothering the parser. The lexer and the parser have to agree what the token codes are. We solve this problem by letting yacc define the token codes. The tokens in our grammar are the parts of speech: NOUN, PRONOUN, **VERB, ADVERB, ADJECTIVE,**PREPOSITION, and CONJUNCTION. Yacc defines each of these as a small integer using a preprocessor ***#define,*** Here are the definitions it used

in this example:

```
# define NOUN 257
# define PRONOUN 258
# define VERB 259
# define ADVERB 260
# define ADJECTIVE 261
# define PREPOSITICN 262
# define cXwUNCTICN 263
```

Token code zero is always returned for the logical end of the input. Yacc doesn't define a symbol for it, but you can yourself if you want.

The Parts of Speech Lexer

Example : shows the declarations and rules sections of the new lexer.

Example : lexer to be called from the parser

```
%{
/*
* We now build a lexical analyzer to be used by a higher-level parser.
* /
#include ""y.tab.hn /* token codes from the parser */
#define LOOKUP 0 /* default - not a defined word type. */
int state;
\n { state = LOOKUP; 1
\.\n I state = LOOKUP;
```

Example : lexer to be called from the parser (continued)

```
return 0; /* end of sentence */
1
lrerb ( state = VERB; 1
^adj { state = ADJECTIVE; 1
"adv { state = ADVERB; 1
"noun { state = NOUN; 1
Prep { state = PREPOSITION; 1
pron { state = FTUXOUN; 3
"conj { state = CONJUNCTI~; 1
[a-zA-Z]+ {
if (state != LOOKUP) {
add-word(state, yytext) ;
) else I
switch (lookUpPword (yytext) ) {
case VERB:
```

```
return (VERB) ;
case ALXEXTIVE:
return (ALUBTIVE) ;
case ADVERB:
return (ADVERB);
case NOUN:
return (NOUN) ;
case PREPOSITION:
return (PREPOSITION);
case PRONOUN:
return (PRONOUN) ;
case CONmION:
return (CONJUWTION) ;
default :
printf("%s: don't reccgnize\nu, yytext):
/* don't return, just ignore it */
}
   }
      }
% %
... same add-word() and lookup.word() as before ...
```

There are several important differences here. We've changed the part of speech names used in the lexer to agree with the token names in the parser. We have also added return statements to pass to the parser the token codes for the words that it recognizes. There aren't any return statements for the tokens that define new words to the lexer, since the parser doesn't care about them.

A Yacc Parser

Example 1-7 introduces our first cut at the yacc grammar.

Example 1-7: Simple yacc sentence parser

```
% t
/*
* A lexer for the basic g r m to use for recognizing mlish sentences.
/
#include <stdio.h>
% 1
%token NOUN PRCXWUN VERB AIXlERB ADJECl'IVE J3EPOSITIM CONJUNCTIM
% %
sentence: subject VERB object( printf("Sentence is valid.\nn); )
subject: NOUN
I PRONOUN
object: NOUN
extern FILE win;
main ( )
(
while ( !f eof (yyin)) {
yparse( ) ;
```

example : Simple yacc sentence parser (continued)

```
yyerror ( s)
char *s;
fprintf (stderr, "%s\na , s) ;
}
```

The structure of a yacc parser is, not by accident, similar to that of a lex lexer. Our first section, the definition section, has a literal code block, enclosed in "%{" and "%I". We use it here for a C comment (as with lex, C comments belong inside C code blocks, at least within the definition section) and a single include file.

The Rules Section

In our grammar we use the special character " I ", which introduces a rule with the same left-hand side as the previous one. It is usually read as "or," e.g., in our grammar a subject can be either a NOUN or a PRONOUN. The action part of a rule consists of a C block, beginning with "{" and ending with "{". The parser executes an action at the end of a rule as soon as the rule matches. In our sentence rule, the action reports that we've successfully parsed a sentence. Since sentence is the top-level symbol, the entire input must match a sentence. The parser returns to its caller, in this case the main program, when the lexer reports the end of the input. Subsequent calls to yyparse() reset the state and begin processing again. Our example prints a message if it sees a "subject VERB object" list of input tokens. What happens if it sees "subject subject" or some other invalid list of tokens? The parser calls yyerroro, which we provide in the user subroutines section, and then recognizes the special rule error. You can provide error recovery code that tries to get the parser back into a state where it can continue parsing, If error recovery fails or: as is the case here, there is no error recovery code, yyparse() returns to the caller after it finds an error.

C. Storage Management

In this section we shall discuss management of storage for collections of objects, including temporary variables,during their lifetimes. The important goals are the most economical use of memory and the simplicity of access functions to individual objects. Source language properties govern the possible approaches, as indicated by the following questions :

- Is the exact number and size of all objects known at compilation time?
- Is the extent of an object restricted, and what relationships hold between the extents of distinct objects (e.g. are they nested)?
- Does the static nesting of the program text control a procedure's access to global objects, or is access dependent upon the dynamic nesting of calls?

Static Storage Management

We speak of static storage management if the compiler can provide fixed addresses for all objects at the time the program is translated (here we assume that translation includes binding), i.e. we can answer the first question above with 'yes'. Arrays with dynamic bounds, recursive procedures and the use of anonymous objects are prohibited. The condition is fulfilled for languages like FORTRAN and BASIC, and for the objects lying on the outermost contour of an ALGOL 60 or Pascal program. (In contrast, arrays with dynamic bounds can occur even in the outer block of an ALGOL 68 program.) If the storage for the elements of an array with dynamic bounds is managed separately, the condition can be forced to hold in this case also. That is particularly interesting when we have additional information that certain procedures are not recursive, for example because recursivity must be noted specially (as in PL/1) or because we have determined it from analysis of the procedure calls. We can

then allocate storage statically for contours other than the outermost. Static storage allocation is particularly valuable on computers that allow access to any location in main memory via an absolute address in the instruction. Here, static storage corresponds exactly to the class of objects with direct access paths .If, however, it is unknown during code generation whether or not an object is directly addressable (as on the IBM 370) because this depends upon the _nal addressing carried out during binding, then we must also access statically-allocated objects via a base register. The only advantage of static allocation then consists of the fact that no operations for storage reservation or release need be generated at block or procedure entry and exit.

Dynamic Storage Management Using a Stack

All declared values in languages such as Pascal and SIMULA have restricted lifetimes. Further, the environments in these languages are nested: The extent of all objects belonging to the contour of a block or procedure ends before that of objects from the dynamically enclosing contour. Thus we can use a stack discipline to manage these objects: Upon procedure call or block entry, the activation record containing storage for the local objects of the procedure or block is pushed onto the stack. At block end, procedure return or a jump out of these constructs the activation record is popped of the stack. (The entire activation record is stacked, we do not deal with single objects individually!) An object of automatic extent occupies storage in the activation record of the syntactic construct with which it is associated. The position of the object is characterized by the base address, b, of the activation record and the relative location offset), R, of its storage within the activation record. R must be known at compile time but b cannot be known (otherwise we would have static storage allocation). To access the object, b must be determined at run time and placed in a register. R is then either added to the register and the result used as an indirect address, or R appears as the constant in a direct access function of the form 'register+constant'.The extension, which may vary in size from activation to activation, is often called the second order storage of the activation record. Storage within the extension is always accessed indirectly via information held in the static part; in fact, the static part of an object may consist solely of a pointer to the dynamic part.

Dynamic Storage Management Using a Heap

The last resort is to allocate storage on a heap: The objects are allocated storage arbitrarily within an area of memory. Their addresses are determined at the time of allocation, and they can only be accessed indirectly. Examples of objects requiring heap storage are anonymous objects such as those created by the Pascal new function and objects whose size changes unpredictably during their lifetime. (Linked lists and the exible arrays of ALGOL 68 belong to the latter class.). The use of a stack storage discipline is not required, but simply provides a convenient mechanism for reclaiming storage when a contour is no longer relevant. By storing the activation records on a heap, we broaden the possibilities for specifying the lifetimes of objects. Storage for an activation record is analyze and understand all the provided review comments thoroughly. Now make the required amendments in your paper. If you are not confident about any review comment, then don't forget to get clarity about that comment. And in some cases there could be chances where your paper receives number of critical remarks. In that cases don't get disheartened and try to improvise the maximum. released only if the program fragment (block, procedure, class) to which it belongs has been left and no pointers to objects within this activation record exist. Heap allocation is particularly simple if all objects required during execution can ‘t into the designated area at the same time. In most cases, however, this is not possible. Either the area is not large enough or, in the case of virtual storage, the working set becomes too large. We shall only sketch three possible recycling strategies for storage and indicate the support requirements placed upon the compiler by these strategies.

Storage can be recycled automatically by a process known as garbage collection, which operates in two steps:

- Mark. All accessible objects on the heap are marked as being accessible.
- Collect. All heap storage is scanned. The storage for unmarked objects is recycled, and
  all marks are erased.

This has the advantage that no access paths can exist to recycled storage, but it requires considerable support from the compiler and leads to periodic pauses in program execution. In order to carry out the mark and collect steps, it must be possible for the run-time system to find all pointers into the heap from outside, find all heap pointers held within a given object on the heap, mark an object without destroying information, and find all heap objects on a linear sweep through the heap. Only the questions of finding pointers affect the compiler;

there are three principal possibilities for doing this:

1. The locations of all pointers are known beforehand and coded into the marking algorithm.
2. Pointers are discovered by a dynamic type check. (In other words, by examining a

storage location we can discover whether or not it contains a pointer.)

3. The compiler creates a template for each activation record and for the type of every

object that can appear on the heap. Pointer locations and (if necessary) the object

length can be determined from the template.

Pointers in the stack can also be indicated by linking them together into a chain, but this would certainly take too much storage on the heap

.

D. *Error Handling*

Error Handling is concerned with failures due to many causes: errors in the compiler or its environment (hardware, operating system), design errors in the program being compiled, an incomplete understanding of the source language, transcription errors, incorrect data, etc. The tasks of the error handling process are to detect each error, report it to the user, and possibly make some repair to allow processing to continue. It cannot generally determine the cause of the error, but can only diagnose the visible symptoms. Similarly, any repair cannot be considered a

correction (in the sense that it carries out the user's intent); it merely neutralizes the symptom so that processing may continue. The purpose of error handling is to aid the programmer by highlighting inconsistencies. It has a low frequency in comparison with other compiler tasks, and hence the time required to complete it is largely irrelevant, but it cannot be regarded as an 'add-on' feature of a compiler. Its inuence upon the overall design is pervasive, and it is a necessary debugging tool during construction of the compiler itself. Proper design and implementation of an error handler, however, depends strongly upon complete understanding of the compilation process. This is why we have deferred consideration of error handling until now

Errors, Symptoms, Anomalies and Limitations

We distinguish between the actual error and its symptoms. Like a physician, the error handler sees only symptoms. From these symptoms, it may attempt to diagnose the underlying error. The diagnosis always involves some uncertainty, so we may choose simply to report the symptoms with no further attempt at diagnosis. Thus the word 'error' is often used when 'symptom' would be more appropriate. A simple example of the symptom/error distinction is the use of an undeclared identifier in LAX. The use is only a symptom, and could have arisen in several ways:

- The identifier was misspelled on this use.
- The declaration was misspelled or omitted.
- The syntactic structure has been corrupted, causing this use to fall outside of the scope of the declaration.

Most compilers simply report the symptom and let the user perform the diagnosis. An error is detectable if and only if it results in a symptom that violates the definition of the language. This means that the error handling procedure is dependent upon the language definition, but independent of the particular source program being analyzed. For example, the spelling errors in an identifier will be detectable in LAX (provided that they do not result in another declared identifier) but not in FORTRAN, which will simply treat the misspelling as a new implicit declaration.

We shall use the term anomaly to denote something that appears suspicious, but that we cannot be certain is an error. Anomalies cannot be derived mechanically from the language definition, but require some exercise of judgement on the part of the implementor. As experience is gained with users of a particular language, one can spot frequently-occurring errors and report them as anomalies before their symptoms arise.

III. CONCLUSION

This report outlines a course in compiler construction. The implementation and source language is Scheme, and the target language is assembly code. This choice of languages allows a direct-style ,stack-based compiler to be implemented by an undergraduate in one semester that touches on more aspects of compilation than a student is likely to see in a compiler course for more tradition all Languages. Furthermore, expressiveness is barely sacrificed; the compiler can be boot strapped provided there is enough run-time support. Besides covering basic compilation issues, the course yields an implemented compiler that can serve as a test bed for coursework language implementation. The compiler has been used, for example, to study advanced topics such as the implementation of first-class continuations and register allocation.

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Influence of Stand Establishment Techniques on Yield and Economics of Rice Cultivation in Kuttanad

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Abstract- Field experiment was conducted at Rice Research Station, Moncompu during *kharif* 2012 to study the effect of different stand establishment techniques on yield and economics of rice cultivation in Kuttanad. The experiment was laid out in RBD replicated four times with five treatments in plots of 60 m^2. Mechanized transplanting at 22 x 14 cm with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAT (days after transplanting) followed by hand weeding at 40 DAT significantly influenced the growth and yield attributes and recorded higher grain yield (5350 kg ha^{-1}), straw yield (9140 kg ha^{-1}), gross returns (100090 Rs ha^{-1}), net returns (73303 Rs ha^{-1}) and B: C ratio (3.74). Among the direct seeding techniques, drum seeding with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAS (days after sowing) followed by hand weeding at 40 DAS recorded higher gross returns, net returns and B: C ratio. The lowest grain yield (4566 kg ha^{-1}), gross returns (86127 Rs ha^{-1}) , net returns (56777 Rs ha^{-1}) were recorded in broad casting of pre-germinated weeds with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAS followed by hand weeding at 40 DAS. The dry weight of weeds were found to be less in treatments with post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} followed by hand weeding at 40 DAS/DAT than conoweeding at 15, 30 and 40 DAT. Conoweeding alone failed to control weeds effectively, but it influences the crop growth and yield by stimulating aeration and root growth.

Index Terms- Broad casting, pre-germinated seeds, manual transplanting, mechanized transplanting, pyrazosulfuron ethyl, conoweeding

I. INTRODUCTION

Kuttanad, the rice bowl of Kerala is a unique wetland ecosystem, lies at an altitude 0.6 to 2.2 m below mean level situated in southern Kerala. About 55,000 hectares, reclaimed from the surrounding backwaters for paddy cultivation is known as puncha lands, these reclaimed holding with an outer bund are made in to cluster of fields called polders or padashekarams. Broadcasting of pre-germinated seeds is the common practice of rice growing in Kuttanad. Since it is a direct seeding technique, both crop and weed seeds germinate together and weeds compete with crop for nutrients, water, space and light and became a biological constraint. For the last few years, infestation of *Leptochloa*, weedy rice and *Echinochlo*a became serious and caused considerable reduction in yield. As a result farmers are forced to skip rice cultivation for one or two seasons. In addition, to maintain the plant population farmers go for higher seed rate lead to severe pest and disease problem which ultimately increase the cost of cultivation.

Comparative trials conducted in India revealed that transplanted crop yield more than the direct seeded crop. Transplanting ensures uniform crop stand, better control of weeds, uniform ripening and less lodging. In spite of these advantages, manual transplanting is quite expensive, laborious, time consuming and causes of lot of drudgery. Manual transplanting takes about 300 to 350 man hours/ha which is roughly 25 % of the labour requirement of crop (Goel *et al*., 2008). In addition to that, lack of skilled labour at the time of transplanting results in to low plant population and eventually low yield (Aslam *et a*l., 2008). Optimizing plant density and timeliness of operation in paddy is considered essential for optimizing yield (Chaudhary *et al*., 2005). In order to get maximum returns from the rice crop, cultivation cost has to be reduced through minimizing the dependence on labour for transplanting. Under such condition mechanized transplanting can be considered as the most promising option as it saves labour, ensures timely planting and attain optimum plant density. Keeping these views in mind an experiment was conducted at Rice Research Station, Moncompu with an objective to study the influence of stand establishment techniques on yield and economics of rice cultivation in Kuttanad.

II. MATERIALS AND METHODS

The field experiment was conducted at Rice Research Station, Moncompu, Kerala (geographically situated at 9 º 5' N latitude and 76 º 5' E longitude and at an altitude 1m below MSL) during *kharif* 2012. The soil is silty clay with pH 5.76, organic carbon 4.6 %, available P and K 58.24 and 172.72 kg ha^1, respectively. The experiment was laid out in randomized block design replicated four times in plots of 60 m^2 size with five treatments. The treatments combinations were mechanized transplanting with Yanji 8 row self-propelled transplanter at 22 x 14 cm with early post emergence application of pyrazosulfuron ethyl 10% WP @ 20 g ai ha^{-1} on seven DAT followed by hand weeding at 40 DAT (T_1), mechanized transplanting with Mahindra walk behind four row transplanter at 30 x 10 cm with conoweeding at 15, 30 and 45 DAT (T_2), drum seeding at 20 x 10 cm with early post emergence application of pyrazosulfuron ethyl 10% WP @ 20 g ai ha^{-1} on seven DAS followed by hand weeding at 40 DAS (T_3), Manual transplanting at 20 x10 cm with early post emergence application of pyrazosulfuron ethyl 10% WP @ 20 g ai ha^{-1} on seven DAT followed by hand weeding at 40 DAT (T_4) and broad casting of pre-germinated seeds with early post emergence application of pyrazosulfuronethyl 10% WP @ 20 g ai ha^{-1} on seven DAS

followed by hand weeding at 40 DAS (T_5). In mechanized and manual transplanting, seedlings were raised by dapog and wet nursery methods, respectively. Fifteen day old seedlings were used for transplanting in both these methods. In drum seeding and broadcasting methods, pre-germinated seeds were used for sowing on the same day when seedlings were transplanted by machine and manual methods. After the land preparation and leveling, sedimentation period of four days was given to avoid the float sinkage in machine transplanted plots with Yanji eight row self- propelled transplanter. The variety used was Uma (MO 16), a medium duration variety. The crop was fertilized with 90:45:45 kg ha^{-1} of N, P_2O_5 and K_2O, respectively. $1/3^{rd}$ dose of N and K_2O and half dose of P_2O_5 were applied at 15 DAT, $1/3^{rd}$ dose of N and K_2O and half dose of P_2O_5 at 35 DAT and remaining $1/3^{rd}$ dose of N and K_2O were applied at 55 DAT.

Observations on growth parameters viz., hills per square meter, plant height, LAI and tillers per square meter were recorded at flowering stage and yield parameters viz., productive tillers per square meter, panicle weight, fertile grains per panicle, 1000 grain weight, grain and straw yield were recorded at harvest. Observations on total weed dry weight were recorded with quadrat of 0.25 m^2 size placed randomly at two representative sites in each plot at 60 and 90 DAS/DAT. Weed samples were collected from each spot, washed in tap water, sun dried first followed by oven drying at 65° until constant weights were obtained. The data on weed dry weight were subjected to square root transformation to normalize their distribution. The cost of cultivation was worked out based on the labour and input cost incurred towards rice cultivation in different establishment techniques. Economics of cultivation was worked out based on the minimum support price for paddy given by the Government of Kerala during *kharif* 2012. All data except gross returns, net returns and B: C ratio was analyzed using ANOVA.

III. RESULTS AND DISCUSSION

Growth attributes

Table 1: Effect of stand establishment techniques on growth attributes of rice at flowering stage

Treatment	Hills per square meter	Plant height (cm)	LAI	Tillers per square meter
Mechanized transplanting at 22 x14 cm with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAT followed by hand weeding at 40 DAT	35	100.2	4.14	350
Mechanized transplanting at 30x10 cm with conoweeding at 15, 30 and 45 DAT	33	102.1	4.13	316
Drum seeding at 20x10 cm with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAS followed by hand weeding at 40 DAS	50	98.5	3.67	312
Manual planting at 20 x10 cm with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAT followed by hand weeding at 40 DAT	48	94.9	3.78	326
Broadcasting of pre-germinated seeds with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAS followed by hand weeding at 40 DAS	98	92.3	3.21	296
CD	14	5.9	0.25	34

DAT- days after transplanting, DAS-days after sowing

Hills per square meter were significantly influenced by the establishment techniques (Table 1). Broad casting of pre-germinated seeds recorded the maximum number of hills per square meter. This was attributed to the fact that in direct seeding uniformity in plant population cannot be assured due to spread of

more number of seeds per unit area. This was followed by drum seeding and manual planning at 20x 10 cm spacing. However, in mechanized transplanting methods, no significant difference was found. Mechanized transplanting at 30x10 cm spacing recorded taller plants (102.1) closely followed by mechanized planting at 22x 14 cm spacing (100.2). This was due to the reason that plants were planted at proper row to row and plant to plant spacing leading to reduced competition (Awan *et al*., 2007) and the plants got sufficient space to grow and increased light transmission in the canopy led to increased plant height (Vijayakumar *et al*., 2006). The plants with the lowest height were observed in broadcasting method might be due to higher competition among plants for space, light and nutrients. Similarly the highest LAI (leaf area index) was also recorded in mechanized transplanting (4.14) followed by manual transplanting (3.78) and the lowest LAI was recorded in broadcasting method (3.21). This was owing to the fact that, younger seedlings planted at proper spacing enhanced the root growth which facilitated increased cell division and cell enlargement and more number of tillers with more leaves and subsequently higher photosynthetic rate for increased LAI (Shrirame *et al*., 2000). Tillers per square meter were also significantly influenced by the stand establishment techniques. The highest number of tillers were recorded in mechanized planting at 22 x14 cm (350 No.m^{-2}) followed by manual planting at 20 x10 cm might be due to reduced weed growth (Table 2) and due to the availability of sufficient amount of nutrients and moisture at tiller initiation stage and better establishment of roots (Aslam *et al*., 2008).

Total Weed dry matter

Table 2: Effect of stand establishment techniques on total weed dry matter at 60 and 90 DAT/DAS

Treatment	Total Weed dry matter at 60 DAT/DAS	Total Weed dry matter at 90 DAT/DAS
Mechanized transplanting at 22 x14 cm with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAT followed by hand weeding at 40 DAT	19.39 (4.46)	37.32 (6.15)
Mechanized transplanting at 30x10 cm with conoweeding at 15, 30 and 45 DAT	52.64 (7.29)	114.85 (10.74)
Drum seeding at 20x10 cm with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAS followed by hand weeding at 40 DAS	26.85 (5.23)	67.07 (8.22)
Manual planting at 20 x10 cm with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAT followed by hand weeding at 40 DAT	20.38 (4.57)	53.82 (7.37)
Broadcasting of pre-germinated seeds with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAS followed by hand weeding at 40 DAS	46.42 (6.85)	66.25 (8.17)
CD	1.00	2.57

DAT- days after transplanting, DAS-days after sowing

Total weed dry weight was significantly influenced by the stand establishment techniques. Among the five different stand establishment techniques studied, mechanized transplanting at 30 x 10 cm with conoweeding at 15, 30 and 45 DAT (T_2) recorded the highest weed dry matter accumulation. The results indicated that conoweeding alone failed to control weeds effectively. But it improved the root growth by pruning of some upper roots and increasing soil aeration which influenced the plant growth (Table1). Early post emergence application of herbicide had brought out effective weed control at early stage of crop weed competition resulted in low weed biomass production in T_1. While comparing the treatments T_1, T_3, T_4 and T_5, it was observed that the total weed dry matter was found to be more in direct seeding than transplanting. This was due to the fact that in direct seeding both crop and weed seeds emerge simultaneously and the roots of the crop plant could not penetrate deep enough to exploit the soil resources fully, giving adequate chance to the weeds to compete with crop plant (Ehsanullah *et al*., 2007). Among the direct seeding techniques, the total weed dry matter was found to be more in broadcasting of pre-germinated seeds (Table 2). The reason was that in treatment T_4, early emerged weeds were controlled by early post emergence application of pyrazosulfuron ethyl 10 % WP and the seeds were planted at proper spacing the later emerged weeds could be better removed by hand weeding. The findings are in agreement with the findings of Budhar and Tamilselvan (2002) and Singh and Singh (2010).

Yield attributes

Table 3: Effect of stand establishment techniques on the yield attributes of rice

Treatment	Panicles m^{-2}	Fertile grains per panicle	Panicle weight (g)	1000 grain weight (g)
Mechanized transplanting at 22 x14 cm with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAT followed by hand weeding at 40 DAT	343	116.5	3.68	25.88
Mechanized transplanting at 30x10 cm with conoweeding at 15, 30 and 45 DAT	311	127.8	3.75	25.58
Drum seeding at 20x10 cm with early post emergence application of pyrazosulfuron ethyl 10 % WP 20 g ai ha^{-1} on seven DAS followed by hand weeding at 40 DAS	306	105.9	3.40	26.03
Manual planting at 20 x10 cm with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAT followed by hand weeding at 40 DAT	318	111.0	3.37	25.63
Broadcasting of pre-germinated seeds with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAS followed by hand weeding at 40 DAS	281	89.4	2.62	24.68
CD	28	15.2	0.44	NS

DAT- days after transplanting, DAS-days after sowing

Yield attributes of rice were significantly influenced by stand establishment techniques (Table 3). The highest number of panicles per square meter were recorded in mechanized planting at 22 x 14 cm with early post emergence application of pyrazosulfuron ethyl at seven DAT followed by hand weeding at 40 DAT (T_1). This was followed by manual planting at 20 x 10 cm (T_4) and was statistically on par with drum seeding at 20 x10 cm (T3) and mechanized planting at 30 x 10 cm with conoweeding (T_2). Broad casting of pre-germinated seeds (T_5) recorded the lowest number of panicles. This might due to less availability of nutrients and moisture to the crop plants at panicle initiation stage resulting from the crop-weed competition. In the treatment T_2, even though the total weed dry weight was found to be more, it produced significantly higher number of panicles (311 No.m^{-2}) than broadcasting of pre-germinated seeds. The reason was that in T_2, the plants were planted at proper spacing and conoweeding at 15, 30 and 45 days after transplanting enhanced the soil aeration and root growth and would favorably influenced the plant growth (Table.1). Significantly higher number of fertile grains per panicle was recorded in the treatment T_2 and was statistically at par with other transplanting methods of crop establishment. The lowest number of grains per panicle was recorded in broadcasting method might be due to less photosynthesis as reflected in LAI at flowering (Table 1). The result is in conformity with the findings of Aslam *et al*. 2008 and Raj *et al*, 2012. Similarly panicle weight was also found to be influenced by establishment techniques. Panicles with more weight was recorded in treatment T_2 which was statistically at par with mechanized planting at 22 x14 cm (T_1), drum seeding at 20 x10 cm (T_3) and manual planting at 20 x10 cm (T_4). Being a varietal character test grain weight was not significantly influenced by the establishment techniques. Similar result is also reported by Ehsanullah *et al*., 2007. However, among the treatments it varied between 24.68 to 26.03 g. .

Yield and economics

Table 4: Effect of stand establishment techniques on yield and economics of rice

Treatment	Grain yield kg ha^{-1}	Straw yield kg ha^{-1}	Gross returns Rs ha^{-1}	Net returns Rs ha^{-1}	B:C ratio
Mechanized transplanting at 22 x14 cm with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAT followed by hand weeding at 40 DAT	5350	9140	100090	73303	3.74
Mechanized transplanting at 30x10 cm with conoweeding at 15, 30 and 45 DAT	5000	8905	93905	64397	3.18
Drum seeding at 20x10 cm with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAS followed by hand weeding at 40 DAS	4969	8956	93429	68164	3.69
Manual planting at 20 x10 cm with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAT followed by hand weeding at 40 DAT	5131	9135	96362	63307	2.92
Broadcasting of pre-germinated seeds with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAS followed by hand weeding at 40 DAS	4566	8505	86127	56777	2.93
CD	484	NS	-	-	

DAT- days after transplanting, DAS-days after sowing

Grain yield was significantly influenced by the stand establishment techniques. Grain production, which is the final product of growth and development, is controlled by two factors. First is the potential ability of the plant population to photosynthesize (source) and secondly, the capacity of the spike lets to accept the photosynthates (sink). The sink is composed of panicles per unit area (determined during the vegetative phase), the spikelet number per panicle determined during reproductive stage) and the fertility percentage determined during reproductive and ripening stage as reported by De Datta (1981). Mechanized transplanting at 22 x14 cm with early post emergence application of pyrazosulfuron ethyl at seven DAS followed by hand weeding at 40 DAT (T_1) recorded higher grain yield (5350 kg ha^{-1}) than other establishment techniques. The yield increase was due to the favourable influence on number of tillers, LAI and number of panicles. It was statistically at par with manual planting (5131 kg ha^{-1}), mechanized transplanting at 30 x10 cm with conoweeding at 15, 30 and 45 DAT (5000 kg ha^{1}) and drum seeding (4969 kg ha^{-1}). Javaid *et al.* (2012) also reported higher grain yield in transplanting compared to drill sowing and broadcasting. Broadcasting method of crop establishment failed to enhance the growth and yield attributes (Table 1 and 3) and their cumulative effect drastically reduced the grain yield (Table 4).

Straw yield was not significantly influenced by the stand establishment techniques. The broadcasting of pre-germinated seeds recorded the lowest straw yield (8505 kg ha^{-1}) due to reduction in growth components such as plant height, LAI and tillers per unit area (Table 4). The highest straw yield (9140 kg ha^{-1}) was recorded in treatment T_1 and it was followed by manual planting (T_4) and drum seeding (T_3). Economic analysis revealed that maximum net income and B:C ratio were recorded in mechanized transplanting at 22x 14 cm with early post emergence application of pyrazosulfuron ethyl 10 % WP followed by hand weeding at 40 DAT (T_1). This might be due to higher grain and straw yield and low cost of cultivation. This was followed by drum seeding at 20 x10 cm with early post emergence application of pyrazosulfuron ethyl 10 % WP followed by hand weeding at 40 DAT. Mechanized transplanting at 30 x 10cm with conoweeding recorded B: C ratio of 3.18 which was lower than T_3 might be due to comparatively high labour cost involved in conoweeding. The lowest benefit cost ratio recorded in manual transplanting might be due to high labour cost involved.

IV. Conclusion

Mechanized transplanting at 22 x14 cm with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g

ai ha^{-1} on seven DAT followed by hand weeding at 40 DAT was found to be the best establishment techniques for maintaining optimum plant population, reduced weed growth, higher grain yield, gross returns, net returns and B: C ratio. Mechanized transplanting at 30 x10 cm with conoweeding at 15, 30 and 45 DAT gave better yield than drum seeding but its net returns and B: C ratio was low due to high cost involved in conoweeding. Since the conoweeding enhanced the aeration and rooting its operation could be limited at one time at 40 DAT as an alternative to manual weeding and early emerged weeds could be controlled by the application of herbicide. Drum seeding at 20 x10 cm with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAS followed by hand weeding at 40 DAS gave better returns and B: C ratio than broadcasting of pre-germinated seeds with early post emergence application of pyrazosulfuron ethyl 10 % WP @ 20 g ai ha^{-1} on seven DAS followed by hand weeding at 40 DAS. So as an alternative stand establishment technique for sustaining the yield of rice cultivation, mechanized transplanting and drum seeding can be recommended in Kuttanad.

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Biochemistry of metal absorption in Human Body: Reference to check Impact of Nano Particles on Human Being

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Abstract- Living is made comfortable by continues and the enormous advancement of Science and Technology so far. Nano technology is now days a favorite tool for technologists from academia as well as industrial sector. Nano particles were present in earlier days also, but up to limited extent. The prevailing use of Nano particles now days, to serve almost every sector of human life, would result in to increased concentration of Nano particles in the environment; eventually to term them as a pollutant. Metal absorption in human body is a very complicated phenomenon itself. For trace element absorption a protein, namely, metallothionein is responsible. Metal particles, when enter the human body, are likely to accumulate in kidney, lungs or brain. Greater surface to volume ratio of Nano particles makes them highly reactive and may act as catalyst even. This gives a call to understand the biochemistry behind metal absorption in human body and to define a strategy for the use and production of nano particles for their discrete use.

Index Terms- Nanotechnology, Nanoparticles, Biochemistry of metal absorption in human body

I. Introduction

Nano has become a very favorite term that each of us, -from academia, industry or research institutes, coin it frequently during routine conversations. Nano particles have high surface to volume ratio i.e. they contain high number of atoms per unit volume. As the size reduces, the proportion of atoms found at the surface increases relative to the proportion inside its volume. This increases reactivity of material at nano level. The nano particles are more reactive and hence adverse effect caused by them at lower dose level is greater than their fine counterparts. Due to expanding use of nano particles and commercialization of nano technology products, exposure of the environment and humans to nano particles is bound to increase and an evaluation of their potential toxicity is highly essential. Once in the system, the metal nanoparticles accumulate in different organs because the body has no way to eliminate them. And because they are so small, they can go everywhere in the body, even through cells, and may interfere with sub-cellular mechanisms [1-3].

II. Which are the metals present in human body?

Common Elements present in Animal and Human body

Animals and humans have similar evolutionary backgrounds. Specific elements play critical roles in the structures of proteins and the activities of enzymes. The purpose of this page is to outline some of the uses of elements in the structure of animals and humans and to illustrate why these elements are essential in the body and for optimal health.

III. Macro Elements

- **Calcium (Ca)** Structure of bone and teeth.
- **Phosphorous (Ph)** Structure of bone and teeth. Required for ATP, the energy carrier in animals.
- **Magnesium (Mg)** Important in bone structure. Deficiency results in tetany (muscle spasms) and can lead to a calcium deficiency.
- **Sodium (Na)** Major electrolyte of blood and extracellular fluid. Required for maintenance of pH and osmotic balance.
- **Potassium (K)** Major electrolyte of blood and intracellular fluid. Required for maintenance of pH and osmotic balance.
- **Chlorine (Cl)** Major electrolyte of blood and extracellular and intracellular fluid. Required for maintenance of pH and osmotic balance.
- **Sulfur (S)** Element of the essential amino acids **methionine** and **cysteine**. Contained in the vitamins thiamin and biotin. As part of glutathione it is required for detoxification. Poor growth due to reduced protein synthesis and lower glutathione levels potentially increasing oxidative or xenobiotic damage are consequences of low sulfur and methionine and/or cysteine intake.

IV. Micro Elements

- **Iron (Fe)** Contained in **hemoglobin** and **myoglobin** which are required for oxygen transport in the body. Part of the **cytochrome p450** family of enzymes. Anemia is the primary consequence of iron deficiency. Excess iron levels can enlarge the liver, may provoke diabetes and cardiac falurer. The genetic disease **hemochromatosis** results from excess iron absorption. Similar symptoms can be produced through excessive transfusions required for the treatment of other diseases.
- **Copper (Cu)** Contained in enzymes of the **ferroxidase** (ceruloplasmin?) system which regulates iron transport and facilitates release from storage. A structural element

in the enzymes **tyrosinase**, **cytochrome c oxidase**, **ascorbic acid oxidase**, **amine oxidases**, and the antioxidant enzyme **copper zinc superoxide dismutase**. A copper deficiency can result in anemia from reduced ferroxidase function. Excess copper levels cause liver malfunction and are associated with genetic disorder **Wilson's Disease**

- **Manganese (Mn)** Major component of the mitochondrial antioxidant enzyme **manganese superoxide dismutase**. A manganese deficiency can lead to improper bone formation and reproductive disorders. An excess of manganese can lead to poor iron absorption.
- **Iodine (I)** Required for production of thyroxine which plays an important role in metabolic rate. Deficient or excessive iodine intake can cause goiter (an enlarged thyroid gland).
- **Zinc (Zn)** Important for reproductive function due to its use in FSH (follicle stimulating hormone) and LH (leutinizing hormone). Required for DNA binding of zinc finger proteins which regulate a variety of activities. A component of the enzymes **alcohol dehydrogenase**, **lactic dehydrogenase carbonic anhydrase**, **ribonuclease**, **DNA Polymerase** and the antioxidant **copper zinc superoxide dismutase**. An excess of zinc may cause anemia or reduced bone formation.
- **Selenium (Se)** Contained in the antioxidant enzyme **glutathione peroxidase** and **heme oxidase**. Deficiency results in oxidative membrane damage with different effects in different species. Human deficiency causes cardiomyopathy (heart damage) and is known as Keshan's disease.
- **Fluorine (Fl)** Constituent of bones and teeth. Important for tooth development and prevention of dental caries. Derives from water, tea, and fish.
- **Cobolt (Co)** Contained in vitamin B12. An excess may cause cardiac failure.
- **Molybdenum (Mo)** Contained in the enzyme **xanthine oxidase**. Required for the excretion of nitrogen in uric acid in birds. An excess can cause diarrhea and growth reduction.
- **Chromium (Cr)** A cofactor in the regulation of sugar levels. Chromium deficiency may cause hyperglycemia (elevated blood sugar) and glucosuria (glucose in the urine).

Other Elements

Rats have been shown to have improved growth on diets which contain other micro nutrients. These include:

- **Lead (Pb)**
- **Nickel (Ni)**
- **Silicon (Si)**
- **Vanadium (Vn)**

These elements are all toxic at high levels.

At elevated levels, heavy metals can cause health problems.

Some metals, such as iron, chromium and copper, are needed in small quantities to keep people and animals healthy. Problems can occur with these metals if the body receives too much of them. Heavy metals such as lead and mercury are never desirable in any amount. Once inside our bodies different metals can build up in different body parts, including the kidney, liver and spleen.

The metal particles (ion or molecule) are normally found in human body. Metals are entering to the human body through food (vegetable, Spinach - iron); air (inhalation of ultrafine particles) and water (ground water contains dissolved metal salts and minerals from earth strata). Few of the metals entered, are essential for life; whereas, others are useless and even may be life threatening. However through following mechanism they remain accumulated in the various parts like kidney, liver, brain, blood of human body.

V. Entry of Metals to Human Body

1. Digestive system: Entry of these metals and the nano materials to human body can take place through digestive system. Titanium Dioxide is used in cosmetics and vitamins and other drugs as a filler to construct tablets. But it is also used in skimmed milk, cheeses, yogurts, mayonnaise, chocolate marshmallows, manufactured fast foods, tomato ketchup. It is fed to animals that humans eat.

Titanium is used as food additive in European countries as a food coloring, it has E number E171. Human digestive system starts with mouth and ends with anus (Fig 1). In mouth mechanical disintegration of food takes place. Then it enters esophagus and liver then stomach. Along the way the necessary elements are absorbed.

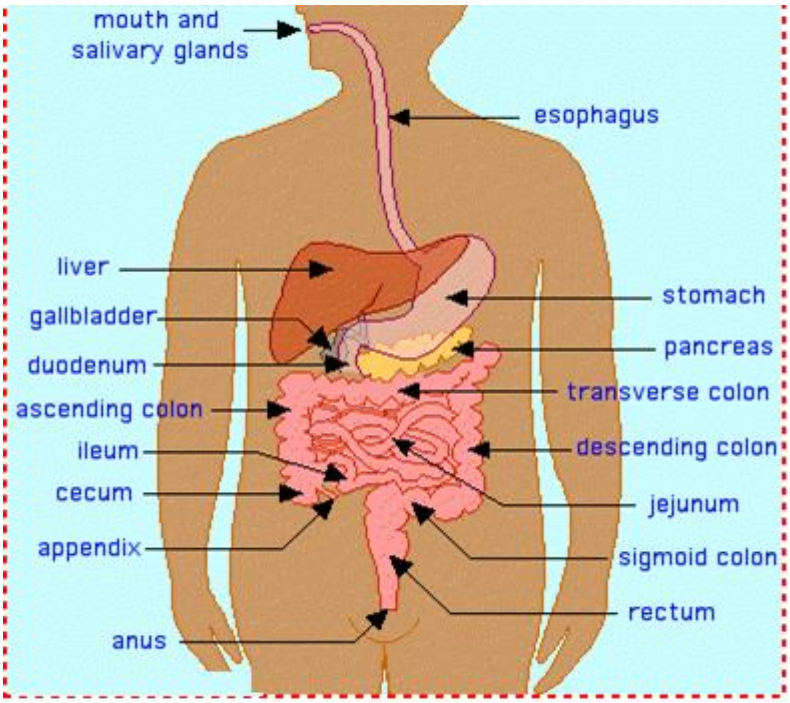

Fig. 1 Human Digestive System

Mechanism of absorption: Essential Metals are absorbed by the human body by developing homeostasis. Gastrointestinal absorption of trace and toxic elements is known to occur in three different phases:

1. The intraluminal phase with its chemical reactions and interaction with the content of stomach and intestine.
2. The translocation phase i.e. diffusion of trace elements across the cell membrane of the enterocytes and
3. Mobilization phase including mobilization and transport of the intracellular elements in to the blood stream or their sequestration back into the intestinal lumen.

The constant concentration of necessary metals in human body is maintained by the mechanism of homeostasis. E.g. iron absorption (Fig.2).

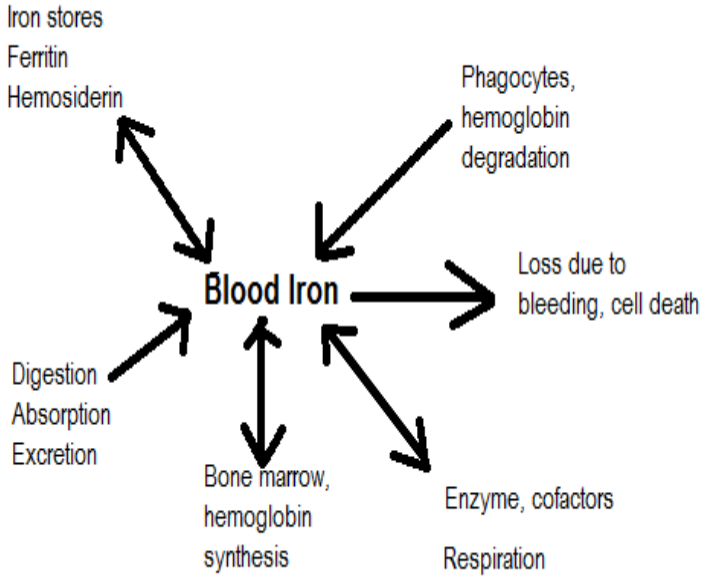

Fig.2 Homeostasis of iron in blood

VI. Entry of Metals and Nanoparticles through Respiration

The metal processing units are the sites where very fine particles of metal are floating as particulate matter in and around the units in close vicinity. Similarly the nano material processing units are rich with ultrafine, light weight nano particles floating in and out the units or laboratories and are likely to get inhaled by the workers over there. Titanium dioxide dust, when inhaled, has been classified by the International Agency for Research on Cancer (IARC) as an IARC Group 2B carcinogen, meaning it is possibly carcinogenic to humans [4].

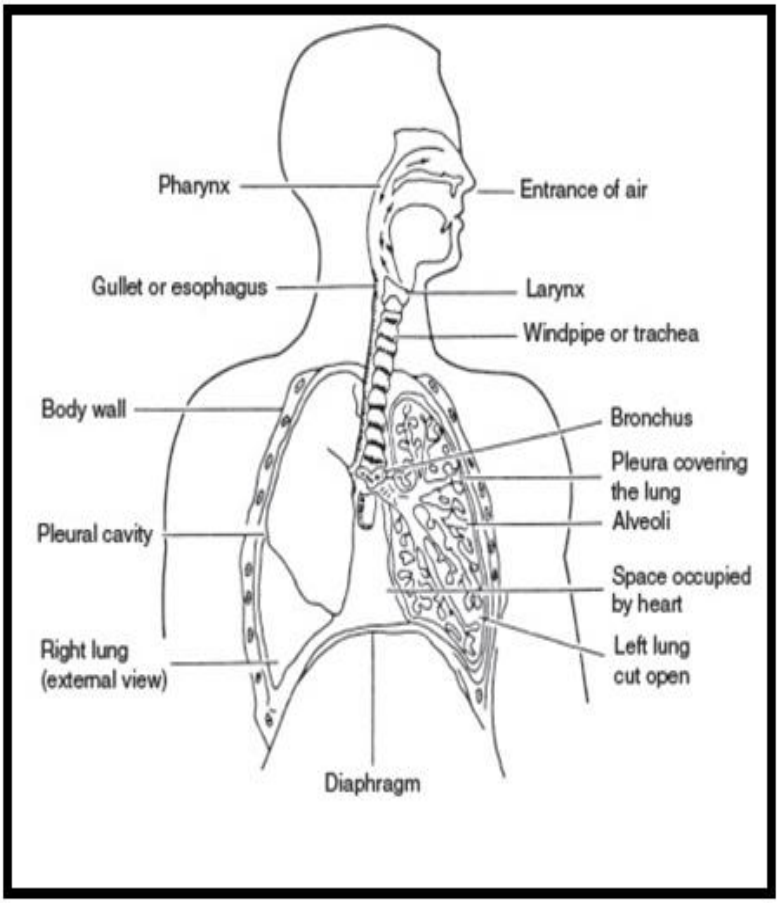

Fig.3 Human respiratory system

Excretion: Non- absorbed and Waste material (solid, liquid and gaseous form) is removed from the body. However in case of nanoparticles, it is observed that they are cytotoxic, can enter the blood stream and even can damage the DNA [5]. Studies on animals have investigated damage to the lungs (Fig. 3) by nano-TiO_2 varying from species to species of guineapigs [6]. The toxicity of nanoparticles is found to be dose dependent when the animal study was followed. Even the effect of size and shape is also pronounced one. In their study report, Ge et al [7] relates the toxicity reduction to bio-compatibility of the CNTs. By using both experimental and theoretical approaches they investigated the interactions of single-wall carbon nanotubes (SWCNTs) with human serum proteins, and find that morphology as well as chemical conformation played vital role in defining absorptivity of the CNTs. Sustainable energy harvesting, such as solar energy, depends increasingly on nanotechnology components. The principles of photovoltaic units and the toxicological aspects of its principal components, namely fullerenes and carbon nanotubes are well studied and discussed by Sergio Manzetti, Otto Andersen [8].

Accumulation: Non-absorbed and non-excreted material remains in the body and may get bio-accumulated in organs like kidney, liver, epithelial cells, and olfactory nerves. Biocidal effects and cellular internalization of ZnO nanoparticles on Escherichia coli bacteria are well studied by Brayner et al [9] in their study they used di(ethylene glycol) as medium to interact Zn nanoparticles with gram negative bacteria. The E. coli cells

after contact with DEG and ZnO were damaged showing a Gram-negative triple membrane disorganization. This behavior causes the increase of membrane permeability leading to accumulation of ZnO nanoparticles in the bacterial membrane and also cellular internalization of these nanoparticles.

Metal Human body interactions: There are environmental (water, air, soil, dust), occupational, medicinal and dietary sources of metal exposure. All the metals are cytotoxicity (ability to inactivate microbes by rupturing the cell wall). Few of them are beneficial to human bioactivity whereas most of them are neither beneficial nor essential to the human. The metals might be carcinogenic [10] for longer exposure of human to them (Table 1).

Table 1 Classification of metals based on characteristics of health effects

Nutritionally essential Metals	Metals with possible beneficial effects	Metals with no known beneficial effects	Metals with Toxic effects
Cobalt	Boron	Aluminum	Arsenic
Chromium III	Nickel	Antimony	Cadmium
Copper	Silicon	Barium	Lead
Iron	Vanadium	Beryllium	Mercury
Manganese		Silver	Beryllium
Molybdenum		Strontium	Nickel
Selenium		Thallium	
Zinc			

Chemical speciation has an impact on solubility, bioavailability, and persistence of metals and metal compounds in the environment; e.g. solubility is one of the major factors influencing bioavailability and absorption of metal and metal compounds.

Human body response to the metals and inorganic metal compounds: The human body responds differently to the inorganic materials. Table No. 2 reports the typical responses of human body to the metals.

Table 2 Classification of metals based on characteristics of health effects

Interaction	**Response**
Metabolism	It is limited to change in oxidation state, transition and pH alterations.
Presence	Sequestered, bound to specific plasma or tissue proteins or bone
Elimination	Being hydrophilic, eliminated in urine and bile
Absorption	Being in ionized state, membrane transport is the mean for tissue uptake.

Absorption of organic xenobiotics in the gastrointestinal track is favored by the lipid nature of intestinal cell membrane. Metals appear to diffuse through the outer surface of protein filament of the stratum conium which is hydrated; whereas lipophilic non polar organic molecules diffuse through the lipid matrix between the protein filaments [11].

Mechanism of metal absorption: Metallothioneins are a group of low molecular weight (aprrox. 6000 Daltons) proteins, rich in sulfydryl groups that serve as a legend for several essential and non essential metals. Metal absorption varies with its affinity towards the metallothioniens as researched by [12]. Table No. 3 enlist the metals and metal absorbing proteins.

Table 3 Metal selective metallothioneins

Metal	**Metallothionein Protein**	**Action**
Iron, Al, Mn	Transferrin, - glycoprotein	Transportation of Fe across cell membrane by binding with iron through plasma
Iron	Ferritin	Storage
Copper	Ceruloplasmin, -glycoprotein	Conversion of ferrous to ferric iron for transport using transferring
Lead	Lead binding protein	Blanketing of Pb in kidney and liver [13]
Metals	Membrane carrier Protein	Transport (multi metal specific) [14]

Replacement of essential metals during absorption with non-essential metals is possible. e.g. lead replaces zinc in hem synthesis by inhibiting the function of hem synthesizing enzyme [13]. Diminished iron absorption is observed in the presence of zinc. Substitution of calcium by lead results in toxicity of several vital enzyme systems in central nervous system [15, 16]. Arsenate replaces phosphate in mitochondria impairing the synthesis of ATP and energy metabolism.

Many of the processes controlling the disposition of metals are intrinsically capacity limited and highly specific. This makes it necessary to understand physiology well enough to model these processes and method to estimate binding constant. Another overarching theme is that metal – metal interactions of

multiple types. (e.g. competition antagonism and synergism as well as essential – nonessential interactions) commonly occur at multiple points during the process of absorption, distribution, metabolism and excretion.

VII. Conclusion

As mentioned in the introductory section of this paper, nano particles are highly reactive species with subatomic sizes i.e. easy to enter through breathing and along with food. Like skin, the gastro intestinal track and the lungs are in direct contact with the environment. Nano particles like Ti or Zn oxides are used in sunscreen lotions. CNTs and fullerence are the members of electronic industries. Nano particles from hair dye (pigments) accumulate around the hair roots. During hair growth, these follicles are opened and in turn provide a route for nano particles to reach dipper layers. No uptake into the blood through healthy skin has yet been reported. Animal experiments show that nano particles can cross the air blood tissue barrier and thus enter in to the body circulatory system. As homeostasis is significant to retain or omit the metals from human body and the metal – metal interactions are of multiple types, it becomes necessary to understand biochemistry of metal absorption, retention, excretion and bio-accumulation to understand the effects of nanoparticles over human body. Thus it can be said that the regulatory strategy over use of nanomaterials, till the fatality of nanomaterials in environment is unknown, is highly necessary.

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Modeling and Simulation of Frequency Converter Used In Speed Control of Asynchronous Motor

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Abstract- Controlling the speed of the Induction Motor is very essential in today's industry. Conventionally Gear box is used for speed control which consumes more energy. This Paper presents the Modeling and Simulation of frequency converter used for speed control of asynchronous motor through digital approach which has high reliability and energy conservation by using MATLAB/ SimulinkTM. The Frequency Converter is mainly used as variable frequency source to adjust the speed of asynchronous motors. The Frequency Converter circuit consists of three-phase Rectifier, IGBT module, DC link and the three phase inverter. The Transistor three-phase inverter converts the DC voltage into variable voltage and variable frequency (VVVF) power source. The commonly used Capacitors bank is replaced by the LC filter. The Inductor and the capacitor are used to reduce switching stresses so that during power up the voltage source inverter is protected by the inrush currents. The circuit is simulated and output results are obtained.

Index Terms- Asynchronous Motor, Pulse Width Modulation (PWM), MATLAB/SimulinkTM, variable voltage and variable frequency (VVVF) power source

I. INTRODUCTION

The three phase asynchronous motors are the primary movers in all the industrial applications in each stage of processing and manufacturing. Generally these asynchronous motors are implemented due to low cost and reliability. This converter transforms fixed frequency of power line into constant DC voltage via diode three-phase rectifier. Then the inverter converts fixed frequency of power line into variable voltage and variable frequency (VVVF) power source. A VVVF AC drive is a power electronic controller used to control the speed of 3hp AC motors (synchronous or induction) by varying the frequency and the voltage applied to the motor terminals.

Moreover this converter can keep high power factor and it greatly reduces the interference of variable frequency supply to the network power line. The project is simulated using MATLAB/ SimulinkTM by using different components such as three-phase rectifier, DC link and the IGBT module which constitute the inverter operation. The inverter has to produce a motor current within acceptable level of current ripple, which should be lower than 5%for most of the applications. A MOSFET can be employed for low power (up to a bound of few KW's) and work efficiently up to 50 KHz. This handles low inductance up to few hundred µH. In order to sustain high power levels such as tens of kW's, IGBT modules are used as switching devices and can switch up to 20 kHz. These are available in two- or six-pack modules. The modulation technique used here is the Pulse Width Modulation (PWM).

II. PROTOTYPE MODELLING

The prototype mainly constitutes of Three-phase Rectifier, Inverter DC link, Three-phase Inverter and asynchronous motor. The design aspect consists of:

A. Three Phase Rectifier

Rectifiers are often found serving as components of DC power supply and high-voltage direct current power transmission system. The input three phase voltage (1)-(3) are given below:

$$Vsp = Vs\,sin(\omega t) \quad (1)$$

$$Vsq = Vs\,sin\left(\omega t - \frac{2\pi}{3}\right) \quad (2)$$

$$Vsr = Vs\,sin\left(\omega t + \frac{2\pi}{3}\right) \quad (3)$$

Where the Vsp, Vsq, Vsr are the input voltages to the three phase rectifier with magnitude Vs and angular frequency ω and the above equations can be approximated to

$$Vpn \approx Vpn1 = Vpeak\ sin(\omega t - \emptyset) \quad (4)$$

$$Vqn \approx Vqn1 = Vpeak\ sin(\omega t - \emptyset - \frac{2\pi}{3}) \quad (5)$$

$$Vrn \approx Vrn1 = Vpeak\ sin(\omega t - \emptyset + \frac{2\pi}{3}) \quad (6)$$

Where Vpeak = (4)/ (π) (Vo/2+ Vd) and Vo is the constant-voltage output of the rectifier and ϕ is the phase angle between each voltage source (Vsp; Vsq; Vsr) and their line-to-neutral voltages are Vpn, Vqn, Vrn for each phase. From the above given (4)-(6) equations the current equations can be as (7)-(9) as shown below and the phase currents will have the form

$$Ia \approx Ip1 = Is1\ sin(\omega t - \emptyset) \quad (7)$$

$$Ib \approx Iq1 = Is1\ sin(\omega t - \emptyset - \frac{2\pi}{3}) \quad (8)$$

$$Ic \approx Ir1 = Is1\ sin(\omega t - \emptyset + \frac{2\pi}{3}) \quad (9)$$

The line-to-neutral voltages Vpn1, Vqn1, Vrn1 are in phase with their respective line currents ia1, ib1, and ic1.The phasor of the line current in phase 'a' is given by:

$$Ia = \frac{Vs}{\sqrt{(R)^2+(\omega Ls)^2}} e^{-jtan-1(\omega Ls/R)} = Is1\ e^{-j\phi} \quad (10)$$

The average and root-mean-square output voltages of an ideal single-phase full-wave rectifier are:

$$Vdc = Vav = \frac{2Vpeak}{\pi}$$

$$Vrms = \frac{Vpeak}{\sqrt{2}}$$

For a three-phase full-wave rectifier the average output voltage is

$$Vdc = Vav = \frac{(3\sqrt{3})Vpeak}{\pi} \cos\alpha$$

Where:

α = firing angle of the thyristor (0 if diodes are used to perform rectification)

The rectifier is modeled with the following parameters:

Table 1: rectifier parameters

No of Bridges	3
Snubber Resistance	100Ω
Snubber Capacitance	0.1e-6 F
Device used	Diodes
Forward Voltage	0.8V
Ron	1e-3 Ω

B. DC Link

An AC-AC converter with approximately sinusoidal input currents and bidirectional power flow can be realized by coupling a rectifier and a PWM inverter to the DC-link. The DC-link quantity is then impressed by an energy storage element that is common to both stages, which is a capacitor C for the voltage DC-link or an inductor L for the current DC-link.

Due to the DC-link storage element, there is the advantage that both converter stages are to a large extent decoupled for control purposes. Furthermore, a constant, AC line independent input quantity exists for the PWM inverter stage, which results in high utilization of the converter's power capability. On the other hand, the DC-link energy storage element has a relatively large physical volume, and when electrolytic capacitors are used, in the case of a voltage DC-link, there is potentially a reduced system lifetime. The large electrolytic capacitors can be replaced by the LC Filter. During the switching of the inverter the higher order harmonics are removed by the LC Filter.

The LC Filter costs up to one sixth of the electrolytic capacitor bank and also occupies less space. Another advantage of the LC Filter circuit is that it shields the Voltage source inverter from current inrush during power up. When the DC link current is continuous and if U is the RMS value of line-to-line voltage of mains then the DC link voltage will be the solution of equation (11):

$$Udc = \frac{3\sqrt{2}}{\pi} U = 1.35U \quad (11)$$

The DC link is modeled with the following parameters:

Table2: DC link parameter

Inductance	200e-6 H
Capacitance	4700e-6 F

C. Three-Phase Inverter

After the signal is rectified the energy is stored in the DC link inductors for the current source inverter and the DC link capacitor for the voltage source inverter. The IGBT of the inverter bridge takes the energy in the form of the DC mode and has to deliver the sinusoidal wave to the induction motor. The speed of the motor is controlled by the fundamental components frequency of this AC waveform.

The prototype of the inverter is shown in Figure1 below:

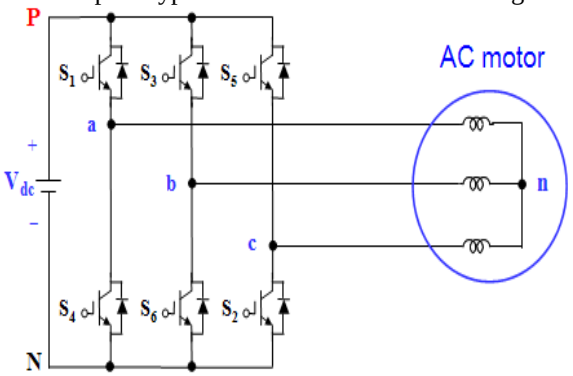

Figure1: Inverter Block Diagram

The line-to-line voltage of the output of the inverter can be expressed as:

$$Vl = \sqrt{\left[\int_0^{\frac{2\pi}{3}} (Vs)^2 d(\omega t)\right]}$$

$$Vl = \sqrt{\frac{2}{3}} Vs = 0.8165Vs \quad (12)$$

The Three-phase Inverter is modeled with the following parameters:

Table3: Three phase inverter parameter

Bridges	3
Snubber Resistance	1e5 Ω
Snubber Capacitance	∞
Device used	IGBT/Diodes
Ron	0.2e-3
Tf(s), Tt(s)	1e-6,2e-6

III. OPERATING PRINCIPLE

The inverter is fed by a dc voltage and has three phase-legs each consisting of two transistors and two diodes. With PWM control, the switches of the inverter are controlled based on a comparison of a sinusoidal control signal and a triangular switching signal. The sinusoidal control waveform establishes the desired fundamental frequency of the inverter output, while the triangular waveform establishes the switching frequency of the inverter. The ratio between the frequencies of the triangle wave and the sinusoid is referred to as the modulation frequency ratio. The switches of the phase legs are controlled based on the following comparison is shown in Figure2:

$$v_{control(phase-a)} > v_{triangle},\ T_{a+} \text{ is on}$$
$$v_{control(phase-a)} < v_{triangle},\ T_{a-} \text{ is on}$$
$$v_{control(phase-b)} > v_{triangle},\ T_{b+} \text{ is on}$$
$$v_{control(phase-b)} < v_{triangle},\ T_{b-} \text{ is on}$$
$$v_{control(phase-c)} > v_{triangle},\ T_{c+} \text{ is on}$$
$$v_{control(phase-c)} < v_{triangle},\ T_{c-} \text{ is on}$$

Figure 2: Switching of Transistors in inverter

In most instances the magnitude of the triangle wave is held fixed. The amplitude of the inverter output voltages is therefore controlled by adjusting the amplitude of the sinusoidal control voltages.

The ratio of the amplitude of the sinusoidal waveforms relative to the amplitude of the triangle wave is the amplitude modulation ratio. The diodes provide paths for current when a transistor is gated on but cannot conduct the polarity of the load current. For example, if the load current is negative at the instant the the upper transistor is gated on, the diode in parallel with the upper transistor will conduct until the load current becomes positive at which time the upper transistor will begin to conduct.

By controlling the switches in this manner, the line-line inverter output voltages are ac, with a fundamental frequency corresponding to the frequency of the sinusoidal control voltage. In most instances the magnitude of the triangle wave is held fixed. The amplitude of the inverter output voltages is therefore controlled by adjusting the amplitude of the sinusoidal control voltages. The output waveforms are shown below in figure.3:

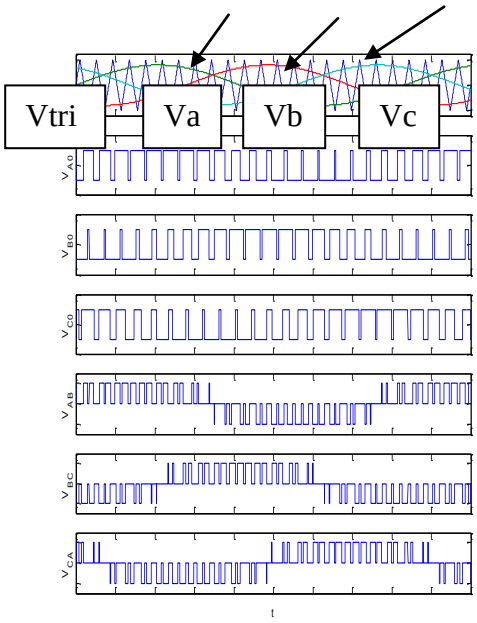

Figure 3: Waveforms of three-phase sine PWM inverter

When $v_{control} > v_{tri}$, $V_{A0} = V_{dc}/2$

When $_{vcontrol} < v_{tri}$, $V_{A0} = -V_{dc}/2$

Where $V_{AB} = V_{A0} - V_{B0}$

$V_{BC} = V_{B0} - V_{C0}$

$V_{CA} = V_{C0} - V_{A0}$

The gating sequence is shown below in Figure4:

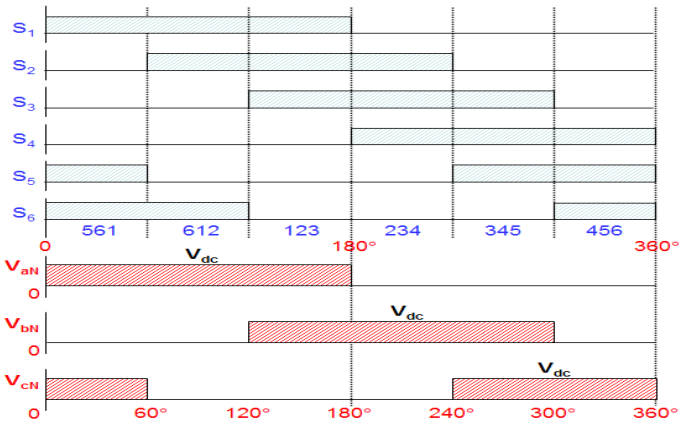

Figure.4: Waveforms of gating signals, switching sequence, line to negative voltages

The switching sequence is shown in Figure5:

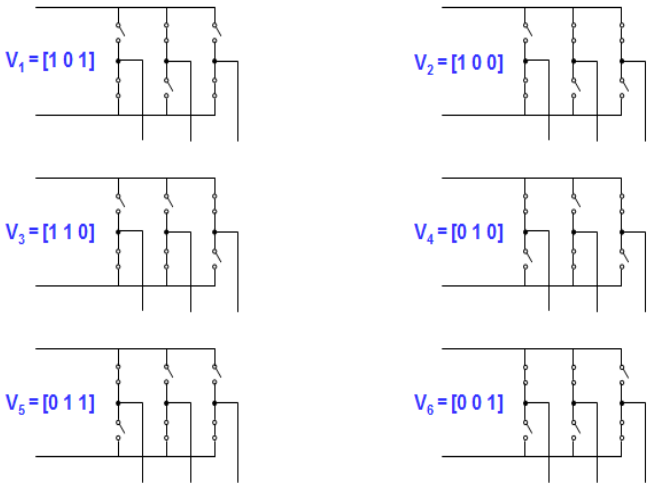

Figure 5: Six inverter voltage vectors

The output of the inverter is the PWM signal in the form of the square wave. This square wave constitutes voltages Vab, Vbc, Vac which are given by the equations (13) and (14).

$$Vab = \frac{Ao}{2} + \sum_{n=1}^{\infty} (An\cos(n\omega t) + Bn\sin(n\omega t))$$

$$Bn = \frac{1}{\pi}\left[\int_{\frac{-5\pi}{6}}^{\frac{5\pi}{6}} -Vsd(\omega t) + \int_{\frac{\pi}{6}}^{\frac{5\pi}{6}} -Vsd(\omega t)\right]$$

$$Bn = \frac{4\pi}{n\pi}\sin(\frac{n\pi}{2})\sin(\frac{n\pi}{3})$$

$$Vab = \sum_{n=1,3,5\ldots}^{\infty} \frac{4Vs}{n\pi}\sin(\frac{n\pi}{3})\sin n(\omega t + \frac{\pi}{6})$$

(13)

$$Vbc = \sum_{n=1,3,5\ldots}^{\infty} \frac{4\pi}{n\pi}\sin\frac{n\pi}{3}\sin n(\omega t - \frac{\pi}{2})$$

(14)

The voltage between the output terminals '**a**' and '**c**' is given by:

$$Vca = \sum_{n=1,3,5,\ldots}^{\infty} \frac{4Vs}{n\pi}\sin\frac{n\pi}{3}\sin n(\omega t - \frac{7\pi}{6})$$

(15)

IV. SIMULATION RESULTS

The circuit consists of rectifier, DC link, three phase inverter and an asynchronous motor connected to the output of the three phase inverter which is used to drive the motor. The whole circuit is simulated using MATLAB/Simulink and the respective circuit diagram and the waveforms are visualized in the below figures.

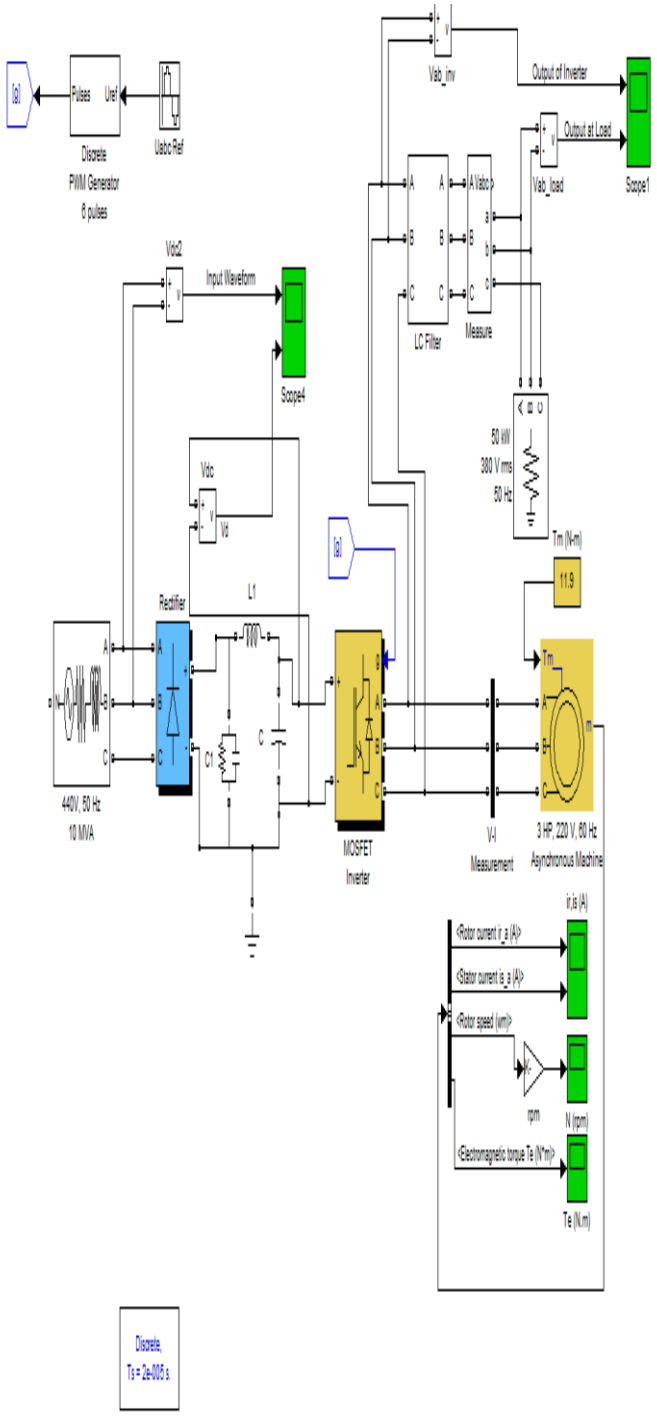

Figure 6: Proposed Circuit diagram

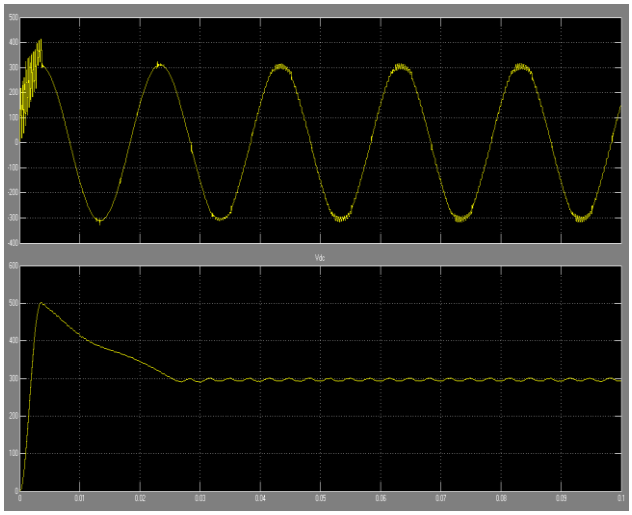

Figure 7 :a) one Input waveform of three-phase
b) DC link output

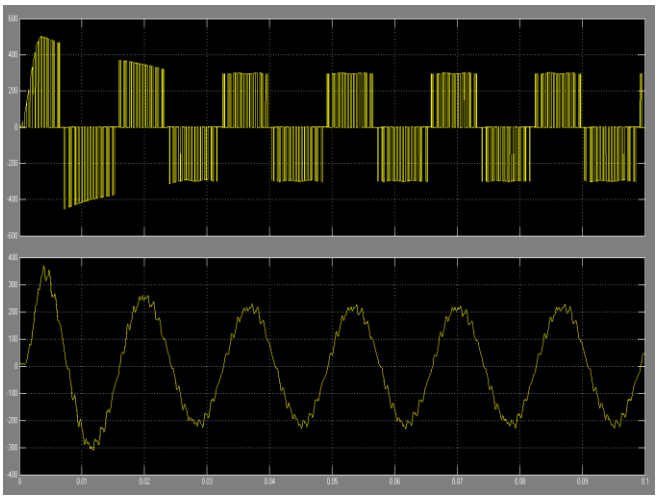

Figure 8: a) Inverter output b) Load output

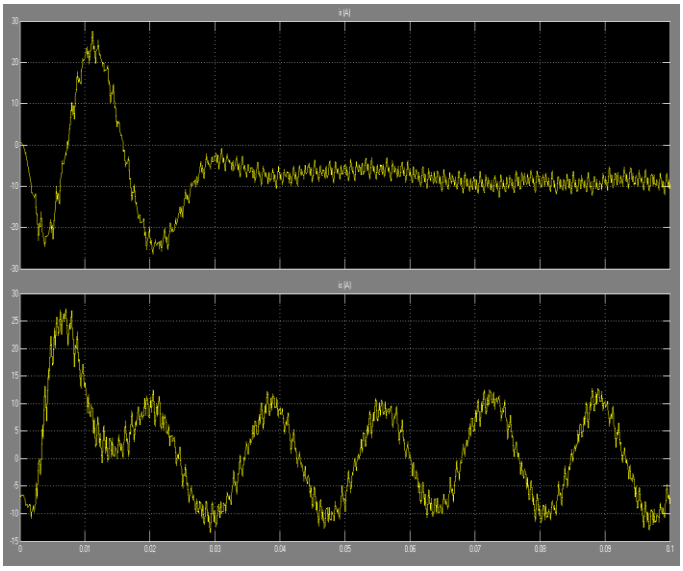

Figure 9: a) Rotor Current b) Stator Current

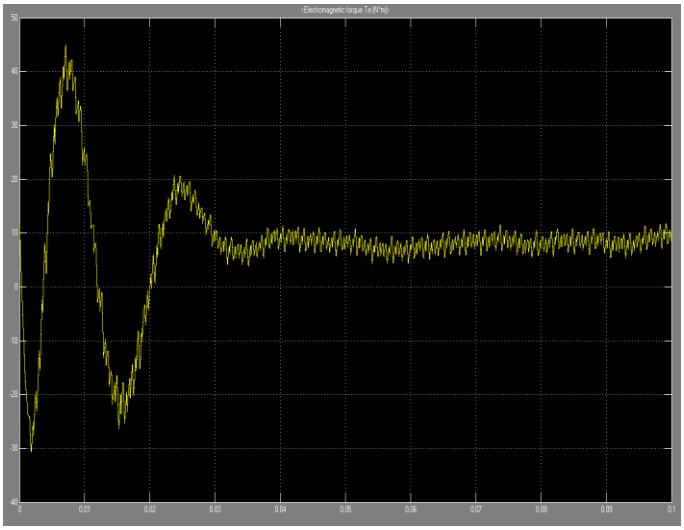

Figure 10: Electromechanical Torque (N.M)

V. CONCLUSION

This paper presents the modeling of the circuit which can be used in Frequency Converters. The circuit described above has been simulated using MATLAB/Simulink and the output graphs such as voltage at the inverter, voltage at the load, Rotor current, Stator current and the Torque waveform are obtained and which are in sync with the desired output. This circuit eliminates the use of large electrolytic capacitors by using LC filter for lower input and output distorted currents. This circuit uses IGBT which have better input and output characteristics than BJT. This circuit can be used in Frequency converters which are generally used in different Industrial applications to cater operations such as Grinding.

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Classification and Grading Rice Using Multi-Class SVM

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Abstract- This paper proposes machine algorithm to grade (Premium, Grade A, Grade B and Grade C) the rice kernels using Multi-Class SVM. Maximum Variance method was applied to extract the rice kernels from background, then, after the chalk has been extracted from rice. The percentage of Head rice, broken rice and Brewers in rice samples were determined using ten geometric features. Multi-Class SVM classified the rice kernel by examining the Shape, Chalkiness and Percentage of Broken (Head Rice, Broken and Brewers) kernels. The SVM classify accurately more than 90% .Based on the results, it was concluded that the system was enough to use for classifying and grading the different varieties of rice grains based on their interior and exterior quality.

Index Terms- Support Vector Machine, Binarization, Kernel, Threshold, Machine Vision, Segmentation.

I. INTRODUCTION

Rice is grown in many regions across India. India is the second leading producer of rice in the entire world, preceded only by China. In Southeast Asia alone, rice is staple food for 80% of the population. As countries reach self-sufficiency in rice production, the demand by the consumer for better quality rice has increased. The Quality of rice is based upon many parameters such as color, size, shape and no of broken kernels. Rice Quality inspection by humans is neither objective nor efficient. In view of this automated rice quality inspection using machine vision is desirable. Recently, machine vision and image processing are widely used in biological and agricultural research with improvement of computer technology a significant reduction of cost of hardware and software of digital imaging. Many studies have applied image processing to grain quality inspection. Machine vision (MV) is a rapid, economic, consistent and objective inspection and evaluation technique [1].

There are many researches applied machine vision to estimate rice appearance quality inspection [2-16]. Sapirstein et al. (1987)employed the image processing to identify grains such as wheat, oats and rye using the parameters such as area, length, width, perimeter, color and shape [2]J.Aulakh et al.(2012) find the percentage purity of hulled rice grain sample by image processing technique[3]. K.Ding et al., 1994 applied the back propagation neural classifier for food shape classification [4]. J.A. Patindol, 2000 suggested grading criteria for rice in terms of defects found in rice samples such as chalky, broken, immature and damaged kernels [5]. T.Brosnan et al., 2003 reported that quality of rice is based on variety of properties such as size, shape, color, chalkiness and no. of broken rice kernels, color and chalkiness [6]. Guzman et al., 2008 investigated the use of machine vision system and multilayer neural networks to classify Philippine rice grains including shape and size parameters [7]. C.Agustin et al., 2008 suggested probabilistic neural network to sort rice kernels into head rice, broken rice and brewer's kernels [8]. Q.Yao et al., 2009 developed an automatic system to inspect the rice appearance quality including rice chalkiness and shape [9]. L.Guangrong, 2011 proposed a method to calculate the chalky area ratio and chalky grain rate and then automatic detection of chalk degree has been realized [10]. S.Shantaiya et al., 2010 used the morphological and color features to identify different varieties of rice using feed forward neural network [11]. B.Verma, 2010 used the back propagation through time neural to sort the rice into chalky, sound, and broken kernels [12]. M. Yao et al., 2010 develop an inspection system of rice exterior quality (head rice rate, chalk rice, crackle rice) based on computer vision [13].

Burges et al.,1998 reported that the Support Vector Machine (SVM) is a new pattern classifier, set of supervised generalized linear classifiers that have often been found to provide higher classification accuracies than other widely used pattern classification techniques, such as multilayer perceptron neural networks[14]. S Noble, 2006 reported that Support Vector Machine (SVM) is a computer algorithm that learns by example to assign labels to objects [15]. M. Galar et al., 2011 suggested the ensemble methods for binary classifiers in multi-class problems, one-vs.-one and one-vs.-all schemes [16].

These researches provided some new ideas and image processing methods for evaluating rice appearance quality. The effectiveness and accuracy of inspection have been improved through these methods.

The objective of this study is to develop algorithm to grade the rice grains into Premium Quality, Grade A, Grade B and Grade C, with help of support vector machine by distinguishing between healthy and unhealthy rice by rough rice (Head Rice, Broken, Brewers, chalky, Long, Short, Selender, Bold and Round).

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II. METHODOLOGY

Indian Basmati rice seed samples were taken up for grading and classification. The basic steps in the classification and grading of rice are as follows:

- Scanning the Rice grain sample.
- Image smoothening
- Image Segmentation.
- Image Binarization.
- Calculation of chalk Volume.
- Analysis of Shape of Rice Grains.
- Analysis of purity of Rice Grain.
- Classification Module

2.1 Scanning the Rice grain sample

Firstly, we put the rice grains on the plane of the scanner which is covered with black plastic sheet as background. Then we scan those rice grains to obtain an original image shown in figure.

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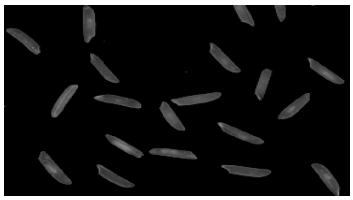

Fig.1 Rice Image

2.2 Image Smoothening

First the image will be converted into grayscale image[0,256] and the image will preprocessed using a smoothing filter (median filter) that included operations which can enhance and smooth images, accentuating image edges and remove noise from an image.

2.3 Image Segmentation

Image segmentation is the first step of image analysis. An image is subdivided into its constituent parts or objects. The image is usually subdivided until the objects of interest are isolated. There are so many approaches for segmentation algorithms. In this study, our main goal is to extract the rice from the background. Maximum variance method is used to segment the image into foreground and background regions.

2.4 Image Binarization

The result of segmentation is usually a binary image. A binary image contains only two types of pixels: the pixels having a gray level value of either 0 or 1.So after segmentation the rice image will be converted into the binary image [0, 1].

2.5 Calculation of Chalky Volume of Rice Grain

Chalk volume is defined as the opaque volume in the rice grain. The procedure to calculate the chalky volume of rice is as follows:

- The number of connected regions in the image is obtained after the segmented binary image of rice is counted.
- Then do the labeling of rice grains.
- Compute the best chalky segmentation threshold.
- Process the connected region of rice with threshold to segment the chalky rice grain.
- Compute the boundary values of each object with help of Harris Corner Algorithm.
- Compute the Convex Hull using the corner points and we will be with volume of chalky rice kernel.

Chalk degree is defined as the percentage of chalky volume of chalky rice and total volume of rice grains is calculated as follows:

$$\text{Chalk Degree (\%)} = \frac{Chalky\ Volume\ of\ Chalky\ Rice}{Total\ Volume\ of\ Rice\ Grains}$$

Chalky grain rate is defined as a percentage of chalky rice grains and total of rice grains namely:

$$\frac{Chalky\ Rice\ Grains}{Total\ of\ Rice\ Grains}$$

Chalky Grain Rate (%) =

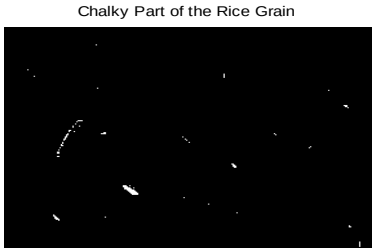

Fig.2 Chalky Part Segmentation

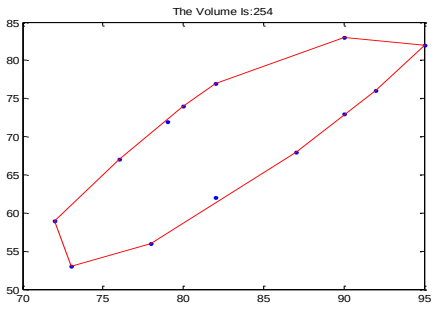

Fig.3 Volume of 1st Grain in Rice Image

2.6 Analysis of Shape of Rice Grain

Rice grain shape is described by length, width and their ratio and are categorized as slender (ratio greater than 3.0), medium (ratio within 2.1-3.0), bold (ratio within 1.1-2.0) and round (ratio less than 1.1).

2.7 Analysis of purity of Rice Grain

Purity test of rice sample is done according to the size of the grain kernel (full, half or broken).Head rice is a kernel or piece of kernel with its length equal to or greater than 75% of the average length (grain size) of unbroken kernel. Broken kernels are kernels whose lengths are 75% of the grain size. Brewers are small pieces or particles of
kernels whose length equal to or less than 25% of the grain size.

Fig.4 Tracing Shape of Rice Grains

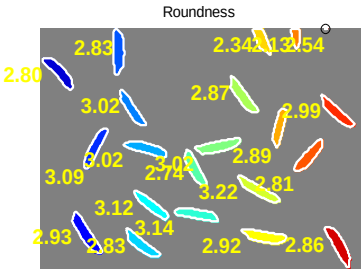

Fig.5 Roundness Ratio

8) Classification Module

A support vector machine (SVM) a new pattern classifier was trained for classification of the samples into the grades (Premium, Grade A, Grade B and Grade C).

A support vector machine (SVM) performs classification by mapping input vectors into a higher-dimensional space and constructing a hyper-plane that optimally separates the data in the higher-dimensional space. SVM models are closely related Using a kernel function, SVMs are alternative training methods for polynomial, radial basis function, and multi-layer perceptron classifiers in which the weights of the network are found by solving a quadratic programming problem with linear constraints, rather than by solving a non-convex, unconstrained minimization problem as in standard neural network training. Given a training set of instance-label pairs x_i, y_i ,i =1, 2,3…..l, $x_i \in R^n$, $y_i \in \{-1,1\}$,the SVM require the solution of the following optimization problem:

$$\min_{w,b,\xi} \left\langle \frac{1}{2} w^T w + C \sum_{i=1}^{l} \xi_i \right\rangle \quad (1)$$

$$\text{subject to } y_i(w^T \phi(x_i) + b) \geq 1 - \xi_i, \xi_i \geq 0$$

Here, training vectors x_i (i=1, 2, 3…….l) are mapped into a higher-dimensional space by the function f . Then the SVM finds a linear separating hyperplane (*w, b*) with the maximal margin in this higher-dimensional space. *C* > 0 is the penalty parameter of the error term. The slack variables, y_i (i=1,2,3……..l) measure the degree of misclassification of x_i .**T**he SVM does not require an estimation of the statistical distributions of classes to carry out the classification task, but it defines the classification model by exploiting the concept of margin maximization.

Furthermore, $K(x_i, x_j) = \phi(x_i)^T \phi(x_j)$ is called the kernel function, with which the computational problem of many dimensions is solved. Although new kernels are being proposed by researchers, the radial basis function (RBF) is a reasonable first choice. This kernel nonlinearly maps samples into a higher dimensional space so it, unlike the linear kernel, can handle the case when the relation between class labels and attributes is nonlinear.

$$K(x_i, x_j) = \exp(-\gamma \| x_i - x_j \|^2), \gamma > 0 \quad \textbf{(2)}$$

The RBF kernel nonlinearly maps samples into a higher-dimensional space, so it can handle nonlinear classification issues. With certain parameters (C, γ) the RBF kernel has the same performance as the linear kernel or the sigmoid kernel.There are two parameters while using RBF kernels: C and γ .It is not known beforehand which C and γ are the best for one problem; consequently, some kind of parameter search must be done. Cross-validation is commonly utilized to identify good (C, γ) so that the classifier can accurately predict unknown (independent) data.A common strategy is to separate the data set into two parts, of which one is considered unknown. The prediction accuracy obtained from the "unknown" set more precisely reflects the performance on classifying an independent data set. An improved version of this procedure is known as cross-validation. In n-fold cross-validation, we first divide the training set into n subsets of equal size. Sequentially one subset is tested using the classifier trained on the remaining n-1 subsets. Thus, each instance of the whole training set is predicted once so the cross-validation accuracy is the percentage of data which are correctly classified.

However, a major drawback of SVMs is that, from a theoretical point of view, they were originally developed to solve binary classification problems. Multi-class classifiers are typically constructed by combining several binary classifiers.

In this study, 400 kernels were randomly selected, with 100 kernels from each group of chalky, broken, long & round and sound rice samples, which were divided into a training set with 75 kernels and a testing set with 25 kernels.

The training set was used to train the SVM model, while the independent test set was used to test the model performance. Ten order statistical characteristics (length, area, perimeter, etc.) provided effective classification features. A one-against-one multi-class support vector machine with kernel of RBF was trained by the transformed training pattern vectors and tested by the transformed independent test feature vectors. The image processing methods were implemented in Matlab R2011a and the SVM algorithms in LIBSVM.

III. RESULTS

The Support Vector Machine helped grade and classify rice kernels accurately (better than 90%) and that too at a nominal cost. A few of the grading reports generated by the system are listed below. (Figure 6 to 9).

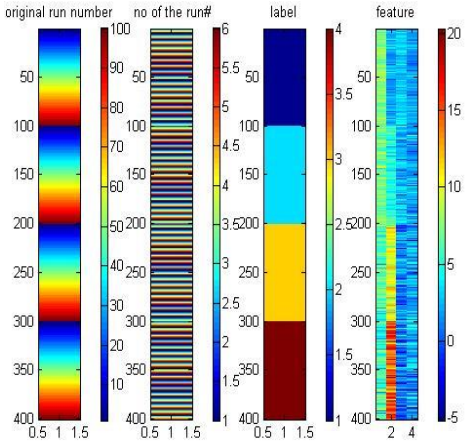

Fig.6 Data before Scaling & Sorting

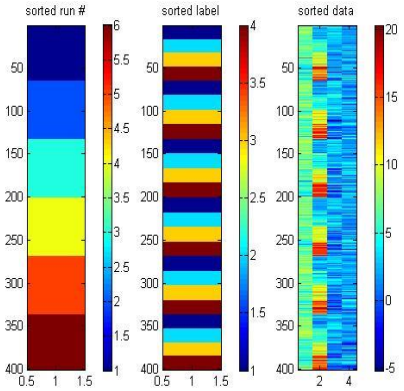

Fig.7 Data after Sorting

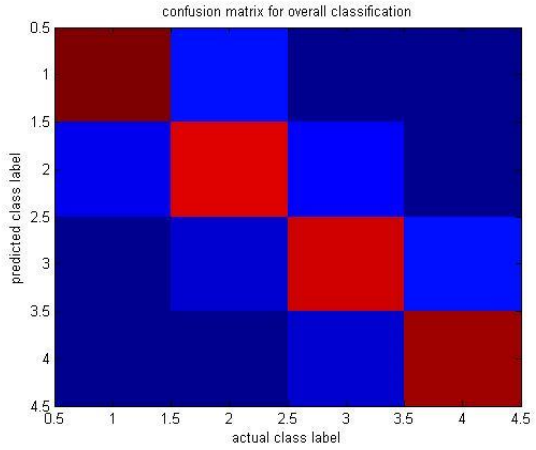

Fig.8 Confusion Matrix of Overall Classification

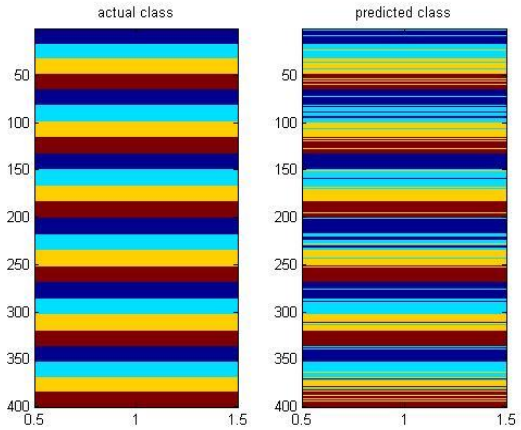

Fig.9 Comparison of Actual Class & Predicted Class

IV. CONCLUSION

This study shows the use of support vector machine in classifying and grading the rice grain. By improving the speed classification systems, we would focus on online quality measurement with this technique in the future

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Blotches and Impulses Removal in Colourscale Images Using Non-Linear Decision Based Algorithm

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Abstract- In this paper proposed blotches and impulse removal in color scale images using nonlinear decision based algorithm. The implementation of this algorithm can be obtained by two stages. In first stage the pixel are detected as corrupted / uncorrupted using decision rules. In second stage estimate the new pixel value for corrupted pixels. The algorithm used as a adaptive length window whose maximum size is 5X5 to avoid blurring due to large window sizes. The mean filtering is automatically switched in this proposed algorithm. It also tests the different images. To analyze the performance of this algorithm as mean square error, peak signal to noise ratio, computation time and image enhancement factor compare to other algorithm. The noise level is effectively removal without any loss it produces the better result in quantitative and qualitative measures of the image and also provides better performance.

Index Terms- Median Filter, MSE, Decision based algorithm, Impulse Noise, Blotches, PSNR.

I. INTRODUCTION

An image is an array or a matrix of square pixels (picture elements) arranged in columns and rows. An image (from Latin: image) is an artifact, for the example of a two-dimensional picture, that has a similar appearance of some subject usually a physical object or a person.

Images may be two-dimensional, such as a photograph, and screen display, and as well as a three-dimensional, such as for a statue or hologram. They may be captured in optical devices such as telescopes, lenses, mirrors, cameras, microscopes, etc., and natural objects and phenomena, such as the human eye or the water surfaces. The word image is also used in the broader sense of any two-dimensional figure such as a map, a pie chart, a graph or an abstract painting. In this wider sense and images can also be rendered manually, such as by painting, drawing, carving, rendered automatically by computer graphics technology, or printing, or developed by a combination of methods and especially in a pseudo-photograph.

Image enhancement is the improvement of digital image quality (wanted e.g. for visual inspection or for machine analysis), without knowledge about the source of the degradation. If the source of degradation is known, one calls the process of the image restoration. Both are conical processes, viz. input and outputs are images.

In remote sensing, artifacts such as drop lines, strip lines, blotches, impulse noise. In this Method the removal of botches and the presence of impulse noise is addressed. Blotches are characterized for being impulsive and randomly distributed, with irregular shapes of approximately constant intensity. There are originated by warping of the substrate or emulsion, dust, mould, dirt or other unknown causes.

Linear filters are not quite effective in the presence of non-Gaussian noise. In the last decade, it has been shown that nonlinear digital filters can overcome some of the limitations of linear digital filters. Mean filtering is a simple, intuitive and easy to implement method of smoothing images,

That is reducing the amount of intensity variation between one pixel and the next. It is often used to reduce the noise in images. The idea of mean filtering is to simply replace each pixel value in an image with the mean (‘average') value of its neighbors and including itself. This has been effect of eliminating pixel values which are unrepresentative of their surroundings. Mean filtering is most commonly used as a simple method for reducing noise in an image.

Median filters are a class of nonlinear filters and have produced good results where linear filters generally fail. Median filters are known to remove impulse noise and preserve edges. There are a wide variety of median filters in the literature. In remote sensing, artifacts such as strip lines, drop lines, blotches, band missing occur along with impulse noise. Standard median filters reported in the literature do not address these artifacts.

In the approach is using for median filter of adequate window size is used for the detection of corrupted pixel. The difference between the pixel of interest and the median filtered output is obtained and compared with the threshold obtained from the minimum and maximum pixel values in the chosen window. A binary flag image is obtained with its values 1 for the corrupted pixels and 0 for the uncorrupted pixels. The corrupted pixel values are estimated for the new values using the median/mean filtering.

II. PREVIOUS ARCHITECTURE

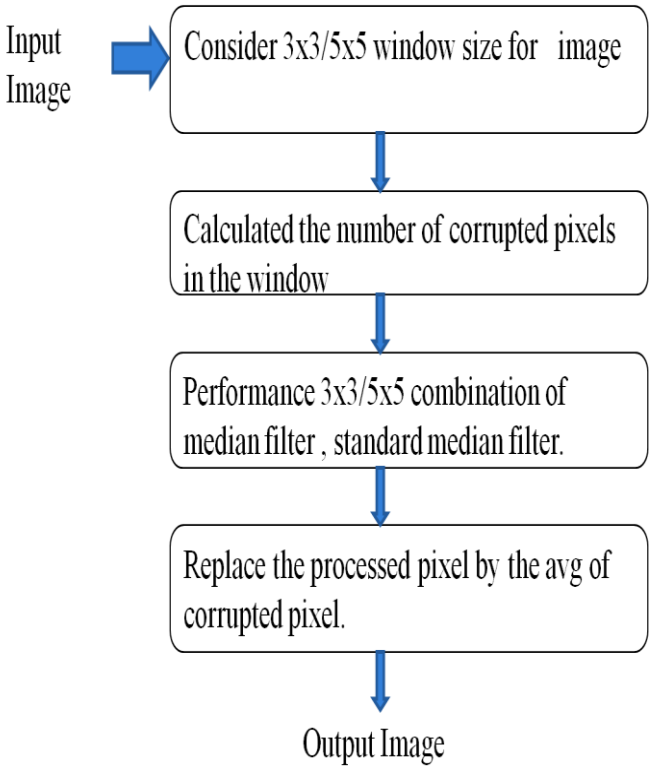


Figure 1: Flow chart of Previous Architecture

Image processing is a subset of the electronic domain wherein the image is converted to an array of small integers, called pixels, representing a physical quantity such as scene radiance, stored in a digital memory and processed by computer or other digital hardware.

Suppose the image, a photo, say. For the moment, let's make things easy and suppose the photo is black and white (that is, lots of shades of grey), so no color. Consider this image as being a two dimensional function, where the function values give brightness of the image at any given point. Assume that in such an image brightness values can be any real numbers in the range 0.0 (black) to 1.0 (white). The ranges of x and y will clearly depend on the image, but they can take all real values between their minima and maxima.

A digital image differs from a photo in that the x, y, and f(x,y)values are all discrete. Usually they are take on only integer values, so the image will have x and y ranging from 1to 256 each, and the brightness values also ranging from 0 (black) to 255 (white). A digital image can be considered by a large array of discrete dots and each of which has a brightness associated with it. These dots are called by picture elements, or more simply pixels. The pixels are surrounding a given pixel constitute its neighborhood.

A neighborhood can be characterized by its shape in the same way as a matrix of a 3*3 neighborhood, or of a 5*5 neighborhood. Except in very special circumstances, neighborhoods have odd numbers of rows and columns; this ensures that the current pixel is in the centre of the neighborhood.

A. Noise Model

An image is mention by ***I*** and the image size for M x N and x(i,j) is its pixel value at position (i,j).it having 8-bit gray scale pixel resolution. The blotches are regions of usually different homogenous gray levels. The blotches are looks like small coherent image area of pixels with almost similar gray values. Distortion of the blotches can be well modeled as burst of impulsive distortion. Assume that each pixel at x(i,j) is corrupted with probability *p* independent of whether other pixels are corrupted or not.

Considered as a noise model,

$$\boldsymbol{y(i,j) = x(i,j) \times (1 - b(i,j)) + b(i,j) \times n(i,j)} \qquad \textbf{(1)}$$

where, $y(i,j)$ is the observed intensity in the corrupted region
$x(i,j)$ is the original pixel value
$b(i,j)$ is a binary flag pixel value which is set to 1 whenever pixels are corrupted and 0 otherwise
$n(i,j)$ is the corrupted pixel intensity

If $b(i,j) = 1$,
then, $\boldsymbol{y(i,j) = x(i,j)(1 - 1) + 1 \times n(i,j) = n(i,j)}$ **(2)**

If $b(i,j) = 0$,
then $\boldsymbol{y(i,j) = x(i,j)(1 - 0) + 0 \times n(i,j) = x(i,j)}$ **(3)**

The image corrupted with blotches noise and impulse noise is now
Modeled as,

$$y(i,j) = \begin{cases} n(i,j) \text{ with } p \\ x(i,j) \text{ with } 1\text{-}p \end{cases}$$

III. PROPOSED ARCHITECTURE

The implementation of the algorithm is developed in two stages, the first stage is detection of corrupted pixels in the image, and the second stage for the replacement of only corrupted pixels with the estimated values, with the uncorrupted pixels left unaltered.

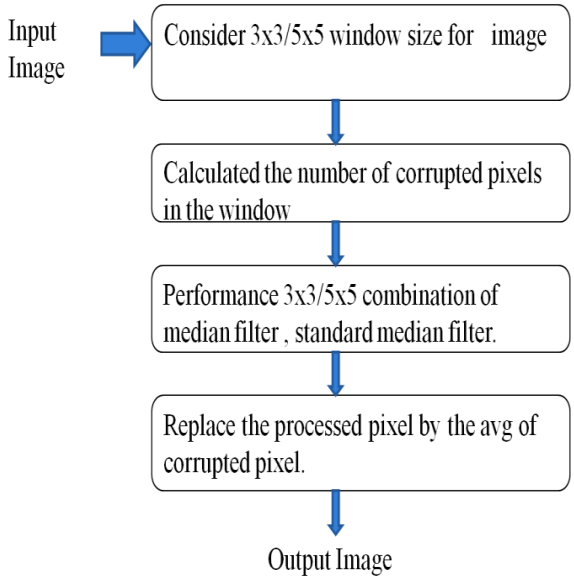


Figure 2: Flow chart of Proposed Architecture

A. Detection of corrupted pixels

The corrupted pixel takes RGB value which is substantially different than the neighboring pixels in the filtering window as applied in standard median filter. To calculate the threshold value in such a way that the edges or finer details are not modified as in standard median filters. The calculated threshold value is (*xmin* + *xmax*)/4, where *x minimum* and *x maximum* are the minimum and maximum values of the pixels in the window *Wx ij* respectively.

B. Estimation of the new pixel value for corrupted pixels

To obtained the corrupted pixels when the binary flag image *b*(*i*,*j*) is 1. The replaced pixels are estimated pixel value using median filter for lesser noise density are mean filter for higher noise density. The pixels may also other values in the intensity range for [0,255].

(1) Case 1

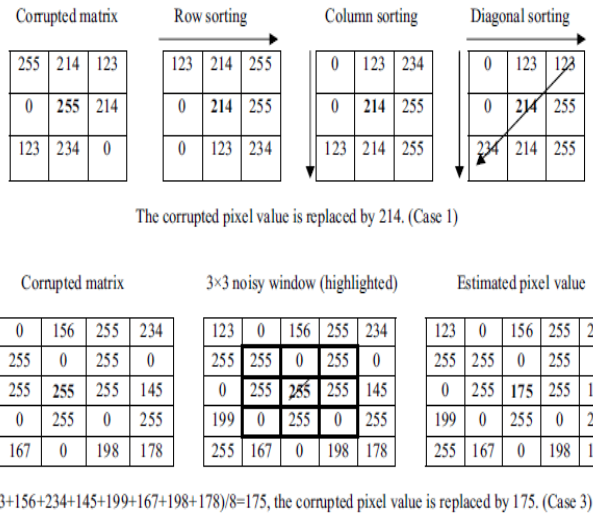
If the number of corrupted pixels " n" in the window size is less than or equal to 4, ie., n≤4, that two dimensional window of size 3X3 sorting, column sorting, and diagonal sorting. The corrupted pixel value *y*(*i*,*j*) (highlighted) is replaced by the median value .

(2) Case 2

If the number of corrupted pixels "*n*" in the window *Wx i,j* is between 5 and 12, that is, $5 \leq n \leq 12$, then 5×5 median filtering is performed and the corrupted values are replaced by the median value.

(3) Case 3

If the number of corrupted pixels "*n*" in the window *Wx i,j* is greater than 13, that is, *n* _ 13 increasing the window size leads to blurring, even with the smaller window sizes, the output may happen to be noise pixels whenever the noise is excessive. In this case, the average of uncorrupted pixels in the window is found and the corrupted value is replaced by the average value instead of median value.



IV. SIMULATION RESULTS

The different varitey of color image tested in the proposed algoritham. The performance of algoritham is evaluated quantitatively using the measures viz. peak signal to noise ratio (PSNR in db), mean square error (MSE), image enhancement factor (IEF) and computation time in seconds.

$$MSE = \frac{1}{MN}\sum_{i=1}^{M}\sum_{j=1}^{N}\{x(i,j) - y(i,j)\}^2 \quad (4)$$

$$PSNR = 20 \times log_{10}\,(255 / \sqrt{MSE}\,) \quad (5)$$

$$IEF = \frac{\sum_{i=1}^{M}\sum_{j=1}^{N}\{x(i,j) - y(i,j)\}^2}{\sum_{i=1}^{M}\sum_{j=1}^{N}\{y(i,j) - x(i,j)\}^2} \quad (6)$$

where *x*(*i*,*j*) = original image pixel value, *y*(*i*,*j*) = processed image pixel value, *n*(*i*,*j*) = noisy image pixel value. The obtained results are compared with outputs of standard median filter (SMF) with 5×5 window size, decision based algorithm (DBF). The proposed algorithm removes blotches and impulse noise simultaneously with edge preservation and reduced blurring in the output image. As seen the proposed algorithm has better subjective quality when compared with the other algorithms.

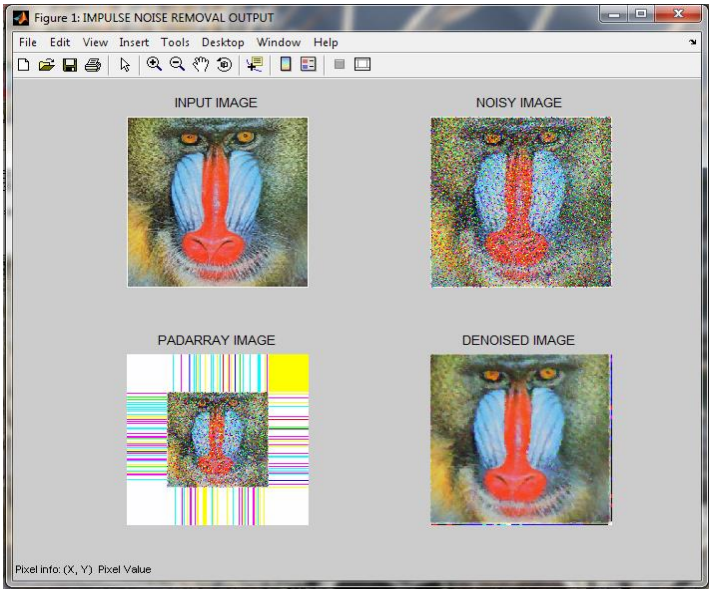


Figure 3: Output of the Color Image

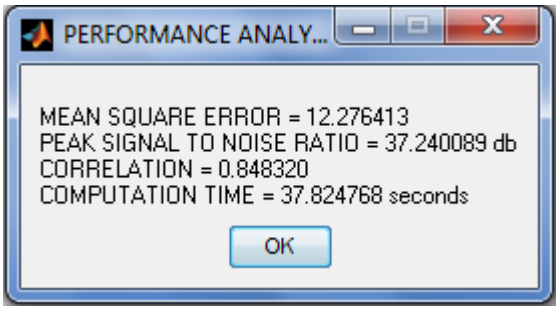


Figure 4: Output Performance Analysis

Table 1: Different Noise Densities Performance For color Image

Noise density	MSE	PSNR(db)	Correlation	Computation time in seconds
10%	12.27	37.24	0.84	37.82
20%	16.23	36.56	0.81	49.82
30%	18.55	35.21	0.79	52.08
40%	19.28	35.05	0.78	53.87
50%	20.47	35.01	0.77	54.23

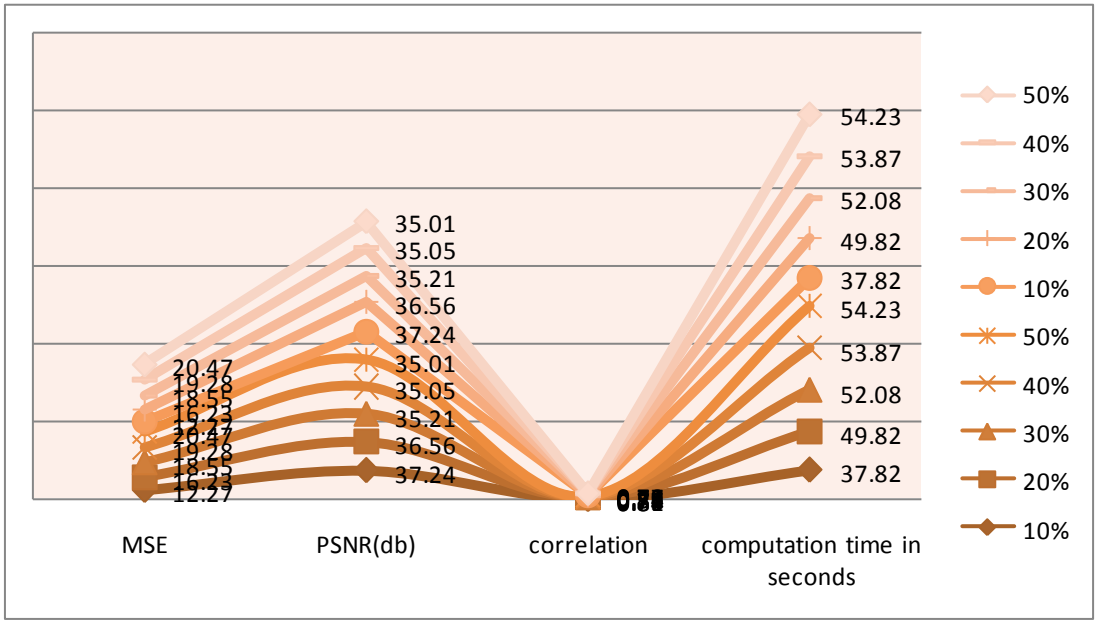


Figure 3: Noise Densities Performance for Color Image

V. CONCLUSION

The removal of blotches and impulse noise in the images is developed using non linear based algorithm. it is using two different stage algorithm, In first stage pixels are detected as corrupted or uncorrupted. The detection is based on the concept of switching threshold technique with help of binary flag. The pixels are uncorrupted mean they are left unaltered; the processing is done only if the pixels are corrupted in the second stage. The estimate of the new pixel value is calculated using median or mean filter depending on the number of corrupted pixels. The algorithm of performance is analyzed using PSNR, MSE, IEF and computation time. The obtained results are compared with other algorithms. The simulation results and subjective analysis so that the proposed algorithm gives better performance as compared to other algorithms.

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Effect of size and morphology on UV-blocking property of nano ZnO in epoxy coating

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Abstract **-** Nano zinc oxide (ZnO) is extensively used as a UV-blocking material and finds application as a UV-blocker in cosmetics, textiles and protective coatings. However, with varying sizes and morphologies in the nano scale length, the UV-blocking behavior of nano ZnO gets varied. In the present study a DGEBA based epoxy coating on MS substrate was formulated and modified with nano ZnO with two different sizes and morphologies. Flake-like and spherical nano ZnO were synthesized by chemical routes using different precursors and characterized for their structure, size and shape. The nano ZnO modified epoxy coatings were exposed in UVB weathering conditions and the UV-blocking efficiency of the two types of nano ZnO particles were studied by evaluating the color change (dE) and yellowness index (YI) for the UVB exposed epoxy coatings with and without nano ZnO. It was observed that the flake-like nano ZnO was more efficient in lowering yellowing resistance of the epoxy coating on weathering compared to the spherical nano ZnO particles.

Index Terms - Nano ZnO, UV-blocking, yellowing, epoxy

I. INTRODUCTION

In recent years, zinc oxide (ZnO) ultrafine particles are well known as UV blocking materials, which make it feasible to be widely used in polymers, fabrics and cosmetic materials [1]. Zinc oxide has a band-gap at around 3.37 eV corresponding to 376 nm and thus it absorbs light that matches or exceeds this band gap energy [1-2]. UV-range of solar spectrum lies within this range and thus UV light gets absorbed by zinc oxide particles. The UV-blocking property of zinc oxide is only enhanced when it comes to nano-sized ultra-fine ZnO particles. As they are inorganic and particulate, they have added advantages of being stable and non-migratory within a matrix and thus potentially impart better effectiveness and a longer service life.

With the on-going development in nano-technology, different nano-structures can be synthesized via different synthesis routes. An equally known fact is that the unique and fundamental properties of nano particles change with their size and shape [3-5]. The phase stability of nano particles depends significantly on its surface free energies and surface stress. Nano-particles either can exist as isolated species or they can form aggregates or they can dissolve into ions in solutions. These processes result in different sizes and sizes regimes and with this variance in size, the properties also vary in nano scale length [3]. The electronic properties for semiconductor nano-particles, localized surface plasmon resonance (LSPR) of noble nanoparticles are cited to be size dependent. LSPR for Ag nanoparticles shifts across the electromagnetic spectrum as the shape of the particle changes from a sphere to cylinder to cube to prism to pyramid [3]. Similarly, the optical characteristics of zinc oxide materials have been found to depend on their micro structures, morphologies and particle size [6]. The UV-blocking property of nano ZnO also thus depends on its size and morphology i.e, nano ZnO particles with different sizes and morphologies would have varied UV-blocking properties.

In the present study nano ZnO particles with two different sizes and morphologies were prepared via chemical route using different precursors. The synthesized nano particles were then incorporated in to a di-glycidyl bisphenol A (DGEBA) epoxy coating based on micron-sized titanium dioxide (TiO_2) and then subjected to UVB weathering conditions for two cycles and the effect of synthesized nano ZnO particles in lowering yellowing of epoxy on weathering was evaluated. It was found that the nano ZnO particles with flake-like morphology compared to the spherical nano ZnO were more efficient in imparting yellowing resistance to the formulated epoxy coating.

II. EXPERIMENTAL

Synthesis of nano ZnO particles

1. Flake-like nano ZnO

Materials
Zinc sulfate heptahydrate ($ZnSO_4 \cdot 7H_2O$), ammonia solution (NH_4OH), ammonium bicarbonate (NH_4HCO_3), anhydrous alcohol and deionized water

Methodology
The first stage involved synthesis of the precursor, zinc carbonate hydroxide (ZCH) using a 1:10:1 (ratio by volume) of aqueous ammonia (7 mol/L): $ZnSO_4 \cdot 7H_2O$ solution (0.5 mol/L): NH_4HCO_3 (3 mol/L). Aqueous ammonia followed by NH_4HCO_3 solution, both, were drop-wise added to a continuously stirred solution of ZnSO4·7H2O at room temperature. The reaction mixture was then heated up to 60^0C and stirred at this temperature for 30 minutes. The ZCH precipitate was then filtered and washed with deionized water to ensure complete removal of sulphate ions ($SO4^{2-}$). The ZCH precursor was air dried and calcined at 400^0C for two hours to obtain nano sized zinc oxide particles.

2. Spherical nano ZnO

Materials
Zinc acetate dihydrate, methanol, KOH

Methodology
The following two solutions of appropriate concentrations were prepared which were mixed at room temperature with continuous stirring:
Solution A - Zinc acetate dehydrate in methanol
Solution B - KOH in methanol
A bulky white precipitate arose immediately which became translucent after stirring for approximately 70 minutes. The solution was then heated to boiling point, after which the heat source was switched off. The solution was then allowed to stand over-night. The sediment thus obtained was stirred and centrifuged off. The white gel like product obtained was washed with de-ionized water and then dried to obtain the nano-sized zinc oxide particles.

The synthesized nano-powders were investigated by Transmission Electron Microscopy (TEM, Philips CM200 electron microscope), Scanning Electron Microscopy (SEM, Model no.S3400, Hitachi) and X-ray diffraction (XRD, X'Pert Pro Philip) for its shape, size and crystallinity. FTIR (JASCO FTIR 6100) studies of nano-powders were carried out to study the structure of the oxides.

Preparation of epoxy coated samples

Materials
Di-glycidyl ether of bis-phenol A epoxy (EEW=185), Diethylene-tetramine (DETA), xylene, titanium dioxide (TiO_2), BYK additives

The mild steel (MS) samples to be coated were de-greased, cleaned and then roughened mechanically with abrasive paper (emery paper grade no.: 100). A TiO_2 based white coating was formulated with 35% pigment concentration, 50% epoxy resin concentration, ssssDETA as hardener and rest with the appropriate concentrations of additives and xylene as solvent. The nano ZnO modified epoxy coatings were prepared by addition of firstly the micron sized TiO_2 pigment along with the required amount of nano ZnO at different loading levels, maintaining the 35% pigment concentration in the coating followed by the additives. Nano zinc oxide was added to the epoxy resin using ultra-sonnication at 1%, 2% and 5% (by weight) of the total coating formulation. The only purpose to take TiO_2 as the pigment was to formulate a white coating

and hence evaluate discoloration/ yellowness of the coating on weathering. The pigment concentration was always maintained at 35% with and without nano ZnO while formulating the coatings. The coating formulation for different loading of nano ZnO in the coating is as tabulated as follows:

Table.1. Epoxy coating formulation with and without nano ZnO

Component	**Neat Epoxy (%)**	**1% nano ZnO + Epoxy (%)**	**2% nano ZnO + Epoxy (%)**	**5% nano ZnO + Epoxy (%)**
TiO_2 pigment	35.0	34	33	30
Nano ZnO	0	1	2	5
Resin	50.0	50.0	50.0	50.0
BYK 530	2.0	2.0	2.0	2.0
BYK 333	1.0	1.0	1.0	1.0
BYK 9076	1.5	1.5	1.5	1.5
BYK 320	0.5	0.5	0.5	0.5
Xylene	10	10	10	10

As mentioned in the table above, the resin content was kept constant throughout in the coatings, with different nano ZnO concentrations. The resulting blend was ultrasonnicated for 20 minutes and the shear time via ultrasonication was kept constant for all loadings of nano ZnO. After mixing TiO_2 pigment, nano ZnO and additives, at the end, hardener and solvent were mixed to obtain the nano ZnO modified epoxy coatings. The formulated coating was applied using brush on the surface treated MS panels so as to achieve uniformly coated panels with good finish and then allowed to hard cure.

Weathering Test

The coated MS panels were subjected to accelerated weathering in a UV weatherometer (QUV Weatherometer, Q-Lab Products & Services) equipped with UVB-313 nm lamps. The test cycle in UVB-weatherometer comprised of 4 hours UVB simulation at 60^0C followed by 4 h of condensation (UVB-lights off during condensation) at 50^0C as per accordance of ASTM G-154. The epoxy coated MS panels were exposed to two test cycles, i.e, 16 hours in the weatherometer.
The coatings exposed in UVB weatherometer were characterized for color change (dE) and yellowness index (YI) using a spectrometer (BYK-Gardener Spectrometer) equipped with Color-Lab Quality Control software. The data were reported on L, a, b scales and overall color difference was given in the following equation:

$$\mathbf{dE = [\Delta L^2 + \Delta a^2 + \Delta b^2]^{1/2}}$$

where; $\Delta L = L_2 - L_1$, $\Delta a = a_2 - a_1$, and $\Delta b = b_2 - b_1$

Respective numbers 1 and 2 are denoted to samples before and after the exposure test.
The yellowness index was measured as per ASTM D 1925, formulated as:

$$\mathbf{YI = 100\ (1.28\ X_{CIE} - 1.06\ Z_{CIE}) / Y_{CIE}}$$

As per ASTM D 1925, the conditions for measurement are as follows: Illuminant: C, Standard Observer function: 2^0, denoted as $C/2^0$. X, Y and Z are the tristimulus values and 1.28 and 1.06 are the coefficients as per $C/2^0$ conditions.
The surface morphology of the coatings was studied using SEM microscopy (SEM, Model no.S3400, Hitachi) to look for the distribution of nano zinc oxide at different loading levels in the coatings.

III. Results & Discussion

Characterization of nano ZnO

1. Flake-like nano ZnO

The synthesis methodology involving zinc sulfate heptahydrate ($ZnSO_4 \cdot 7H_2O$), ammonia solution (NH_4OH) and ammonium bicarbonate (NH_4HCO_3) as starting materials, resulted in formation of zinc carbonated hydroxide (ZCH) which on calcination gave nano ZnO. The purity of nano ZnO was checked by FTIR and XRD analysis and their shape and size were established by SEM, TEM and HRTEM analysis. Zinc oxide shows typical characteristic Zn-O peak at around 470 and 480 cm^{-1} [7-9]. This was confirmed by FTIR analysis as seen in figure.1 which showed Zn-O peaks at 468 and 480 cm-1.

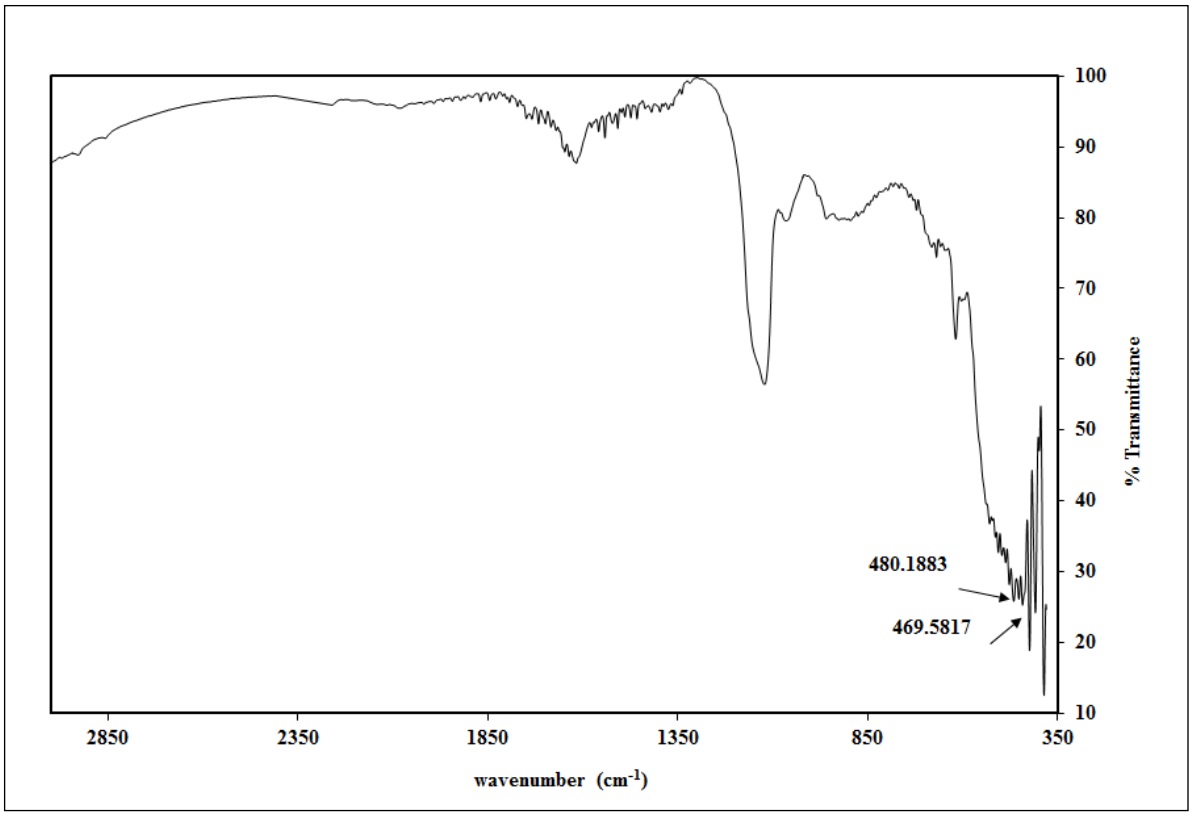


Figure.1. FTIR spectrum of flake-like nano zinc oxide

Figure.2 shows the XRD spectra of the synthesized nano zinc oxide. The XRD analysis identified well indexed diffraction peaks, in good agreement with those of a wurtzite hexagonal structure of ZnO (ICSD Reference code 01-075-0576) [10-12]. The flake-like nano ZnO particles showed the characteristic peaks at 320, 340 and 36 resulting from the 100, 002 and 101 planes respectively.

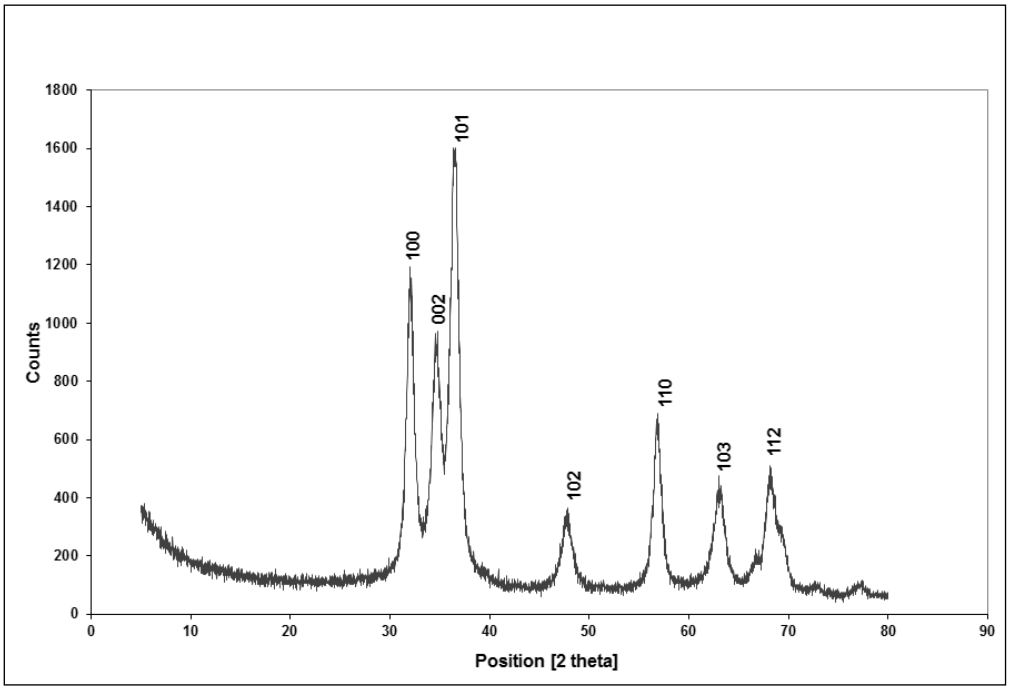


Figure.2. XRD spectrum of flake-like nano ZnO

The SEM micrograph (figure.3), showed the nano ZnO particles to have flake-like morphology.

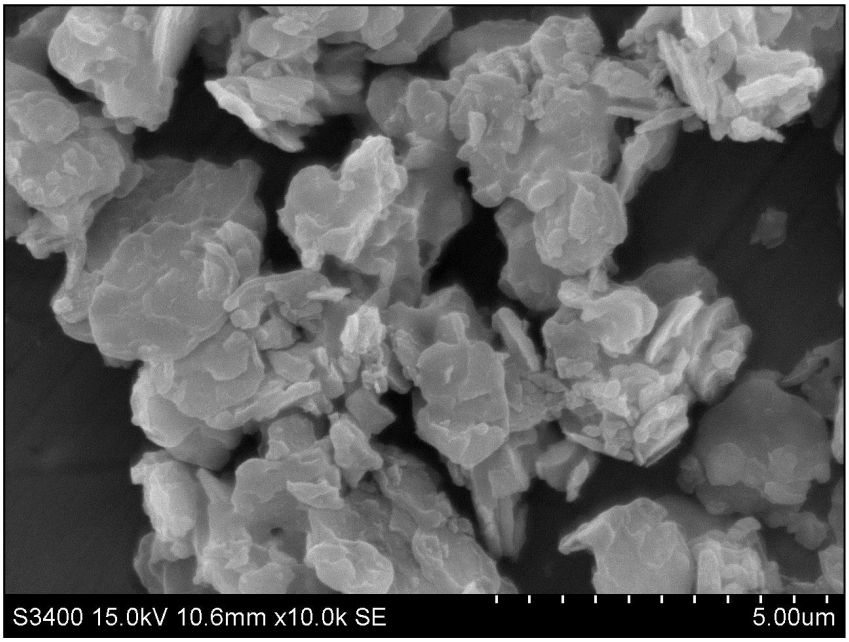
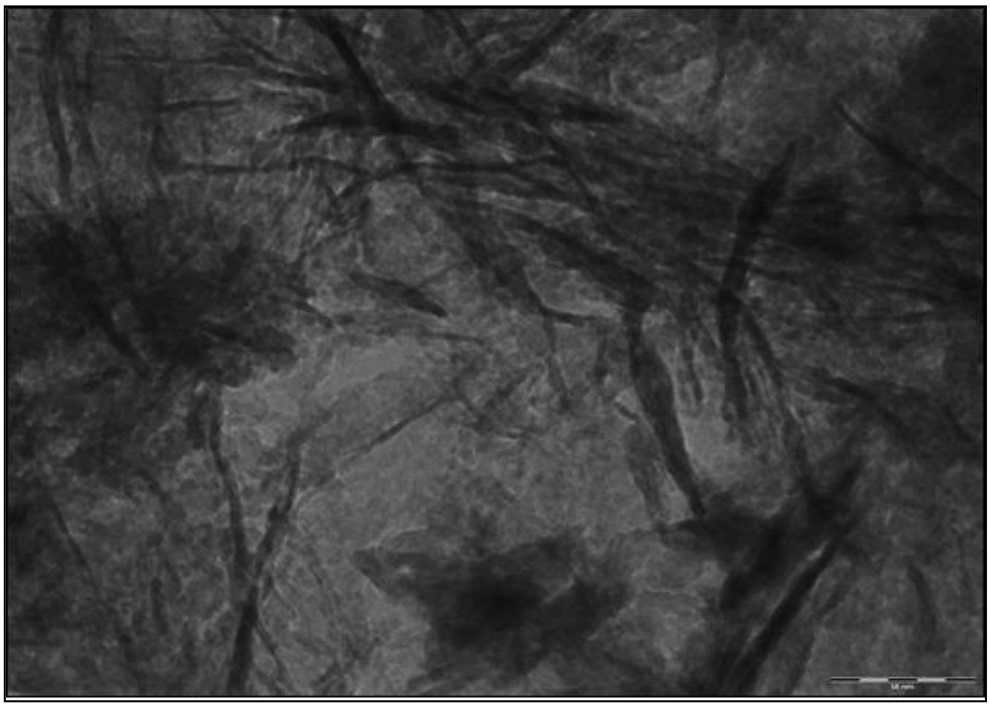


Figure.3. SEM micrograph of flake-like nano ZnO

The flake-like morphology was confirmed by TEM.As shown in figure.4, the flake-like morphology of ZnO composed of small spherical particles with diameter ranging from 7 to 20 nm.



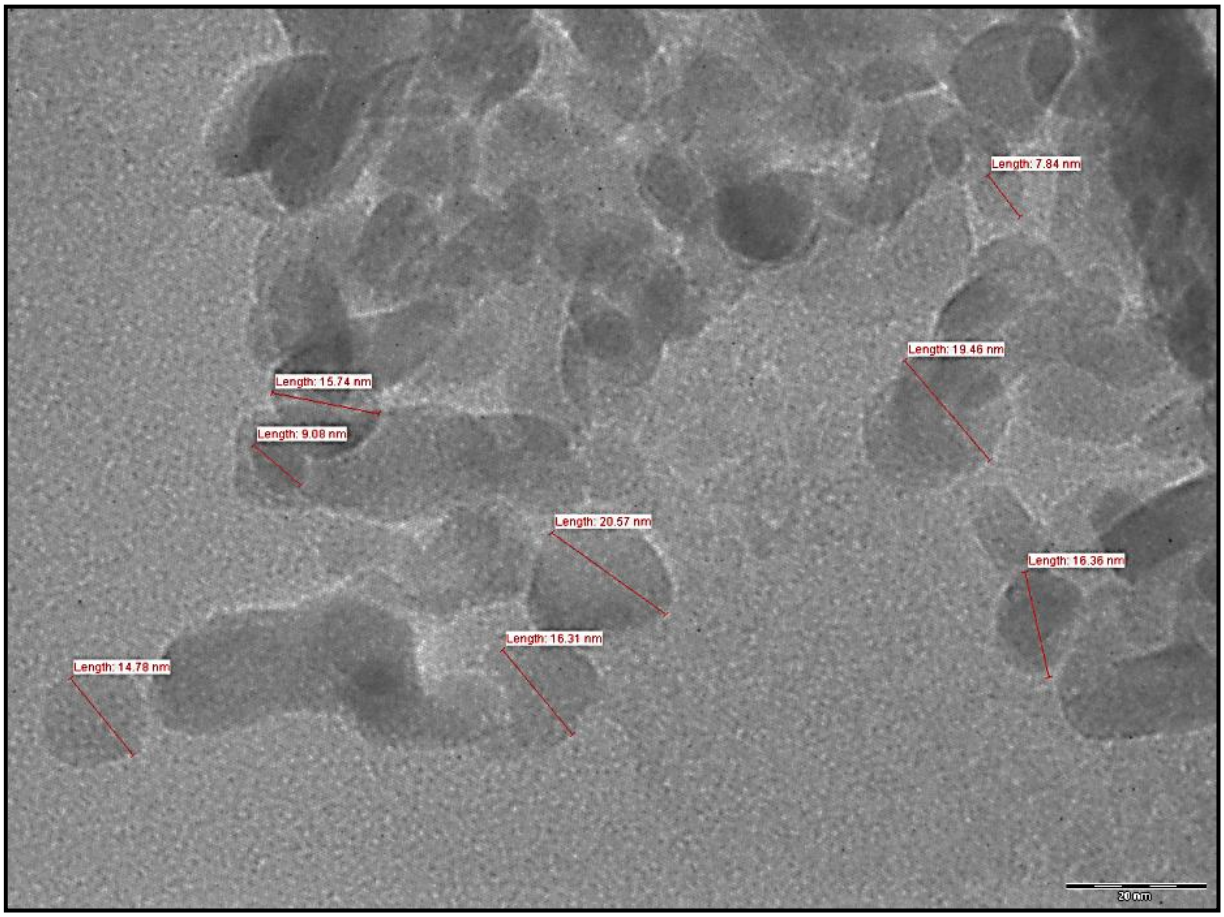


Figure.**4.** TEM image of flake-like nano ZnO

The lattice fringes of synthesized zinc oxide were clearly identified in the HR-TEM image as shown in figure.5. The inter-planar spacing i.e. d-spacing was obtained as 0.29 nm which closely matched to the d-spacing of 100 plane (0.281nm) of zinc oxide [13-15].

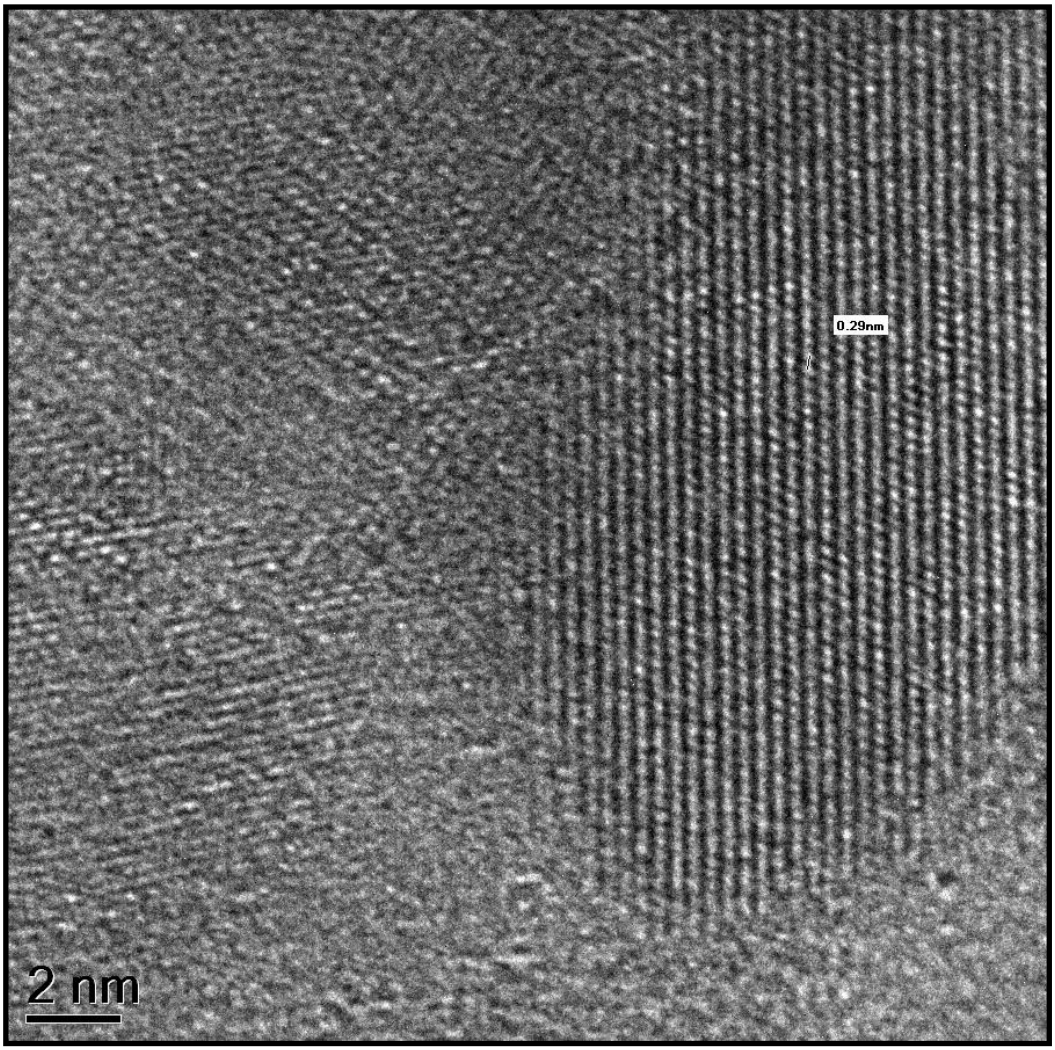


Figure.5. HR-TEM image of flake-like nano-ZnO

2. Spherical nano ZnO

Nano ZnO was found to have spherical particle shape when prepared by chemical route using zinc acetate dehydrate and methanol as starting materials. The FTI spectrum with characteristic Zn-O peaks at 468 cm^{-1} and 476 cm^{-1} is as shown in figure.4.6. [i]The XRD spectrum of the nano ZnO particles is shown in figure.7. The XRD pattern observed was assigned to the pure phase of zinc oxide (Joint Committee on Powder Diffraction Standards (JCPDS card number: 01-079-0205) which confirmed a wurtzite hexagonal structure.

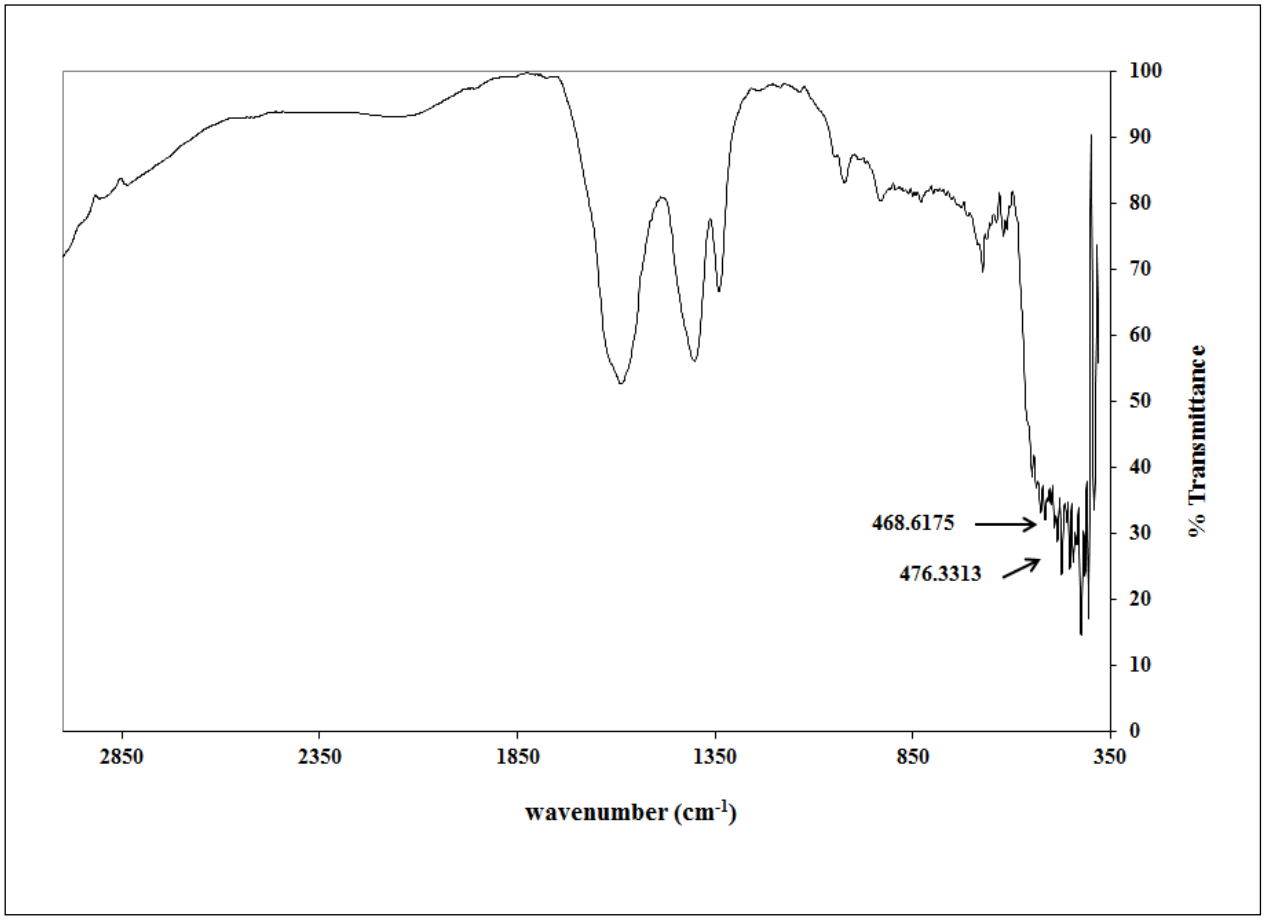


Figure.6. FTIR spectrum of spherical nano ZnO

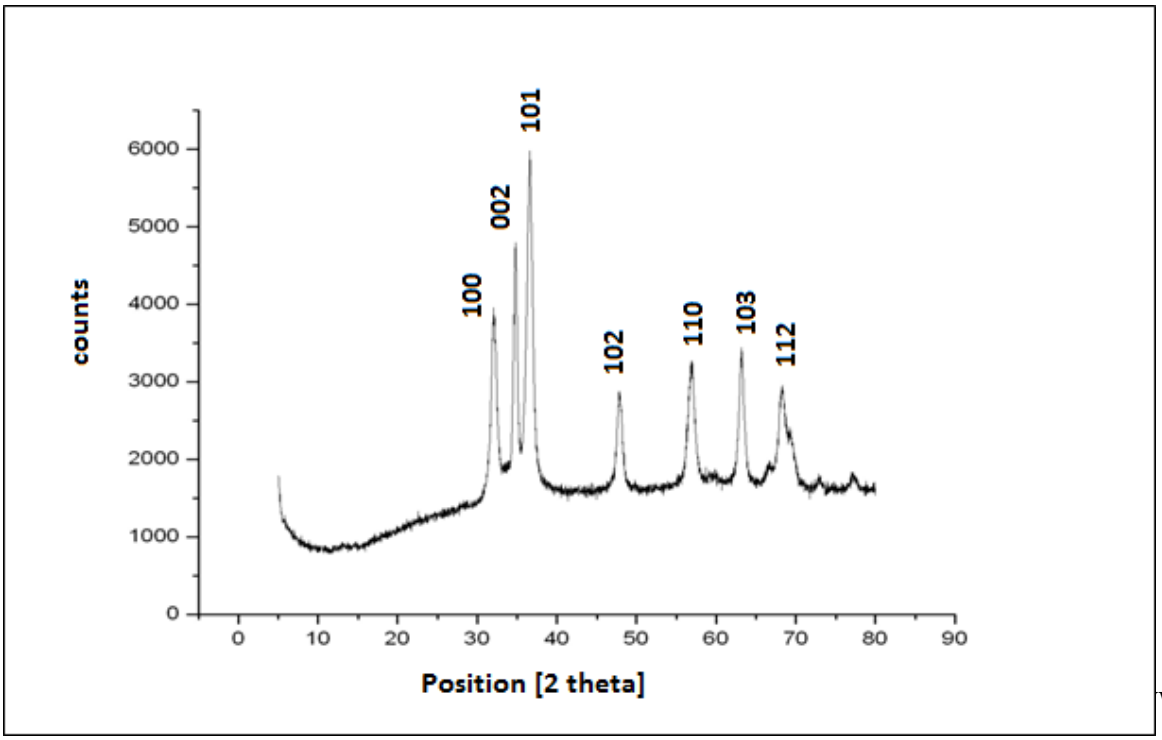


Figure.7. XRD spectrum of spherical nano ZnO

The SEM micrograph of the nano-ZnO particles showed single phase primary particles with spherical shape. TEM monograph showed that the product consisted of particles with diameter ranging from 20-25 nm.

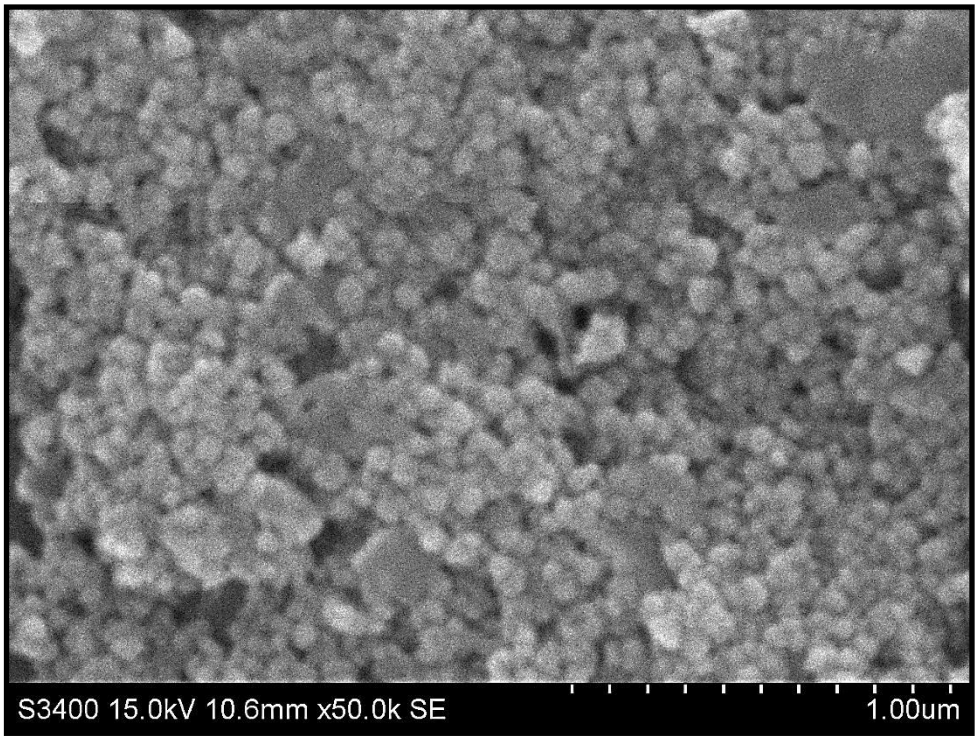


Figure.8. SEM micrograph of spherical nano ZnO

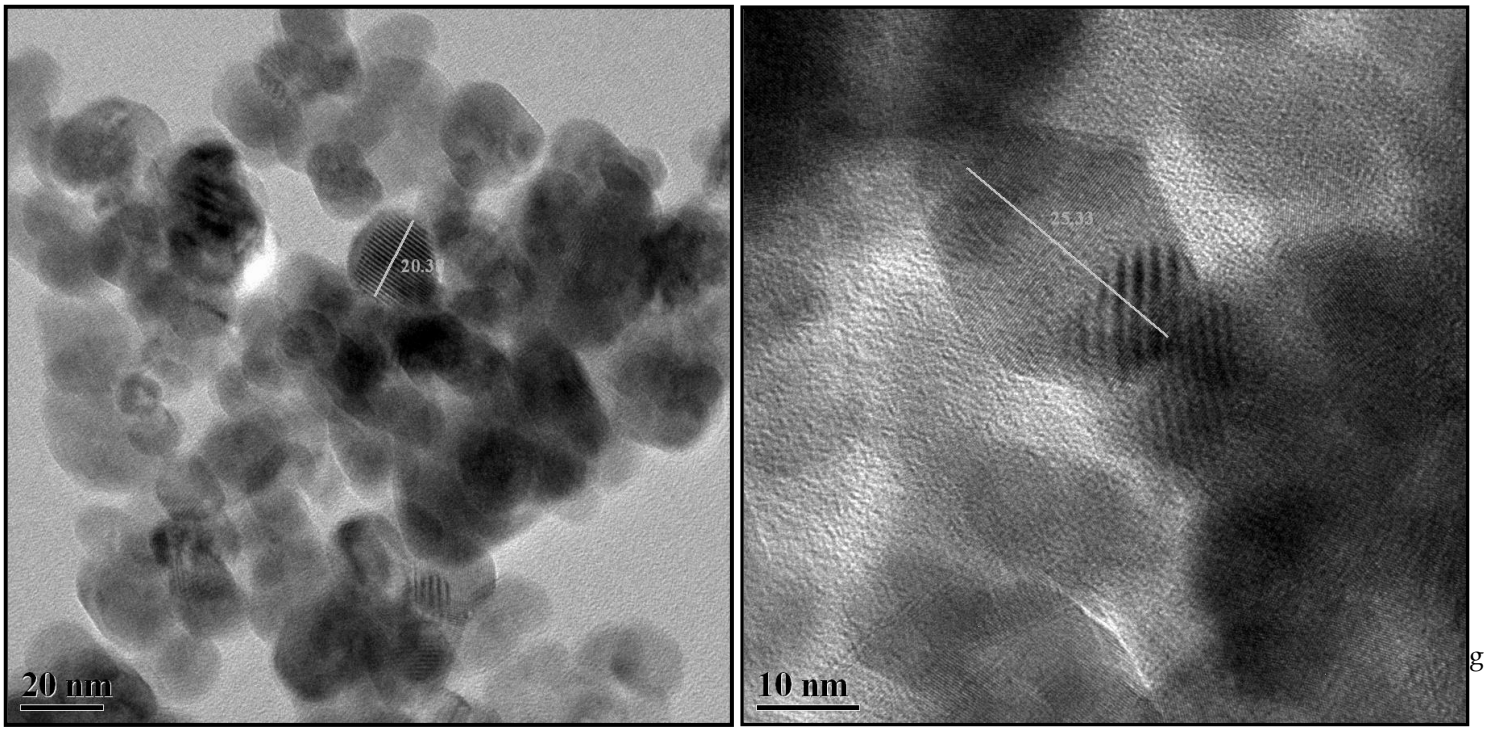


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Figure.9. TEM micrograph of spherical nano ZnO

Weathering Study of nano ZnO modified Epoxy Coatings

The effect of nano-ZnO addition to epoxy was studied by formulating epoxy coating with different concentrations of nano ZnO. A TiO_2 based coating with DGEBA epoxy resin and DETA hardener, as mentioned in the previous chapter, was formulated with different concentrations of nano ZnO particles to optimize the concentration of nano-ZnO in the coating. Nano ZnO, both flake-like and spherical shaped, was added at 1%, 2% and 5% by weight of total coating formulation. The pigment concentration was maintained as 35% during the loading of nano ZnO particles in the coating.

Evaluation of color change and yellowing of weathered nano ZnO modified epoxy coatings

The neat epoxy and both flake-like and spherical shaped nano ZnO modified epoxy coatings formulated as per the above tabulated values were exposed to accelerated UVB exposure for two cycles (16 hours). The color change (dE) and yellowness index (YI) values obtained by spectrophotometric analysis are as shown in figure.10-11

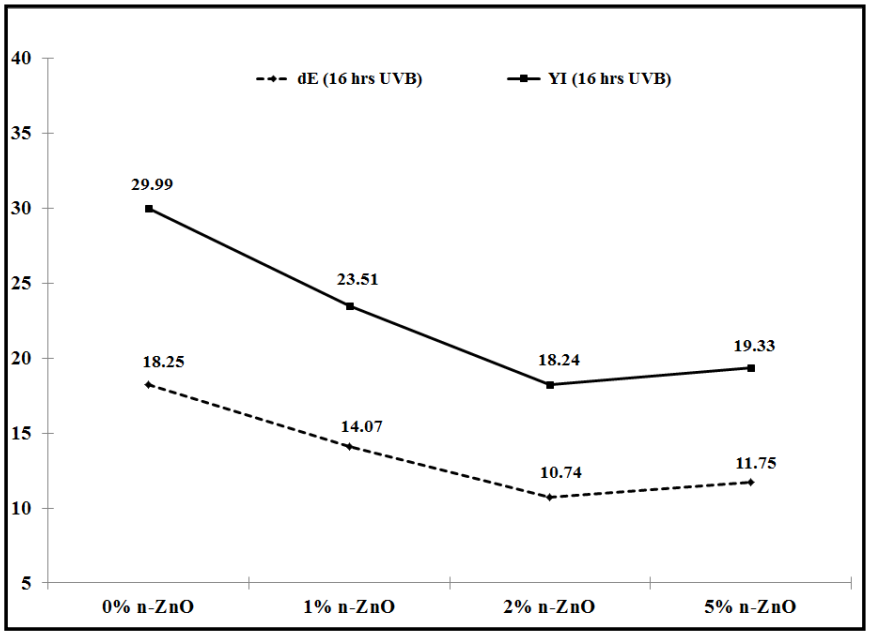

Figure.10. dE and YI plot for flake-like nano ZnO + epoxy coating

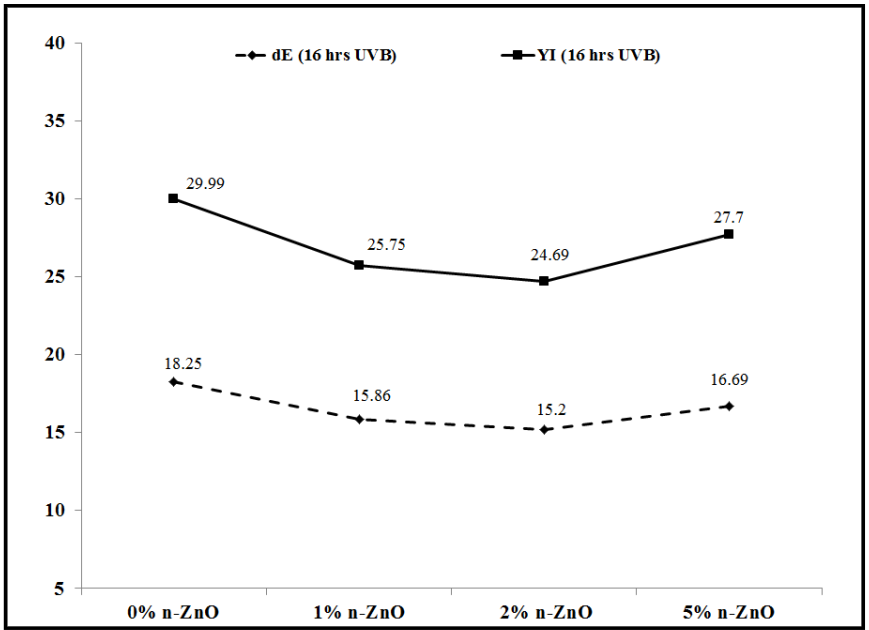

Figure.11. dE and YI plot for spherical nano ZnO + epoxy coating

The epoxy coatings formulated with and without nano-ZnO, were applied on 0.7 mm thick MS panels by brush. The coating thickness was determined using a thickness measuring gauge (Elcometer DFT Gauge) and the coating thickness for the coating systems were found to be in the range of 85-100 microns. The color change (dE) and yellowness index (YI) results were evaluated after 16 hours of UVB exposure. i.e, two cycles of UV (B) weathering were each cycle of 8 hours comprises of 4 hour UVB light followed by 4 hour condensation. The discoloration observed in epoxies on weathering is attributed to the formation of a quinone- methide structure and has been confirmed in many cited studies by both infrared and UV-Vis analysis [16].

As observed from the dE and YI plots, both flake-like and spherical shaped nano ZnO exhibited lower dE and YI values at all loading levels compared to the neat epoxy coating system. The optimized loading concentration of both flake-like and spherical nano ZnO was observed to be 2 wt% of the total coating formulation. Flake-like nano ZnO lowered yellowing and color change by 12 and 8 units respectively while the spherical nano ZnO lowered the same by only 5 and 3 units respectively. The increase in dE and YI values for both flake-like and spherical nano ZnO beyond the optimized 2 wt% loading is attributed to the agglomeration of nano particles in the coating matrix. Agglomeration of nano-particles is un-desirable as it makes nano-particles lose their high surface area and thus the desired functional properties [17]. The result of agglomeration is that the nano ZnO would not be evenly distributed and so there will be regions in the coating that would be starved of zinc oxide and not protected, by zinc oxide in the coating.

The extent of weathering resistance, offered by the nano ZnO modified epoxy coating systems were determined by evaluating the percentage reductions in color change (dE) and yellowing (YI) during the exposure. The percentage reduction values were calculated as follows:

Suppose after a test cycle,

Neat Epoxy = A

Modified Epoxy = B

A and B are dE/YI values after the weathering cycle, then;

% Reduction = [(A-B) / A] x 100

The percentage reduction values in dE and YI, thus calculated for the nano ZnO modified epoxy coatings are as shown in figure.12 – 13.

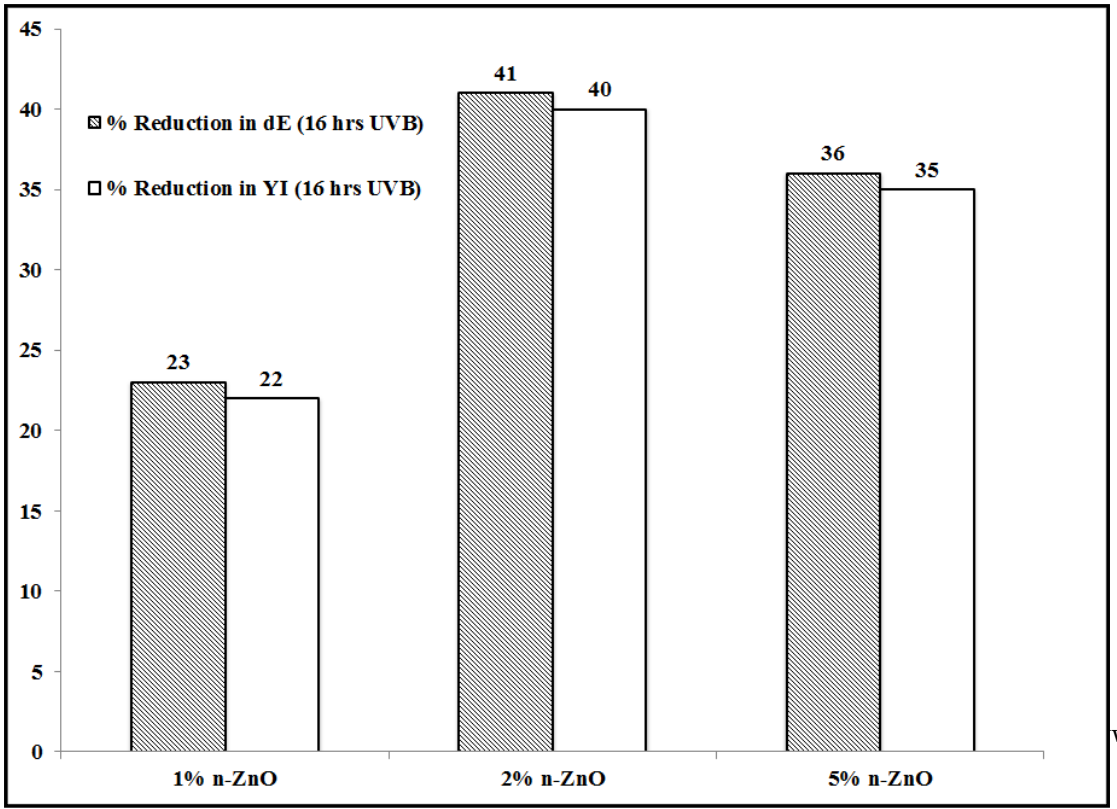

Figure.12. % Reduction in dE and YI plot for flake-like nano ZnO + epoxy coating

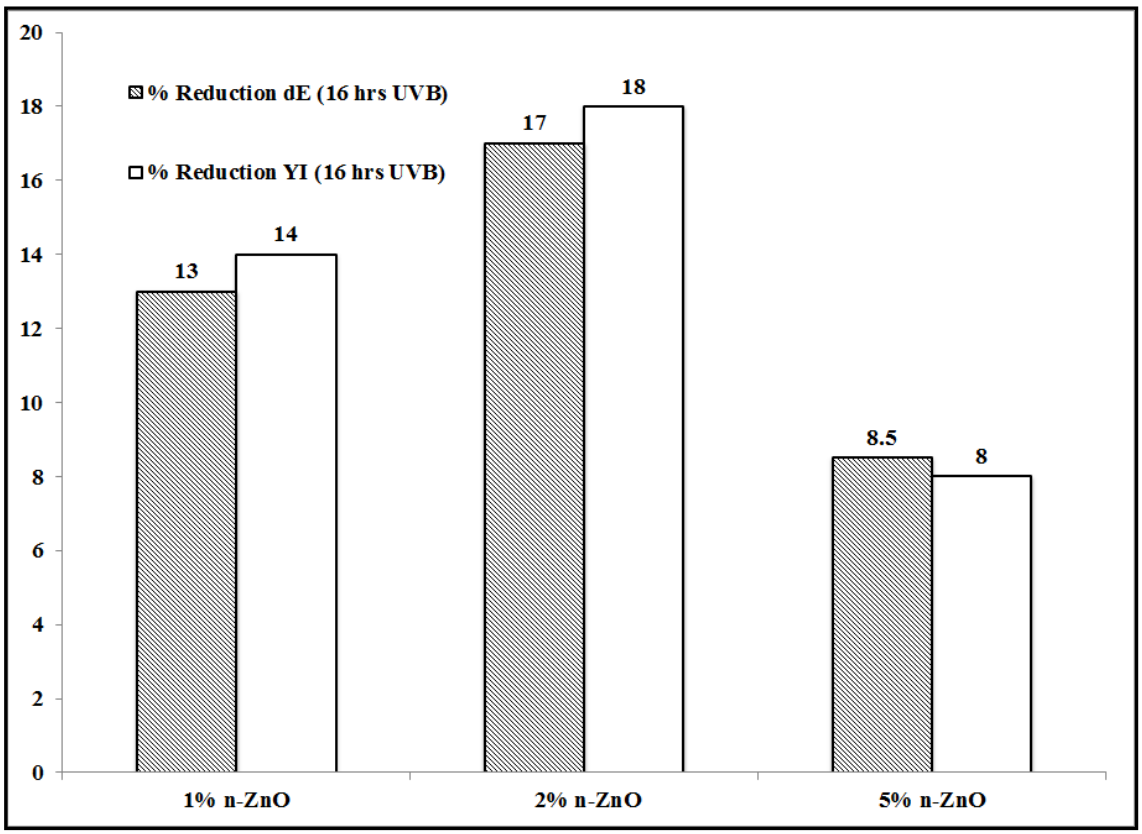

Figure.4.13. % Reduction in dE and YI plot for spherical nano ZnO + epoxy

Table.2. dE, YI, % Reduction in dE and YI for nano ZnO modified epoxy coatings

Systems	**Flake-like nano ZnO + Epoxy**				**Spherical nano ZnO + Epoxy**			
	dE	**YI**	**% Reduction dE**	**% Reduction YI**	**dE**	**YI**	**% Reduction dE**	**% Reduction YI**
0% nano ZnO +Epoxy	18.25	29.99			18.25	29.99		
1% nano ZnO + Epoxy	14.07	23.51	23	22	15.86	25.75	13	14

2% nano ZnO + Epoxy	**10.74**	**18.24**	**41**	**40**	**15.2**	**24.69**	**17**	**18**
5% nano ZnO + Epoxy	11.75	19.33	36	35	16.69	27.7	8.5	8

High percentage reduction values in dE and YI indicates more resistance to color change and yellowing. The highest percentage reduction values were observed at 2% loading level respectively for flake-like and spherical nano ZnO modified epoxy coatings compared to 1% and 5%. Flake-like nano ZnO reduced color change and yellowing by 40% while spherical nano ZnO reduced the same by 17%, at 2% loading in the coating formulation.

The enhanced weathering resistance for the nano-ZnO incorporated epoxy coatings is due to UV-absorption property of ZnO owing to its large band gap energy. Zinc oxide has a band-gap at around 3.37 eV corresponding to 376 nm and thus it absorbs light that matches or exceeds this band gap energy [1, 2]. UV-range of solar spectrum lies within this range and thus UV light gets absorbed by zinc oxide particles. The UV-blocking property of zinc oxide is only enhanced when it comes to nano-sized ultra-fine ZnO particles.

Though, both flake-like and spherical nano ZnO offered resistance to color change and yellowing in the nano ZnO modified epoxy coatings, the flake-like nano ZnO outperformed spherical nano ZnO. Nano ZnO with flake-like morphology acted as a better UV-blocker than the spherical nano-ZnO and hence imparted more efficient weathering resistance when incorporated in the epoxy coatings. It was thus inferred that nano ZnO with different shape and size performed differently when incorporated in epoxy coatings and subjected to UVB exposures. The better performance of the flake-like nano-ZnO is attributed to its morphology and size. Nano particles possess unique properties owing to their nano size dimensions however these properties change with their shape and size and the causes are different for different materials [3, 4]. Roshidah Rusdi et al. (2011) have reported in their work that ZnO nano-tubes have a wider band gap compared to that of spherical nano ZnO particles and have explained this by the characteristics of materials at nano dimensions [6]. At nano scale, increase in pressure results in strong forces in the interiors of crystallites and hence the lattice spacing of ZnO nanotube decreases. It was proposed then, that due to the strong internal forces, the electrons taking part in transition from the valence band to the conduction band need greater energy, accounting for the observed wider band gap [3, 4]. All these changes eventually change the properties of the nano particles. It was also seen that the absorption profile for nano ZnO with nanotube and nano rod structures were a little different from that of the spherical particles. Also, from thermodynamic considerations, the total free energy is a sum of free energy of the bulk and the surface of nano particle.

$$G_{nanoparticle} = G_{bulk} + G_{surface}$$

For nanoparticles, $G_{surface}$ is no longer a minor component and hence all fundamental properties of nanoparticles are size dependent as well as shape dependent [5].

The flake like nano ZnO particles diameter ranged from 7 – 20 nm while the spherical shaped nano particles had diameter size of 20 – 25 nm. The different behavior of flake-like and spherical nano ZnO in imparting weathering resistance to epoxy coating can be thus explained on the basis of size and morphology difference of the nano ZnO particles. As the nano size decreases, the space in which charge carriers move decreases and additional quantum confinement is imposed on their motion. This leads to increase in the band gap energy, electron-hole kinetic energy and the density of the charge carriers within and at the nanoparticle surface. The enhanced reduction in color change and yellowing for the flake-like nano ZnO compared to spherical nano ZnO in the epoxy coating at all loadings in the epoxy coating is thus attributed to the difference in their size, shape and morphology.

IV. CONCLUSION

Nano zinc oxide (ZnO) particles were synthesized with two different morphologies namely, flake-like and spherical with different sizes. Flake-like nano ZnO had diameter ranging from 7-20nm while spherical nano ZnO had diameter in the

range of 20-25nm. Both flake-like and spherical nano ZnO when incorporated in a DGEBA based epoxy coating lowered yellowing of the epoxy coating on weathering. The optimized loading concentration was 2 wt% of the total coating formulation. The efficiency of the nano ZnO particles were found to be size and morphology dependent as the flake-like nano ZnO lowered yellowing and color change by 40% while the spherical nano ZnO particles lowered the same by 18%.

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A Floristic Study of Koria District (Chhattisgarh) India

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Abstract- The State of Chhattisgarh has about 44% of its geographical Area Covered with forests. The Koria district in Chhattisgarh lies between $22^0$58' to $23^0$49' North latitudes and 81^0 33' to $82^0$45' East longitude. The average rainfall is 121.36 cm. The forest area is 81.23% of Total dist. area. The annual mean temperature is 24^0C. The temperature varies between 16.2^0C to 31^0C. Geologically the area is dominated by upper Gondwana rocks. Which are rich in coal deposit. The highest mountain ranges of the region occupy the northern part of the district.

Koria District of Chhattisgarh state has a very rich flora exhibiting diversity of flowering plants. There is no comprehensive description of the flora as well as vascular cryptogams of the district. Some plant species are on the verge of extinction . Keeping these points in view, the present investigation was planned. Present paper deals with floristic account of flowering plants and provides a preliminary analysis of the flowering plants of Koria district.

The district harbours 657 species under 439 genera and 119 families of Seed plants (Angiosperms). Of these, 364 species are wild and 293 species are cultivated/planted. D:M (Dicot:Monocot) ratio was found to be 6.00 family-wise, 4.85 genera-wise and 5.19 species-wise. Generic coefficient was found to be 66.81. The Genera:Species ratio was 1:1.496.

Index Terms- Chhattisgarh , Floristic study , Generic coefficient Koria district

I. INTRODUCTION

India contains about 8% of world's biodiversity. Chhattisgarh, the 26th state of the country, has ample variation in physical and cultural features. It has about 44% of its total geographical area covered with forests. It enjoys hot and humid climate and gains rainfall from both north-east and south-west monsoon. It has about 30 small and big drainage systems. These features have an important contribution to its biological wealth.

Koria district in Chhattisgarh lies between 22o58' and 23o51' North Latitude and 81o59' and 82o45' East Longitude and has a forest area of 81.23%. Average rainfall is 121.36 cm. and annual mean temperature is 24oC. The district is dominated by Upper Gondwana rocks which are rich in deposition of coal.the highest hill ranges occupy the northern part of Deogarh, Sonhat and Manendragarh.

The district has very rich in plant diversity, including medicinal plants. The flora of the district is not very comprehensively described. therefore There is an urgent need for the systematic enumeration, authentic identification and documentation of the flora of the district. Present paper deal with the floristic diversity of the Koria district in Chhattisgarh.

II. RESULT AND DISCUSSION

Result in the Table -1 indicate the florisitc analysis of the Koria district. It shows a total of 119 families belonging to 102 dicot and 17 monocot angiosperms, Seven Gymnosperms and Nine Pteridophytes were also recorded. Ten most dominating families were Poaceae, Papilionaceae (Fabaceae), Asteraceae, Acanthaceae, Euphorbiaceae, Rubiaceae, Apocynaceae, Lamiaceae, Anacardiaceae, and Malvaceae.

The Dicot / Monocot ratio was 6.00 (family wise), 4.85 (genera wise) and 5.19 (species wise) The generic coefficient was found to be 66.81. The Genera : species ratio was 1:1.496. which is less as compared to previous report of Chipde (1980) who has reported 1:1.623 in case of flora of Bilaspur (Chhattisgarh).

Thakur *et.al.* (2009) have reported Generic coefficient as 86.3% in forest vegetation of Sagar district in M.P. They have reported 31 dicot and 1 monocot families distributed in 63 genera and 73 species of trees.

Suresh *et.al.* (2008) have enumerated 67 species from disturbed area of Thaniparai hills and 72 species from undisturbed area of Sundaramahalingan hills under Grizzled Giant squirrel wildlife sanctuary forest of Virudhunagar district.They have reported 51 families from the two sites with 125 genera and 139 species.

T.S. Nayar. *et.al.* 2008 has also provided a preliminary analysis of flowering plants of Kerala based as 1303 publication appeared until 2008. They have stated that the state harbours 4694 species under 1418 genera and 188 families.

Silar Mohammad *et. al.* (2008) have reported florisitc diversity of Ahobilam forest in Andhara Pradesh.. They have collected a total of 250 wild and naturalized important plant species belonging to 71 families. These include many medicinal rare endemic and threatened categories of plants.

Awasthi *et. al.* (2007) have also reported floristic diversity of Bandhavgarh national park, enumerating 47 plant species.

Inamati *et. al.(*2007) have reported 43 families represented by 130 spp. across four altitudinal zones in Devimane, (Western Ghats) Karnataka.

Ganeshaiah *et.al.* (2002) described a total of 178 families composed of 1408 genera and 4758 species in Karnataka. They have taken into consideration the major climatic zones of the state.

Uma Shankar (2001) described the floristic composition of tropical deciduous *Sal* forest in Darjeeling.

III. MATERIALS AND METHODS

The study was carried out during 2003 to 2007. The study encompasses an area of 20 km to 200 km radius around the Head Quarter town of the Koria district. The district comprises of 5 development blocks viz. :- Baikunthpur, Sonhat, Manendragarh Khadgawan, Bharatpur. The sampling sites were selected radomly. Some of them to quite are Shivpur, Katghodi, Pahadpara, Itga, Rakiya, Deori,Orgai, Amhar, Paradol, Sonari, Narayanpur, Kothari, Khongapani, Dubchhola, Boridand, Podi, Jilda, Bhagwanpur, Patwahi and, Semaria.The study includes an extensive and thorough field survey, data was analysed floristically. Dicot/Monocot ratio and generic coefficient was calculated following Jacord (1912) as under :-

$$\text{Generic Coefficient (G) \%} = \frac{\text{Total no. of genera}}{\text{Total no. of species}} \times 100$$

Ten most dominating families were also found .

IV. CONCLUSION

The result in the present study clearly show that the flora is very rich floristically which may be attributed to its varied topography and variation in climatic conditions.

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Table -1 Floristic Analysis of the Koria district (Chhattisgarh) India

DICOTYLEDONS:-

S.No.	Family	Wild		Cultivated	
		Genea	Species	Genera	Species
1	Ranunculaceae	01	01	01	02
2	Magnoliaceae	-	-	01	02
3	Annonaceae	02	03	01	02
4	Menispermaceae	02	02	-	-
5	Fumariaceae	01	01	01	01
6	Brassicaceae	01	01	02	04
7	Capparidaceae	03	05	-	-
8	Cleomaceae	-	-	01	02
9	Violaceae	-	-	01	01
10	Polygalaceae	-	-	01	01
11	Caryophyllaceae	-	-	01	01
12	Tamaricaceae	01	01	-	-
13	Elatinaceae	-	-	01	02
14	Malvaceae	06	06	02	05
15	Bombacaceae	01	01	01	02
16	Sterculiaceae	05	05	01	03
17	Tiliaceae	04	06	-	-
18	Linaceae	02	03	-	-
19	Zygophyllaceae	02	02	-	-
20	Oxalidaceae	01	01	-	-
21	Tropaeolaceae	-	-	01	01
22	Rutaceae	04	05	02	06
23	Simaroubaceae	01	01	-	-
24	Burseraceae	02	02	-	-
25	Meliaceae	05	05	-	-
26	Celastraceae	03	03	-	-
27	Rhaminaceae	02	04	-	-
29	Vitaceae	02	02	01	02
30	Sapindaceae	03	03	01	02
31	Anacardiaceae	06	06	03	04
32	Moringaceae	-	-	01	02
33	Papilionaceae	16	20	04	10
34	Caesalpiniaceae	05	12	02	06
35	Mimosaceae	05	13	01	03
36	Rosaceae	-	-	03	05
37	Crassulaceae	01	01	01	02
38	Saxifragaceae	03	03	01	01
39	Droseraceae	-	-	01	02
40	Combretaceae	05	10	01	03
41	Myrtaceae	02	03	03	04
42	Lythraceae	03	03	-	-
43	Onagraceae	01	01	-	-
44	Trapaceae	-	-	01	02
45	Passifloraceae	01	01	02	03
46	Caricaceae	-	-	01	01
47	Cucurbitaceae	02	02	08	15

48	Cactaceae	01	01	01	02
49	Umbelliferae	01	01	02	04
50	Cornaceae	01	01	-	-
51	Rubiaceae	11	15	03	06
52	Asteraceae	11	13	05	13
53	Plumbaginaceae	01	02	01	03
54	Primulaceae	-	-	01	01
55	Sapotaceae	02	02	01	01
56	Ebenaceae	01	03	-	-
57	Oleaceae	01	01	01	04
58	Apocynaceae	09	13	04	10
59	Asclepiadaceae	07	07	03	04
60	Polemoniaceae	-	-	01	01
61	Convolvulaceae	05	06	02	03
62	Cuscutaceae	-	-	01	01
63	Solanaceae	02	03	06	10
64	Scrophulariaceae	02	02	03	04
65	Bignoniaceae	04	06	03	05
66	Pedialiaceae	02	02	01	01
67	Martyniaceae	-	-	01	01
68	Acanthaceae	11	12	05	09
69	Verbenaceae	07	09	04	06
70	Lamiaceae	07	10	04	09
71	Nyctaginaceae	02	02	02	07
72	Amaranthaceae	01	01	01	02
73	Chenopodiaceae	-	-	01	02
74	Polygonaceae	-	-	02	05
75	Aristolochiacee	01	01	01	02
76	Proteaceae	01	01	-	-
77	Lauraceae	02	04	-	-
78	Loranthaceae	01	01	-	-
79	Euphorbiaceae	10	15	06	10
80	Ulmaceae	01	01	01	01
81	Cannabinaceae	01	02	01	02
82	Urticaceae	-	-	01	01
83	Moraceae	03	06	01	03
84	Casurarinaceae	-	-	01	01
85	Ampelidaceae	01	01	-	-
86	Berberidaceae	01	01	-	-
87	Bixaceae	03	03	01	02
88	Boraginaceae	02	02	-	-
89	Clusiaceae	01	01	-	-
90	Cochlospermaceae	01	01	-	-
91	Dilleniaceae	01	01	01	01
92	Dipterocarpaceae	01	01	-	-
93	Myriaceae	01	01	-	-
94	Myrsinaceae	02	02	01	02
95	Nyctanthaceae	01	01	01	01
96	Salicaceae	01	01	-	-
97	Salvadoraceae	01	01	01	02
98	Samydaceae	01	02	-	-
99	Santalaceae	01	01	-	01
100	Piperaceae	01	01	-	02
101	Styreceae	01	01	-	-
102	Thymelaeaceae	01	01	-	-

Monocotyledons:-

S.No.	Family	Wild		Cultivated	
		Genera	Species	Genera	Species
1	Hydrocharitaceae	-	-	01	01
2	Orchidaceae	03	04	-	-
3	Zingiberaceae	05	05	03	04
4	Cannanaceae	02	02	01	02
5	Musaceae	-	-	01	02
6	Hypoxidaceae	-	-	01	02
7	Amaryllidaceae	01	01	02	06
8	Dioscoreaceae	01	02	01	02
9	Liliaceae	06	08	02	06
10	Commelinaceae	01	02	01	01
11	Typhaceae	-	-	01	01
12	Arecaceae	01	02	02	02
13	Pandanaceae	01	01	01	02
14	Araceae	02	02	02	05
15	Potamogetonaceae	-	-	01	02
16	Cyperaceae	01	02	02	02
17	Poaceae	25	29	04	06
	Total	284	364	155	293

D:M Ratio :- 6.00 (Family-wise) 4.85 (Genera-wise)
5.19 (Species wise)

Gymnosperrm :-

S/no.	Family	Wild		Cultivated	
		Genera	Species	Genera	Species
1	Pinaceae	01	01	-	03
2	Cycadaceae	-	-	01	02
3	Ephedraceae	-	-	01	02
	Total	01	01	02	07

Pteridophyta:-

S/no	Family	Wild		Cultivated	
		Genera	Species	Genera	Species
1	Adiantaceae	01	02	-	05
2	Dryopteridaceae	01	02	01	04
3	Martittiaceae	01	01	-	-
	Total	03	05	01	09

Review on Research for Removal of Phenol from Wastewater

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Abstract **-** Phenol is a major pollutant in the wastewater because of its presence in the effluent of major processing and refining plants. It has severe effect on human being, both short term and long term. Various methods are used for removal of the phenol from wastewater such as adsorption, photodecomposition, volatilization and other various biological and non-biological methods. In the present study attempt is done to present the survey of the research on the phenol removal by various methods. The methods such as Polymerization, electrocoagulation, extraction, photodecomposition, advanced oxidation and ion exchange were used effectively by various investigators. These methods are reported to be efficient for the phenol removal. Suitable method for phenol removal can be selected based on availability of the material, extent of separation required and properties of phenolic effluent

Index Terms **-** phenol, effluent, concentration, enzyme, adsorbent

I. INTRODUCTION

Developing green and sustainable technology for the effluent treatment is very important research area in this era of industrial and social development. Many researchers have carried out the research in this field. One of the important pollutants from this point of view is phenol. It has its presence in the effluent from major chemical and pharmaceutical industries such as petrochemical industries, petroleum refineries, coal gasification operations, liquefaction process, resin manufacturing industries, dye synthesis units, pulp and paper mills and pharmaceutical industries. It is a highly corrosive and nerve poisoning agent. Phenol causes harmful side effects such as sour mouth, diarrhea, impaired vision, excretion of dark urine. It is also toxic for fishes.

The toxic levels usually range between the concentrations of 10-24 mg/L for human and the toxicity level for fish between 9-25 mg/l. Lethal blood concentration of phenol is around 150-mg/100 ml. Various treatment processes used for the removal and/or recovery of phenols are hot gas or steam stripping, adsorption, ion exchange solvent extraction, oxidation, phase transfer catalysis and biological treatment processes. Phenolic waste water is treated using activated carbon in the fixed bed/ moving bed/ fluidized bed.

II. VARIOUS METHODS FOR PHENOL REMOVAL

A.Polymerization

Stanisavljević and Nedić have carried out a research on phenol removal by polymerization. They carried out the phenol removal in a reaction step which was polymerization of phenol in presence of an enzyme horseradish peroxidise (HRP)[1]. They used hydrogen peroxide along with the peroxidase enzyme. The native enzyme (E) is oxidized by peroxide (H2O2) to an active intermediate enzymatic form. This accepts an aromatic compound into its active site and carries out its oxidation. A free radical is produced and released into solution leaving the enzyme in the compound state. This compound oxidizes a second aromatic molecule, releasing another free radical product and returning the enzyme to its native state, thereby completing the cycle. The Free radicals formed during the cycle diffuse from the enzyme into the bulk solution where they react to form polyaromatic products. These polymers are water-insoluble and may be removed by solid liquid operations. It was observed that phenol conversion in all experimented conditions was greater than 90%. The high efficiency observed is in accordance with conditions optimized to guarantee 90% polymerization using purified HRP. There is a compromise between the reduction in variable costs by the use of fewer enzymes and the increase in capital investment at the time of building the treatment facility. The effect of horseradish peroxidase (HRP) and H2O2 concentrations on the removal efficiency of phenol, defined as the percentage of phenol removed from solution as a function of time, has been investigated by Vasudevan and Li[2]. They found that the phenol is almost completely precipitated within 10 minutes. . The removal efficiency increases with an increase in the concentration of HRP, but an increase in the time of treatment cannot be used to offset the reduction in removal efficiency at low concentrations of the enzyme, because of inactivation of the enzyme. Nicell et. al. have carried out the search on phenol polymerization and precipitation using polyethylene glycol as additive [3]. They observed that the presence of polyethylene glycol enhance enzyme performance for the treatment of synthetic wastewaters containing a range of phenol concentrations between 0·5 and 16 millimolar (47 to 1500 mg/l) at neutral pH. Results demonstrated a significant improvement in catalyst lifetime in the presence polyethylene glycol in both batch and continuous stirred tank reactor configurations. Free and immobilized horseradish peroxidase was used for phenol polymerization by Pradip et. al. [4]. They used Horse Radish roots for the extraction of peroxidase enzyme in the laboratory. Buffer of pH 6.8 was used to soak the roots in order to avoid the enzyme inactivation during crushing and extraction. They observed that phenol polymerization of 84% was achieved at 100 mg/l using free HRP enzyme and phenol degradation of 62% was observed using Immobilized HRP enzyme bed reactor. They also observed that as the concentration of phenol increased, there was reduction in phenol polymerization efficiency. Effect of reaction condition on phenol polymerization was studied by Masuda et.al. [5]. They investigated the quantitative relationships between removal efficiency of phenol and reaction conditions using coprinus cinereus peroxidase. At an adequate enzyme dose, they observed that the most effective ratio of hydrogen peroxide to phenol was nearly 1/1 (mol/mol). They also

observed that 12.2 U of the enzyme was needed to remove 1 mg of phenol.

B. Electro coagulation

Abdelvahab et.al. have carried out the research on electrochemical removal of phenol from oil refinery waste [6]. They explored the possibility of using electro coagulation for phenol removal. They used the cell with horizontally oriented aluminum cathode and a horizontal aluminium screen anode. They studied the phenol removal with respect to various parameters such as pH, operating time, current density, initial phenol concentration and addition of NaCl. According to them removal of phenol during electro coagulation was due to combined effect of sweep coagulation and adsorption. The results showed that, at high current density and solution pH of 7, remarkable removal of 97% of phenol after 2hour can be achieved. Also they observed that the rate of electro coagulation increases with decrease in phenol concentration and the maximum removal rate was attained at 30 mg /l phenol concentration. The study showed that, electro coagulation of phenol using aluminum electrodes is a promising process. Effect of the variables like pH, operating time, current density, initial phenol concentration and addition of NaCl on phenol removal by coagulation was studied by Zazouli and Taghavi [7]. They observed maximum phenol removal percentage at a pH value of 7. They also observed that increasing the concentration of phenol led to decrease in the removal efficiency. They also observed that the removal rate of phenol increased with increasing current density, and the highest current density (25 mA/cm2) had the maximum removal efficiency. Ashtoukhy et. al. have carried out the research on treatment of petrochemical wastewater containing phenolic compounds by electro coagulation using a fixed bed electrochemical reactor[8]. They explored the possibility of using electro coagulation to remove phenolic compounds from oil refinery waste effluent using an electrochemical reactor with a fixed bed anode made of randomly oriented aluminium rasching rings packed in a perforated plastic basket located above the horizontal cathode.

C. Extraction

The extraction of phenol from simulated sebacic acid waste water was tried by Rao et. al.[9]. They used 1-hexanol, 1-heptanol and 1 octanol as solvent for phenol removal. It was observed that 1 octanol showed lesser phenol removal efficiency compared to other two solvents. Xu et. al. have carried out investigation on extraction of phenol in wastewater with annular centrifugal contactors [10]. They carried out the experimental study on treating the wastewater containing phenol with QH-1extractant (the amine mixture) and annular centrifugal contactors. They observed that the extraction rate of the three-stage cascade was more than 99%. When 15% NaOH was used for stripping of phenol in QH-1(the amine mixture), the stripping efficiency of the three-stage cascade was also more than 99% under the experimental conditions.

D. Photodecomposition

Investigation of the photodecomposition of phenol in near-UV-irradiated aqueous TiO_2 suspensions was carried out by Ilisz et. al.[11]. They investigated The effects of charge-trapping species on the kinetics of phenol decomposition. They observed that the heterogeneous degradation of phenol followed apparently zero-order kinetics up to 70% conversion. The results of the experiments in the presence of Ag^+ indicated that the phototransformation of phenol can proceed via direct electron transfer, neither dissolved O_2 nor its reduced forms play a significant role in the degradation mechanism. Akbal and Onar have studied photocatalytic degradation of phenol [12]. They carried out the investigation to study photocatalytic degradation of phenol in the presence of UV irradiated TiO_2 catalyst and H_2O_2. They concluded that photocatalytic degradation can be effective method for phenol removal.

E. Biological Methods

Marrot et.al have carried out the research on biodegradation of high phenol concentration by activated sludge in an immersed membrane bioreactor.[13]. They investigated the effect of adaption of mixed culture on phenol degradation. They found that biological treatment was economical and practical for removal of phenol. High concentrations of phenol are inhibitory for growth. Biological and enzymettric treatments were used for phenol removal by Bevilaqua et. al[14]. The systems studied were conventional batch aerobic biological followed or preceded by enzymatic treatment. They employed the Tyrosinase as enzyme.. They observed that biological treatment effectively degrades phenol upto concentration of 420 mg/l. Enzymatic polishing of biotreated effluent removed up to 75% of the remaining phenol in a four-hour reaction with 46 $U.mL^{-1}$ of tyrosinase and 50 $mg.L^{-1}$ of chitosan (used as coagulant). The research on detection of phenol degrading bacteria and pseudomonas putida in activated sludge by polymerase chain reaction was carried out by Movahedyan[15]. According to results in this study, the best phenol-degrading bacteria that can utilize 500 – 600 mg/l phenol completely after 48 hours incubation belong to Pseudomonas Putida strains. It is clear that use of isolated bacteria can lead to considerable decrease of treatment time as well as promotion of phenol removal rate. Tziotzios et. al. have reported research on biological phenol removal[16]. They carried out research to study the efficiency of packed bed reactor on phenol biodegradation and effect of specific area of specific material on biodegradation efficiency. They used the indogeneous bacteria from olive pulp. The maximum phenol removal rate was observed to be 12.65 gram per litre per day, when filled with gravel support material under draw fill operation. Same operation with plastic packings showed maximum phenol removal of 4.3 grams per litre per day. A review on recent advances in biodegradation of phenol was carried out by Basha et al[17]. Comparative study of free and immobilized growth was carried out by Pishgar et. al[18]. They observed that the culture was able to degrade the phenol upto 700 mg/l. The immobilized cells were able to remove phenol at concentration of 100 to700 mg/l in a slightly shorter time period. The biodegradation rate of phenol improved when immobilized cells were applied. Kinetic Studies for an aerobic packed bed biofilm reactor for treatment of organic wastewater with and without phenol was carried out by Dey and Mukharjee[19]. They carried out studies on removal efficiency of COD and phenol in a mixture of carbohydrate and phenol.The aerobic bioreactor with the glass beads was used for the purpose. They concluded that the performance of the bioreactor decreased marginally under 50 mgL^{-1} phenol charging along with other carbohydrate in the influent wastewater.

At this condition, phenol removal rate of 89% was achieved. Almsi et. al. have carried out the work related to anaerobic wastewater stabilization pond for phenol removal[20]. The phenol removal efficiency was found to 90 percent. According to the study, anaerobic ponds with petroleum wastewater is attractive alternative than many other methods. Removal of phenol by using rotating biological contactors was tried by Pradip et.al.[21]. They examined the impact of process variables, viz concentration, rotational speed and percentage submergence of the disc on phenol removal.

F. Electro-Fenton (EF-Fere) method

An improved Electro-Fenton (EF-Fere) method using H_2O_2 amendments and electrogenerated ferrous ions was investigated to treat phenol-containing wastewater by Jiang et.al.[22]. The degradation process of phenol was carried out in an EF-Fere system, which was composed of a power source, a cylindrical electrolytic cell and a H_2O_2 dosing system. The electrolysis was controlled by an electrochemical working station. For the phenol degradation experiments conducted in the EF-Fere electrolytic system, the maximum COD removal efficiency of phenol-containing wastewater is achieved at the condition of 800 mg/L initial ferric ions concentration, 1.0A electric current with continuous H_2O_2 addition mode. Furthermore, SnO_2 film anode and UV irradiation in the EF-Fere system are beneficial to COD removal efficiency.

G. Advanced oxidation processes

Rubalcaba et. al. have used advanced oxidation processes coupled to a biological treatment for phenol remediation[23]. Results showed promising research ways for the development of efficient coupled processes for the treatment of wastewater containing toxic or biologically non-degradable compounds. Similar research has been carried out by Esplugas et. al. [24]. Though Fenton reagent was most effective for degradation of phenol,. lower costs were obtained with ozonation. In the ozone combinations, the best results were achieved with single ozonation. Phenol degradation in presence of chlorides and sulphates was carried out by Siedlecka et. al.[25]. They studied the degradation of three representatives of phenolic compounds in presence of chlorides and sulphates, namely phenol, 2 chlorophenol and 2 nitrophenol. The presence of anions influenced the degradation rates. Relatively low degradation rates were observed for 2 chlorophenol. The biodegradability of phenol was increased by chloride while that of other two derivatives was increased by presence of sulphates.

H. Adsorption and Ion exchange

Experiments were conducted to examine the liquid-phase adsorption of phenol from water by silica gel, activated alumina, activated carbon by Roostaei and Tezel [26]. Experiments were carried out for the analysis of adsorption equilibrium capacities and kinetics. They found it to be a promising method for phenol removal. The research on phenol removal from aqueous solution by adsorption and ion exchange mechanisms onto polymeric resins was done by Caetano et.al. [27]. They evaluated the removal of phenol from aqueous solution by using a nonfunctionalized hyper-cross-linked polymer macronet and two ion exchange resins. The nonfunctionalized resin reported the maximum loading adsorption under acidic conditions, where the molecular phenol form predominates. Ion exchange resins showed the maximum removal in alkaline medium. Desorption of nonfunctionalized resin was achieved by using the solution (50% v/v) of methanol/water with a recovery close to 90%. In the case of the ion exchange resins the desorption process was performed at different pHs. Qadir and Rhan have investigated the removal of phenol by using adsorption [28]. The treatment of waste with active carbon is considered to be an effective method for the removal of phenol from waste solution because of its large surface area. The experimentation on adsorption isotherms for phenol removal on activated carbon was carie out by Maarof et.al[29]. The adsorption isotherm parameters for the Langmuir and Freundlich models were determined using the adsorption data. It was found that both the Langmuir and the Freundlich isotherms described well the adsorption behavior of phenol on Norit Granular Activated Carbon) NAC 010, while the Freundlich isotherm described very well the adsorption of phenol on NAC 1240. Jadhav and Vanjara have carried out the research on adsorption of phenol sawdust, polymerized saw dust and sawdust carbon[30]. They studied the influence of the parameters like concentration, agitation speed, amount of adsorbent and the pH on adsorption capacity. Hycinthe as adsorbent was used for phenol removal by Uddin et.al[31]. They carried out the batch kinetic and isotherm studies under varying experimental conditions of contact time, phenol concentration, adsorbent dosage and pH. The infered that the adsorption of phenol decreased with increasing pH. The results also showed that kinetic data followed closely to the pseudo-second-order model. Activated carbons prepared from date stones was used for removal of phenolic compunds from aqueous solutions by adsoption by Dhidan[32]. He used activated carbon prepared from date stones by chemical activation with ferric chloride (FAC) as an adsorbent to remove phenolic compounds such as phenol (Ph) and p-nitro phenol (PNPh) from aqueous solutions. He achieved the maximum phenol removal of 98% at pH value of 5 and 90 minutes of contact time. Kadhim and Al-Seroury have carried out research on characterization the removal of phenol from aqueous solution in fluidized bed column by rice husk adsorbent.[33].They found that the pretreatment of rice husk increase the specific surface area and changed the functional groups , therefore leads to increase in the capacity of adsorption. Kulkarni et. al. have investigated the phenol removal from the effluent by using activated carbon in batch and fluidized bed experimentation[34]. In batch studies they studied the effects of various parameters like adsorbent dose, pH, particle size on rate of adsorption . In case of fluidized bed, the effect of various parameters like concentration, fluid flow rate and adsorbent particle size were studied. The % decrease in phenol concentration increases with increase in adsorbent dose. With a reduction in particle size, initially steep increase in % removal of phenol is observed and it becomes more significant for finer particles.In case of fludized beds, Increase in fluid flow rate gives better adsorption in case of activated carbon. In the present study particle size of 0.420 mm is found more beneficial.Girish and Murti have studied the potential of various low cost adsorbents for phenol removal[35]. This review indicated that these agricultur-al materials have equivalent or even more adsorption capacity to activated carbon. Lua and Jia have used oil palm shell activated carbon in a fixed bed adsorber for phenol removal.[36]. The used the adsorption using the activated carbon derived from oil palm shells for the phenol, which they found to be very effective.

A research on mass transfer coefficients in a packed bed using tamarind nut shell activated carbon to remove phenol was done by Goud et. al[37]. They reported work on the preparation and characterization of activated carbon from tamarind nutshell, an agricultural waste byproduct, and its use in a packed bed for the removal of phenol. They found that that the sorption of phenol is dependent on both the flow rate and the particle size of the adsorbent, and that the breakpoint time and phenol removal yield decrease with increasing flow rate and particle size. For mass transfer coefficient, experimental values were in excellent agreement with the predicted values from the correlation. The experimentation for examining the application of adsorption packed-bed reactor model for phenol removal was done by Sorour et.al[38].They conducted the experiments to determine the Langmuir equilibrium coefficients and to determine the bulk sorbate solution concentration for different adsorption column depths and different time as well. They predicted a packed bed model which was in well agreement with the laboratory data. Batch and column studies for phenol removal on to onto nano iron oxide and alginate microspheres was carried out by Soni et.al[39]. They studied the adsorption behavior of o-nitrophenol from aqueous medium, using nano iron oxide loaded calcium alginate beads using equilibrium batch and column flow techniques. They reported the sorption capacity per unit bed volume and rate constant to be 578.4 mg/L and 1.18 L/mg/min, respectively. Kulkarni et.al. have used tamarind bean activated carbon for phenol removal in fluidized bed[40]. They studied the effect of the effect of various parameters like

concentration, fluid flow rate and adsorbent particle size.They observed that It is observed that as the concentration increases the percent saturation of adsorbent increases. Also increase in fluid flow rate gives better adsorption.

I. Membrane based separation

Membrane-Based Separation of phenol/water mixtures using ionically and covalently cross-linked ethylene-methacrylic acid copolymers was tried succesfully by Mixa and Staudt[41]. They performed the Membrane-based separation of phenol/water mixtures with concentrations of phenol between 3 wt% and 8 wt% in the feed with nonmodified as well as cross-linked ethylene-methacrylic acid (E-MAA) copolymers with different amounts of methacrylic acid. They concluded that using nonmodified membranes with higher methacrylic acid monomer content in the polymer, lower fluxes and higher enrichment factors were observed. Also the Investigation of different cross-linked membranes showed that with high phenol concentration in the feed, ionic cross-linking seems to be very promising. Polyurethane urea as pervaporation membrane for the selective removal of phenol was tried by Gupta et.al[42]. They reacted the Hydroxyterminated polybutadiene (HTPB) with 2,4-toluene diisocyanate (TDI) followed by the addition of a diamine chain extender (prepared by the condensation reaction of 4,4′-diaminodiphenylsulfone and terepthalaldehyde) to prepare an imine containing polyurethane urea (PIUU).About 88 percent of phenol was seperated as a condensed permeate. Diaconu et. al have carried out research on use of membrane techniques for phenol separation[43].They presented a separation study of some phenolic compounds frequently encountered in the environment: m-nitrophenol, p-nitrophenol, m-cresol, p-cresol, using the bulk liquid membranes technique. They used chloroform as the organic solvent for the membrane and studied the operational parameters of the transport and also established the optimum separation conditions (the feed phase pH, the receiving phase pH, the time period of the transport). The best transport efficiencies were obtained for m-nitrophenol and p-nitrophenol. The transport efficiencies in the case of these phenolic derivatives were 92% for nitrophenol and 98% for p-nitrophenol.

III.CONCLUSION

For removal of phenol extensive research has been carried out to study the removal potential of each of the method. Many biological and nonbiological processes such as polymerization, electro coagulation, extraction, photodecomposition, advanced oxidation process and adsorption were employed various researchers for phenol removal. By using the enzyme horseradish peroxide, for phenol polymerization. At optimal conditions, phenol polymerization efficiency of about 90 percent was observed. The removal of phenol during electro coagulation was due to combined effect of sweep coagulation and adsorption. At high current density and solution pH of 7, remarkable removal of 97% of phenol after 2 hour was achieved.

Photoreactor for wastewater treatment using titanium dioxide nanoparticles is also promising alternative. Various suspended and attached growth processes were also used effectively and phenol removal efficiency of 90 % was achieved. By using electro-Fenton process the removal efficiency of 96% was obtained. Adsorption operation by various adsorbents is also attractive alternative. It has the advantage of wide range of adsorbents available. Many adsorbents were waste materials from agricultural and other industrial operations such as distilleries. Phenol removal ranging from 88 % to 95 % was observed using various adsorbent. Adsorption seems to be possibly most widely studied operation for phenol.

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Ecological Impacts of Weed (*Parthenium hysterophorus* L.) Invasion in Saline Soil

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Abstract- *Parthenium* density and species richness were found to be more in non-invaded (*N-I*) sites as compared to invaded (*I*) sites of Jaunpur district. pH, OC and %N of the soil indicates that the plant survive naturally at higher level of salinity. In pot experiment, plant growth, proline content, TSS content, chlorophyll, N, P and K was not significantly affected up to 1% NaCl (w/w) over the control after 30 and 60DAS. BCF for parthenium plant was >1 up to 2% NaCl treatment. Bacterial population was more in rhizospheric soil then rhizoplane soil under NaCl after 30 and 60DAS, and indicates bacterial-rhizoadaptation. Hence, the *parthenium* plant able to improve soil health and might be useful in the field of restoration ecology.

Index Terms- Parthenium growth, Nutrients, Sodium, Proline, Bacterial population.

I. INTRODUCTION

Parthenium was thought to be initially introduced in India before 1950s [1,2] with food grains imported from the USA, but heavily introduced after 1956 through the transport of Milo (red wheat) from Mexico. After 1956, the weed spread throughout India, and invades in all disturbed land, including farms, pastures, and roadsides. In some areas, outbreaks have been of almost epidemic proportions, affecting crop production, livestock and human health. *Parthenium* is found mostly in hot climates. High temperature (28 to 33 °C) is favorable because it increases the dry matter production of the plant due to maximum photosynthesis [3]. Low winter temperature inhibits the growth of the plant and the seed production [1,3,4]. *Parthenium* density and biomass varies with soil type [5]. The most common soil types on which it is found are alkaline clay, loam soil to heavy black clay soils [6]. Soil with rich clay produces a smaller number of larger seeds, whereas seed mass declines in relatively coarser soils with large number of small seeds [5]. Invasive plants are threat to biodiversity, leading to change in natural habitat and nutrient cycles [7], but only little work has been conducted on change in the soil. Therefore, keeping this view the present study was based on to evaluate the *Parthenium* growth under saline soil.

II. MATERIAL AND METHODS

Study area description and weed status

Surveys on five study areas were done by "walk transect" with stratified & systematic surveying methods for the inspection of *P. hysterophorus* invasion into Jaunpur distric. Phyto-sociological studies were conducted by applying Quadrat method. The species richness (number of species per sampling unit) and the density of *Parthenium* (pl/m^2) were determined for each quadrate sampled. The frequency of each species in *Parthenium* non-invaded and invaded sites was calculated. Jaccard's Similarity Index (JSI) and Sorenson's Similarity Index (SSI) between *Parthenium* non-invaded and invaded area of each site were calculated as JSI=C X 100 /A+B-C, SIS=2C X 100/A+B. Where A= Total number of species in non-invaded plots and B= Total number of species in invaded plots. C= Total number of species common to invaded and non-invaded plots. Plot wise data of vegetation and soil attributes were used in statistical analysis. Mean values of *Parthenium* density, species richness, soil pH, soil organic carbon (OC), soil nitrogen content (N) in invaded sites were compared by using Mann-Whitney U test. The statistical analysis was conducted using the Statistical Package for Social Sciences (SPSS) version 11.5.

Experimental set-up for pot experiment

A growth chamber experiment was conducted according to [8] by using small; autoclaved plastic pots (each filled with 200 g of soil) supplemented with NaCl (w/w) with different concentration 0.5, 1.0, 1.5, 2.0 and 2.5% and equilibrated for 20 days. The similar height of small plants were uprooted from natural terrestrial eco-system and showing in the prepared pots. The treatments were repeated three times with six replicates, and plants were harvested after 30 and 60DAS for further analysis.

Analysis of Na, chlorophylls and osmolytes

Sodium, Nitrogen, Phosphorous and potassium analysis were examined by the method earlier described by [8,9] in leaf and soil. The leaf /soil samples were dried in an oven at 68^0 C for 48 h. The dried leaves were ground and digested in an acid mixture (HNO_3: H_2SO_4: $HClO_4$ = 10:1:3) for determination of Na, N, PO_4 and K. Bioconcentration factor (BCF) of the sodium was calculated by the method [10,11] as BCF=Metal concentration in plant/Metal concentration in soil. Leaf chlorophyll a and b were estimated by spectrophotometrically [12] at 663 nm and 645 nm. Proline and TSS were determined by the method earlier described by [8]. For proline, 0.5 g of leaf tissue was homogenized in a pestle and mortar with 10 ml 3% (w ⁄ v) aqueous sulphosalicylic acid and filtered through Whatman No. 2 filter paper; 1 ml of filtrate was mixed with 2 ml glacial acetic acid and 2 ml ninhydrin in a test tube; boiled at 100^0C for 1 h and placed on ice to stop the reaction. Thereafter, 4 mL toluene was mixed vigorously for 20–30s, and the aspirated chromospheres (toluene) layer brought to room temperature. Absorbance was read at 520 nm against a reagent blank and the

amount of proline (µg/g fresh weight) was calculated from a standard curve for proline [13]. 100 mg of fresh leaves were placed in a test tube containing 5 ml 2.5 N HCl, boiled for 3 h then cooled to room temperature for TSS (total soluble sugar) analysis using the anthrone method [8].

III. RESULTS AND DISCUSSION

Five major sites were selected in the present study of *pathenium* plant status in non-invaded (*N-I*) and invaded (*I*) sites. *Parthenium* density and species richness were found more in *N-I* sites as compared to *I* sites (Table 1). Kohli et al. [14] reported a decline in species richness from 25 to 12 from *Parthenium* non invaded site to high invaded site of Lower Himalaya (India). Similarly, Belz et al., [2] reported that *Parthenium* have allelopathic effect in its root and shoot leachates and thus has the ability to reduce the growth and germination of numerous associated species. The mechanism of decrease of species richness is elaborated [1]. Soil pH and electrical conductivity (ECe=dSm^{-1}) was more in *I* sites over the *N-I* sites. Earlier researcher reported that densities of parthenium and biomass vary with soil type [5]. The most common soil types on which it is found are alkaline, clay, loam soil to heavy black clay soils [6]. In the present study, soil pH was with range 6.9 to 7.7 in *N-I* and 7.2 to 7.8 in *I* sites, while electrical conductivity was with range 9 to 23 dSm^{-1} in *N-I* sites and 18 to 42 dSm^{-1} in *I* sites (Table 1). Temperature (^{0}C) and rain fall was recorded (data not shown) over the (December 2011 to July 2012) of the *parthenium* grown sites, indicates that the *parthenium* plant survive at maximum range of temperature as well as rain fall. Pandey et al. [3] reported that *parthenium* found mostly in hot climates with range of temperature (28-33°C) are favorable because it increases the dry matter production of the plant due to maximum photosynthesis, while low winter temperature inhibits the growth of the plant and the seed production [1,3,4]. In the present study, with a novel approach to examine the organic carbon (OC) and percent nitrogen (%N) in the soil of *parthenium* grown sites and result revealed that no any clear relation was observed for OC and %N in soil under *N-I* and *I* sites (Table 1), while the level of OC and %N was sufficient to support the growth of the plant [9]. Among results indicate that the *parthenium* plant survive at adverse condition like high ECe, pH and able to maintain the OC and %N in natural environment. Keeping these views, we observed that the growth of *parthenium* plant and soil health in pot experiment under saline condition. Root, shoot length and dry biomass of *parthenium* plant was not significantly affected up to 1% NaCl (w/w) concentration over the control, while reduction was observed from 1 to 2.5% NaCl concentration after 30 and 60DAS (Table 2). In the present study, the proline content increased with increased concentration of NaCl. Plant proline content was not much affected at 1% NaCl (w/w), while at 1.5% NaCl plant proline was sudden increased *ie* 54% and 65.8% over the control after 30 and 60DAS respectively.TSS was increased up to 1.5% NaCl concentration after that TSS was sudden decreased. Proline may act as a mediator of osmotic adjustment, protects macromolecules during dehydration and serve as a hydroxyl radical scavenger [15]. Proline and TSS was increased under saline soil was earlier reported [8,9]. Chlorophyll-a (Ch-a) and Chlorophyll-b (Ch-b) was almost not affected up to 1% NaCl (w/w) concentration, while affected beyond 1% to 2.5% NaCl over the control after both 30 and 60DAS (Table 2). In the present study, percent nitrogen (%N) and percent potassium (%K) in the plant was little increased up to 1% NaCl treated soil with plant, while percent phosphorous (% PO_4) increased up to only 0.5% NaCl treated plant. N, P and K content in the plant was more in 60DAS as compared with 30DAS in all the NaCl treated pot. Earlier report supported that *parthenium* plant induces changes in the physical and chemical properties of soil such as soil texture, soil pH, soil organic matter, soil nitrogen, soil potassium, soil phosphorus etc [16,17].

Table 1: Comparison of mean of *parthenium* density, species richness and soil attributes of parthenium non-invaded (*N-I*) and invaded (*I*) sites of Jaunpur.

Sites	*Parthenium* density		Species richness		Soil pH		ECe of soil		OC% in soil		N% in soil	
	N-I	***I***	***N-I***	***I***	***N-I***	***I***	***N-I***	***I***	***N-I***	***I***	***N-I***	***I***
Chaukiya	29	52	27	12	7.1	7.4	22	24	2.2	2.55	0.38	0.52
Khetsarai	63	108	16	9	7.0	7.2	9	18	2.5	2.02	0.52	0.57
Kuthan	117	212	22	11	7.7	7.6	18	35	3.4	3.06	0.32	0.31
Sahgang	72	157	13	8	7.1	7.5	23	31	2.4	3.00	0.31	0.57
Sipah	21	47	12	8	6.9	7.8	12	42	2.1	1.53	0.42	0.59
SL	<0.001	<0.001	1.21	0.925	<0.001	<0.001	<0.001	<0.001	0.572	.821	0.011	0.018

Data are the average mean of 50 quadrate and soil samples (n =10 X 50) form each location, ECe=(electrical conductivity in dSm^{-1}), and SL (Significance level) based on Mann-Whitney U test.

Sodium accumulation in the plant was increased with increased concentration of NaCl at 30DAS, while maximum 67% sodium accumulates at 1% NaCl (w/w) treated plant after 60DAS (Table 3). Bioconcentration factor (BCF) was >1 indicates that plant behave as a good accumulator of sodium, in the present study BCF was >1 up to 1.5% NaCl (w/w) treated plant, hence the parthenium plant might use in phytoremadiation of metal (*ie*. Na) [10,11]. Plant microbe interaction was observed on the basis of population count (cfu/gm soil) of rhizospheric and rhizoplane

soil in the treated pot. Bacterial population was greatly influenced under higher salinity level; maximum bacterial population was found in rhizospheric region as compared with rhizoplane. This showed rhizo-adaptation of bacteria, most of the rhizobacteria were earlier reported for their plant growth promoting activities, and they help to growth in plant with direct and indirect mechanisms under saline condition [8, 9]. It is interesting finding that bacterial population was not detected at 2 and 2.5% NaCl treated pot in rhizoplane, while bacterial population 5 X 10^5 and 8 X 10^5 cfu/gm soil was found in rhizospheric region at 2.5% NaCl treated pot after 30 and 60DAS respectively (Table 3). Overall rhizospheric bacterial population was unaffected up to 1% NaCl (w/w) treated plant in pot. Bacterial population at 1% NaCl treated pot was 12 X 10^6 cfu/gm soil and 20 X 10^6 cfu/gm soil after 30 and 60 DAS respectively in the rhizospheric soil and around 8 X 10^6 cfu/gm soil was in rhizoplane soil after both DAS (Table 3). Upadhyay et al. [8,9,18] earlier reported that the N, P and K concentrations was increased in wheat plant inoculated by PGPRs under salinity stress conditions, and more bacterial population were found in rhizospheric region as compared with rhizoplane for wheat plant under salinity [9]. The present study revealed that parthenium is an exotic weed, help to mitigate salinity in soil and improve soil health under saline environment. Therefore, the parthenium plant might be use in the field of restoration ecology.

Table 2: Plant growth and biochemical analysis of *Parthenium* plant under different salt concentrations [T1=0.5%, T2=1.0%, T3=1.5%, T4=2.0 and T5=2.5% (NaCl:w/w)] in pot.

TT	Proline		TSS		Ch-a		Ch-b		R L (cm)		S L (cm)		DBM (g)	
	30D	**60D**	**30D**	**60D**	**30D**	**60D**	**30D**	**60D**	**30D**	**60D**	**30D**	**60D**	**30D**	**60D**
C	0.88	0.94	88	124	2.8*	4.3*	2.2*	3.1*	6.2*	8.4*	15.6*	18.5*	4.8*	6.3*
T1	1.14*	1.26*	93	132*	2.6**	4.3*	2.1*	3.0*	6.1*	8.6*	15.2*	18.2*	4.6**	6.3*
T2	1.25*	1.22**	95*	137*	2.5**	4.1*	2.0*	3.0**	6.3*	8.2*	15.7*	18.3**	4.3	6.1**
T3	1.94	2.75	98*	140**	1.7	2.6	1.3	1.5	4.2**	6.3	11.6	12.5	3.2	4.2
T4	2.62**	2.84	72	112*	1.5	2.1	1.1	0.9	4.1**	5.7	11.4	11.6	2.8	2.9
T5	2.68**	2.80	57	81	1.2	1.5	0.9	0.5	3.8	4.0	9.2	10.1	2.4	2.6

Data are average mean of six replicates, * = Significant from control at 0.05 level (t-test), ** = Highly significant from control at 0.01 level (t-test), C (Control), Proline (µg/g Fresh weight), TSS (Total soluble sugars= µg/g Fresh weight), Ch-Chlorophyll (mg/g Fresh leaf), RL (Root length), SL (Shoot length) and DBM (Dry Biomass).

Table 3: Ions and bacterial population status in different NaCl (w/w) treatments (TT) [T1=0.5%, T2=1.0%, T3=1.5%, T4=2.0 and T5=2.5%] *parthenium* plant in pot.

TT	N (%)		PO4 (%)		K (%)		Na (%)		BCF of Na		[a]Cfu/gm Soil		[b]Cfu/gm Soil	
	30D	**60D**	**30D**	**60D**	**30D**	**60D**	**30D**	**60D**	**30 D**	**60 D**	**30 D**	**60 D**	**30 D**	**60 D**
C	0.14*	0.19*	0.2*	0.32*	0.22*	0.24*	0.6	0.9	-	-	125	201	108	181
T1	0.16*	0.21*	0.6*	0.39*	0.21*	0.30	1.4*	1.9*	2.8	3.8	122	205	106	116
T2	0.25*	0.29*	0.6*	0.29*	0.32*	0.28	1.5*	2.8*	1.5	2.8	120	204	80	84
T3	0.22	0.26	0.7	0.16	0.24	0.19*	1.8	2.5*	1.2	1.7	58	64	21	10
T4	0.15*	0.17	0.4	0.12	0.18	0.16	2.1	2.0	1.0	1.0	9	14	Nd	nd
T5	0.11	0.08	0.2	0.3	0.14	0.14	2.2	2.0	0.8	0.8	5	8	Nd	Nd

Data are average mean of six replicates, C =Control), * = Significant from control at 0.05 level (t-test), D= (Days) BCF=Bioconcentration factor, Cfu= colony forming unit ([a])= Bacterial population (1 X10^6) in rhizospheric region,([b])= Bacterial population (1 X10^6) in rhizoplane region and nd=not detected.

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An Overview of Network Architecture and security framework of Asynchronous Transfer Mode

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Abstract- Asynchronous Transfer Mode (ATM) has emerged as most promising technology in supporting broadband multimedia communications. Congestion control plays important role in the effective and stable operation of ATM networks. The management of these networks creates new challenges for both private network operators and public telecommunications service provider communities due to the heterogeneous mix of ATM Switch equipment, and the need to establish, control and monitor end-to-end connections (virtual circuits) through a network. ATM is a standard for carriage of a complete range of user traffic, including voice, data, and video signals. It is designed to unify telecommunication and computer networks. It uses asynchronous time-division multiplexing and it encodes data into small, fixed-sized cells. ATM provides data link layer services that run over a wide range of OSI physical Layer links. ATM has functional similarity with both circuit switched networking and small packet switched networking. It was designed for a network that must handle both traditional high-throughput data traffic (e.g., file transfers), and real-time, low-latency content such as voice and video.

Index Terms- ATM Networks, Congestion, VP, VC, VPI, VCI, OSI model

I. INTRODUCTION

There are lot of high speed networking technologies used to transfer data between two or more devices. Asynchronous Transfer Mode is one of these technologies. ATM's speed, low latency and ability to handle all types of traffic over a single network make it ideal for a range of bandwidth-intensive applications, including multimedia, medical imaging and more.One difference between ATM and other networking technologies is the concept of "virtual circuits".Negotiating a virtual circuit with nominated bandwidth and quality of service parameters is a feature of ATM circuit provisioning.VP's in asynchronous transfer mode (ATM) networks provide substantial speedup during the connection establishment phase at the expense of bandwidth loss due to the end-to-end reservation of network resources. Thus, VP's can be used to tune the fundamental tradeoff between the network call throughput and the processing load on the signaling system. They can also be used to provide dedicated connection services to customers such as virtual networks (VN's).ATM operates as a channel-based transport layer, using virtual circuits (VCs).ATM became popular with telephone companies and many computer makers in the 1990s. ATM operates at the data link layer (Layer 2 in the OSI model) over either fiber or twisted-pair cable.Asynchronous Transfer Mode (ATM) is a technology that has the potential of revolutionizing data communications and telecommunications.

II. NECESSITY

2.1 Frame Network

Before ATM, data communications at the data link layer had been based on frame switching and frame networks. Different protocols use frames of varying size and intricacy. As networks become more complex, the information that must be carried in the header becomes more extensive. The result is larger and larger headers relative to the size of the data unit. In response, some protocols have enlarged the size of the data unit to make header use more efficient (sending more data with the same size header). Unfortunately, large data fields create waste. If there is not much information to transmit, much of the field goes unused. To improve utilization, some protocols provide variable frame sizes to users.

2.2 Mixed Network Traffic

As we can imagine, the variety of frame sizes makes traffic unpredictable. Switches, multiplexers, and routers must elaborate software systems to manage the various sizes of frames. A great deal of header information must be read, and each bit counted and evaluated to ensure the integrity of every frame. Internetworking among the different frame networks is slow and expensive at best, and impossible at worst. Another problem is that of providing consistent data rate delivery when frame sizes are unpredictable and can vary so dramatically. To get the most out of broadband technology, traffic must be time-division multiplexed onto shared paths.

2.3 Cell Networks

Many of the problems associated with frame internetworking are solved by adopting a concept called cell networking. A cell is a small data unit of fixed size. In a **cell** network, which uses the **cell** as the basic unit of data exchange, all data are loaded into identical cells that can be transmitted with complete predictability and uniformity. As frames of different sizes and formats reach the cell network from a tributary network, they are split into multiple small data units of equal length and are loaded into cells. The cells are then multiplexed with other cells and routed through the cell network. Because each cell is the same size and all are small,

the problems associated with multiplexing different-sized frames are avoided.

2.4 Asynchronous TDM

ATM uses asynchronous time-division multiplexing-that is why it is called Asynchronous Transfer Mode-to multiplex cells corning from different channels. It uses fixed-size slots (size of a cell). ATM multiplexers fill a slot with a cell from any input channel that has a cell; the slot is empty if none of the channels has a cell to send. When all the cells from all the channels are multiplexed, the output slots are empty.

III. ARCHITECTURE

ATM network is switched network. It's architecture uses a logical model to describe the functionality it supports. ATM functionality corresponds to the physical layer and the data link layer of the OSI reference model.
The ATM reference model is composed of the following planes, which span all layers:

- *Control*-This plane is responsible for generating and managing signaling requests.
- *User*- This plane is responsible for managing the transfer of data.
- *Management*-This plane contains two components:
  -Layer management manages layer-specific functions, such as the detection of failures and protocol problems.
  -Plane management manages and coordinates functions related to the complete system.

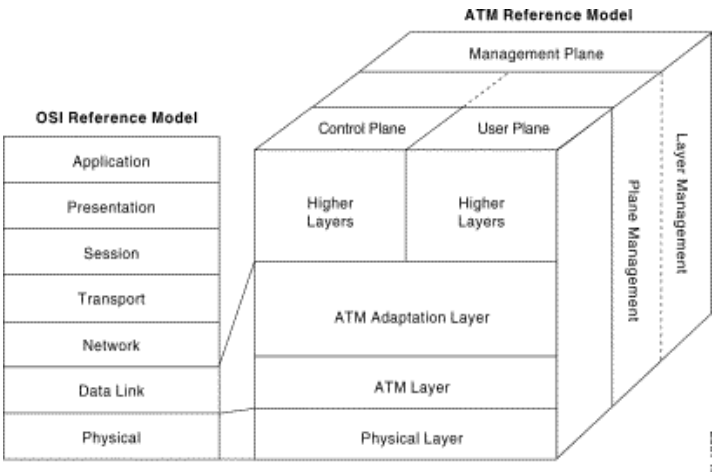

Figure 1:Relating layers of OSI and ATM model.

The ATM reference model as shown in figure 1,is composed of the following ATM layers:

- *Physical layer*-Analogous to the physical layer of the OSI reference model, the ATM physical layer manages the medium-dependent transmission.
- *ATM layer*-Combined with the ATM adaptation layer, the ATM layer is roughly analogous to the data link layer of the OSI reference model. The ATM layer is responsible for establishing connections and passing cells through the ATM network. To do this, it uses information in the header of each ATM cell.
- *ATM adaptation layer (AAL)*-Combined with the ATM layer, the AAL is roughly analogous to the data data-link layer of the OSI model. The AAL is responsible for isolating higher-layer protocols from the details of the ATM processes.

3.1 The ATM Physical Layer

The ATM physical layer has four functions: bits are converted into cells, the transmission and receipt of bits on the physical medium are controlled, ATM cell boundaries are tracked, and cells are packaged into the appropriate types of frames for the physical medium. Examples of physical medium standards for ATM include Synchronous Optical Network/Synchronous Digital Hierarchy (SONET/SDH), DS-3/E3, 155 Mbps over multimode fiber (MMF) using the 8B/10B encoding scheme, and 155 Mbps 8B/10B over shielded twisted-pair (STP) cabling.

3.2 ATM Adaptation Layers: AAL1

AAL1, a connection-oriented service, is suitable for handling circuit-emulation applications, such as voice and video conferencing. Circuit-emulation service also accommodates the attachment of equipment currently using leased lines to an ATM backbone network. AAL1 requires timing synchronization between the source and destination. For this reason, AAL1 depends on a medium, such as SONET, that supports clocking. The AAL1 process prepares a cell for transmission in three steps. First, synchronous samples are inserted into the Payload field. Second, *Sequence Number* (SN) and *Sequence Number Protection* (SNP) fields are added to provide information that the receiving AAL1 uses to verify that it has received cells in the correct order. Third, the remainder of the Payload field is filled with enough single bytes to equal 48 bytes.

3.3 ATM Adaptation Layers: AAL2

Originally AAL2 was intended to support a variable-data-rate bit stream, but it has been redesigned. It is now used for low-bit-rate traffic and short-frame traffic such as audio (compressed or uncompressed), video, or fax. A good example ofAAL2 use is in mobile telephony. AAL2 allows the multiplexing of short frames into one cell. The CS layer overhead consists of five fields:

- Channel identifier (CID). The 8-bit CID field defines the channel (user) of the short packet.
- Length indicator (LI). The 6-bit LI field indicates how much of the final packet is data.
- Packet payload type (PPT). The PPT field defines the type of packet.
- User-to-user indicator (UUI). The UUI field can be used by end-to-end users.
- Header error control (HEC). The last 5 bits is used to correct errors in the header.

The only overhead at the SAR layer is the start field (SF) that defines the offset from the beginning of the packet.

3.4 ATM Adaptation Layers: AAL3/4

Initially, AAL3 was intended to support connection-oriented data services and AAL4 to support connectionless services. As they evolved, however, it became evident that the fundamental issues of the two protocols were the same. They have therefore been combined into a single format calledAAL3/4.AAL3/4 provides comprehensive sequencing and error control mechanisms.AAL3/4
prepares a cell for transmission in four steps. First, the *convergence sublayer* (CS) creates a *protocol data unit* (PDU) by prepending a beginning/end tag header to the frame and appending a length field as a trailer. Second, the *segmentation and reassembly* (SAR) sublayer fragments the PDU and prepends a header to it. Then, the SAR sublayer appends a CRC-10 trailer to each PDU fragment for error control. Finally, the completed SAR PDU becomes the Payload field of an ATM cell to which the ATM layer prepends the standard ATM header.

3.5 ATM Adaptation Layers: AAL5

AAL5 is the primary AAL for data and supports both connection-oriented and connectionless data. AAL5 also is known as the simple and efficient adaptation layer (SEAL) because the SAR sublayer simply accepts the CS-PDU and segments it into 48-octet SAR-PDUs without adding any additional fields. AAL5 prepares a cell for transmission in three steps. First, the CS sublayer appends a variable-length pad and an 8-byte trailer to a frame. The pad ensures that the resulting PDU falls on the 48-byte boundary of an ATMcell. The trailer includes the length of the frame and a 32-bit cyclic redundancy check (CRC) computed across the entire PDU. This allows the AAL5 receiving process to detect bit errors, lost cells, or cells that are out of sequence. Second, the SAR sublayer segments the CS-PDU into 48-byte blocks. A header and trailer are not added (as is in AAL3/4), so messages cannot be interleaved. Finally, the ATM layer places each block into the Payload field of an ATM cell. For all cells except the last, a bit in the *Payload Type* (PT) field is set to zero to indicate that the cell is not the last cell in a series that represents a single frame. For the last cell, the bit in the PT field is set to one.

IV. THREATS

As other networks, ATM networks will suffer a lot of threats. Typical ones are eavesdropping, spoofing, service denial, VC stealing and traffic analysis etc. Notice that VC stealing and traffic analysis happen only in ATM networks.

4.1 Eavesdropping

Eavesdropping refers to the threat that the attacker connects or taps into the transmission media and gain unauthorized access to the data. It is one of the most common attacks to the network. Since most ATM networks are connected with optic cables, some people might get the wrong impression that is not so easy to tap a ATM network.

4.2 Spoofing

Spoofing attack means that an attacker tries to impersonate another user to the third part therefore can get access to resources belonging to the victim to take advantages or just destroy them. Spoofing might need special tools to manipulate the protocol data unit. And sometimes it might require the attacker has special access permission, say,must be the super user in UNIX environment. However, since a network will be connected to many untrusted networks via the Internet, it's impossible to prevent a hacker from getting this access permission or even trace the people with this particular access permission. ATM is being implemented in public domain. Therefore, it is subject to this kind of attack also.

4.3 Service Denial

ATM is a connection-oriented technique. A connection, which is called Virtual Circuit(VC) in ATM, is managed by a set of signals. VC is established by SETUP signals and can be disconnected by RELEASE or DROP PARTY signals. If an attacker sends RELEASE or DROP PARTY signal to any intermediate switch on the way of a VC,then the VC will be disconnected. By sending these signals frequently, the attacker can greatly disturb the communication between one user to another, therefore will disable the Quality of Service(QoS) in ATM. Combining this technique with other tricks like eavesdropping, the attacker can even completely block one user from another.

4.4 Stealing of VCs

If two switches in an ATM network compromise, the attacker can even steal a VC from another user. Say VC1 and VC2 are two virtual channels which will go through switch A and switch B. VC1 is owned by user U1 and VC2 is owned user U2. If A and B have compromised, then A can switch VC1's cells going from A to B through VC2 and B will switch back those cells to VC1. Since switches will forward cells based on the VCI(Virtual Channel Identifier) or VPI(Virtual Path Identifier) in the cell header, A and B can just alter these fields back and forth. Switches between A and B won't notice these changes and will switch the assumed VC2's cells just like the authentic VC2's cells. In public packet-switching network, U1 won't gain too much by this trick. However, in ATM network, if quality of service is guaranteed, then user 1 can gain a lot by stealing a higher quality channel which user 1 is not entitled to use according to the access control policy. User 1 can gain even more if every user has to pay for the communication. In both cases, user 2 will be hurt. Someone maybe argues that the possibility that the switches will compromise is pretty low. That will true if the ATM network is owned by one organization. when we consider ATM internetworking, in which case cells will travel through different ATM networks, it will be very easy for two switches to compromise.

4.5 Traffic Analysis

Traffic analysis refers to a threat that the hacker can get information by collecting and analyzing the information like the volume, timing and the communication parties of a VC. Volume and timing can reveal a lot of information to the hacker even though the data is encrypted, because encryption won't affect the volume and timing of information. And also the source and destination parties can be obtained from the cell header (normally is in clear text) and some knowledge about the routing table. Another related threat is called convert channels. In this technique, the attacker can encode the information in the timing and volume of data, VCI, or even session key to release information to another people without being monitored. Normally, these two attack won't happen. However,when ATM is used in a environment requiring stringent security, it might happen.

V. SECURITY FRAMEWORK

People have practiced security for a long time. In the past, security services were considered only after the network service was totally designed. These ad hoc approaches turn out to be unsatisfactory. ATM Forum tries to avoid such pitfalls by considering the security as one integrated part of ATM. Recently, ATM Forum Security Working Group proposes a draft of Security Framework for ATM to address the basic requirements for ATM security. Main security objectives for ATM security:

- Confidentiality
- Data Integrity
- Accountability
- Availability

Confidentiality and data integrity are obvious. Accountability means that all ATM network service invocations and network management activities should accountable. And any entity should be responsible for the actions it initiates.Accountability includes both authentication and non-repudiation. It is extremely important for operators to manage the system and bill the services. Availability means all legitimate entities should be able to access to ATM facilities correctly, no service denial should happen. That is important for QoS operation. According to these main objectives, the draft proposes the principal functions which a ATM security system should provide:

- **Verification of Identities:** Security system should be able to establish and verify the claimed identity of any actor in an ATM network.
- **Controlled Access and Authorization:** The actors should not be able to gain access to information or resources if they are not authorized to.
- **Protection of Confidentiality:** Stored and communicated data should be confidential.
- **Protection of Data Integrity:** The security system should guarantee the integrity of the stored and communicated data.
- **Strong Accountability:** An entity can not deny the responsibility of its performed actions as well as their effects.
- **Activities Logging:** The security system should support the capability to retrieve information about security activities in the Network Elements with the possibility of tracing this information to individuals or entities.
- **Alarm reporting:** The security system should be able to generate alarm notification about certain adjustable and selective security related events.
- **Audit:** When violations of security happen, the system should be able to analyze the logged data relevant to security.
- **Security Recovery:** The security system should be able recover from successful or attempted breaches of security.
- **Security Management:** The security system should be able to manage the security services derived from the above requirements.Among the ten requirements, the last two won't provide security services. However, they are necessary to support the maintenance of security services. If the security system can not be recovered from attacks and can not provide security services any more, then the system won't be secure after these attacks. On the other hand, security services and the information about security have to be managed securely. They are the foundations of the security system.

VI. CONCLUSION

ATM technology perhaps is the most complex networking technology we ever have. To secure such a complex system will be even more difficult than design it. And now people just begin to discuss some issues of ATM security. It will take times for us to figure out how to completely achieve our security objectives. Because the goal of ATM is to provide a unified networking platform and communication infrastructure, ATM security , as a part of this infrastructure, has to be flexible and compatible with other technology. That will introduce more difficulties to ATM security. ATM Forum Security Working Group has come up with drafts for security specification and security framework. A lot of other security issues have been discussed in ATM Forum.ATM is capable of transporting multiple types of services simultaneously on the same network. All data is placed in cells of uniform size. The cell header contains information concerning cell routing using VCI's and VPI's. Cells from various applications with the same destination can be interleaved to share physical facilities. This allows network providers to transport different types of services using the same physical facilities. This is an advantage for network providers in that facilities can be fully utilized. It is an advantage for end users since they can connect their various networks and only pay for the data they are sending.

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Software Development Using Agile Methodology Using Scrum Framework

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***Abstract*-** Agile methodologies are enhancement in SDLC with the aim to provide more efficientsoftware on time. In this, where the universally accepted AGILE Methodology calledSCRUM came into existence, almost the 65% of the industry moved to SCRUM model,from the existing regular Software Life Cycle Model. SCRUM assumes that the systemsdevelopment process is an unpredictable, complicated process that can only be roughly described as an overall progression. SCRUM defines the systems development process as aloose set of activities that combines known, workable tools and techniques with the best thata development team can devise to build systems. Since these activities are loose, controls tomanage the process and inherent risk are used. SCRUM is an enhancement of the commonlyused iterative/incremental object-oriented development cycle.

***Index Terms*-** AGILE Methodologies, Product Backlog, SCRUM, SCRUM Frame work,SPRINT.

I. INTRODUCTION

TheSCRUM based software development life cycle is the latest trend in the industry where thereWill be no hierarchy for the process and the development will be done iteratively by interactingWith the people and the concept of conventional life cycle process is not followed. Majorly SCRUM deals with individual interaction the process and tools which we use for a normal life cycle process, customer collaboration, responding to the changes immediately..

SCRUM INTRODUCTION

Software Development Life Cycle Processes can be achieved by different process models and methodologies which are existing in the market. The main aim of these AGILE Methodologies is reduce risk by breaking projects into small, time-limited modules or time boxes ("iterations") with each iteration being approached like a small, self-contained mini-project, each lasting only a few weeks. The major difference which we can make between the traditional SDLC and the AGILE methodology is customer collaboration, immediate response to the change. We have different types of methodologies and Frameworks in AGILE like Dynamic systems development method (DSDM), Rapid Application Development (RAD) & SCRUM. Where SCRUM is a process that allows us to focus on delivering the highest business value in shortest time in this process the development team works as a unit to reach a common goal as opposed to a "traditional, sequential approach". There are 3 core roles in SCRUM like “Product owner” “Development Team” & “scrum master”. The ancillary role in the SCRUM team is taken care by stake holders and the managers. The SCRUM process is like first the PLAN is estimated and the product is INSPECTED after the completion of the Inspection the product is materialized and the SPRINT is adapted. The SPRINT is piece of work which is executed is given particular time(like 2 - week 3 week) the features which are to be implemented in the given application are divided in to number of task and each task is categorized into SPRINTS in turn these sprints are developed into a project according to the needs of the customer.

II. SPRINT

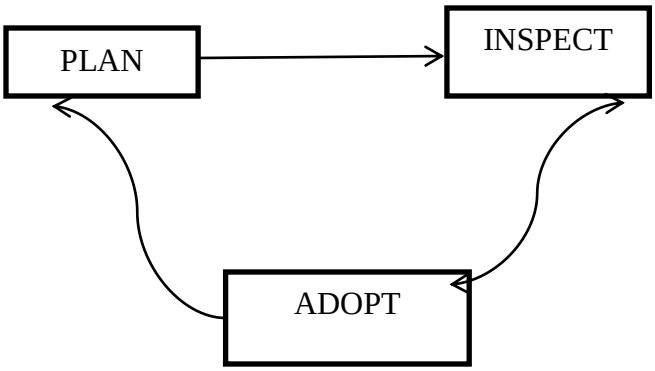
Sprint is a piece of work done with in a stipulated time like Two – Four weeks. Each sprint is preceded by a planning meeting, where the tasks for the sprint are identified and an estimated commitment for the sprint goal is made, and followed by a review or retrospective where the progress is reviewed and lessons for the next sprint are identified. Sprint is done with the help of product backlog where we find all the requirements for the execution of SPRINT. Product owner inform the team member the items to be identified and developed accordingly he also gives the priority to the list of items to be developed. This sprint backlog is the property of the development team where no one can come and edit the backlog only the sprint team/development team can do the changes and these changes are to be done in a fixed time or in time box. Once if the development is done the team demonstrates or trains the end user how to work on the product.

The best process SCRUM does is if self organizes the teams by encouraging verbal –communication between all the team members and good discipline in the project. KEY principle of the SCRUM is if any customer changes his mind during the development process and asks for any changes they are ready for accepting it but in conventional or traditional manner it is not acceptable. For this SCRUM adopts an empirical approach by responding & delivering to the product backlogs quickly

III. SCRUM PROCESS

The SCRUM approach assumes that the analysis, design, and Development processes in the Sprint phase are unpredictable. A control mechanism isused to manage the unpredictability and control the risk. Flexibility, responsiveness, andreliability are the results. Sprints are categorized like SPRINT-1 SPRINT -2 AND SPRINT-3 …….. ETC

SPRINT-1: Potentially Shippable: FEATURE A
SPRINT 2: FEATURE A + FEATURE B
SPRINT 3: FEATURE A + FEATURE B + FEATURE C.



SCRUM TEAMS are cross functional where these mainly deals with product backlogs by regularly checking the updates and building up the product .SCRUM MASTER plays important role with 3-P's
1. Process owner
2. Problem solver
3. Protector
But Scrum Master is not considered as Project Manager but he acts as barrier between the team and product owner. Scrum Master's role is to protect the Development Team and keep it focused on the work done by the development team..

IV. SCRUM VALUES

The Five Scrum Values:
1. Focus – "Concentrate all your thoughts upon the work at hand. The sun's rays do not burn until brought to a focus." – Alexander Graham Bell

Team focus is the domain of the Scrum Master. The SM removes work impediments to the Team, shields them from external influence and is responsible for making the Team fully functional and effective. The nature of Scrum means that the PO aids the focus of the Team by making sure that all work is prioritized in a backlog. Finally the Team must be focused on finishing the sprint User Stories while adhering to the Definition of Done.

2. Courage – "Fortes fortunaadiuvat – fortune favours the brave" – Latin proverb

The SM needs the courage to protect and guide the Team. Standing up to the PO and Stakeholders at the right time, really takes guts. The PO must have the courage to entrust the Sprint Backlog to the Team, a giant leap of faith as it is the PO who answers to the Stakeholders at the end of the sprint. Finally the Team must have the courage to aggressively commit to as much work as they think they can do each sprint.
3. Openness – "It is impossible for a man to learn what he thinks he already knows." – Epictetus

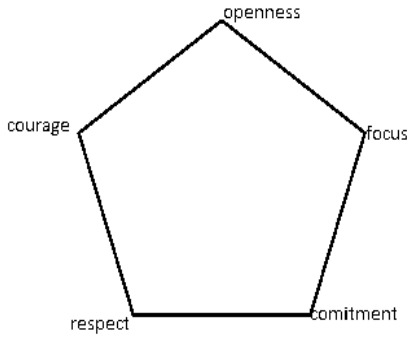
The PO must be open to accepting change, alternatives and new ideas, both from the Team and Stakeholders. By providing a qualified backlog with priorities and value, the PO is transparent about what is coming up next and the Team knows what to expect. The Team needs to be open to find the best solution to any problems from within. Scrum also pushes openness with the Retrospective Meeting, where any problems are pushed to light and dealt with in an open environment.

4. Commitment – "Do, or do not. There is no try." – Master Yoda

The whole scrum process is a commitment to a new way of working, to be more adaptable. The Team commits to what they will do each sprint by choosing the Sprint Backlog and they also commit to how the work will be 'done' in the Definition of Done. This means the Team commits to doing whatever is necessary in order to meet their goals. The SM commits to actively guiding the Team and takes a weight of responsibility in making the Team adhere to the Scrum process. The PO commits to having a certain fraction of his Product Backlog ready for the Stakeholders every sprint, and also commits on the priorities of what the Team will do in each sprint.

5. Respect – "I speak to everyone in the same way, whether he is the garbage man or the president of the university." – Albert Einstein

In Scrum, the limits and boundaries of the Scrum roles really need to be transparent, and respected. Everyone on a scrum project needs to be aware that the PO is in charge of what the Team works on, but not how they do their work, and that the Team is responsible for getting the work done, but not questioning what work gets done. The SM also needs to be aware that though he has more responsibility than a Team member, he is an equal member of the Team, and not a leader. In the ideal case, the SM is a gentle shepherd, or quiet guide to the Team.



V. TESTING

Testing itself has more important and immediate role in Scrum based Software development process? Testing isn't just providing results. Testing is important part of the Software

development process. In this it will provide feedback to the development, testing challenges supports and finally verifies developers work.

DoD (Definition of Done): The feature is done when code is ready, testing is done, no regression exists, and feature is ready to be released

Popular Testing methods in SCRUM Framework are

1. Regression testing
2. Exploratory testing
3. Automated testing
4. Integration testing
5. NFT (non-functional testing)

VI. CONCLUSION

The Scrum Frame work offers a high degree of flexibility promises a high probability of success. The major benefits are an anticipating cultureincreases, the sense of urgency promotes the sharing of knowledge encourages dense communications facilitates honesty among developers.

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Review on Parking Brake Lateral Play in Four Wheeler

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Abstract- A parking brake (PB) system is a type of mechanical brake-by-wire system that is the conventional lever parking system by generating a clamping force for parking using lever system. At the push of a button, a driver can easily apply or release the parking brake; this enables elderly or disabled persons to easily apply a full braking load. The PB system operates quickly and over a wide force range through the use of electrical components It is sometimes also used to prevent a vehicle from rolling when the operator needs both feet to operate the clutch and throttle pedals. Automobile hand brakes usually consist of a cable directly connected to the brake mechanism on one end and to a lever or foot pedal at the driver's position. The mechanism is often a hand-operated lever (hence the hand brake), on the floor on either side of the driver or a pull handle located below and near the steering wheel column, or a (foot-operated) pedal located far apart from the other pedals. Although sometimes known as an emergency brake, using it in any emergency where the footbrake is still operational is likely to badly upset the brake balance of the car and vastly increase the likelihood of loss of control of the vehicle, for example by initiating a rear-wheel skid. Additionally, the stopping force provided by using the handbrake is small and would not significantly aid in stopping the vehicle.

Index TermsBy compensating for the inertia effect through the novel on–off control method, our control logic can be implemented using a simple control unit without a PWM driver. After whole study on parking brake system we will work on lateral play in parking brake which is requirement of customer we are facing problem in the parking brake system regarding the lateral plays requirements. Customerdesires the lateral play in parking brake (at 0th notch and at 11th notch of ratchet) should be in the specified limits. For that they want to modify the existing assembly of parking brake.

I. INTRODUCTION

The most common use for a parking brake is to keep the vehicle motionless when it is parked. Parking brakes have a ratchet locking mechanism that will keep them engaged until a release button is pressed. They are recommended always to be left with the handbrake engaged, in concert with their lowest gear (usually either first or reverse). It is operated by pushing the lever down with one's hand to apply the brake, and pulling it upwards to release it. However, this has been known to cause severe back problems in drivers who do this regularly and many choose to push it up with their feet. Some cars with automatic transmission are fitted with automatically releasing parking brakes. For Large vehicles are usually fitted with power operated or power assisted handbrakes. Power assisted handbrakes are usually found on large vans as well as some older heavy vehicles. These operate in the same way as a conventional handbrake, but pulling the lever will operate a valve that allows air or hydraulic pressure or vacuum into a cylinder which applies force to the brake shoes and makes applying the handbrake easier. A recent variation is the electric parking brake. It is expected that these systems will incorporate other features in the future. BMW, Renault, already have a system where the emergency brake initiates when the car stops and then goes off as soon as the gas pedal is pressed preventing the car from rolling. The new feature is called a hill hold. The vehicle operator can easily turn off the system. However, this method requires an additional electric circuit to drive a DC motor here use an on–off control method for the force control logic to supply the maximum voltage until the clamping force reaches the desired final force. Using this method, the DC motor Continues to rotate after the power is cut-off due to its momentum, resulting in an excessive clamping force. For functions that need fast or repeated apply-release operations such as anti-lock brake systems (ABS) and drive-away release, the excessive clamping force may cause a longer release time. Thus, the excessive force caused by the inertia effect should be compensated for.

The one of the up growing industry in developing and producing brake systems and body parts for their valuable customers such as Tata Motors, General Motors, Volkswagenetc. Company is facing problem in the parking brake system regarding the lateral play requirements. They are unable to meet the following requirements of the customer.

A load of 4.5 N to be applied laterally from both left to right and right to left sides at 40 mm grip point ‘G’ when the lever is at initial position and the lateral play should be within 5 mm.

i. A load of 222 N to be applied laterally from both left and right sides at 40 mm grip point ‘G’ when the lever is at 75% full apply means 11th notch (consider initial position as 0th notch) and the lateral play should be within 20 mm. (While perform this test do not connect the rope to the lever)

Now it is the time to articulate the research work with ideas gathered in above steps by adopting any of below suitable approaches:

In this approach combine all your researched information in form of a journal or research paper. In this researcher can take the reference of already accomplished work as a starting building block of its paper.

Jump Start

This approach works the best in guidance of fellow researchers. In this the authors continuously receives or asks inputs from their fellows. It enriches the information pool of your

paper with expert comments or up gradations. And the researcher feels confident about their work and takes a jump to start the paper writing.

There are numbers of software available which can mimic the process involved in your research work and can produce the possible result. One of such type of software is Matlab. You can readily find Mfiles related to your research work on internet or in some cases these can require few modifications. Once these Mfiles are uploaded in software, you can get the simulated results of your paper and it easies the process of paper writing.

As by adopting the above practices all major constructs of a research paper can be written and together compiled to form a completeresearch ready for Peer review.

Fig.1 parking brake G'Point

Customer desires the lateral play in parking brake (at 0th notch and at 11th notch of ratchet) should be in the specified limits. For that they want to modify theexisting assembly of parkingbrakeAccording to the policy of the company, they have permitted for the sponsorship for this project to solve the above problem for their customer. In this project, the modification of parts will be carried out by undergoing different stages of design and analysis. The FEA analysis software for simulation will be used during the project work. Validation and testing through available test rig will be carried out at the industry. The existing 3D model of parking brake which is already used in customer vehicle having problem in lateral play of brake which affect on performance of braking and ergonomics aspect during operation. Existing 3D model is created in catiav5 (we can also make the model in Pro-E). Model and actual part snap of this part shown below is checked for lateral play in both side and which is not meeting customer's requirement. We have taken this part as a base model for solving the customer requirement.

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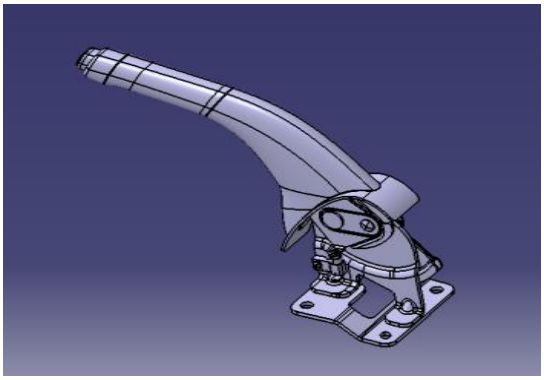

Fig.2 3D & Actual Part model for the existing parking brake assembly

The existing assembly has been tested for the above requirements of the customer by using test rig shown in the fig.4which is able to measure the lateral play for the engagement of each teeth of the ratchet with the lever of the parking brake for certain applied load during each engagement of teeth. Lateral Play test rig consists of load cell, sensor and digital force display unit.

II. DMU AND ANALYSIS

DMU it is digital mock up in which we can check the model with main assembly. In this, the designed model will be checked in software for any fouling with other components and fitment of the part in mechanism stage from that we can get theoretical idea about the operation and expected result of part or the checking possible clearance is available or not after completion of total checking through DMU.We can move further for the analysis of design to the different load condition in Ansys or any other analysis software for simulation results will give the optimized design. Validation will be carried out by checking the samplepiece on test rig.

III. EXPERIMENTAL ENVIRONMENTS

This topic focuses on a low cost and simple mechanism for both Force estimation and control of a power brake system. The importance of force sensor is to take measurement of initial contact point (where the force starts increasing) Measuring the

lateral displacement of brake in both side displacements will be measured. So need to propose an initial contact point detection method. They demonstrate through theoretical analysis that contact occurs when the angular velocity of the DC motor reaches its maximum value. The clamping force can then be estimated as a function of the effective angular displacement, of lever from the initial contact point. According to requirement of load application we can give the load to the brake for movement of brake lever in angular displacement at that time we can get displacement readings with help of force sensor for that particular load applied through the load cell system it will shown on the digital display screen it is connected to the controller to control the displacement of brake in lateral for the given control load such set development is needed for checking the lateral play in parking brake assembly For low-cost control, we use a simple on–off control, which applies the maximum input signal until the estimated clamping force reaches the target force for braking force application. The input signal becomes zero when the estimated force reaches or exceeds the target force.

Where u(t) is the input voltage for the motor, umax is the maximum voltage of the battery, and e is the error between the target force, and the estimated clamping force,

This topic focuses on a low cost and simple mechanism for both Force estimation and control of a power brake system. The omission of a force sensor leads to the problem that the initial contact point (where the clamping force starts increasing) between the brake pads and the brake disk cannot be sensed. Measuring the displacement is not sufficient to complete clamping force estimation without knowledge of the brake pads' So need to propose an initial contact point detection method using only the angular Velocity of the DC motor. They demonstrate through theoretical analysis that contact occurs when the angular velocity of the DC motor nears its maximum value. The clamping force can then be estimated as a function of the effective angular displacement, from the initial contact point.

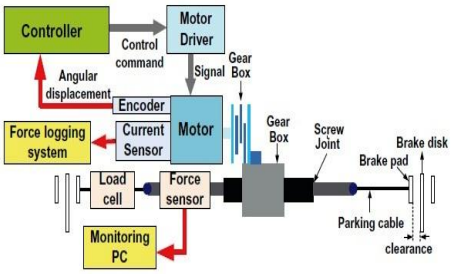

Fig.3 General Layout of lateral play checking set up.

IV. EXPERIMENTATION

a) Initial Reading

The brake is fitted in test rig. for the purpose of checking the lateral play, it is properly fitted in instrument and ready to take the load this instrument consist of sensor for measurement of lateral displacement of brake and the load cell gives the required load to the lever operation, with the help of digital display we can seen the actual readings of load and displacement of brake laterally.

Below picture shows the test rig set at initial reading.

Fig.6 Initial & Final Reading set up

b Final Reading

At 11th notch (i.e. 222 N force and 75% full apply load), the reading shown on display screen for the lateral play at one side is 15.5.mm. Similarly, for other side, the reading was 18 mm. cumulative lateral play is 15.5+18=33.5 mm, which is not met the requirement of the customer (i.e. 20 mm).

c Readings for lateral play

Table 1 Readings for play checking on lateral play test rig

Part name & Auto line part number E2 0048/49 G000001	**sample Number LP-1**	**Measured Value LP-2**		**Pass or Fail**
1		18.4	15.5	Fail
2		18	15.4	Fail
3		18.4	15.4	Fail

LP1: Force applied from left to right side ,LP2: Force applied from right to left side.Above mentioned readings were taken when customerwas given requirement for lateral play limitation but the readings were not met the 20mm lateral play condition at 222N load which shows fail readings, so need to study and workout on this existing brake assembly, we will take it as a base assembly for further modification after modification

we can take readings on same test rig in same way and compare with requirement of customer.

V. RESULTS AND DISCUSSIONS

To compare the estimated forces with the measured forces, which are obtained from the force sensor embedded in the module of the system,. We repeated the application and release 30 times for five different target forces using the two on–off control methods (the simple on–off and the novel on–off). The negative error indicates that over-clamping force was applied. [14]The differences in the deviation of the error are due to mechanical vibrations of the test bench. We observed that the misalignment between the gear and screw could cause vibrations. These vibrations appear in the current and force sensors signals as low frequency signals. As the braking force increased, this vibration and the error deviation also increased. However, no significant vibration was observed in the Manufacturing Verification Test (MVT) PB systems. Therefore, use of our novel on–off controller in manufactured parking brake systems could provide more regular and decreased error deviation compared to that in the test Rig.

VI. FUTURE PLAN OF ACTION

Parking brake system is very essential for the safe parking of any vehicle and also used as a emergency brake at critical conditions. Lateral play is the factor which affects the performance of the parking brake system. Hence, it is required to design the parking brake system such a way that it should not give more lateral play than the desired. In this project, the emphasis will be on the following requirements of the Auto line Customer

1. To maintain the lateral play within specified limit of 20 mm (cum.).
2. By using the data available with the existing model of Auto line and its results, we need to change the design for the parking brake assembly.
3. The methodology may involve DFMEA and FEA analysis.
4. The available test rig facility will be used to measure the lateral play for the further design modification.

In the further stage of the project, we will be focusing on the preparation of DFMEA and Design modification. In order to satisfy the customer's requirement, we need to modify the basic structure of the product and to carry out such modifications; we need to depend on trial and error basis. These changes in design will be carried out in 3D model created in either CATIA v5 or Pro-E.

The modified 3D models will be analyzed by using FE simulation software like Ansys. Based on the results obtained from the simulations, the design parameters will be finalized.
Sample piece will be produced based on the optimized design and the experiments will be carried out on the sample piece with the help of lateral play test rig. While carrying out these experiments, the lateral play for the position of 0^{th} notch and 11^{th} notch of ratchet will be checked and will be compared with the customer's expected lateral play.

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Canal Rehabilitation Focus Group Discussion - A special study on Purba Putiary Kolkata, Borough: XI/ Ward: 114, Venue: Purba Putiary Upohar Abashan , Kolkata

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Abstract- The area which has been assessed is Purba Putiary Unnoyon Abashan Samaboy Samiti Ltd. The FGD was conducted in the Purba Putiary Kundghat Uopohar Abashan which consists of 192 single room flats divided into 6 blocks namely A, B, C, D, E and F respectively. The people mainly belonging from Hindu community and are primarily Bengali. The Rehabilitated dwellers are mostly belonging from areas namely Enny Sarani, Thakurpukur Cancer Hospital area and Keora Pukur Khalpar area as well as from Kabardanga area. According to the committee of this abashan , 40% of the dwellers came from Enny Sarani / 30% Thakurpukur Cancer Hospital area / 30 % Keora Pukur Khalpar area.

Index Terms- Purba Putiary, Single Room Flats, Rehabilitation of Dwellers, Change in Living

I. KNOWLEDGE OF KEIP

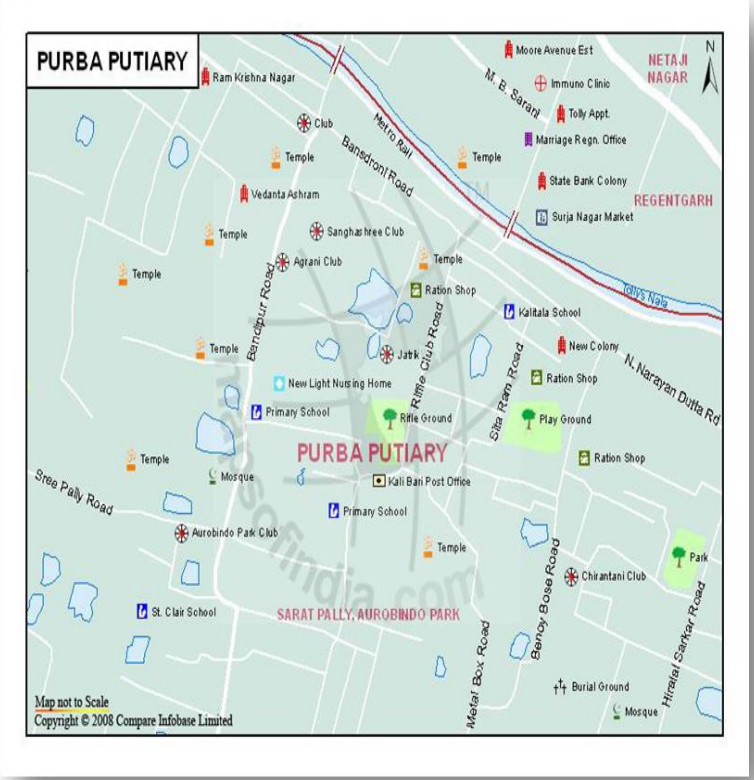
The Kolkata Poribesh Unnayan Prakalpa commonly known as KEIP was initiated to work for the resettlement of the Canal Dwellers whose initial work was started in around 2008 and still is in the process of completion. The people were known about the issue for their resettlement from the Canal Area namely Kheora Pukur Khal and accordingly Genuine Card holders (persons below poverty line) were substituted with these flats.

" According to the Abashan Secretary Mr. Amitava Dutta , the KEIP work was pre informed to them and the entire settlement costs and building construction was borne by the authority of KEIP , the residents had to pay a nominal Rs 5000/ during the stretched time allotted by the KEIP body."

The people came to know regarding the project of Canal Rehabilitation with the general news of the nominal money deposition to the Kolkata environmental improvement project body. The ladies were also not aware of the developments as a result was in a very mixed situation. However with the initiation of the work in the form of single room cooperative society's people became more aware about their living strategies for near future and KEIP played a major role in their part.

II. METHOD OF RESEARCH

The research work was executed in the form of focus group discussion as well as household surveys in the form of social assessment of the KEIP's work implemented in Kolkata. Social factors like gender, household conditions as well as income were mostly viewed for the conclusion of the research. It has been supplemented with adequate diagrams, case studies, verbal lectures of the common residents of the studied area. Major living necessities like urban conditions ,roads drainage has also been taken into consideration.



III. LOCATION MAP

The major occupation of the people are mainly small individual grocery shops, stationary shops small vehicle repairing shops or Cycle repairing shops as well Rickshaw Pullers and people engaged in small services are also found. The area is very well connected to the general amenities and easily accessible to the Metro Rail Service (Netaji metro station) as well Schools, Colleges and Major Hospitals like M.R Bangur which signifies the apt location of the resettled Canal Dwellers.

The super built area of the flat is 225sq ft where as 190 sq ft is the paper area.

IV. Situation before Pre KEIP

The major resettled Canal dwellers belonged from the places namely Enny Sarani , Kheora Pukur Khal , Kabardanga and Thakur pukur Cancer Hospital Area . They shifted the original location mostly 3-4 years ago due to the renovation of the Canal namely the Khaora Pukur Khal and the Khal that joins the Diamond Harbor Khal. The initial processing was started by KMC which were handed off to the KEIP body. The house types were basically semi pucca types and kuccha jhupri types. The housing materials were mainly plastic, roof was made by cement in some cases (parish). The source of drinking water was the nearby Tube wells and wells and they had to wait in long queue for the daily collection of the waters. They basically used Tube well waters for bathing as well as for household purposes. There was a huge problem in terms of the access to basic facilities like electricity, Toilets, solid waste disposal, drainage facility etc. The electric connection was provided to them and basic health care facility to access to schools, markets was very problematic in the previous area (according to the respondents).

Photos of the Community Society of Kolkata – A case study on Purba Putiary

V. Situation of the Area (Post KEIP)

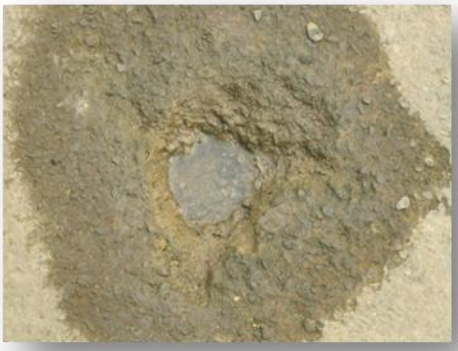
The dwellers were settled to the Purba Putiary Kundghat Upohar Abashan Samiti around 3-4 years ago based on the general BPL card holder facility of the mentioned renovation canal area. The respondents are very satisfied with the present development of the area, facilities are provided easily, proximity to the schools, Bus stand, Metro Rail Service, Hospitals, Auto Services, Market Services are the prominent ones. The respondents have their voter ID card and Ration Card also. According to them " We are lucky that we got a flat in a prime area only against Rs 5000/ ".

There has been a gross transformed lifestyle change in the area. The respondents are now quite happy with their present place of residence since they are getting the urban basic facilities at a very low cost and joint living under one Abashan also helps the people to look after their daily problems. However with the interactions between the respondents some problems are found which are discussed namely

- The safety security has been better but still a lot has to be done since there are no street lights facilities inside the cooperative/ Abashan area. The women and children feel very insecure to move around the night in the premise as well as outside it.
- The drainage system of the flats is very poor and no initiative has been taken in this regard which leads to the overflowing of the septic tanks after 1 year which is cleaned by the flat owners.
- There is a chronic problem of portable drinking water since the one which is supplied to the flat owners through the submersible pump is not at all suitable for drinking purpose and people have to access the nearby tube well for having portable drinking water. There is only one tap in Bathroom and one in the verandah for household purpose.
- 32 people chamber outlet is not defined and the existing chamber is not suitable for the mass sewerage wastes. Block E has back pond outlet, Block B has high Drain Outlet and Block D wastes overflows its own chamber. Sewerage system to be looked very minutely since a major problem of the area.
- Community Hall has not been handed to the committee people; there has been appeal for 2 rooms, Toilet facility. Electric Meter has not been installed where as electric wire has been.

- Cracks are observed all over the buildings and the roofs are totally damaged along with imbalance slope in the bathroom

area. Roof Parish Repair is a must.



- The concept of stealing and thieves are very common social problem in the area.
- Block A, B and C does not have working Mother Meter. As a result 32 flat owners in each block but only some people have been allotted sub meter and others take electricity from them as a result bill charge have to be borne by the borne Mother Meter Blocks.

Rent and selling of flats is very common and the unclaimed flats (59) are not locked due to the political reasons and are a place of nuisance and illegal activities at times.

" Pressure exists on the residents and they cannot control the adverse situation due to the political influence and physical people have been hackled by the political groups said Mr. Liton Das, President and Amitava Dutta, Secretary of the Abashan Committee."

They have also mentioned that, by phase's maintenance Fund Allocated in 3 phases by KEIP in the ratio 4:4:2. But only part of the total amount is received by the respondents (Rs 47000/)

VI. A Special Study (On Vocational Training)

BOROUGH - 11 / WARD 114.

The FGD was conducted in the Purba Putiary Kundghat Uopohar Abashan.

The Kolkata Poribesh Unnayan Prakalpa commonly known as KEIP was initiated to work for the resettlement of the Canal Dwellers whose initial work was started in around 2008 and still is in the process of completion. The people were known about the issue for their resettlement from the Canal Area namely Kheora Pukur Khal and accordingly Genuine Card holders (persons below p The major resettled Canal dwellers belonged from the places namely Enny Sarani , Kheora Pukur Khal , Kabardanga and Thakur pukur Cancer Hospital Area . They shifted the original location mostly 3-4 years ago due to the renovation of the Canal namely the Khaora Pukur Khal and the Khal that joins the Diamond Harbor Khal. The initial processing was started by KMC which were handed off to the KEIP body.overty line) were substituted with these flats.

The dwellers were settled to the Purba Putiary kundghat Upohar Abashan Samiti around 3-4 years ago based on the general BPL card holder facility of the mentioned renovation canal area. The respondents are very satisfied with the present development of the area, facilities are provided easily, proximity to the schools, Bus stand, Metro Rail Service, Hospitals, Auto Services, Market Services are the prominent ones.

There has been a gross transformed lifestyle change in the area. The respondents are now quite happy with their present place of residence since they are getting the urban basic facilities at a very low cost and joint living under one Abashan also helps the people to look after their daily problems.

It was also a part of the responsibility of KEIP that to organize some vocational courses for the dwellers to rehabilitate them properly. In this context , KEIP has organized 3 type of vocational course in Purba Putiary kundghat Upohar Abashan Samiti.

1) Beautician course (for female dwellers.) (1yr)
2) Mobile repairing course, (both male and female.) (1yr)
3) Automobile repairing course (for male.) (1yr)

But only the females has joined in the Beautician course and two others were not materialized here. For mobile repair and TV repair , the trainees has to go outside and far away from the locality to take the course. So all of the dwellers were denied to take those two courses. Only the beautician course was organized inside the premises and total 11 female trainees has received the course. All were women ranging from 14- 25 yrs respectively. The training was divided into 2 phases. In the first phase girls were called and all those who were interested were allots for a basic training for 6 months and those who were interested after that were provided by an advance course for 6 months. The training was provided from Monday to Friday for a period of 2 hours. Initially it was decided that the training will be organized mainly to the Women Head Family / Handicapped Special Cases for allotment of Individual frame out. But after that it was decided by the Purba Putiary Kundghat Uophar Abashan Committee that anybody can take the training . Other trainings which were aimed to be provided were mobile repairing, automobile repairing which were not received any positive feedbacks. Training was provided to them free of costs by KEIP.

The course was started on 2010 June and finished on June 2011. The girls are now earning fair amount of

money which is a good help for their family. They took the assignment to do such job nearby their locality. According to the girls " Now most of the women are beauty mindful so it is a very good profession for us and we are getting good number of clients every day. " Some of the girls are also involved with some Beauty Parlor and doing home visits also . According to them they can earn monthly 3000/ with this service and they are very thankful to KEIP for this initiative.

- Community hall and poor drainage outlet has also been observed.
- The rooms have also been observed respectively.

VII. Local Stationary shop – (Community Society Building)

Suggestion

If KEIP can organize mobile repairing or automobile repairing course here inside the premises then many boys can take it and time should be in evening. So that boys can take this course and can continue their present occupation to carry their family properly.

VIII. Overall Conclusion

The local residents of the Abashan / Cooperative Society are demanding for certain facilities like

- Clear drinking water,
- Roof repair of the top and their flat interior walls,
- Closing of the Open drains which have led to the abrupt increase of the mosquitoes.
- There are 2 Vats one temporary and the other permanent; however around the premise of the Flats dumping of garbage's is seen.
- Another major suggestion has been the locking of the empty flats, which is the most factors according to the dwellers.

Field observation:

- The Unnnoyon abashan has been observed (photo attached).
- Cracked Roof tops have also been observed.
- Shop has been observed within the premises of the cooperative along with its Register.

Acknowledgement

We would like to offer our deepest sense of gratitude to the community mobilizers of the KEIP giving valuable suggestion and supervising the entire work. Without their help and guidance this Research Report would not have taken its present shape.

We also would offer my deepest sense of gratitude to Institute of Social Science for their heartiest co-operation.

Lastly we would also like to give a big thanks to our parents and seniors for ensuring our presence in the field focus group discussion.

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An Approach Secret Sharing Algorithm in Cloud Computing Security over Single to Multi Clouds

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Abstract- Now a day's rapidly increased use of cloud computing in the many organization and IT industries and provides new software with low cost [1]. So the cloud computing give us lot of benefits with low cost and of data accessibility through Internet. The ensuring security risks of the cloud computing is the main factor in the cloud computing environment, for example sensitive information with cloud storage providers may be entrusted. But 'single cloud' providers is a less popular with customers due to risks service availability failure and possibly of malicious insiders in the 'single cloud'. A towards movement of 'multi clouds' or 'multiple clouds' or 'cloud-of-clouds' has emerged currently using Shamir's Secret Sharing Algorithm.

This paper surveys to many running research related paper to single cloud and multi clouds security using Shamir's Secret Sharing algorithm and addresses possible solutions and methodology. Main focus of this paper use of multi clouds and data security and reduce security risks and affect the cloud computing user using Shamir's Secret sharing algorithm. It is a form of secret sharing, where a secret is divided into parts, which is giving each participant its own unique part, where some of the parts or all of them are required in order to reconstruct the secret. If we're going to Count all participants to combine together the secret might be impractical, and therefore sometimes the threshold scheme is used where any 'k' of the parts are sufficient to reconstruct the original secret [7].

Index Terms- Shamir's Secret Sharing Algorithm, data integrity, cloud storage, data intrusion, service availability.

I. INTRODUCTION

The cloud computing is a cost-effective, service availability, flexible and on demand service delivery platform for providing business through the internet [2]. Cloud computing resources can be quickly extracted and effortlessly scaled with all the processes, services and applications provisioned on demand service despite the consequences of the user location or device. Hence, the opportunity for an organization to enhance their service deliverance efficiencies is achieved through cloud computing. The issues in cloud security series from substantial security of the cloud fixing and hardware infrastructure, through the architectural security of function and data deployments, to the actual security of the cloud framework in the presence of peripheral attacks and the mechanisms accessible to respond to and recuperate from these attacks [3]. The use of cloud computing Subashini and Kavitha argue services for many reasons including because this service provide fast access the applications and reduce service costs [6]. Cloud computing providers should address privacy and security as matter for higher and urgent priorities. The dealing with 'single cloud' providers is becoming less popular service with customers due to potential problems such as service availability failure for some time and malicious insider's attacks in the single cloud. So now single cloud move towards 'multi clouds', 'interclouds', or 'cloud of clouds'.

Aim of the paper the data security aspect of cloud computing, data and information will be shared with a third party without any hacks. Every cloud users want to avoid untrusted cloud provider for personal and important documents such as debit/credit cards details or medical report from hackers or malicious insiders is the importance. It supply secure cloud database that will prevent security risks. We apply multi clouds concept using Shamir's Secret Sharing algorithm that is reduce risk of data intrusion and loss of service availability for ensuring data.

II. OBJECTIVE

Cloud computing concept is relatively new concept but it is based on not so many new technologies. Many of the features that makes cloud computing attractive, however has to meet certain basic security criteria. In our paper, we have briefed on various measure ion cloud computing security challenges from single to multi clouds. While making a cloud secure, the following objectives are to be met:

- Understanding the cloud computing environment provided by the cloud service provider.
- The cloud computing solution should meet the basic security and privacy requirements of any firm deploying it.
- Maintain an account of the privacy of the cloud and data security and applications that are deployed in cloud computing environment.
- Data Integrity.
- Service Availability.
- The user runs customer applications using the service provider's resources

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III. ALGORITHM USED

3.1 Shamir's Secret Sharing Algorithms:

Data stored in the cloud can be compromised or lost. So, we have to come up with a way to secure those files. We can

encrypt them before storing them in the cloud, which sorts out the disclosure aspects [7].However, what if the data is lost due to some catastrophe befalling the cloud service provider? We could store it on more than one cloud service and encrypt it before we send it off. Each of them will have the same file. What if we use an insecure, easily guessable password to protect the 2012 45th Hawaii International Conference on System Sciences file, or the same one to protect all files? I have often thought that secret sharing algorithms could be employed to good effect in these circumstances instead [8].

Mathemetical Definition given below:

Our goal is to divide some data D (e.g., the safe combination) into n pieces D1, D2…,Dn in such a way that:

1. The Knowledge of any k or more Di pieces makes D easily computable.
2. The Knowledge of any k-1 or fewer Di pieces leaves D completely undetermined (in the sense that all its possible values are equally likely).

This scheme is called (k,n) threshold scheme. If k=n then all participants are required to reconstruct the secret original data.

The essential idea of Adi Shamir's threshold scheme is that 2 points are sufficient to define a line, 3 points are sufficient to define a parabola, 4 points to define a cubic curve and so forth. That is, it takes k points to define a polynomial of degree $k-1$.

Suppose we want to use a (k,n) threshold scheme to share our secret S, without loss of generality assumed to be an element in a finite field F.

Choose at random $k-1$ coefficients $a_1,\cdots,a_{k-1}$ in F, and let $a_0 = S$. Build the polynomial $f(x) = a_0 + a_1 x + a_2 x^2 + a_3 x^3 + \cdots + a_{k-1}x^{k-1}$. Let us construct any n points out of it, for instance set $i = 1,\cdots,n$ to retrieve $(i, f(i))$. Every participant is given a point (a pair of input to the polynomial and output). Given any subset of k of these pairs, we can find the coefficients of the polynomial using interpolation and the secret is the constant term a_0 [27].

Shamir Approach:

We divide our secret into pieces by picking a random degree polynomial $q(x) = a_0 + a_1 x + a_2 x^2 + \mathrm{L} + a_{k-1}x^{k-1}$
in which $a_0 = S$ $S_1 = q(1), S_2 = q(2), \mathrm{K}, S_n = q(n)$
and represent each share as a point $(x_i, q(x_i) = y_i)$

Example:

The following example illustrates the basic idea. Note, however, that calculations in the example are done using integer arithmetic rather than using finite field arithmetic. Therefore the example below does not provide perfect secrecy, and is not a true example of Shamir's scheme.

Preparation:

Suppose that our secret is 1234 $(S = 1234)$.
We wish to divide the secret into 6 parts $(n = 6)$, where any subset of 3 parts $(k = 3)$ is sufficient to reconstruct the secret. At random we obtain 2 numbers: 166, 94.
$(a_1 = 166; a_2 = 94)$

Our polynomial to produce secret shares (points) is therefore:
$$f(x) = 1234 + 166x + 94x^2$$

We construct 6 points from the polynomial:
$$(1,1494);(2,1942);(3,2578);(4,3402);(5,4414);(6,5614)$$

We give each participant a different single point (both x and $f(x)$ [27].

Reconstruction:

In order to reconstruct the secret any 3 points will be enough.
Let us consider

$$(x_0, y_0) = (2,1942);(x_1, y_1) = (4,3402);(x_2, y_2) = (5,4414).$$
We will compute Lagrange basis polynomials:

$$\ell_0 = \frac{x-x_1}{x_0-x_1}\cdot\frac{x-x_2}{x_0-x_2} = \frac{x-4}{2-4}\cdot\frac{x-5}{2-5} = \frac{1}{6}x^2 - \frac{3}{2}x + \frac{10}{3}$$

$$\ell_1 = \frac{x-x_0}{x_1-x_0}\cdot\frac{x-x_2}{x_1-x_2} = \frac{x-2}{4-2}\cdot\frac{x-5}{4-5} = -\frac{1}{2}x^2 + \frac{7}{2}x - 5$$

$$\ell_2 = \frac{x-x_0}{x_2-x_0}\cdot\frac{x-x_1}{x_2-x_1} = \frac{x-2}{5-2}\cdot\frac{x-4}{5-4} = \frac{1}{3}x^2 - 2x + \frac{8}{3}$$

Therefore

$$f(x) = \sum_{j=0}^{2} y_j \cdot \ell_j(x)$$

$$= 1942\cdot\left(\frac{1}{6}x^2 - \frac{3}{2}x + \frac{10}{3}\right) + 3402\cdot\left(-\frac{1}{2}x^2 + \frac{7}{2}x - 5\right) + 4414\cdot\left(\frac{1}{3}x^2 - 2x + \frac{8}{3}\right)$$

$$= 1234 + 166x + 94x^2$$ [27].

IV. SOLUTION METHODOLOGY

Cloud customers may form their expectations based on their past experiences and organizations' needs. They are likely to conduct some sort of survey before choosing a cloud service provider. Customers are expected also to do security checks that

are centered on three security concepts: confidentiality, integrity and availability. On the other hand, cloud service providers may promise a lot to entice a customer to sign a deal, but some gaps may manifest later as overwhelming barriers to keep their promises. Many potential cloud customers are well aware of this, and certainly, still sitting on the sidelines. They will not undertake cloud computing unless they get a clear indication that all gaps are within acceptable limits. All relevant information are visualized into cloud computing security in a snapshot which is presented in Fig.1 [9]. We organized cloud computing security into three sections: security categories, security in service delivery models and security dimensions.

Security in cloud services is based on the following:

- Strong network security is possible around the service delivery platform
- Data encryption: for data in transit (particularly over wide area networks), and sometimes stored data, but it cannot be applied to data in use.
- Access controls to ensure that only authorized users gain access to applications, data and the processing environment and is the primary means of securing cloud-based services

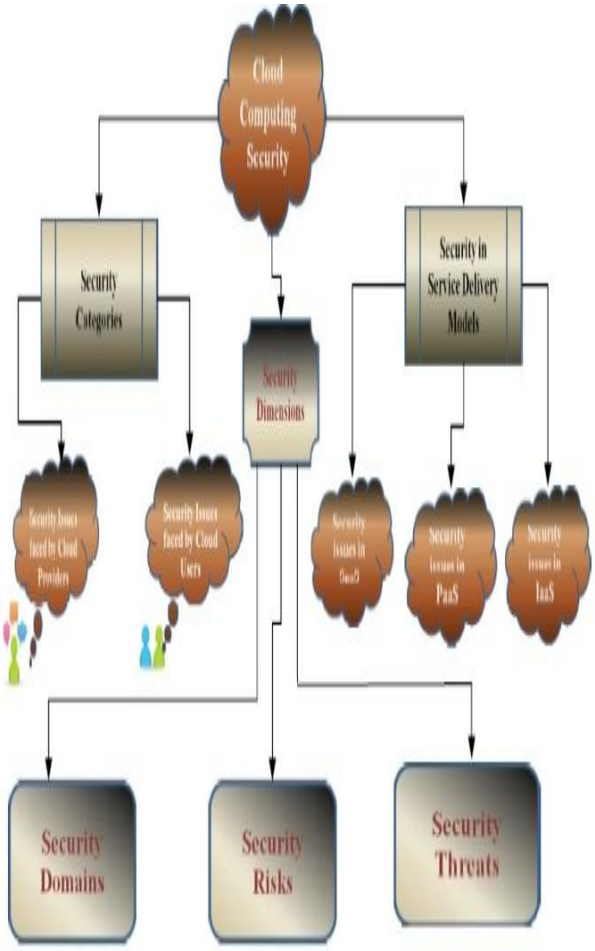

Figure 1: Graphical View of Cloud Computing Security [9]

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- Service providers are able to inspect activity in their environment and provide reports to clients.

Logs need to be carefully constructed to appraisal the actions of their system administrators and other restricted users or risk producing reports that mix events relating to different customers of the service.

Both the organizations seeking cloud solutions and the service providers have to ensure cloud security is addressed [11]. Some of the measures to ensure security in cloud are good governance, compliance, privacy, Identity and Access Management (IAM), Data protection, Availability, Business Continuity and Disaster Recovery plans etc. The figure (fig 2) below depicts the above mentioned security measures in a snapshot:

Figure 2: Measures to ensure Security in Cloud [10]

V. IMPLEMENTATION

5.1 Data Integrity:

It is not an easy task to securely maintain all essential data where it has the need in many applications for clients in cloud computing. To maintain our data in cloud computing, it may not be fully trustworthy because client doesn't have copy of all stored data. But any authors don't tell us data integrity through its user. So we have to establish new proposed system for this using our data reading protocol algorithm to check the integrity of data before and after the data insertion in cloud. Here the security of data before and after is checked by client with the help of CSP using our "effective automatic data reading protocol from user as well as cloud level into the cloud" with truthfulness[8].

5.2 Data Intrusion:

The importance of data intrusion detection systems in a cloud computing environment. We find out how intrusion detection is performed on Software as a Service, Platform as a

Service and Infrastructure as Service offerings, along with the available host, network and hypervisor-based intrusion detection options. Attacks on systems and data are a reality in the world we live in. Detecting and responding to those attacks has become the norm and is considered due diligence when it comes to security[8].

5.3 Service Availability

Service availability is most important in the cloud computing security. Amazon already mentions in its licensing agreement that it is possible that the service might be unavailable from time to time. The user's web service may terminate for any reason at any time if any user's files break the cloud storage policy. In addition, if any damage occurs to any Amazon web service and the service fails, in this case there will be no charge to the Amazon Company for this failure. Companies seeking to protect services from such failure need measures such as backups or use of multiple providers [8][11].

5.4 DepSky System Model Architecture:

The DepSky system model contains three parts: readers, writers, and four cloud storage providers, where readers and writers are the client's tasks. Bessani et al. explain the difference between readers and writers for cloud storage. Readers can fail arbitrarily (for example, they can fail by crashing, they can fail from time to time and then display any behavior) whereas, writers only fail by crashing [8][14].

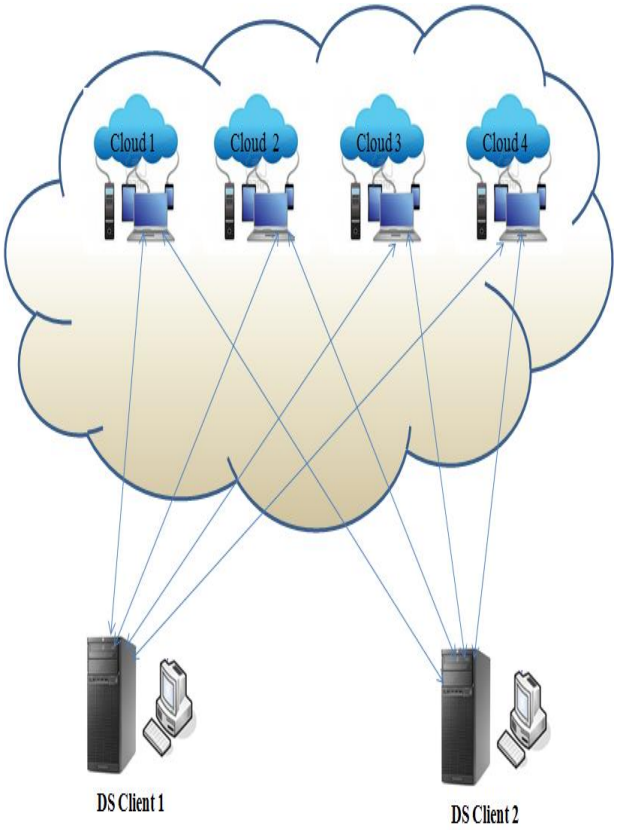

Figure 3: DepSky Architecture [14]

VI. RESULT AND DISCUSSIONS

In any cloud computing environment, the scope of activities can be divided into three major steps as: preliminary activities, initiating activities and concluding activities. The preliminary activities include a wide range of steps from identifying the security, privacy and organizational requirements to analyzing the security and privacy provided by the security provider and the levels of risks involved with respect to control objectives of the organization **.**Reliable distributed storage which utilizes a subset of BFT (Byzantine fault tolerance) techniques was suggested by Vukolic to be used in multi clouds or interclouds. One example for protocol of controls the multiple clouds HAIL (High Availability and Integrity Layer).HAIL is permits set of services to ensure that client's stored data is retrievable and integral and also provides a software layer to address availability and integrity of stored data in the intercloud.

We already discussed before, Bessani et a1 present a virtual storage cloud system called Depsky consisting of a combination of different clouds to build a cloud of clouds. Finally, the Depsky system presents an experimental evaluation with several clouds that is different from other previous work on multi clouds.

VII. CONCLUSION

The purpose of this work is to survey the recent research on single clouds and multi-clouds using secret sharing algorithm and to address the security risks and solutions using Shamir's Secret Sharing algorithm. These algorithms generate their own secret sharing schemes and use secure channels to distribute shares among themselves[7][8]. The Shamir's secret sharing scheme has a good abstract foundation which provides an excellent framework for proofs and applications [28].

We presented algorithms for performing addition, standard and scalar multiplication with shares. We are currently developing a secure computation platform based on a simple secret sharing scheme than Shamir's. Cloud computing is currently the latest trend when it comes to online computing, it may help the enterprise and the end user by providing their needs, but the provider has to make sure that they are valuable and customer data is safe[8][28]. We support the migration to multi clouds due to its ability to decrease security risks that is affect the cloud computing users.

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Mobile Based Electricity Billing System (MoBEBIS)

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Abstract- Billing is a critical function of both the Electricity and the Water Boards towards getting a meter read. Meter reading, even though it looks simple, is far from simple and involves processes that can give various problems. Most problems, currently seen, result from the manual processes followed. Calculation errors, delays in system updating and fault tracking issues are the major problems that companies find difficult to find answers for. This paper suggests a mobile based system to collect, process and notify consumers about consumption. This system will be reliable, efficient and accurate to suit the requirements of these companies. The proposed solution uses evolving Mobile Technologies, over a solution which uses Mobile applications to handle a company's day today work. The burden on the Meter Reader is lessened and other new features have also been introduced. Customer interaction with the company is improved and customers can easily view their current electricity usage using their mobile phones. However, the feasibility of such a project for a third world country like Sri Lanka, is a concern with tregard to the cost factor involved. The project demands substantial investments. Will the country be able to meet the costs involved? Yet, most of the problems related to Electricity Billing are addressed through this system and this might prove to be the best solution for specific companies to optimize services on a low budget

Index Terms- Mobile Based billing, Meter Reading, Electricity Billing

I. INTRODUCTION

The current procedure with regard to the billing process for electricity is not a fully automated system. It involves manual processes from the time the Meter reader starts reading the meter until the system is updated with the current reading. A meter reader visits a house, does the meter reading, and then manually calculates the amount considering the units consumed. Back in the office a data entry officer enters the meter readings into the system manually. The procedure is far from satisfactory and it is believed a better system using available technologies would definitely be an advantage.

To overcome problems with this manual approach a few solutions are identified. 'Smart Meter'[1] and 'Java Based Meter Reading System'[2] are some of the solutions. But they are not much welcome in a third world country like Sri Lanka, because the initial cost is very high.

There is also another major problem. When everything is automated there is no 'Fault tracking', so there will be a lot of illegal issues like illegal power consumption. A further problem is the vast sums of money required to update or maintain this device [3].

However, to meet the problems associated with the major problems related to the Manual Billing process it is suggested to use a mobile device (android mobile phone). The proposed system is mobile and Web based. The System eliminates most of the error prone manual calculations and manual data entering. It increases the interaction between the company and their customers. The System update happens fast and customers can have the flexibility to get to know their electricity usage at any time they want. This will help reduce unwanted power consumption.

The product could be a welcome solution for the Ceylon Electricity Board and the Water Board. They could easily use the system for a faster, easy and error free environment to suit the comfort of customers.

II. RESEARCH METHODOLOGY

Problem Identified

The Electricity Board and the Water Board currently use a manual process for billing purposes due to the following reasons: They have got used to the manual process and they can go along with it even though there are concerns associated with it. They are reluctant to change their current process since it will be an extra effort. The Electricity Board or the Water Board cannot invest a huge amount of money for a new solution. However, the customers face immense problems with the current procedure of using this manual process to calculate Bills.

The Meter reader's perspective is that the reading is collected manually from the meter and calculations for a specific month after which, the data gets manually entered to the system. There is no route map for guidance or a plan covering each and every house. Another issue is that complaints cannot be made at the time of meter reading but back in the office.

From the customer's perspective there is no way of knowing their current electricity usage or calculating it manually with the given formulae. There is also no facility to compare the previous month's electricity usage with the current month. When they need to make complaints about the bills or about the device it is considerably difficult for the customers with the current system.

Finally from the Electricity Board's perspective, all these process are manually maintained. This is a big burden for them. Another serious issue is that they do not have a proper way to communicate with their consumers when they want to notify

about power failure or the latest news related to power consumption.

Solution Proposed.

The suggested solution highlights the following: MoBEBISis and a Web site. MoBEBISis consists of two separate Mobile applications which are given to both Meter reader and the Customer (Customer is able to use the system only as a subscriber to this particular system.) A Web site is maintained basically for administrative purposes.

Meter readers will get the best benefits from this particular system. From the beginning of the day, a mobile phone with a route map called Walk Order Map(fig 4) which has the route of houses that he has to cover within a day, will be in use. Whenever a meter is read, the particular meter is indicated with the red color confirming that the reading is already captured. This leaves no room for 'missed' readings especially for novice Meter Readers. Meter readers are not required to do the calculation manually. All that needs to be done is to get the meter reading and send it to the system as bulk. Then the system does the calculation and proceed bills are sent to the relevant consumers via SMS. Importantly, the Meter reader can make complaints then and there whenever a fault is seen or an illegal power usage is spotted. In such a case, an image of that particular meter can be sent. This option is also provided through the system.

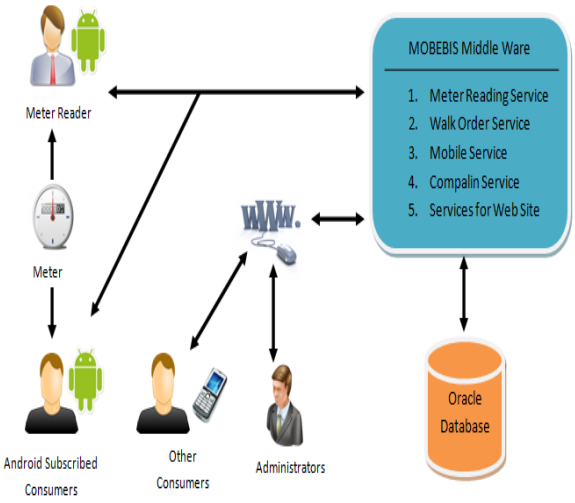

Fig 1. The System Overview

N ill also be benefited using this system. There is a separate Mobile application developed for the customers so that they can be interactive members of this process. If the customers wish to know the current power consumption/ units consumed with the up-to-date bill, they can send the current reading of the meter to the SMS Manager Windows Service (3.2) via customer mobile application itself or SMS via normal mobile phones and the calculated amount will be received via an SMS. Self-Billing Subscribers are given an option where they can submit their monthly meter reading by taking an image (3.5) of the digits of the meter. Instructions are given to the customers and they can take an image of the meter and send it to the system. Then the system idendifies the digits using Image Processing Library (3.1). Then the system gets updated after processing the bill using identified digits and most importantly a meter reader does not have to come and take the reading since it is already provided by the customer.
A customer can view bills when necessary. By selecting the intended month, a customer can view bills. Customers are also given a facility where they can make complaints against the service and a track record of the complaints made by him can be viewed for the status of the complaints via the system. Another feature in this system is where the customer can subscribe/ unsubscribe the functions given. On the other hand, customers can use the Website to make payments, track complaint status and so on.
This application provides the Electricity Board with some features to help with administration as well. The built Web application is for administrative purposes. An area manager/ a responsible administrator can assign Meter readers to a particular route using this system. Technical officers will be able to keep track of the complaints / tasks assigned to them. The Electricity Board can publish any instant messages (Like sudden power drops due to maintenance purposes) without going for public media so that they can save money.

III. Research Findings and Evidence

The system implementation is the arithmetic approach of problem-solving and planning for a software solution after the purpose and specifications have been determined. Under this section it discusses some specific library and service implementation of the system. All these services and library implementations come under sub parts of the MoBEBIS middleware and Android Mobile Application development.

A. Images Processing Library

The image processing library is related to the Android application which is used by the Android Mobile Consumers who are going to subscribe to the Self Billing Service. That application is capable of sending consumers monthly reading as an image to the MoBEBIS Middleware via Web service (3.3).[6,7]

When the system receives the image it is processed in this library to identify the current reading for the particular consumer. The steps to be followed are briefly described below.

First is retrieves the image from a Web Service to the library. This is called Acquisition. Then there is Gray Scaling the image so that the image continues tone image to an image that a computer can manipulate. Next it identifies the threshold value and applies it to the image using Otsu thresholding approach. Thereafter, it isolates the numeric area from the image, using Blob Counter after which it applies the Media Filter to remove noise from the image. Then the system does the segmentation to analyses the pixel in each digit. Finally Neural Network to identifies digits.

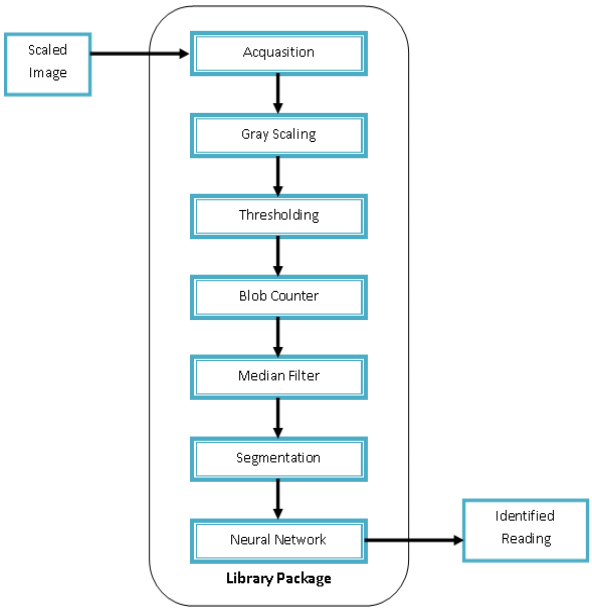

Fig. 2 System Diagram for Image Processing

B. SMS Manager Windows Service

The main idea of having an SMS service is to have better communication with consumers and SMS is the ideal way to achieve it. When the bill is processed by the system, billing details are sent via SMS. Besides this the consumer can make service requests like Get last Month Bill which is published by the MoBEBIS services for a better interaction with MoBEBIS. Some of the requests that consumers can make from the system are as follows:

1. Requesting Bill up to Current Date/ Last Bill
   Consumer can make a request via SMS.
   Eg - REQ TODAY [AccNo][Current Reading].
   REQ LAST [AccNo]

2. Subscribing and Unsubscribing to the Services
   Consumer can subscribe and Unsubscribing to/from services via SMS.

Eg - REQ [SUB/UNSUB] [AccNo] [ServiceNo]

C. WCF REST Services (Meter Reading Service)

Meter Reading Service is the key service given to a Meter Reader to submit daily collection of reading to the system using an android application. Android consumers registered with the monthly reading service, can send their reading to the system in the form of an image. All Services are implemented as WCF REST Service [4].

D. Encryption Library

This application need to be secure when in use. Therefore the Rijndael encryption algorithm Rijndael [5] is used in this application.

E. Customized Adjustable Camera

The customer Android Application is provided with a customized adjustable camera to capture the numeric area of the meter. The normal camera is customized according to the requirement. The camera is provided with an adjustable square to capture the numeric area of the meter. A customer can change it very easily by using blue dots provided with it

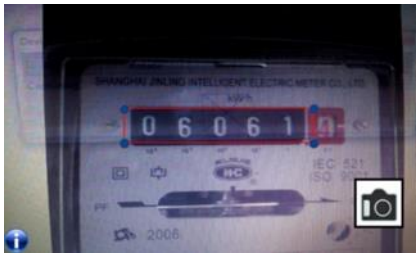

Fig. 3 Customized Image capture Adjustable Camera

F. Walk Order Map Service

The Meter Reader android application is consistent with Walk Order Map. T services. At the beginning walk order from the Elec to view the houses to be ning all the houses are marked as red dots on the map. The Map gets updated by changing the color into blue when a house gets covered. The meter reader can also get a quick understanding about what has been covered and what remains to be covered. To handle this application it is necessary to allow obtaining periodic updates of the device's geographical location when the device enters the proximity of a given geographical location.[9]

G. Android Map Service

Android gives your applications access to the location services supported by the device through the classes in the android.

Location package. The central component of the location framework is the Location Manager System service, which provides APIs to determine location and bearing of the underlying device (if available).

H. Search Consumer Details

Consumer details are embedded into a QR Code. When the meter reader need details about the consumer, QR Reader, which is provided through mobile application can be used to retrieve details from the QR code which is stuck on the meter.[8]

I. Unique Features

The unique features of this system will enable the Electricity Board staff, meter reader, technical officers and specially the consumers to experience many benefits. They are as follows:

1. Optimize time - The system eliminates the need of allocating time for the meter readers to go to the office to take the walk order and also give the readings.

2. 24/7 connectivity - As the system is Web-based (web site), it is available any time.

3. Immediate access - The meter reader application is available for the reader to log in to the system and also to understand the current day's work. The application, being a Web-based system, it is readily available. Therefore, it can be accessed from anywhere, any time. Users access their account online and they are ready to work no matter what user's setup or hardware is.

4. Better security - As a Website is available on the Internet, only the authorized users can access the system according to access privileges. The consumer mobile application needs to be downloaded from the Web site after successful registration. Capturing reading of the meter function will be available once a month and the consumer is capable of capturing one's own meter securely. The System encrypts all sensitive data, so no one can access it in transfer to the system's database via a Web service. Access will be denied for other users.
5. Reliable and effective communication –Web officers and consumers can communicate via SMS and also by E-mails. Consumers can send complaints through e-mails and android application itself. Any service interruption news will be sent to the consumers via SMS.

Wide flexibility - All manual processes are promised through the implementation of the MOBEBIS. The system allows the consumers to choose some specific services if they need or not.

IV. CONCLUSION AND FUTURE WORK

The basic idea of developing MoBEBIS was to address some common issues related to the manual electricity billing process. Before suggesting this solution they were using a manual process on meter reading, amount calculation, and billing customer and so on. The interaction between customer and Electricity Board was very poor and it took much longer to respond to customer queries.

The other major problem which was identified during the literature survey is that customers keep complaining that bills are incorrect. Most of the time bill calculation and system updating are done manually. There can be some resulting human errors to frustrated customers who are not satisfied about the service of the Electricity Board at all. Another difficulty observed was with making complaints against the Electricity Board service via a call.

The suggested system has come up with solutions which address all the above mentioned problems. With this service the burdens of the Meter reader as well as the Electricity Board get lessoned and are made more efficient. A mobile solution is given for the Meter reader so that the day to day work becomes less tiresome. Most of the manual processes and calculations are eliminated so that the meter readings can easily be collected more accurately to be updated to the system.

On the other hand, the Android customers are also given a mobile solution so that they can view their latest bills, make complaints against the Electricity Board's service, and make payments and other important tasks efficiently. Administration tasks of the Electricity Board can be easily done via a Web site which is provided as part of the complete project.

For future research, it is suggested that image processing be done in the mobile device as the image gets captured. The send the numeric value can be sent to the server through Web service.

This particular solution can be further expanded over the Sri Lanka Water Board as well since they are following almost the same process of billing customers. This system is well tested and it is proven that it functions properly and accurately so that it can be a solution for the Electricity Board to serve their customers satisfactorily.

ACKNOWLEDGMENT

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possible:

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An Efficient Group Key Distribution Security Scheme in Wireless Sensor Networks

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Abstract- An Unattended Wireless Sensor Network (UWSN) collects the sensing data by using mobile sinks (MSs). It differs from the traditional multi- hop wireless sensor networks in which unbalanced traffic makes the sensors close to the base station deplete their power earlier than others. An UWSN can save the battery power and prolong the network lifetime. Unfortunately, MSs would be given too much privilege when acting as the collecting base station, which will cause security concern if Replicated. Besides, UWSNs are usually deployed in unreachable and hostile environments, where sensors can be easily Replicated. Thus, their security issues should be carefully addressed to deal with node compromise. In this paper, we present a novel key management scheme to secure UWSNs. We employ the Blundo symmetric polynomial mechanism to guard against the newly Replicated nodes in a period while utilizing the periodic key updating based on the reverse hash chain to block the Replicated nodes and revoke the Replicated MSs if failing the authentication. We show that our scheme is robust against node Replicated attacks and carry out comparison analysis on the intrusion-tolerance ratio, communication and computing overhead.

Index Terms- wireless sensor networks; security; key management; mobile sink.

I. INTRODUCTION

An Unattended Wireless Sensor Network (UWSN) is a kind of hybrid wireless sensor network consisting of mobile sinks (MSs) and static sensing nodes [1], which periodically collects sensed data from static sensors by using the mobile sinks. In such a UWSN, the static nodes are composed of a large number of low-power sensors with limited communication, computing and storage capacity, and are usually deployed in uninhabited environments, such as military sensing, invasion detection, commercial security, and etc. While in the traditional wireless sensor networks, a large number of sensors are required to send the collected data to a designated base station by multi-hop communications, resulting in excessive routing burden to the nodes close to the base station. In UWSNs, on the other hand, it is the MSs that bear most of the communication overheads by moving around and collecting data. The deployment of mobile nodes could reduce the power cost in the multi-hop data transmissions in static WSNs and balance the network energy consumption, and hence prolong the network lifetime. Unfortunately, the introduction of the mobile nodes also brings some security concerns, which will be addressed in this paper.

In some earlier studies, MSs are assumed to have the same capability as the base station, and the efficiency of data collection is the major design consideration [2, 3]. However, security is an unavoidable issue in UWSNs because of the unattended nature and hostile environments. To secure a UWSN, the key management has been investigated lately. It is observed that the key management schemes for conventional WSNs cannot be directly applied in UWSNs because of the participation of MSs. Meanwhile, the battery power in a sensor is more limited than that of a node in a mobile ad hoc network (MANET), thus the key management scheme designed for MANETs is also not effective in UWSNs [4]. Moreover, some key management schemes [5] for UWSN often provide MSs with excessive privileges, even without considering the pinpoint attacks on MS.

High privileges given to MS may cause the security problems. Song, et al. [6] investigated the revocation problem of MS in their key management scheme, which synthesized Blundo symmetric polynomials [7] and Merkle Tree [8]. However, the routing for sensing data collection in their scheme was set up for a single path and a single task, and the MS identity was bond to such a unique path and task. Combined with symmetric polynomial key pool scheme [9] and random key pre-deployment scheme [10], Rasheed et al. [5] proposed a new key management scheme for a UWSN, in which the employment of symmetric polynomial key pools can enable the bonding relation of MS with the moving routing and data collection task.

However, it assigns too important role to the MS. If the MS fails or is Replicated, 90% of the pre-distributed random keys will be useless, which results in very low probability of establishing security channels between nodes, and the resulting scheme was degenerated to basic Blundo scheme. Moreover, they have not taken the security of MS into account.

To address the security over UWSNs, in this paper, we present a novel Modernize key management scheme, in which the symmetric polynomial is used to generate shared keys between nodes and enables the threshold security feature of the network, while the reverse hash chain is utilized for key update or revocation to effectively restrict the privileges of MSs and at the same time for the identity generation of the newly-joined MSs.

II. NETWORK MODEL

A UWSN is a WSN with MSs used to help static BS collect data. Here, we use the general WSN network structure based on group deployment. For simplicity, we study the key management scheme under the network model with one MS.

A. Network Structure

We assume that the common nodes (Nodes) and the Cluster heads (CHs) pre-loaded with the same symmetry polynomial are distributed based on group deployment [11]. According to the principle of group deployment, it is easy for the common nodes to find the nearest sink as their cluster head in two-dimensional space.

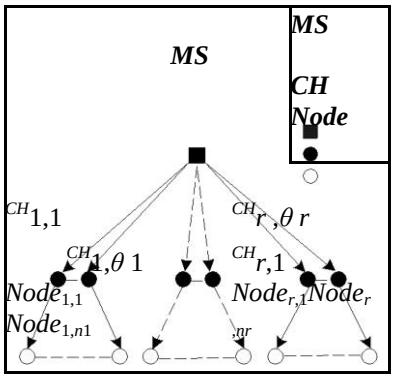

Figure 1. Three-layer UWSN network structure

When the MS joins the WSN, the network forms a three-layer structure, consisting of the MS, CHs and Nodes as shown in Figure 1. Here we assume that MS and the CHs have strong computing power, communication capability and high storage. In this cluster-based UWSN, CHs shall wait for the arrival of the MS after collecting the sensed data.

B. Adversarial Model

The network is assumed to be unattended, with the MS periodically collecting the sensed data and at the same time managing and maintaining the network, which includes security management such as key management, the revocation of Replicated nodes, etc. In the network, the adversary can launch external attack by monitoring the transmitted information or even internal attack by using physical capture to get all the secret information of arbitrary nodes.

We do not assume the MS is equipped with costly hardware security protection, which means that the adversary can obtain all confidential information stored in the MS if captured, but is costly for the adversary to perform physical capture attack. The adversary cannot capture and compromise more than t static nodes in a certain time period T_1 or capture and replicate the MS in time period $T_2(T_2>T_1)$.

III. MODERNIZE INTRUSION-TOLERANT KEY MANAGEMENT

In our Modernize intrusion-tolerant key management (ITKM) scheme, key distribution consists of three phases, namely, key material pre-distribution, shared-key generation, and key update as well as MS revocation.

TABLE I. NOTATION

notation	Meaning
A-B	Node A and B
$f_{A\text{-}B}(ID,y)$	The shared binary symmetric polynomial between A and B
t	The degree of Binary symmetrical polynomial
S_i	The $(i+1)$th hash value of the reverse hash chain
	• The number of symmetric polynomials pre-loaded in MS
	• The number of Sinks in each group

[1]. Key Material Pre-distribution

The BS chooses a random symmetric bivariate polynomial $f(x,y)$ of degree t with coefficients over a finite field F_q, where q is a prime number large enough to accommodate a symmetric key:

$$f(x,y) = \sum_{m,n} a_{mn} x^m y^n \qquad (a_{mn} = a_{nm})$$

$$0 \le m,n \le t$$

which is also called Blundo symmetric polynomial. BS pre-loads the same Blundo symmetric polynomial for MS and CHs while loading CHs and Nodes with different one referred to as $f_{MS\text{-}Node}(ID,y)$ and $f_{CH\text{-}Node}(ID,y)$. BS randomly selects an initial value S_n and calculates the reverse hash sequence $\{S_i\}_{i=0}^{n}$:

$$S_{i-1} = h(S_i)(1 \le i \le n)$$

In which, $h()$ is a collision-free one-way hash function and $\{S_i\}_{i=0}^{n}$ is called the reverse hash chain [12] who is generated through the formula (2).

Similarly, BS chooses a random value r_{MS} and then uses formula (3) to calculate the reverse hash sequence with $MS_n = r_{MS}$.

$$MS_{i-1} = h(MS_i)(1 \le i \le n)$$

At the beginning of the network deployment, BS pre-loads MS and CHs with Blundo symmetric polynomial $f_{MS\text{-}CH}(ID,y)$ and hash value S_0, in which the ID in $f_{MS\text{-}Node}(ID,y)$ represents the node identity and S_0 represents the first value of the reverse hash chain. Meanwhile, BS pre-loads the CHs and Nodes with Blundo symmetric polynomial $f_{CH\text{-}Node}(ID,y)$ and hash S_0.

B. Key Agreement Protocol for Peer Nodes (KAP)

Since the network consists of MS, CHs and Nodes, there are two types of key agreement: one between homogeneous nodes (between the CHs or between Nodes) and one between heterogeneous nodes (between the MS and CHs or between CHs and Nodes) . According to the network model, the MS does not directly communicate with Nodes, and so we do not need to consider key agreement between the MS and Nodes.

Let us first consider the key agreement between peer nodes u and v which are pre-loaded with Blundo symmetric polynomial $f(u,y)$ and $f(v,y)$ respectively. Let the current hash values stored in u and v are S_i and S_j, respectively. If there is an attack or network failure, some nodes may fail to update their hash values, which will result in $S_i \neq S_j$. The node with the latest hash value will identify the other communication party subjected to malicious attack. Otherwise, they can establish the shared key as shown in Figure 2.

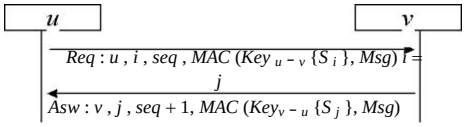

Figure 2. Key agreement protocol between peer nodes

Here, the node u firstly sends a request message *Req* to v

$$Req = \{u, i, seq, MAC(Key_{u-v}\{S_i\}, Msg)\} \quad (4)$$

where u also represents the node identity, i the identity of current hash value, *seq* the sequence number of transmission between node u and v. *Msg* is the current message $Msg = u \mid i \mid seq$, and $Key_{u-v}\{S_i\}$ the shared key calculated by u independently through (5).

$$Key_{u-v}\{S_i\} = h(f(u, v), S_i) \quad (5)$$

Upon receiving *Req*, v firstly compares i with j. Only when $i = j$ can v calculate $f(v,u)$ and $Key_{u-v}\{S_i\}$. According to the symmetry property of Blundo symmetric polynomial, we have

$$f(v,u) = f(u, v) \quad (6)$$

Because $i = j$, we know that $S_i = S_j$. From the definition of $Key_{v-u}\{S_j\}$ and (6), we have

$$Key_{u-v}\{S_i\} = Key_{v-u}\{S_j\} \quad (7)$$

Hence, we obtain

$$MAC(Key_{v-u}\{S_j\}, Msg) = MAC(Key_{u-v}\{S_i\}, Msg) \quad (8)$$

Therefore, v can verify the MAC through $Key_{v-u}\{S_j\}$ in *Req* message and complete the node and message integrity authentication. For legitimate node u, v will reply with *Asw*

$$Asw = \{v, j, seq + 1, MAC(Key_{v-u}\{S_j\}, Msg)\} \quad (9)$$

Upon receiving *Asw*, u will carry out the same procedure to verify the legitimacy of v.

When $i \neq j$, there should be at least one node failing to update its hash value, which will be considered unsecured or Replicated and, as a result, cannot pass the authentication or obtain (7), even if they do have the same Blundo symmetric polynomial and calculate the result in (6). This is because the one-way character of hash function and the node cannot calculate S_i through S_j and can not calculate the right $Key_{v-u}\{S_i\}$. Therefore, though v has $f(u,v)$, S_j and $MAC(Key_{u-v}\{S_i\}, u \mid i \mid seq+1)$, it cannot forge $MAC(Key_{u-v}\{S_i\}, v \mid i \mid seq+1)$ in *Asw*.

The above analysis shows that due to the updating of the hash values in the reverse hash chain, even if a node has a valid network identity, when it fails to obtain the current hash value, it cannot impersonate legitimate nodes and continue to engage in legal network communication.

However, as wireless sensor nodes may be subject to physical capture. When an attacker captures k nodes, the attacker can combine k Replicated nodes to expand the impact of the attack. Blundo et al. [7] showed that the symmetric polynomial (1) is t intrusion-tolerant when $k \leq t$ because the attacker cannot obtain the parameter a_{mn} of (1). Thus, even if k nodes collude, they still cannot find the polynomial $f(ID,y)$ pre-loaded in the nodes and launch node impersonation attack.

Finally, we want to point out that by using information entropy, we can show that the following two results analytically. Due to page limit, we omit the proofs here.

Theorem 1. In the key agreement protocol, any two nodes u and v are able to calculate their shared key $Key_{u-v}\{S\}$ when they are pre-loaded with the same hash value S and Blundo symmetric polynomial as shown in (1).

Theorem 2. The key agreement protocol and key update protocol are both t intrusion-tolerant.

C. Key Update Protocol for Heterogeneous Nodes (KUP)

BS needs to send MS the latest hash value S_{i+1}, through which BS could carry out the update of the secret keys. As it moves, the MS will communicate with CHs and securely send CHs the latest hash value S_{i+1} to replace the current one, namely, S_i. Meanwhile, the MS needs to notify corresponding Nodes that the current hash value has changed. Because key updates between the MS and CHs is similar to that between CHs and Nodes, so we let M be the upper node such as the MS in MS-CH or the CH in CH-Node, and let u be the low-level node such as the CH in MS-CH or the Node in CH-Node. Next, we will describe key update between M and u.

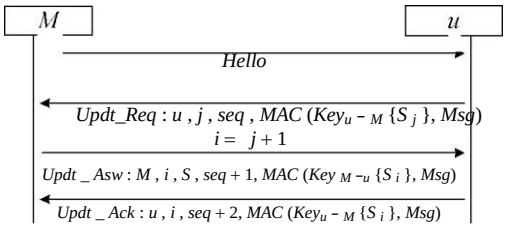

Figure 3. Key update protocol between heterogeneous

nodes

Assume the current hash value stored in M and u is S_i and S_j respectively, between which $i = j+1$ must hold. For $i < j+1$, u needs refuse to update its hash value and conclude that M might be a malicious node. If $i > j+1$, it indicates that the node u missed to update of the latest hash value, and will be marked as a potential malicious node by M. Notice that a malicious node M cannot calculate $S_i = h^{-(i-j)}(S_j)$ through $h^{-1}()$ because of the collision-freeness of one-way hash function. However, the malicious node u , who missed one round of key updating, may still want to join the network again through key updating. To prevent this, we design the KUP as shown in Figure 3.

Let M be pre-loaded with $f(M,y)$ and hash value S_i . Let u be pre-loaded with $f(u,y)$ and hash value S_j. M notifies u that it comes to update the hash value and conducts key agreement through *Hello* message. Otherwise, M and u can communicate with each other using their shared key established through KAP, which does not belong to KUP.

As shown in Figure 3, M first broadcasts *Hello* message to inform u that it comes for key updating. Upon receiving the *Hello* message, u sends the update request message *Updt_Req* to update its key materials.

$$Updt\_Req = \{u, i, seq, MAC(Key_{u-M}\{S_i\}, Msg)\} \quad (10)$$

The elements u, i, seq, and Msg in (10) are similarly defined as in (4) and the updated shared key $Key_{u\text{-}M}\{S_i\}$ between u and M is calculated through equation (5).

When M gets message *Updt_Req* with $j \neq i$-1, M records u as a potential malicious node as discussed before and then quits the update process; Otherwise, when $j = i$-1, M calculates $S_j = h(S_i)$ to obtain S_j and determines whether $MAC(Key_{M\text{-}u}\{S_i\}, Msg) = MAC(Key_{u\text{-}M}\{S_j\}, Msg)$ through processes (5), (6) and (7). After passing the verification, M will send u the message

$$Updt\_Asw = \{M, i, S, seq+1, MAC(Key_{M-u}\{S_i\}, Msg)\} \quad (11)$$

From which M can send S_i to u securely with $E_{key_{M-u}\{S_{i-1}\}}(S) = D_{key_{u-M}\{S_j\}}(S)$ .Upon receiving *Updt_Asw*, u calculates $S_i = h(S_j)$ and verifies the legitimacy of S_i because $S_i = h(S_i)$. Then u calculates the new shared key $Key_{M\text{-}u}\{S_i\}$ and uses it for message authentication as well as node authentication. Finally, u sends *Updt_Ack* message (12) to confirm the establishment of the shared secret key between M and u.

$$Updt\_Ack = \{u, i, seq+2, MAC(Key_{u-M}\{S_i\}, Msg)\} \quad (12)$$

Through the above analysis, we observe that the Blundo symmetric polynomial and reverse hash chain are effectively combined in our KUP. Using the shared key calculated from the pre-loaded key materials between heterogeneous nodes, our KUP can carry out key update and realize securely shared keys establishment and authenticated node revocation.

As a final remark, the MS may be subject to capture. To fight against this attack, effective node revocation and authentication node update are necessary. BS firstly generates a number of identities for MS as shown in (3). All CHs are pre-loaded with MS_0 firstly. After a period of time, the current identity of MS is supposed to be MS_i. If MS_{i+1} is needed, BS will pre-load MS_{i+1} with the relevant symmetric polynomial $f(M_{i+1},y)$ and a new hash value. In this case, MS_{i+1} can send CHs its identity through *Hello* message and CHs can authenticate MS through $MS = h(MS_{i+1})$. If MS_{i+1} is verified, CHs will store it as a legitimate one. Finally, we conclude that the use of reverse hash chain to generate the MS identity, can not only reduce unnecessary storage overheads, but also is resistant to the collusion attacks.

IV. PERFORMANCE ANALYSIS

In this section, we carry out performance for our scheme.

Network Intrusion Tolerant Rate

For the sake of discussion, we assume that the numbers of symmetric polynomials pre-loaded in the MS and cluster heads are a and b, respectively, there are b cluster heads in one subarea, and the total number of subareas is not more than t. Suppose that there are n nodes in each subarea, then the total number of nodes in the network is $N = a \cdot t \cdot n / b$. Our key agreement scheme can resist collusions among no more than t nodes in one subarea. We define the intrusion tolerance rate as the secure links probability shown in (13):

$$p_{Tol}(x) = \frac{L_{All}(x) - L_{Replicated}(x) - L_{Ind\_Replicated}(x)}{L_{All}(x) - L_{Replicated}(x)} \quad (13)$$

where x is the number of nodes Replicated by the adversary through physical capture attack, $L_{All}(x)$ is the total number of links in the network, $L_{Replicated}(x)$ is the number of the Replicated links and $L_{ind\_Replicated}(x)$ is the number of links that are potentially unsecure because of the leaked secrets from the directly captured nodes.

It is clear that, if $n < t$, no matter how many nodes are Replicated, they will not affect the communications security among normal nodes. If $n > t$, the attacker could compromise more than t nodes in one subarea, leading to completely security breach in the subarea. With some mathematical manipulations, we obtain the network intrusion tolerance rate as in equation (14).

$$p_{Tol} = \begin{cases} 1 & 0 \le n \le t+1 \\ g(x) & t+1 < n \le N \end{cases} \quad (14)$$

where $g(x)$ is given by:

$$g(x) = \frac{\frac{N}{n} - \frac{x}{t+1}\left(\frac{n}{2} + \frac{n - [x \bmod (t+1)]}{2}\right)}{\frac{N}{n} - \frac{x}{t+1}\left(\frac{n}{2} - 1 + \frac{n-t-1}{2} + \frac{n - [x \bmod (t+1)]}{2}\right)}$$

In Figure 4, we show the network intrusion tolerance rate for n < 101, n = 200 and n = 400, when the network parameters are a = 2, b = 2, t = 100 and N = 10000. We also compare our scheme with those in [5], [6] and [9].

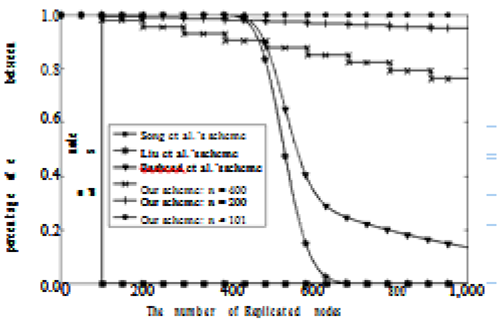

Figure 4. Network intrusion tolerant rate

We observe that, in [6], when the number of Replicated nodes exceeds the threshold t, the entire network will be completely controlled by the attacker. The scheme in [9] can guarantee that when the number of Replicated nodes is not more than 400, the whole network is secure.

B. Network Overhead

Liu et al. [9] point out that the communication overhead of Blundo symmetric polynomial approach is the same as RC5-MAC when the polynomial degree is t = 100 and node identity is 16 bits. Thus, the computation overhead of Blundo symmetric polynomial is 8 times of RC5-MACs when t = 100 and the node identity is 64 bits. From [13], the overhead of computing a MAC is equivalent to the overhead of transferring 1 byte information. It is assumed that the identity of our MS is 64 bytes and each remaining nodes 2 bytes.

We assume that the network is divided into m blocks as done in [6] and k times hash computation are needed to generate a random key in [5]. In addition, we assume that all encryption operations are based on RC5, while the message authentication and hash operations are CBC-MAC based on RC5. Table II shows the overhead of computation and communication in relevant schemes. In Song et al.'s scheme [6] the shared key generation process contains the computing of Merkle tree root, MS identity, symmetric polynomials and *MAC*s, which is equivalent to $11+\log_2{}^m$ bytes data transmission. In [5], the computation overhead consists of the computing of symmetric polynomials and random key and *MAC*s, which is equivalent to k+10 bytes data transmission. In our KAP there are 1 symmetric polynomial communication, 2 hash operations and 2 *MAC*s operations, which is equivalent to 12 bytes communication overhead. However, our KUP contains 1 Blundo symmetric polynomial, 4 hash, 3 *MAC*s, and 1 RC5 operation, which is equivalent to 16 bytes communication overhead. Results are summarized in table II.

TABLE II. COMPUTING AND COMMUNICATION OVERHEADS

Scheme	Computation overhead (RC5-MAC)	Communication overhead (Byte)
Song[6]	$11+\log_2{}^m$	$4\log_2{}^m + 23$
Rasheed[5]	k +10	64
KAP	12	32
KUP	16	78

Song et al.'s scheme needs to send and receive the identities of communicating parties, the authentication on message of MS, *seq* and the *MAC*s. However, Rasheed et al.'s scheme is more complex, requiring ensuring the communicating parties to get two kinds of shared keys. In our KAP, 2 node identities, the identity of current hash value i, 2 *seq* and 2 *MAC*s are needed to transmit, while in KUP, 3 node identities, 3 hash values identification, 3 *seq*, 3 *MAC*s and one encrypted hash value are needed to transmit.

V. CONCLUSION

In this paper, we propose an Modernize and intrusion-tolerant key management scheme for UWSN which uses the Blundo symmetric polynomial and the reverse hash chain technology for key agreement and key updating to prevent Replicated nodes from communicating with static nodes. We conduct detailed analysis and compare the performance of our scheme with related outcomes in terms of intrusion tolerant rate, computing and communications overhead and show the superiority of our proposed schemes.

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A 16-bit MIPS Based Instruction Set Architecture for RISC Processor

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Abstract- Microcontrollers and microprocessors are finding their way into almost every field in today's world, incorporating an element of 'smartness' into conventional devices. Energy efficient, space efficient and optimized microcontrollers are the need of the day. Our paper proposes a new Instruction Set that is a subset of the MIPS architecture. It derives the advantages of MIPS like simplicity and speed. Besides, since it is a smartly optimized subset of MIPS, it is a smaller version consisting of the most commonly required instructions.

Index Terms- ISA, MIPS, Processor design, RISC.

I. INTRODUCTION

MIPS is a reduced instructions set computer (RISC) architecture. It is one of the first RISC Instruction set architectures. MIPS is an acronym for "Microprocessor without interlocked pipeline stages". It was developed by a team led by John Hennessey at Stanford University. MIPS implementations are primarily used in embedded systems such as Windows CE devices, routers, residential gateways, and video game consoles such as the Sony PlayStation 2 and PlayStation Portable. Until late 2006, they were also used in many of SGI's computer products. MIPS implementations were also used by Digital Equipment Corporation, NEC, Pyramid Technology, Siemens Nixdorf, Tandem Computers and others during the late 1980s and 1990s. Since MIPS is a RISC computer it employs less number of transistors and hence decreases the transistor count. Pipelining is thus heavily employed to make use of extra available space on the chip to improve code execution performance. MIPS was defined to be a 32 bit architecture called MIPS32. Later Revisions of this architecture is 64 bit in size and hence called MIPS64. [1]

II. MIPS 16 INSTRUCTION SET DESCRIPTION

A. Motivation

Small scale applications do not require that much of computing power. This paper proposes a reduced version of MIPS instruction set for such small scale applications. This ISA will be called MIPS 16. The main aim of this ISA is to reduce the transistor count of a MIPS processing unit by scaling down the bus and register width and providing less but enough number of instructions for small scale applications. The implementation of such an instruction set would take up less real estate on the chip (or FPGA) and will allow more peripherals to be fabricated on a single chip making it ideal for a System-On-Chip (SOC) implementation of an application. It will also be beneficial in embedded system design where a custom processor core implementation is required with tight instruction requirements so that it takes less space on a FPGA.

B. Instruction Set Specification [3]

MIPS instructions have fixed width. The original MIPS 32 ISA has 32 bits wide instructions. Each instruction in MIPS16 is 16 bits wide. Further, MIPS16 has 8 internal registers as opposed to the 32 registers of MIPS32. As the name suggests, data bus is 16 bits wide and address bus is preferably 16 bits wide too. I/O support is memory mapped. Memory is accessed by LOAD and STORE instructions. The instructions follow an <operand register, register, register> format.

The Instructions can be divided into 4 groups:

a. **Arithmetic**: Basic computational instructions add and subtract.
b. **Logical**: Operations like AND, OR, EXOR
c. **Data Transfer**: Load and Store operations
d. **Branch and control**: Jump, Call, Return, etc.

The ISA supports direct and immediate addressing modes.

C. Instruction Word Format

A MIPS16 instruction is 16 bits wide. Since MIPS uses a Register-Register type of instruction a general instruction specifies two source registers and a destination registers. The format of such an instruction will be

ADD R_{s1}, R_{s2}, R_d

R_{s1} = First source operand register
R_{s2} = Second Source operand register
R_d = Destination register

The instruction word has a **5 bit op-code** specifying the operation to be performed. Number of operands may be variable e.g. ADD requires three operands while NOT requires only two. Format of a three operand instruction word is shown in Table I

TABLE II
THREE OPERAND INSTRUCTION

Op-code	Rs1	Rs2	Rd	Reserved
5 bits	3 bits	3 bits	3 bits	2 bits

In case of ALU instructions, the 2 reserved bits act as function bits where they are used to distinguish between versions of a common instruction. For instance, the instructions ADD and ADC have the same opcode but different function bits. This results in a simpler control logic as the reserved bits are decoded directly by the ALU control logic.

In case of lesser operands appropriate operand is given a constant value. E.g. NOT instruction requires only one source and one destination operand. Therefore, Rs2 field will be made "000" as shown in Table II. Likewise a POP instruction will require only destination and hence both the source operands will be constant and only destination needs to be provided as shown in Table III.

NOT R_s, R_d

TABLE III
TWO OPERAND INSTRUCTION

Op-code	R_s	Constant	R_d	Reserved
5 bits	3 bits	000	3 bits	2 bits

POP R_d

TABLE III
ONE OPERAND INSTRUCTION

Op-code	Constant	Constant	R_d	Reserved
5 bits	000	000	3 bits	2 bits

The unused fields in an instruction are also used to provide immediate input. The size of the immediate field depends on the number of operands instruction uses.

ADDI R_s, R_d, #10

R_s = Source Register

R_d = Destination register

#10 = Immediate value (0-31 in decimal)

TABLE IV
IMMEDIATE INSTRUCTION

Op-code	R_s	R_d	Immediate
5 bits	3 bits	3 bits	5 bits

Instructions like the move immediate(MVIH / MVIL) require a 8-bit value to be specified within the instruction. In such a case, the 8 bt value is split into 2 parts. The higher 3 bits are specified in place of the R_{s1} operand and the next lower 5 bits are specified in the lower 5 bits of the instruction.

TABLE V
MOVE IMMEDIATE INSTRUCTION

Op-code	R_s	R_d	Immediate
5 bits	7:5 bits of imm value	3 bits	4:0 bits of imm value

Jump instructions have two modes viz..PC relative and absolute modes. In PC relative mode. The lower 5 bits of the instruction are used to specify a 5 bit signed value as shown in Table IV. This value is added/subtracted to the PC to get the jump address. The PC relative mode is used for conditional jump instructions. The absolute mode is used for unconditional jumps and jump-&-link instructions. In these instructions the all bits other than the opcode are used to specify a 11 bit signed PC offset value.

The instruction set is so designed so as to simplify the instruction decoding logic and the control logic.

TABLE VI
JUMP (ABSOLUTE MODE) INSTRUCTION

Op-code	Immediate
5 bits	11 it signed PC Relative offset

D. Comparison between MIPS-16 and MIPS-32 [7]

MIPS-16 can be considered to be a derivative of MIPS-32 instruction set. But the philosophies behind their design are different. MIPS-16 provides more flexibility in terms of optimizing the design by keeping only the required instructions. MIPS-16 is designed for small scale applications while MIPS-32 is a high performance 32-bit architecture which can handle large data and perform fast calculations by employing multiple pipelines and multiple registers at the cost of larger chip area and complicated logic design.

Some key differences have been highlighted in TABLE VII

TABLE VII
COMPARISON OF MIPS-16 AND MIPS-32 ISA

Serial No.	MIPS-16 Instruction set	MIPS-32 Instruction set
1	Instruction word length is 16-bit	Instruction word length is 32-bit
2	Supports only 8 general purpose registers	Supports up to 32 general purpose registers
3	Register is 16 bits wide	Register is 32 bits wide
4	Program counter should be incremented by 2 after every instruction (for non-branching instructions)	Program counter should be incremented by 4 after every instruction (for non-branching instructions)
5	ALU is simpler. It does not support operations with bulky logic like Multiplication and Division	ALU is complicated. It supports complicated operations like Multiplication and Division.
6	Floating point instructions are not	Floating Point instructions are included

	included in MIPS-16	and are called SIMD instructions
7	Pipelining is not essential and depends on the application.	Pipelining is a key feature of a MIPS-32 based processor.
8	Transistor count and chip area is less	Transistor count and chip area is more
9	Suitable for small scale applications and applications with low computing requirement	Suitable for high performance high throughput applications.

E. List of Instructions

As the op-code has a 5 bit length there are 32 possible distinct instructions. If the reserved bits at the end of the instruction are utilized for grouping 2 or more similar instructions more op-codes can be incorporated in the instruction set e.g. ADD and ADC can be grouped together as they perform similar function with the difference being inclusion of carry into the sum. A complete list of 37 instructions has been provided in Table VIII with a short description of each instruction.

TABLE VIII
MIPS-16 INSTRUCTION SET

Sr. No.	Mnemonic	Instruction Format	Description
1.	ADD	ADD R_{s1}, R_{s2} ,R_d	Adds R_{s1} and R_{s2} and stores the sum in R_d ignoring carry.
2.	ADC	ADC R_{s1}, R_{s2} ,R_d	Adds R_{s1} and R_{s2} and stores the sum in R_d with previous carry.
3.	SUB	SUB R_{s1}, R_{s2} ,R_d	Subtracts R_{s2} from R_{s1} and stores the difference in Rd ignoring the previous borrow.
4.	SBB	SUB R_{s1}, R_{s2} ,R_d	Subtracts R_{s2} from $R_{s1}$and stores the difference in Rd with the previous borrow.
5.	AND	AND R_{s1}, R_{s2} ,R_d	Performs Bitwise AND of R_{s1} and R_{s2} and stores the result in R_d
6.	OR	OR R_{s1}, R_{s2} ,R_d	Performs Bitwise OR of R_{s1} and R_{s2} and stores the result in R_d
7.	XOR	XOR R_{s1}, R_{s2} ,R_d	Performs Bitwise XOR of R_{s1} and R_{s2} and stores the result in R_d
8.	NOT	NOT R_{s1} ,R_d	Performs Complement of R_{s1} and stores the result in R_d
9.	SHIFTL	SHIFTL R_{s1} ,R_d	Shifts R_{s1} by one place to the left and store it in R_d
10.	SHIFTR	SHIFTR R_{s1} ,R_d	Shifts R_{s1} by one place to the right and store it in R_d
11.	ADDI	ADDI R_{s1} ,R_d,#5-bit	Adds a 5-bit unsigned value to R_{s1} and stores the sum in R_d
12.	SUBI	SUBI R_{s1} ,R_d,#5-bit	Subtracts a 5-bit unsigned value from R_{s1} and stores the difference in R_d
13.	MOV	MOV R_{s1} ,R_d	Copies R_{s1} to R_d
14.	MVIH	MVIH R_d,#8-bit	Copies immediate value into higher byte of R_d
15.	MVIL	MVIL R_d,#8-bit	Copies immediate value into lower byte of R_d
16.	LDIDR	LDIDR R_{s1} ,R_d,#5-bit	Loads R_d with a nibble at address given by [R_{s1} +5 bit immediate value]
17.	STIDR	STIDR R_{s1} ,R_d,#5-bit	Stores R_d with a nibble at address given by [R_{s1} +5 bit immediate value]
18.	LDIDX	LDIDX R_{s1}, R_{s2} ,R_d	Loads R_d with a nibble at address given by [R_{s1} + R_{s2}]
19.	STIDX	STIDX R_{s1}, R_{s2} ,R_d	Stores R_d with a nibble at address given by [R_{s1} + R_{s2}]
20.	JMP	JMP #11-bit	Unconditional jump to address offset by 11 bit signed value from current PC value
21.	JMPI	JMPI R_d,#15	Unconditional jump to address offset by 5 bit signed value added to R_d
22.	JGEO	JGEO R_{s1} ,R_{s2},#5-bit	Conditional Jump to PC + 5 bit signed offset if R_{s1} is greater than or equal to R_{s2}
23.	JLEO	JLEO R_{s1} ,R_{s2},#5-bit	Conditional Jump to PC + 5 bit signed offset if R_{s1} is less than or equal to R_{s2}
24.	JCO	JCO #5-bit	Conditional Jump to PC + 5 bit signed offset if carry is set
25.	JEO	JEO R_{s1} ,R_{s2},#5-bit	Conditional Jump to PC + 5 bit signed offset if R_{s1} equals to R_{s2}
26.	PUSH	PUSH R_{s1}	Push R_{s1} to the stack top and update stack top
27.	POP	POP R_d	Pop from the stack top and store the value to R_d and update stack top
28.	CALL	CALL R_{s1}	Calls a subroutine located at [R_{s1}]. Return address is pushed onto stack
29.	JAL	JAL #11-bit	Calls a subroutine located at [PC + 11 bit signed offset].Return address is pushed onto stack.
30.	MOVSP	MOVSP R_{s1}	Copies value at R_{s1} to stack pointer SP
31.	RET	RET	Return from a function. Return address is popped from the stack
32.	STC	STC	Set the carry flag
33.	NOP	NOP	No operation. Idle machine cycle should be executed
34.	HLT	HLT	Halts the processor.
35.	RST	RST	Resets the processor.
36.	IE	IE	Enables the interrupt.
37.	ID	ID	Disables the interrupt.

F. Implementation Strategies

The implementation strategies that can be employed will depend on the application. Some of the design considerations are listed below:

a. *Single-Cycle or Multi-cycle implementation*: [3], [4], [5]
Implementation can use a single cycle or a multi cycle control system for its datapath.
A single cycle control system performs all the elementary datapath operations in a single cycle. This generally requires dedicated hardware for every phase of instruction fetching, decoding and execution. It is faster at the cost of larger chip area.
A multi cycle implementation divides the execution of an instruction into well-defined time states. The execution happens in a timely order and might require different number of time states for different instruction. The main advantage is the hardware can be shared for similar elementary functions in different time states. Multi cycle should be preferred for applications with smaller chip area requirements.

b. *Pipelining Requirement:*[6]
Pipelining provides performance enhancement by concurrent execution of more than one pair able instructions. The design involves use of multiple datapaths and a logic to check for pairability and hazard removal that occurs due to concurrent execution. This significantly complicates the design and takes a larger chip area. But the performance improvement would be tremendous.

III. CONCLUSION

MIPS-16 is thus a low-cost, compact and hence in effect a low power RISC instruction set architecture as compared to the MIPS-32 architecture. Its compact size and flexibility makes it ideal for an optimized implementation of an embedded system. It provides all the basic instruction and functionality for a small scale embedded system not involving heavy arithmetic calculations. MIPS-16 can be implemented on FPGA by an appropriate strategy as per the application's requirement. Single cycle design should be used for better performance while multi cycle design should be preferred for compactness. Pipelining can further improve the system performance.

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Area Configuration and Link Failure Effect in IP Networks using OSPF Protocol

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Abstract- Open Shortest Path First (OSPF) protocol is widely deployed in IP networks to manage intra-domain routing. An OSPF is a link-state protocol, in which routers establish relationships developing neighbors, enabling each to build a consistent, global view of the routing topology. This paper is based on overview of OSPF comprising different types of routes, routers, networks, areas, and protocols/processes used in OSPF. We used OPNET IT Guru Academic Edition 9.1 network simulator for six different network scenarios to calculate the shortest path from source to destination router. The simulation developed with different parameters and is configured based on OSPF area and link failure using traffic sent in bits/seconds as performance metric. Finally, the research experiment results showed that balance area OSPF is performed better in finding the shortest path in traffic sent after a link fails.

Index Terms- OSPF routes, ABR, ASBR, BR, Packet formats, OPNET

I. INTRODUCTION

Routers routes the packets from source to destinations using a road map. Routing can be performed by two ways: static and dynamic. The paths in routing can be defined manually and routes can be installed within the routers in static routing, but this manual work can be replaced by implementing routing algorithms. Dynamic routing analyzes the network traffic to determine the shortest path from source to destination using various metrics. For network routing there are two types of protocols are needed: routed protocols and routing protocols. Routed protocols are used as a means of communication between the underlying devices used in network. Internet Protocol (IP) is the most widely used protocol under the routed protocol category whereas the other examples are AppleTalk, DECnet. Routing protocol is used to determine routes among devices used in network from source to destinations such as Routing Information Protocol (RIP), Enhanced Interior Gateway Routing Protocol (EIGRP), Open Shortest Path First (OSPF) protocol, Intermediate-System to Intermediate-System (ISIS) and Border Gateway Protocol (BGP). The relationship between the routed and routing protocols is routed protocol relies on routing protocol for transport of packets to determine the optimal path over the Internet. Routing protocol maintains routing tables to determine optimal routes by comparing different metrics such as hop count, cost of transmitting link, bandwidth, Maximum Transfer Unit (MTU), and packet delay [2],[7],[8],[9],[10].

II. OSPF ROUTES

Depending upon the router interface and associated AS, there are three types of OSPF routes used in OSPF areas: intra-area routes, inter-area routes and external routes [8],[10]. This is shown in Figure 1. The first two types of routes are concerned directly with the locations of router used in OSPF areas. Intra-area routes form with the interfaces of routers within the same OSPF area whereas inter-area routes form with the interfaces of routers located at different OSPF areas. These both types of routes are internal to OSPF area generated by itself. As the name suggested, external routes are external in the sense that they connects with routers those running other protocol such as RIP, BGP in remote AS to OSPF area. External routes are classified as E1 route and E2 route. The default external route is E2 route. E1 route calculates the route from source by adding the internal and external cost to reach the destination, but E2 routes only considers external cost.

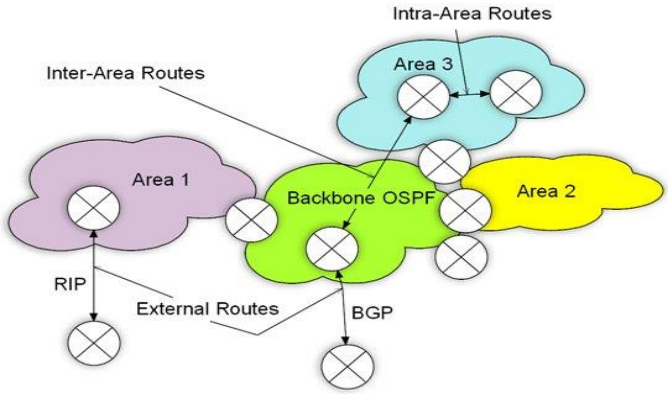


Fig-1: OSPF Route Types

III. OSPF ROUTERS

There are four types of routers used in OSPF as: Internal Routers (IR), Backbone Router (BR), Area Border Router (ARB) and Autonomous System Border Router (ASBR) [7],[8,[10]. This is shown in Figure 2.

IR: Routers those belong to the same OSPF area used for maintaining single current database useful for finding shortest path for other connected areas.

Backbone Router: Routers whose interfaces are connected to the backbone area.

ABR: Router whose interface connects multiple area with backbone area. The responsibility of ABR is to summarize

routing information of each connected area and present to the backbone area for further distribution.

ASBR: Routers those redistribute the routing information between OSPF area and non-OSPF running areas.

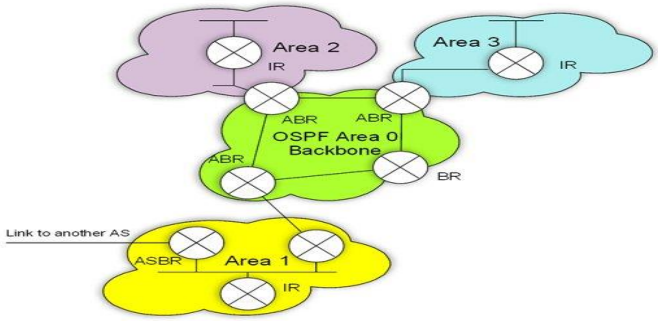


Fig-2: OSPF Router Types

IV. OSPF NETWORK TYPES

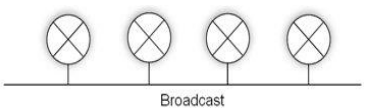
OSPF networks can implement in four different types as Broadcast, Non Broadcast Multi Access (NBMA), Point-to-Point and Point-to-Multipoint [1],[4],[9],[10],[11]. This is shown in Figure 3.

Broadcast: Ethernet or Fiber Distributed Data Interface (FDDI) connects two or more OSPF routers to access multiple end points results in saving bandwidths, faster neighbor discovery and higher scalability. OSPF Hello packets are used to develop neighbor relations. An adjacency can develop with help of electing designated router (DR) and backup designated router (BDR). Neighbor can identify each other by adjacency.

NBMA: Frame relay, ATM or X.25 network use to form NBMA networks. NBMA emulates the function of a broadcast network. OSPF routers need to configure with the IP address of each of its neighbor for establishing relationship. OPPF Hello packets transmitted to each adjacent neighbor forms adjacency.

Point-to-Point: This is the simplest type of network that OSPF can implement. Leased lines or High-level Data Link Control (HDLC) connects two or more routers allowing only single-neighbor relationship. There is no need of DR and BDR in point-to-point networks.

Point-to-Multipoint: Routers can be manually configured, if they connected to a non-broadcast network. In real sense, it is a point-to-point network implemented in multiple manners with Private Virtual Circuits (PVCs). This network can occupy a common subnet without electing DR and BDR.



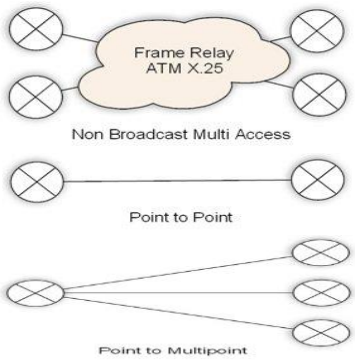


Fig-3: OSPF Network Types

V. OSPF AREAS

OSPF area is identified by a 32-bit number. OSPF uses two broad categories of areas as backbone area and regular area. The backbone area is numbered as 0.0.0.0. The backbone area is the base for other OSPF areas. This area interconnects other areas for fast and efficient transmission of IP packets to find shortest path. Regular area provides user and allows different resources connection. Regular area is configured into different structure as stub area, totally stub area, and Not So-Stuby Area (NSSA). Stub area provides a single exit to outside network forming a dead end. The routers in stub area can reach to outside network using a default route injected by ABR. Totally stub area only allows intra-area routes and default routes within the area. These default routes are used to send traffic outside the area. NSSA allows importing external routes into stub area. ASBR converts stub area into NSSA by connecting an external network running other routing protocol [3],[5],[9]. This is shown in Figure 4.

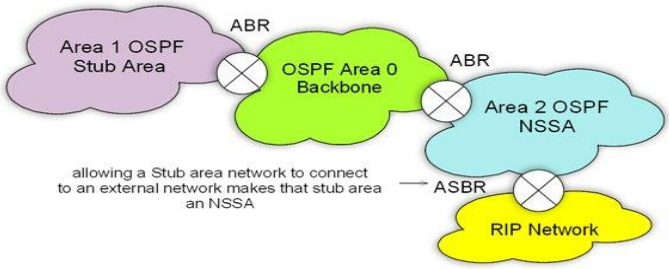


Fig-4: OSPF Area Types

VI. OSPF PACKETS/PROCESS

All OSPF packets uses a 24-byte header contains information that determines the behavior of packet to be processed by every OSPF router. OSPF routers runs on top of IP and communicate with each other by using OSPF sub-protocols for establishing and maintaining neighbor relationships, exchanging their link state information by advertisements (LSAs), validating and distributing link state database (LSDB) updates using flooding [9],[10],[11],[12]. The OSPF packet header is shown in Figure 5.

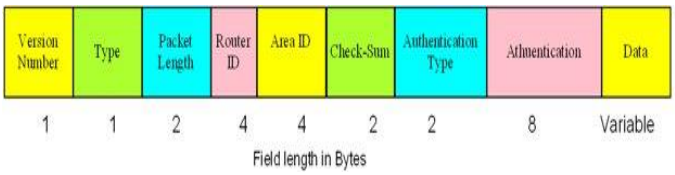


Fig-5: OSPF Packet Header

The OSPF packet header contains eight fields: Version Identifies the OSPF version being used. Type specifies the type of one of the five different types of OSPF packet. The OSPF packet is used for different routine functionality of OSPF. Hello packet (Type 1) is used for discovering and maintaining neighbor routers. Database description (Type 2) summarizes the database contents. Link state request (Type 3) are used to download topological database from neighbor routers. Type 2 and Type 3 are required to form the adjacency. Link State update (Type 4) is used to update LSDB whenever network changes by LSAs. Link state acknowledgment (Type 5) simply floods the acknowledgements to database for updates using Type 4. Type 4 and Type 5 provides reliable update mechanism. Packet Length specifies the length in bytes including header. Router ID is unique 32-bit number assigned to the router. Single OSPF area can be identified by assigning Area ID. Security can be assigned with checksum field for checking any damage to the packet during transmission. Authentication type and Authentication is used to measure the exchange of OSP packets in authenticated manner. Data field is responsible for carrying upper layer information which is encapsulated.

OSPF packet processes are Hello, flooding and exchange of OSPF packets in OSPF networks.

Hello packets are used to discover the neighbors and exchange the initial parameters in strict order to establish bidirectional communication. In order to exchange LSDB, OSPF routers discover Hello packets before validating time. The different fields used in Hello packet is shown in Figure 6.

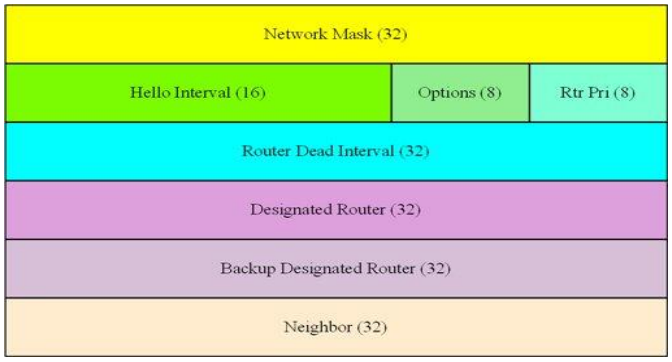


Fig-6: Hello Protocol Packet Format

Network mask is associated with the network interface. HelloeInterval field is used to set a time in seconds from the OSPF router to communicate with the neighbors. Precedence setting in number is required to elect Designated Router (DR) in the subnet. This is stored in the Rtr Pri field. RouterDeadInterval is used to note the state of OSPF router within which it should respond to the Hello request. IP address of the network's DR and Backup designated Router (BDR) using Designated Router and Backup Designated Router fields respectively. An OSPF router maintains list router IDs of each router that has sent a valid Hello packet using Neighbor field.

When any change or update occurs at interfaces of OSPF routers, link state updates are flooded to every OSPF router's LSDB. Flooding is used for LSDB synchronization. The packet layout used in flooding process carries with number of advertisements and LSAs. This is shown in Figure 7.

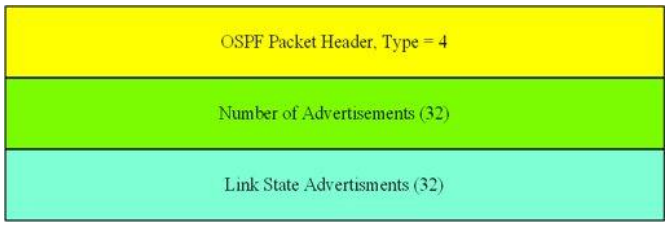


Fig-7: Flooding Process/Protocol

The exchange process is based on bidirectional communication used for synchronizing the LSDB. After synchronization, any changes in the router's links use the flooding protocol to update all the OSPF routers. The requirement for exchange process is, one of the two routers is master and other is slave. The important field of exchange protocol is shown in Figure 8. The DD sequence number is used to sequence the collection of Database Description (DD) packets. The various types of LSAs are Router, Network, summary (IP Networks), summary (ASBR). The time in seconds for LSA is recorded in LS age field since the LSA was originated. Link state checksum is used for storing complete contents of LSA.

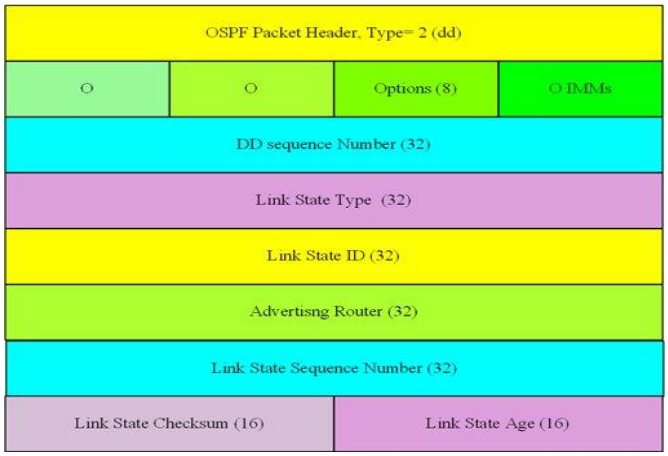


Fig-8: Exchange Process/Protocol Packet Format

VII. Simulation Setup

Implementation is done in OPNET IT Guru Academic Edition 9.1. The simulation scenario setup is given in Table 1.

Table-1: Simulation Setup

Parameters	**Values**
Routing Protocol	OSPF
Network type	Campus

Scale	10 Km x 10 Km	
IP Address Family	IPv4	
Router	Ethernet4-Slip8-Gtwy	
Send Style	Broadcast	
SPF calculation style	Periodic	
	Delay time	Hold time
	5 sec	10sec
Reference Bandwidth	1000 Mbps	
Simulation Time	10 minutes	

There are six network scenarios, which are configured as:
The scenario N0_AREA_OSPF is configured using the parameters given in Table 4 and the cost as the interface metric information. The traffic demand is set between Router1 and Router10 for traffic sent in bits/seconds. This is shown in Figure 9. In AREA_OSPF scenario, we duplicated the prior scenario by breaking OSPF area into three areas as Area1, Area2 and default Area0 called as backbone of OSPF. This is shown in Figure 10 with different colors. Now, we deploy packet based load balancing between the same pair of routers using BALANCED_AREA_OSPF scenario. We establish link failures between the pair of routers Router1-Router4 and Router8-Router10 by introducing a failure object with delay 0f 100 seconds in all the prior three scenarios duplicated as NO_AREA_FAILURE_OSPF, AREA_FAILURE_OSPF and BALANCED_AREA_FAILURE_OSPF shown in Figure 12, Figure 13 and Figure 14 respectively.

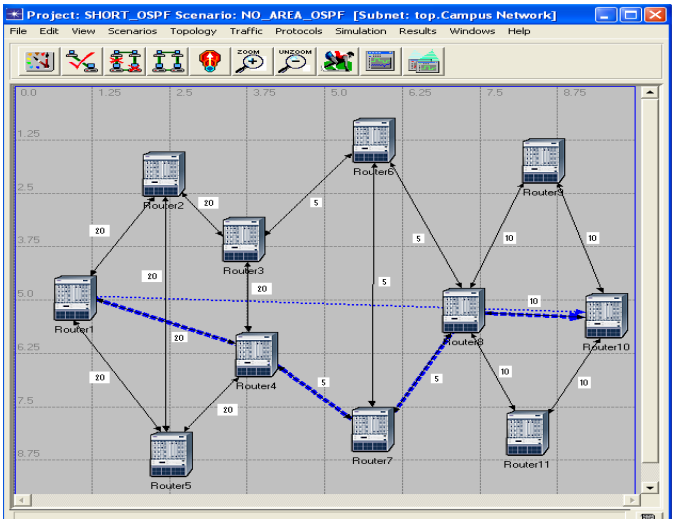


Fig-9: Scenario N0_AREA_OSPF

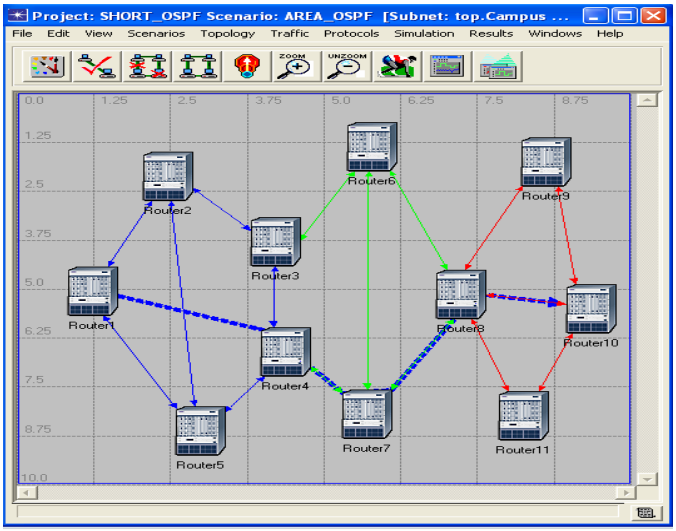


Fig-10: Scenario AREA_OSPF

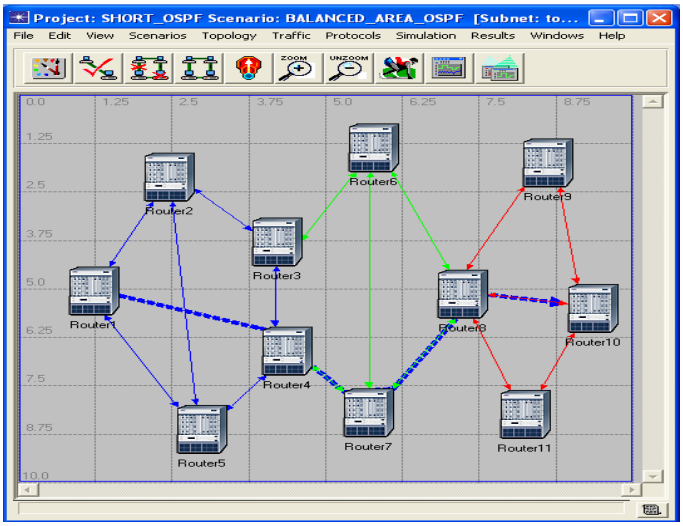


Fig-11: Scenario BALANCED_AREA_OSPF

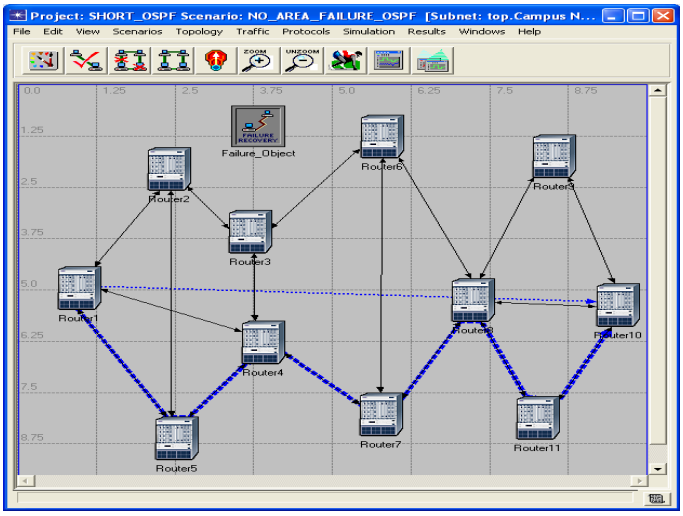


Fig-12: Scenario N0_AREA_FAILURE_OSPF

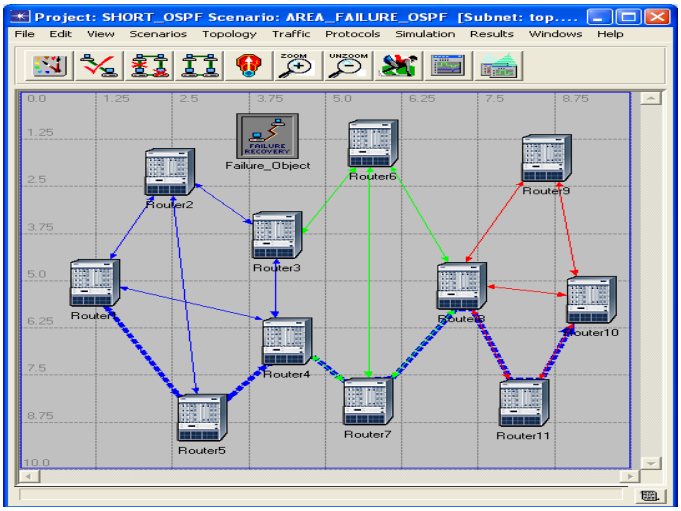


Fig-13: Scenario AREA_FAILURE_OSPF

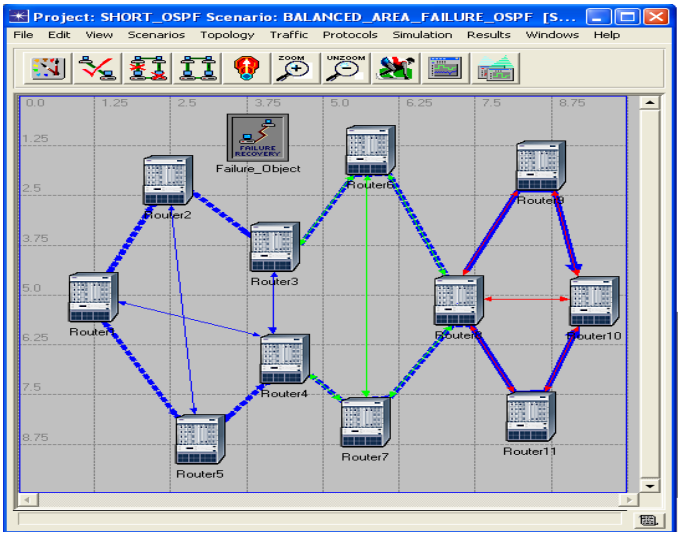


Fig-14: Scenario BALANCED_AREA_FAILURE_OSPF

VIII. SIMULATION RESULTS AND ANALYSIS

We analyzed the obtained result based on shortest path selection and traffic sent (bits/seconds) for six different scenarios [7].

The shortest path from Router1 to Router10 is calculated for all six scenarios with different path selection. In case of the first three scenarios without introducing link failures, there is no change in path selection. There is no scope in balance area OSPF without failure scenario for alternate shortest path due to existence of link pairs between Router1-Router4 and Router8-Router10. This is shown in Figures 8, 9 and 10. But, change in shortest path selection effect is observed in the last three scenarios where we introduced two link failure pairs for Router1-Router4 and Router8-Router10 for delay of 100 seconds. There are two equivalent shortest paths in balance area failure scenario from Router1 to Router10. This is shown in Figure 12, Figure 13 and Figure 14.

The traffic sent (bits/seconds) in balance area failure scenario is highest among all the scenarios started with sharp increase upto 17250 bits and after 40 seconds slowed downward to 3750 bits at the simulation end time. This is due to higher traffic created in searching two equivalent shortest paths from Router1 to Router10.

The area scenario is observed with fine increased traffic sent upto 15250 bits over to no area scenario. This means the OSPF protocol increases its performance by administering via areas. But, once we removed area OSPF, then the traffic sent raised initially upto 10000 bits and then downward to 2000 bits. There is significant similarity between the results obtained in area and balanced area scenarios. This is clearly shown in Figure15 where both graphs are overlaid. Finally, it is found that the traffic sent without link failures is higher than link failures scenarios except for balanced area OSPF.

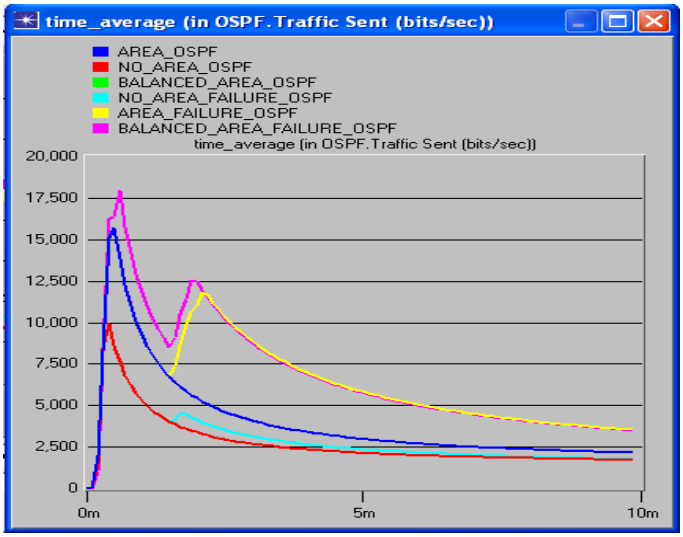


Fig-15: Overlaid Statistics

IX. CONCLUSION

In the start of this paper, we overviewed the concept of OSPF protocol focused on its types of route, networks, areas, routers and subprotocols and their packet formats. An implementation for the OSPF protocol in OPNET IT Guru Academic Edition 9.1 is configured successfully using various parameters and scenarios. We studied the effect of area configuration and link failures in IP networks for shortest path calculation in balanced and unbalance load configuration using OSPF protocol. The comparative results showed that the balance area OSPF is performed better in finding the shortest path for traffic sent after a link fails for a delay of 100 seconds giving two alternate routes from source to destination routers.

ACKNOWLEDGMENT

We are grateful to OPNET Technologies, Inc. for permitted us to download licensed network simulator OPNET IT Guru Academic Edition 9.1.

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The quality of 'egusi' melon [(Citrullus lanatus Thunb.) Matsum and Nakai] seeds derived from fruits harvested at different growth stages and at different positions on the mother plant

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Abstract- A study of two cultivars of 'egusi' melon [Citrullus lanatus (Thunb.) Matsun and Nakai] known as bara and serewe, was undertaken at the Teaching and Research Farm of the University of Agriculture Makurdi in 2010. The aim was to study the effects of variable developmental stages and position of fruits on the mother plant on seed quality. Bulk crop was raised and fruits that developed from date tagged flowers were harvested at three days interval from 20 DAA to 47 DAA to monitor seed development and maturation. In another experiment, basal, middle and apex flowers were date tagged as they developed along the growing twine and fruits that developed were harvested at full maturity to monitor seed quality as influenced by fruit position. Mass maturity was attained between 35 and 38 DAA in the two cultivars and highest germination percentages of 85 and 87 were obtained at the last harvest (47 DAA) for both cultivars. Germination of fruits positioned at the base and middle, which showed no significant differences in themselves were significantly superior to apex fruits. It was therefore concluded that seed crop of the two cultivars should be delayed for up to 47 DAA or beyound and preference should always be given to seeds from base and middle positioned fruits against apex fruits.

Index Terms- Anthesis, fruit position, assimilates, maturation, seed quality, germination.

I. INTRODUCTION

'Egusi' melon is among the most popular African indigenous vegetable crops produced in Nigeria on a large scale. Schippers (2002) acknowledged the crop as occupying a vital role in the income generation ability of subsistence Africans. Ojo *et al.* (2002) reported a seed yield in Nigeria ranging between 131 to 1005kg /ha. Bankole and Adebanjo (2003) acknowledged that melon seeds (*Colocynthis citrullus* L.) are an important indigenous oil seed consumed by many rural communities in West Africa. Dehulled seeds contain oil, protein, amino acids, and some amounts of vitamins (John, 2002). Because of the nutritional and economic importance of 'egusi' melon, it is imparative that adequate studies should have been made to improve on its seed production practices. However, most African researchers generally pay more attention to agronomic crops while vegetable crops suffer neglect due to ignorance about their nutritive value. Bellin-Sesay (1996) reported that consumption of vegetables is far from sufficient in almost all developing countries. According to their report, two billion people, mostly women and children, are deficient in one or more micronutrients, which have been known for centuries as being responsible for diseases such as goitre, critinism, loss of sight, e.t.c. Nevertheless, vegetables still play a subordinate role in the discussion of food security in developing countries.

Farmers have always witnessed a wide variation in germination and emergence of melon seeds in the field and therefore target plant population densities are hardly met. Consequently, high seed rates are used. NIHORT (2000) recommended the sowing of four seeds per hole and that where seedling emergence is impressive, thinning to two plants per hole should be done. This practice is not only labourious but also wasteful.

Melon is a plant with an indeterminate growth, which continues to develop fruits progressively along the length of the stem. This means that melon fruits produced at the base of the stem are older, and continue to reduce in age, as they get closer to the apex of the stem. Melon farmers however, conduct harvest operations the same day and do not grade fruits according to age or fruit possition on the mother plant before seed extraction. A typical melon seed lot is therefore usually composed of seeds whoose individual ages vary widely. Seed age has however been reported, as been a major determinant of seed quality. (Demir and Samit, 2001; Oladiran and Kortse, 2002; Demir *et al*., 2004;). Since the seeds are also derived from fruits at different positions on the mother plant therefore, the food composition of individual seeds in the lot also varies. While seeds from fruits harvested at the base may be fully filled, seeds from the apex fruits may not have had enough time to be fully filled before senescence. Fruit position on the mother plant has however, also been reported as having an effect on seed quality (Alan and Eser, 2007; Ibrahim and Oladiran, 2011; K'Opondo, 2011) This study was therefore conducted to determine if seed age at harvest and fruit position on mother plant also contribute to the poor seed quality usually witnessed by melon farmers.

II. MATERIALS AND METHODS

'Egusi' melon [*Citrullus lanatus* (Thunb.) Matsum and Nakai] of cultivars bara and serewe were produced at the Crop Production Research farm of the University of Agriculture Makurdi in 2010. Planting was done on the flat at a spacing of 2m x 2m on 7th October 2010. Bulk crop was raised and each female flower was date tagged at anthesis to monitor fruit age. Fruits that developed from the tagged flowers were harvested at three days interval starting from 20 to 47 days, i.e. 20, 23, 26, 29, 32, 35, 38, 41, 44 and 47 days after anthesis (DAA). At each harvest, the seeds were extracted, washed, and dried. Dry seed weight per fruit, 100-seed weight and germination percentages were then determined.

In another study, the first five fruits of each plant were tagged as base fruits, the second five as medium, while the rest were left as apex fruits. Harvesting was done for all the fruit positions on the same day at maturity. Post harvest handling of fruits and determination of data was done on the same parameters as in the first study. This study was conducted to ascertain if there are variations in the quality of seeds derived from the three fruit positions.

Seeds samples from both experiments were tested for viability. Germination tests were conducted on four replicates of 50 seeds each, spread over distilled water-moistened absorbent paper in Petri dishes and incubated at 30° C for 28 days. Counts were taken every other day.

III. RESULTS

Fruit age

Although the two cultivars recorded progressive inreases in fruit length with age, significant increases between each succeeding harvests were were only recorded at 23 and 47 DAA respectively in bara and 41 DAA in serewe (Figure 1).

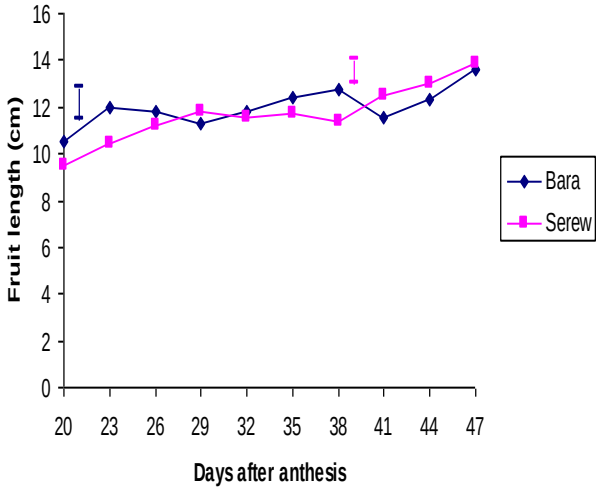

Fig. 1 Variations in average fruit length of bara and serewe harvested at different days after anthesis.
LSD at P = 0.05

Figure 2 shows that though there were progressive increases in diameter of fruits of the two cultivars with age, significant differences in bara were only recorded at the second harvest (23 DAA) and harvest at 41 and 44 DAA which were not significantly different in themselves. Serewe however, recorded significant increases at the second harvest (23 DAA) and harvest at 29 DAA. Thereafter, no other significant differences were observed until the last harvest (44 DAA).

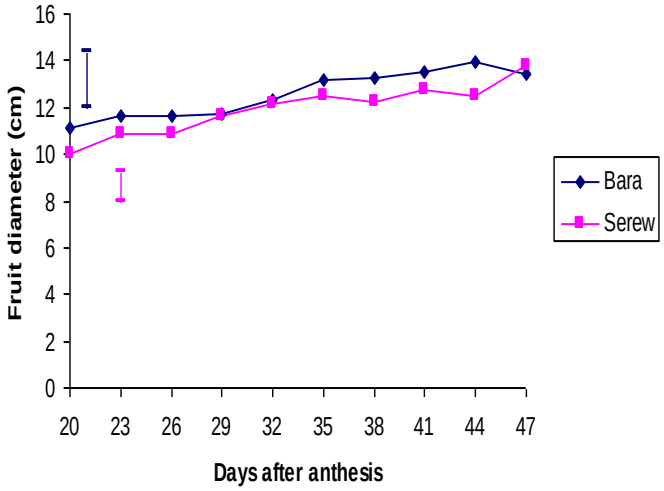

Fig. 2 Variations in average fruit diameter of bara and serewe harvested at different days after anthesis.
LSD at P = 0.05

As shown on Figure 3, no significant increase in fruit weight was observed in bara from 20 to 29 DAA. However, fruits harvested at 32 DAA were significantly heavier than those harvested at 20 DAA. Fruits harvested from 38 to 47 DAA, which showed no significant differences in themselves, also weighed significantly heavier than the later. Serewe fruits harvested at 29 DAA weighed significantly heavier than harvest at 20 DAA. Thereafter, further succeeding harvests, whoose weight increased progressively in succession did not record any significance until the last harvest at 47 DAA.

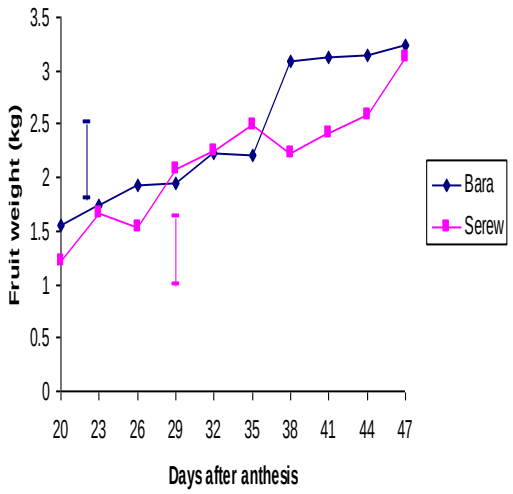

Fig. 3 Variations in average fruit weight (kg) of bara and serewe harvested at different days after anthesis.
I LSD at P = 0.05

Number of seeds per fruit of the two cultivars also increased progressively from inception of harvest without significant differences until harvest at 35 DAA, which recorded a significant difference to the first harvest (20 DAA). Thereafter, further succeeding harvests increased progressively to the last harvest without significant differences (Figure 4). The total number of seeds produced by serewe was greater than those produced by bara.

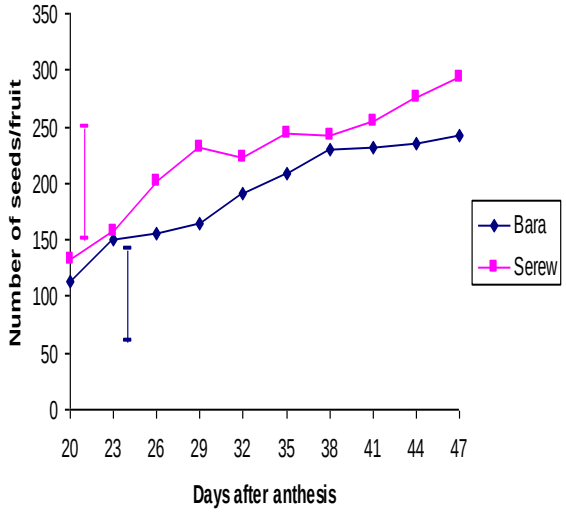

Fig. 4 Variations in average number of seeds per fruit of bara and serewe harvested at different days after anthesis.
I LSD at P = 0.05

Following the same trend as observed in number of seeds per fruit, the dry seed weight per fruit of Bara cultivar also recorded progressive insignificant increases from inception of harvest until 35 DAA when a significant difference was recorded with the first harvest (20 DAA). Further succeeding harvest recorded no significant differences among themselves. Figure 5, however shows that in serewe a slight variation from the foregoing trend was recorded as it was seeds of harvest at 32 DAA that weighed significantly higher than the first harvest (20 DAA). In contrast to the greater seed number produced by serewe, bara produced a greater seed weight per fruit.

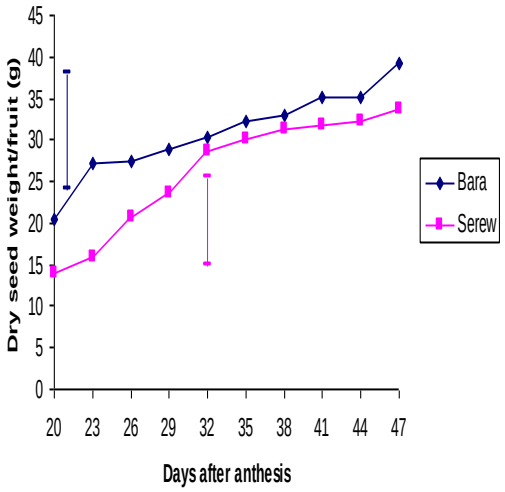

Fig. 5 Variations in average dry seed weight (g) per fruit of bara and serewe harvested at different days after anthesis.
I LSD at P = 0.05

Progressive increases in 100-seed weight of bara were observed which were not significant until at 38 DAA. Further seed weight recorded thereafter until the end of harvest were insignificant. Following the same trend, serewe also produced insignificant progressive increases until 35 DAA.Thereafter, there were no significant differences until the end of harvest. The 100-seed weight of bara cultivar were also greater than those of serewe even though as earlier observed, serewe produced more seeds per fruit.

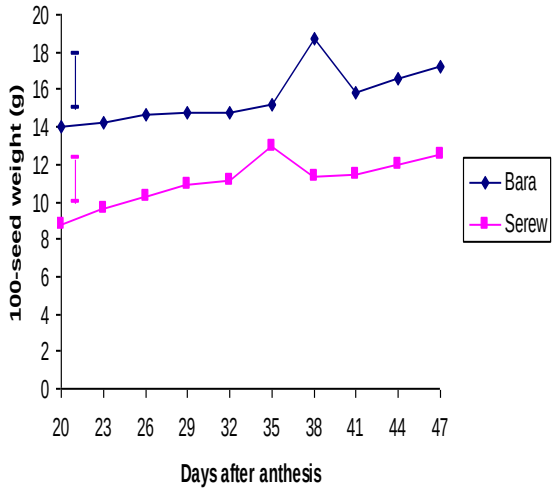

Fig. 6 Variations in 100-seed weight (g) per fruit of bara and serewe harvested at different days after anthesis.

I LSD at P = 0.05

A general progressive improvement in germination was observed in bara throughout seed development period. Significant increase in germination was recorded between 20 and 26 DAA and further between 26 and 32 DAA. Thereafter, no significant increase was observed to the end of production period (Figure 8). Serewe also recorded significant increase in germination between 20 and 26 DAA and further between 23 and 35 DAA. No other significant increase was observed thereafter until the last harvest at 47 DAA in which the improvement in germination was significantly better than all other harvests.

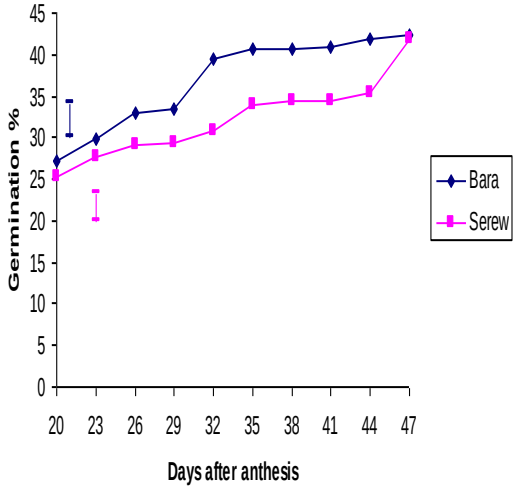

Fig. 7 Variations in 100-seed weight (g) per fruit of bara and serewe harvested at different days after anthesis.

I LSD at P = 0.05

Fruit position on mother plant

Though bara fruits harvested at the base position were longer and decreased retrogressively upwards towards the apex, analysis revealed that significant difference only existed between the base and apex fruits while middle fruits bore no significance with both the base and the apex fruits. In serewe however, base fruits were significantly longer than both middle and apex fruits (Figure 10).

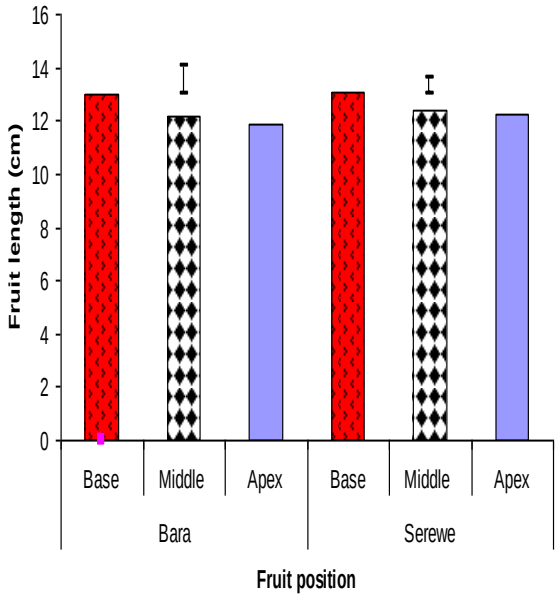

Fig. 10 Variations in the average fruit length (cm) of bara and serewe fruits harvested at different positions on mother plant.

I LSD at P = 0.05

Base fruits of bara significantly recorded greater diameters than middle and apex fruits (Figure 11). However, the differences in diameters of serewe fruits at the different fruit positions were insignificant.

Fig. 11 Variations in the average fruit diameter (cm) of bara and serewe fruits harvested at different positions on mother plant.

LSD at P = 0.05

Although the weights of of both cultivars (bara and serewe) fruits harvested at the base obtained a higher value and progressively retrogressed towards the apex, analysis revealed that there were no significant differences between them (Figure 12).

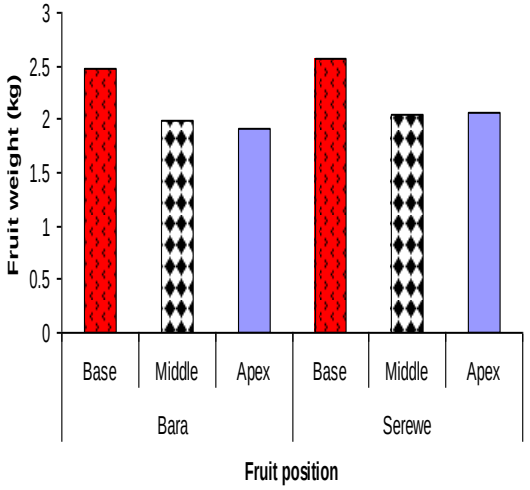

Fig. 12 Variations in the average fruit weight (kg) of bara and serewe fruits harvested at different positions on mother plant.

Number of seeds per fruit of both cultivars (bara and serewe) also did not show any significant differences among the respective fruit positions even though the base fruits maintained a greater number as shown in Figure 13. The total number of seeds produced by serewe cultivar was greater than those produced by bara.

Fig. 13 Variations in the average number of seeds per fruit of bara and serewe harvested at different positions on mother plant.

There were no significant differences in the dry seed weight per fruit of bara harvested at different positions on the mother plant. However, whereas harvest of serewe fruits from base and middle positions yielded no significant differences in dry seed weight, harvest at the apex position resulted in significantly lower dry seed weight to the later two positions (Figure 14). In contrast to the greater seed number produced by serewe, a higher seed weight was obtained from bara cultivar.

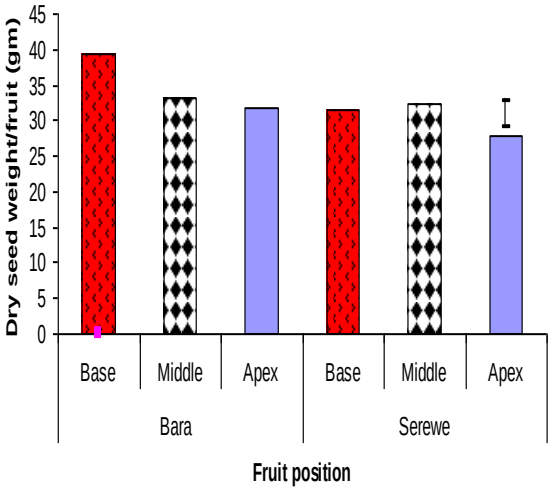

Fig. 14 Variations in the average dry seed weight (gm) per fruit of bara and serewe harvested at different positions on mother plant.

I LSD at P = 0.05

No significant differences were obtained in 100-seed weight of both cultivars (bara and serewe). However, Figure 15 reveals that in contrast to the number of seeds produced by serewe, bara produced a greater 100-seed weight.

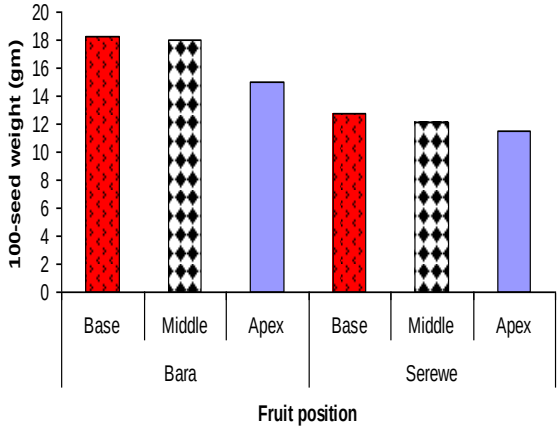

Fig. 15 Variations in the 100-seed weight (gm) of bara and serewe harvested at different positions on mother plant.

There was no significant difference in the germination levels of seeds obtained from fruits of both bara and serewe cultivars harvested at the base and middle positions. However, germination was significantly enhanced by these positions in comparism to the apex position as shown in Figure 16.

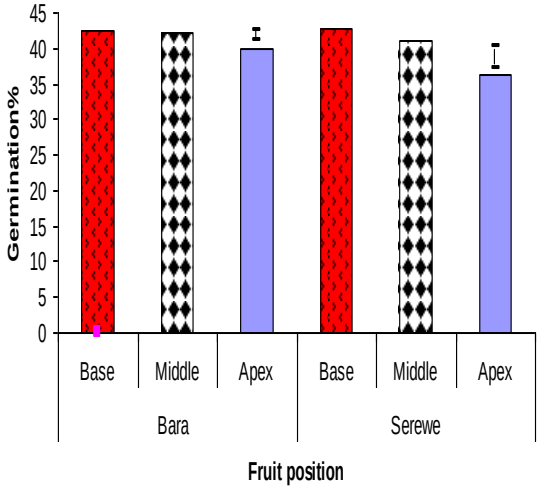

Fig. 16 Variations in percentage germination of bara and serewe harvested at different positions on mother plant.
I LSD at P = 0.05

IV. DISCUSSION

Yadav (1983) and Dhanelappagol *et al.* (1994) named amino acid, phosphorus active substances, dry matter, sugar, water soluble proteins, acids and necotonic acid levels as some of the assimilates usually transported from other plant parts for accumulation into the fruits/seeds as the plant develops progressively until physiological maturity is attained. The progressive increases obtained in fruit/seed attributes (length, diameter and weight) with age in this study is therefore an indication of progressive increase in the accumulation of assimilates during fruit/seed maturation. This finding agrees with that of Bino *et al.* (1996) who categorised cell expansion and accumulation of reserve materials by the embryo and endosperm as a second in the three phases of seed development. Demir and Ellis (1992) reported a similar trend in tomato but stated that maximum seed quality is attained in tomato some time after the end of the seed-filling period. Natrajan and Srimathi (2008) also reported that increase in Petunia pod weight with increase in DAA was supported by increse in pod length and width due to the development from zygote to matured seeds. Furthermore, Kortse *et al.* (2012) reported increases in the values for fruit and seed traits of *Citrullus lanatus* as time progressed up to 26 and 38 DAA in the two seasons of production. The decrease in fruit/seed traits retrogressively from basal through medium to the apex suggests that as an indeterminate plant develops more and more flowers, there is competition for assimilates. The basal fruits, which have an age in the competition because they started first, therefore get a higher share while the later fruits suffer shortages. This finding agrees with the report by Demir and Ellis (1993) that the duration of seed filling in marrow was two days longer within the earlier-formed basal fruits than within the later-formed apical fruits, while the rate of seed-filling was also greater in the former. Consequently, mature seeds from basal fruits were 17% heavier than those from apical fruits. Passam *et al.* (2010) also found that during fruit growth and maturation in eggplant, there was competition between fruits on the plant for essential nutrients and storage reserves, which led to a reduction in flower induction and flower size and adversely affect seed yield.

The number of filled seeds per fruit, which increased with corresponding increase in DAA, is an indication that different seeds within a fruit do not mature at the same rate. This view agrees with that of Nielsen (1996) who reported that seeds on the same fruit may not normally be of the same age. This is expected since pollen grain may not germinate at the same rate and also pollen tube growth, zygote formation and seed development and maturation may not proceed at the same rate (Silvertown, 1984; Delph *et al.*, 1998).

The greater seed number produced by serewe relative to that produced by bara in both experiments which conversely generated values of lower seed weight also supports the competition for assimilates already explained above. This competitive theory suggest that the quantity of assimilates alocated to an individual seed in a fruit would be determined by the total number of seeds the fruit contains. The more seeds a fruit contains the higher the competition and the fewer nutrients individual seeds would get and thus the low seed weight. This finding contradicts the report by Marcelis and Hofman-Eijer (1997) on *Cucurbita pepo* in which they reported that fruits that contained more seeds competed better for available assimilates and therefore achieved greater size.

The result of this study in which values of 100-seed weight recorded significant increases at 38 and 35 DAA respectively in bara and serewe suggests that maximum dry matter accumulation was attained between 35 and 38 DAA in these two cultivars. Germinability however improved until the last harvest. This is in agreement with the report by Demir and Ellis (1992) which found maximum seed quality of tomato to be attained some time after the end of the seed filling period. This finding is also in agreement with those of Nerson (2002) and Demir *et al.* (2004) who both found fully matured melon seed lots (35 – 49 DAA and 40 – 45 DAA respectively) to have better germination and longevity than the less matured ones. The insignificant differrences obtained in the germination of seeds from base and middle positions of both cultivars show that there were no marked changes in provision of assimilates at this stage of development however, the seeds from the apex positions were lacking in supply.

It is therefore recommended that harvest of melon fruits for seed can be delayed for up to 47 DAA and even beyound. For better germination still, basal and middle positioned fruits should be preffered.

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Novel Approach to Increase Efficiency of Engine

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Abstract- The four stroke engine of motorcycles is mainly worked upon on four basic actions of engine induction, compression, power and exhaust. Initially these stokes has the relationship between TDC [top dead curve] to BDC[bottom dead curve] and pressure created in assembled engine, there are many types of engine having different types of cc i.e. cubic centimeter such as 100,110,125 up to 1300cc. The combustion in petrol engine is done using flammable mixture of air and petrol which is ignited by time spark when the charged is compressed. Now a day's many researches are working to increase the fuel efficiency in order to increase the mileage of the vehicle. **Our idea proposed is giving a significant results and mileage is observed to be increase dramatically**. We have modified the engine with petrol consumption mechanism which is working more effectively than the engine available presently and the mileage observed is double that of existing mileage of vehicle. Initially the inlet valve allow to passes the fuel and air via pipe to the engine the spark plug also connected with the engine because petrol engine needed a spark for fuel combustion's the working of four stroke engine the main motto is to give the up and down moment to the cylinder head of the engine so that it can rotate the fly wheel via connecting rod [Con- rod], piston, crank shaft in crank case. The inlet allows to pass the fuel and then the petro was stored in cylinder between piston and the top head of the cylinder and due to the spark the combustion occurs and it produces the chemical reaction of oxygen and carbon mono-oxide only 70% of fuel combustion occur and some fuel is used to keep the piston oily for cooling purpose so that the piston also cooled during maximum load or at long duration working.

Index Terms- Engine, Rpm, Modified, Combustion, Minutes

I. INTRODUCTION

Stage 1

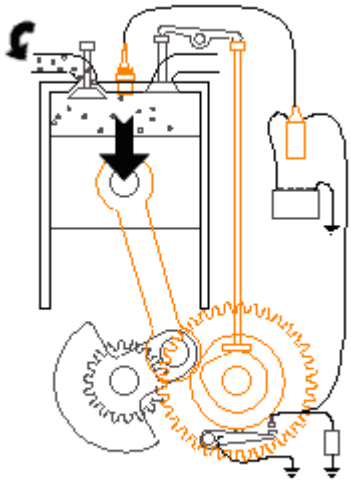


Fig.1[2]

Initially in induction process the petrol get sparks and it start combustion and create the petrol vapor ; these vapor produces the pressure on the head of the cylinder, so some wastage of petrol happens so here some petrol get waste due to incomplete combustion of petrol and some in piston to make it cooled with oil..

Stage 2: Compression

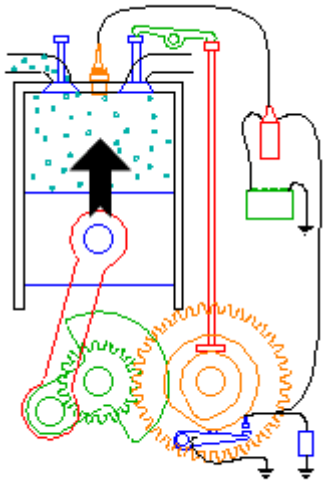


Fig 2 [2]

As per the induction process end at the end of induction process many amount of petrol vapor collected and due to the combustion pressure get induce in cylinder between head and cylinder head due to these the piston get compress and as well as piston compress towards bottom side the connecting rod also moved and it makes the crank shaft to moved with the fly wheel with both intake and exhaust valves closed, the piston returns to the top of the cylinder compressing the air or fuel-air mixture into the_combustion chamber of the cylinder head. During the compression stroke the temperature of the air or fuel-air mixture rises by several hundred degrees. [1]

Stage 3: Power Stroke:

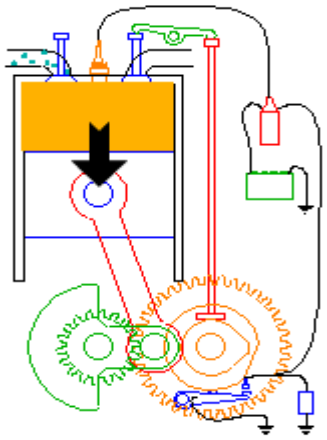


Fig 3 [2]

This is the start of the second revolution of the cycle. While the piston is close to Top Dead Centre, the compressed air–fuel mixture in a gasoline engine is ignited, usually by a_spark plug, or fuel is injected into a diesel engine, which ignites due to the heat generated in the air during the compression stroke. The resulting pressure from the combustion of the compressed fuel-air mixture forces the piston back down toward bottom dead centre.[1]

Stage 4: Exhaust Stroke:

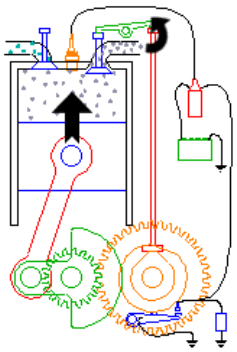
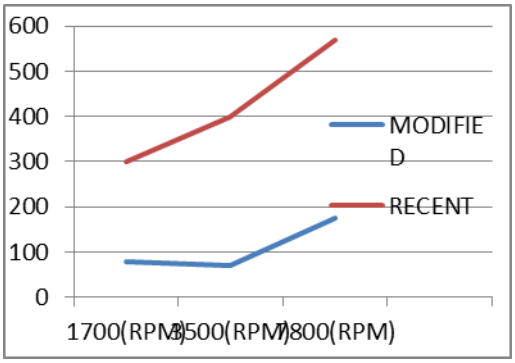


Fig 4 [2]

During the *exhaust* stroke, the piston once again returns to top dead centre while the exhaust valve is open. This action expels the spent fuel-air mixture through the exhaust valve(s). [1]

II. OBSERVATIONS

Readings Taken At 4th Gear:
Comparison between Readings:
Comparison: Rpm Vs Petrol Consumption



On X axis: RPM
On Y axis: PETROL CONSUMED

Effects:
Bike Performance After Modification:
Breaking Effect:

Gear	Effect
1st	Normal
2nd	Normal
3rd	Normal
4th	Normal

In break effect there was no effect occurred while running on petrol vapor , it works same as the petrol engine.

Light Efficiency:

Gear	Light Effect
1st	Normal
2nd	Normal
3rd	Normal
4th	Normal

The light effect is also normal means same as the petrol running engine.

Gear Shifting:
The gear shifting that means the shifting up and shifting down of gear is mainly effect to the load of the engine according to that the performance is depended when running the vapor engine there was smoothly up shifting and down shifting occurs means no problem of hardness or roughness .

Other Effect:

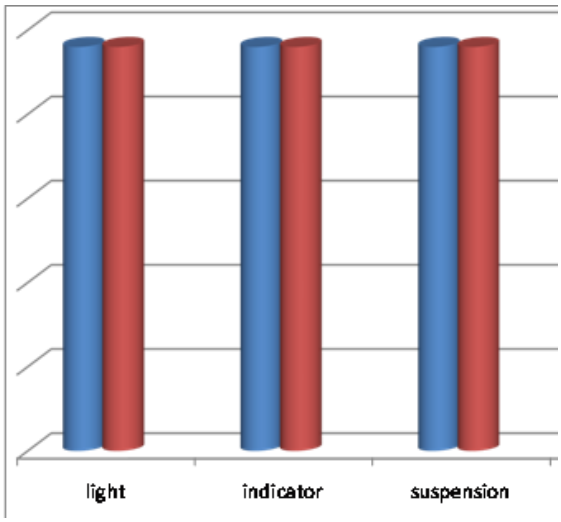
The other effect such as the vibration effect in engine cannot feels.

Indicators:

The indicator also perform the same as it works in petrol engine

Suspension:

The suspension of the vehicle also same.comparision between both engines:



REDUCTION COST:

In modified assemble weight of engine is also reduces which also reduces cost.

III. CONCLUSION

Above results and data for the standard engine reflects that the efficiency of fuel combustion and ultimately the mileage of a vehicle is found to be increase significantly

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[2] www.animatedengines.com/otto.html

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Effects of Stubble Management on Yield of Tomato

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Abstract- Field experiments were conducted on-farm in the forest zone of Ghana to determine the effects of stubble mulch (*in-situ* residues from *Mucuna pruriens* var utilis or natural grass fallow) and method of land preparation on the yield of dry season tomato (*Lycopersicon esculentum*). Land preparation methods studied were slash burn and ridge; ridge and mulching with the existing residue; and zero tillage without burning. The design was randomized complete block with four replications. Tomato seedlings were transplanted at a spacing of 50 cm within plants and 90 cm between rows; 2 seedlings per hill. Results showed that stubble mulch resulted in significant reduction in soil temperature and an increase in soil moisture content. Ridging resulted in a higher content of soil nutrients. Mulching especially with Mucuna residue, was however associated with insect pest problems. Despite this problem, stubble mulching with mucuna resulted in up to 100% increase in tomato yield and resulted in the highest yields among the management options. Result of the studies showed that soil moisture and temperature might be more important than soil fertility in the short term in determining yield in dry season tomato in the study area.

Index Terms- Stubble management, tomato, dry season, vegetables

I. INTRODUCTION

Vegetable production in the dry season is an important source of income for many peri-urban farmers in Ghana. Although yields are low at this season, this is compensated for by high price they command during the dry period due to scarcity of vegetables during this season. Dry season vegetables are usually grown under manual irrigation close to streams. Due to the difficulties involved in this form of irrigation, farm sizes are usually smaller than the two main seasons and the amount of water applied to the crops is often inadequate. The low humidity during this time of the year couple with dry north easterly winds result in high rate of water loss through evapo-transpiration; the vegetables often suffer from water stress. Mulch farming is recommended as a sustainable management option for water conservation (Lar, 1993) especially for high value crops such as vegetables (Carsky *et al*., 1998). Nakashima *et al*. (1991) reported up to 20 times increase in yield of pepper grown with stubble mulch compared with those grown on bare ground. Increased yield of tomato due to mulching has also been reported (Kwapata, 1991). Shajari *et al*. (1990) observed that mulching was very efficient soil temperature regulator and that it resulted in improved water-use efficiency of Brassica. Legume cover crops provide a source of nitrogen for subsequent crop (People *et al*., 1995) reduces erosion (Flach, 1990), reduce runoff and contamination of soil water (Hoyt *et al*., 1994), utilize soil nitrogen that might otherwise be lost to leaching (Stivers and Shennan , 1989), improve soil physical properties (Barber and Navarro 1994), suppress nematode population (Crow *et al*., 1996) and reduce cost of weed management (Nancy *et al*., 1996). Mucuna cover crop produces large quantities of biomass that can be used as mulch. Furthermore, the legume is capable of restoring the fertility of degraded soils (Hulugalle *et al*., 1986, Osei-Bonsu and Buckles 1993). Tillage practices have been reported to have significant impact on crop production (Awe and Abegunrin, 2009). Adekiya *et al*. (2009) reported that ploughing plus harrowing and ridging increased tomato fruit yields by 40, 16, 24 and 62% over zero tillage, manual mounding, ploughing only and ploughing and harrowing respectively.

The objective of this study was to determine the effect of stubble mulch and method of land preparation on the yield of dry-season tomato

II. MATERIALS AND METHODS

Farmer managed on-farm trials were conducted between 1998 and 1999 in the forest zone of Ghana. The study area has a bimodal rainfall pattern. The major season begins in April and ends in July; the minor season begins in September and ends in mid-November. Mid-November to March is the dry season which is the harmattan period.

Four farmers participated in the study, one at Apatrapa and Duase, and two at Darko in the Ashanti Region of Ghana. The experimental plots had been cropped to maize in the major season followed by natural fallow in the minor season and vegetables in the dry season between 3 and 5 years depending on the site. The predominant weed at all sites was spear grass (*Imperata cylindrical*). Green maize was planted on the experimental fields during the major season of 1998. After harvesting the maize, the experimental fields were laid out (five plots/ site) and Mucuna was randomly planted in mid August 1998 on 3 of the plots after spraying the entire field with glyphosate at a rate of 3 l/ha. After planting, the Mucuna and the 2 other plots were left under Mucuna and grass (*Imperata cylindrical*) fallow respectively in the minor season.

The quantity of Mucuna and grass biomass was assessed by cutting all plant residues from four randomly selected areas before the whole area was slashed using cutlass in mid-November, 1998. The samples were dried in a forced oven at 80 ^{o}C for 48 hours and their masses determined. The following treatments were imposed: Mucuna residue burnt and the soil ridged (Mucuna burn and ridged), Grass residue burnt and soil ridged (Grass burn and ridged), soil ridged and mulched with in-situ Mucuna residue (Mucuna ridge and mulched), soil ridged and mulched with in-situ grass residue (Grass ridged and

mulched), no burning and no tillage for Mucuna only (Zero till Mucuna).

Soil samples for nutrient analysis were taken at a depth of 0-15 cm from the ridges (and on the flat for the zero till option) immediately after land preparation at 5 cores per plot and analysed by standard laboratory procedures. Available phosphorus was determined using the Bray-1 method (Bray and Kurtz, 1945), organic carbon by the Walkey-Black wet dichromate method (Walkley and Black, 1965) and total nitrogen by the Kjeldahl digestion method. The experimental design was a randomized complete block with five treatments and sites as replication. Tomato seedlings were transplanted in the last week of November. Each plot measured 5.4 m x 5 m, having 6 rows of tomato spaced 50 cm within plants and 90 cm apart. Water was applied to the crops at 3-day intervals at a rate of 400 ml per hill in the evenings except during the days of rainfall. Fertilizer was applied at a rate of 125 kg/ha of NPK 15-15-15 1 week after transplanting followed by sulfate of ammonium at the same rate at flower bud stage. Weed assessment (5samples/plot) was done before weed control at 4 weeks after transplanting. Weed control was carried out once because dry weather conditions after the first weeding retarded weed growth. Soil temperatures and moisture content were measured at tomato flowering stage, which was the time most stress was observed. Soil temperature and moisture content at a depth of 5 cm was assessed four times at weekly interval during the morning at 8:00 GMT and during the afternoon at 1:00 GMT. In each case, six samples were taken per plot and the means calculated.

III. RESULTS

Results of the soil analysis are presented in Table 1. Potassium content ranged from 0.16 to 0.52 cmol/kg and organic carbon from 1.250 to 1.530. Potassium and organic carbon contents of the ridged and burnt plots were significantly ($p<0.05$) higher than the zero tilled plot but no difference was observed in phosphorous and nitrogen. For the same method of land preparation, soil nutrient status was consistently not significantly higher on the burned plots than unburned. There was also no significant effect on the type of stubble on soil nutrients. There were however, no significant differences between nitrogen and phosphorus content.

The soil temperature and moisture content of the stubble management options are presented in Table 2. There were no significant differences ($p<0.05$) in morning temperatures and moisture content of the soil soon after manual watering. However, afternoon temperatures were significantly ($p<0.05$) lower and moisture content just before manual watering was higher on the mulched than the burned plots. Soil under zero tillage had the highest moisture content (13.5%).

By November 1998, Mucuna and grass had accumulated mean dry matter of 4.2 t/ha 4.7 t/ha respectively (Data not shown). There was no significant difference ($p<0.05$) between the grass and Mucuna biomass.

Insect pests attacked tomato seedlings which resulted in 13% - 42 % destruction of the transplanted seedlings which necessitated refilling of vacancies (Table 3). Highest attack occurred on the plots mulched with Mucuna residues, compared with grass mulch and burned residues. After refilling, tomato plant stand remained statistically equal among the treatments.

Weed pressure in tomato was highest (883 kg/ha) on the plots mulched with grass and lowest (66kg/ha) on the zero till plot (Table 4).

Yields of tomato was low which is normal during this period. It ranged from 508 kg/ha to 1272 kg/ha (Table 4). Burning grass residues and ridging resulted in lowest yield but this was not significantly ($p<0.05$) different from yields obtained from grass ridge and mulch and Mucuna burned and ridge plots. Zero till Mucuna resulted in more than 100% increase in tomato yields compared with tomato after grass fallow either burned or unburned.

IV. DISCUSSION

The pilling up of top soil to form ridges might have accounted for increased levels of K and organic carbon on the ridged plots compared with zero tilled plots. The results partly agree with studies by Agebede *et al*. (2009) who observed that manual ridging resulted in higher soil N, P, K and Ca compared with untilled plots. Fallow on the other hand had no effect on soil fertility probably because the residues were not fully decomposed at the time of soil sampling.

The reason for weed buildup in the glass mulch may include infestation of rhizomes from the soil and weed seed from the mulch material. Fire might have destroyed some of the weed seed and contributed to the reduction in weed incidence on the burned plots. Mucuna on the other hand effectively controlled the weeds during the fellow period (Hulugalle *at al*., 1986, Osei Bonsu and Buckles 1993) and this might have reduced infestation and weed build up on Mucuna plots.

Mucuna mulch was however associated with insect pest problems. The insects (mainly grasshoppers and crickets) hid in the mulch and were responsible for the more than 40% seedling damaged on the plots mulched with Mucuna, which necessitated refilling. The relatively high incidence of pests in Mucuna mulch may be due to the fact that the mulch was more compact resulting in conditions more favorable to the insects. Zero till Mucuna resulted in more than 100% increase in tomato yields compared with tomato after grass fallow both burned and unburned. Low afternoon soil temperatures and high soil moisture content may have contributed to high yield on the zero till plots. High weed pressure on the spear grass plots might have contributed to the low yields from these plots.

V. CONCLUSION

The study revealed that soil moisture and temperature rather than soil nutrients may be most important factors determining yields of dry season vegetables in the study area especially when inorganic fertilizer is applied. Mucuna stubble suppressed weeds more than grass and resulted in higher soil moisture retention and tomato fruit yield.

Table 1: Effects of stubble management on soil nutrient status (0-15 cm depth) in 4 on-farm tomato trials.

Treatment	Ex K (cmol/kg)	Org C (%)	Total N (%)	Bray-1 P (mg/kg)
Grass burn & ridge	0.52	1.517	0.123	7.63
Grass ridge and mulch	0.34	1.395	0.119	8.80
Mucuna burn & ridge	0.42	1.530	0.129	7.45
Mucuna ridge and mulch	0.40	1.475	0.126	7.50
Mucuna zero till	0.16	1.250	0.105	8.30
CV%	36.4	7.5	10.3	32.5
LSD (0.05)	0.21	0.165	ns	ns

Table 2: Soil temperature (^{O}C) and moisture content (%) at 5 cm depth as affected by stubble management.

Treatment	Soil temperature		Soil moisture*	
	8:00GMT	13:00 GMT	A	B
Grass burn & ridge	20.0	35.0	14.3	8.4
Grass ridge and mulch	21.8	26.5	16.5	11.3
Mucuna burn & ridge	19.8	31.0	14.5	10.7
Mucuna ridge and mulch	20.8	27.0	14.4	11.2
Mucuna zero till	21.5	25.5	15.9	13.5
CV%	2.5	5.5	9.2	8.2
LSD (0.05)	ns	4.5	ns	2.4

*Soil moisture A assessed soon after rain fall and soil moisture B assessed after 2 weeks without rainfall

Table 3. Tomato establishment and weed dry weight as affected by stubble management.

Treatment	% stand vacancy filled	Plants/m^2 at harvest
Grass burn & ridge	14	4.3
Grass ridge and mulch	13	3.8
Mucuna burn & ridge	19	4.4
Mucuna ridge and mulch	42	4.3
Mucuna zero till	39	3.9
CV%	29	12.7
LSD (0.05)	18	ns

Table 4: Effects of stubble mulch and method of land preparation on weeds and fruit yield of tomato

Treatment	Weed dry weight (kg/ha)	Fruit Yield (kg/ha)
Grass burn & ridge	269	508
Grass ridge and mulch	833	579
Mucuna burn & ridge	207	648
Mucuna ridge and mulch	97	1013
Mucuna zero till	66	1272
CV%	23.6	31.4
LSD (0.05)	191	450

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Moringaoleifera: Nature is Most Nutritious and Multi-Purpose Tree

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Abstract- Moringaoleifera, commonly called the horseradish tree, drumstick tree, ben oil tree, miracle tree, and Mother's Best Friend, known for its multi-purpose attributes, wide adaptability, and ease of establishment. Its leaves, pods and flowers packed with nutrients important to both humans and animals. This review gives information regarding cultivation, nutrients and uses potentials of Moringaoleifera. Where diets lack in these essential nutrients the Moringa tree makes a major contribution to human health.

Index Terms- Moringaoleifera- multi-purpose- uses- cultivation- ecology

I. INTRODUCTION

In recent years, interest has grown in the utilization of what have come known as 'multipurpose' plants. One such plant is *Moringa oleifera* Lam (syn. *Moringa pterygosperm*a), the most widely cultivated species of a monogeneric family Moringaceae. *Moringaoleifera* is one of the 14 species of family Moringaceae, native to India, Africa, Arabia, Southeast Asia, South America, and the Pacific andCaribbean Islands (Iqbal *et al*., 2006). Because M. oleifera naturalized in many tropic and subtropics regions worldwide, a number of names such as horseradish tree, drumstick tree, ben oil tree, miracle tree, and Mother's Best Friend refer to the plant. (jahn, 1981). It called 'Shagara al Rauwaq' in Nilevalley (von Maydell, 1986). According to Muluvi *et.al* (1999), the Moringa tree introduced to Africa from India at the turn of the twentieth century where used as a health supplement and it was originally an ornamental tree in the Sudan, planted during British rule in the alleys along the Nile, public parks, and the gardens of foreigners. It seems likely that the Arab women of Sudan discovered this remarkable clarifier tree (Jahn1986). The Moringa consumed by humans throughout the century in diverse culinary ways (Iqbal *et al.*, 2006). Almost all parts of the plant used for taste, flavor or as vegetable and seed activities (Chumark et al, 2008) culturally for it's nutritional value, purported medicinal properties (DanMalam et al, 2001; Dahiru et al, 2006). The investigation of the different parts of the plant is multidisciplinary, including but not limited to nutrition, ethnobotany, medicine, analytical chemistry, phytochemistry and anthropology (McBurney et al, 2004). The purpose of this review is to summarize the uses and benefit of *Moringaoleifera*

II. IDENTIFICATION

Moringa is a small, fast-growing, drought deciduous tree or shrub that reaches 12 m in height at maturity. It has a wide-open, typically umbrella- shaped crown, straight trunk (10-30 cm thick) and a corky, whitish bark. The plant (depending on climate) has leaflets 1-2 cm in diameter and 1.5-2.5 cm in length its leaves are impair pinnate ,rachis 3 to 6 cm long with 2 to 6 pairs of pinnules. Each pinnule has 3 to 5 obovate leaflets that are 1 to 2 cm long (Von Maydell, 1986). The terminal leaflet is often slightly larger. Its leaflets are quite pale when young, but become richer in color with maturity. Cream-colored flowers emerge in sweet-smelling panicles during periods of drought or water stress when the tree loses its leaves. The pods are triangular in cross-section-30 to 50 cm long and legume-like in appearance. The oily seeds are black and winged. . The tree produces a tuberous taproot, which explains its tolerance to drought conditions (F/FRED, 1992).

III. ECOLOGY

Originally Moringa considered as a tree of hot semi-arid regions (annual rainfall 250-1500 mm), which is adaptable to a wide range of environmental conditions; from hot dry to hot, humid and wet conditions. The tree is tolerant to light frosts, but does not survive as a perenniall under freezing condition. Moringa is quite drought tolerant and is well suited for a wide range of adverse environments that would not be suitable for other fruit, nut and tree crops. Moringa grows more rapidly, reaching higher heights, when found in well-drained soils with ample water, but tolerates sandy soils, heavier clay soils and water-limited conditions. The tree can be established in slightly alkaline soils up to pH 9 as well as acidic soils as low as pH 4.5 but it yields much less foliage (chumark et al 2008). In Oaxaca State, Mexico, traditional cultivation of *Moringaoleifera* carried out in the highlands even up to altitudes of 800-1200 m, if the sites protected by mountains. The highest site with abundantly flowering and fruiting *Moringa* trees was at San Juan Gegoyache in the valley of the Totolepan River. In Oaxaca, *Moringaoleifera* transferred from the lowlands of the Pacific coast toward this region because the tree provided poor people's white flowers for religious feasts in churches and houses. This successful migration also provides a good indication that *Moringa* trees can be grown in other micro-climates at similar altitudes even if up

until now it has been maintained that the tree is only suitable for lowlands under 600 m(Jahn1986).

IV. OCCURRENCE THROUGHOUT THE WORLD

Moringa trees though native in the sub-Himalayan tracts, it is widely cultivated in Africa, Central and South America, Sri Lanka, India, Mexico, Malaysia, Indonesia and the Philippines (Anwar and Bhanger, 2003). According to Muluvi *et al* (1999), the Moringa tree wide natural spread in the world and introduced to Africa from India where it used as a health supplement and it was originally an ornamental tree in the Sudan, planted during British rule in the alleys along the Nile, public parks, and the gardens of foreigners. It seems likely that the Arab women of Sudan discovered this remarkable clarifier tree (Jahn, 1986). Table (1) illustrates the uses of different moriga species at different part of the warld.

V. GROWING MORINGA

Moringa planted by direct seeding, transplanting, or using hard stem cuttings. Direct seeding is preferred when plenty of seed is available and labor is limited Transplanting allows flexibility in field planting but requires extra labor and cost in raising seedlings, Stem cuttings are used when the availability of seed is limited but labor is plentiful (pM.C. Palada and L.C. Chang 2003). Transplanting entails collection of seeds from the tree, development of plantlets in the greenhouse for two to three months and transplantation of mature stems (1-1.5 m long) to the main fields (chumark et al, 2008).Moringaoleifera is easily established by seed and no seed treatment is required (chumark et al, 2008). The rapidly germinating seedlings, can reach 5 m in one year if sheltered from drying winds and provided with enough water, however it normally cut back annually to one meter or less and allowed to regrow, so that pods and leaves remain within arm's reach (Shen et al 2006). Plants rose from 1 m cuttings beat pods from the second year of growth onwards, with maximum production at four to five years. An individual tree can yield 50 to 70 kg of pods in one year (Shen et al 2006). Moringa is an ideal tree for Accompanying/ Inter-row crops and agro-forestry uses, as the branches can be easily trimmed to regulate shade effects; It's open crown allows plenty of sunlight to reach under-story crops often garden vegetables that benefit from some shading. Frequent pruning, lopping, coppicing or pollarding will increase and maintain leaf production. Moringa will sprout back repeatedly and vigorously when lope or pollard (chumark et al 2008) it is best to keep trees at a height that facilitates leaf harvesting. The leaves are attractive to all livestock however, so Limitation (Chen et al 2006) though quite tolerant to drought, the tree is deciduous, and it loses most of its leaves in periods of mended water-stress and susceptible to breakage in high winds. (pM.C. Palada and L.C. Chang 2003) Moringa is relatively short-lived reaching only 20 years on average (vonMaydell, 1986) it is so easy to establish, however, this limitation does not discourage cultivation of this very useful and adaptable tree

VI. USES AND NUTRITIONAL BENEFIT

Every part of the Moringa tree used as medicine and food commodity, which has received enormous attention as the 'natural nutrition of the tropic'. According to Jed, et .al (2005) and Anwar *et al* (.2005), Moringa trees have been used to combat malnutrition especially among infants and breastfeeding woman in many developing countries, particularly in India, Pakistan, the Philippines, Hawaii and many parts of Africa. Elkhalif, et.al (2007) studies the nutritive values of the leaves of M. Olaeifera tree (Rawag) which is available in Sudan and the results showed that, moisture content was 74.42%, protein 16.7% ,fiber 3.5%, ash 8%, and oil 1.7%. In addition, the minerals content were determined and they found that the calcium content was 0.20 mg/100g, magnesium 0.13mg/100g, potassium 0.075mg/100g, and phosphorus 0.031 mg/100g.

The pod, seed and oil

The pods are often cooked and eaten like green beans. The whole seeds also ate green, roasted or powdered, and steamed in tea and curries (Fahey, 2005). The pods and seeds, often referred to as Moringa kernels, have a taste that ranges from sweet to bitter and are most popularly consumed after frying to get a peanut-like taste (Makkar et al, 1996). The pods are generally prepared in a similar fashion to green beans and have a slight asparagus taste. The pods are highly nutritious containing all the essential amino acids (Ramachandran et al., 1980). Although primarily utilized worldwide by the Asian Asia, Africa, America population as a vegetable, usage by other peoples is increasing. An international market already exists for both fresh and tinned pods (Jahn, 1986)

The leaves

Moringa leaf has been purported to be a good source of nutrition and a naturally organic health supplement that used in many therapeutic ways (McBurney et al, 2004; Fahey, 2005; DanMalam et al, 2001). The leaves are a very rich source of nutrients and contain the essential vitamins A, C and E. Table 2 illustrates the nutritional value of M. Oleifera leaves compare with other food (Fahey 2005).

Leaves rich in biologically active carotenoids, tocopherols and vitamin C have health-promoting potential in maintaining a balanced diet and preventing free-radical damage that can initiate many illnesses (Smolin et al, 2007) succulent leaves are harvested daily for soups, sauces, or salads. Can ate fresh, cooked, or stored as a dried powder for many months reportedly without any major loss of its nutritional value fresh leaves are picked, shade dried, ground to a powder, and then stored for later as a food flavoring or additive. Dried or fresh leaves used in foods such as soups and porridges (Lockett et al, 2000). .Also used as curry gravy in noodles, rice or wheat (Abilgos et al, 1999). Farmers have added the leaves to animal feed to maintain a healthy livestock (Sarwatt et al, 2002; Fahey, 2005; Sáncheza et al, 2006). Newer applications include. In addition, the use of Moringa powder as a fish food in aqua cultural systems (Dongmeza et al, 2006) and the Moringa leaves as a protein supplement for animals, such as cows. The feeding value of Moringa reported to be similar to that of soybeans and rapeseed meal (Soliva et al 2005). Pregnant women and lactating mothers use the powdered leaves to enhance their child or children's

nourishment. especially, in underdeveloped countries mothers suffering from malnutrition (McBurney et al, 2004; Lockett et al, 2000; WHO Readers Forum, 1999). Kasolo et al (2011) stated that *Moringa oleifera* leaves were safe for human consumption because no serious side effects have been observed by the people using them. However, the toxicity of biologically active agents has been found to depend on the dose, they study the acute toxicity profile of *Moringa oleifera* leaves, they concluded that, *Moringa oleifera* leaves extracts of ether, ethanol and water contains phytochemical compounds which when given orally as a single dose in 24 hours to mice were relatively non-toxic.

The stem and bark

Planted as a hedge in courtyards, Moringaoleifera provides wind protection, shade and support for climbing garden plants. One can easily understand the household/garden component. The wood of Moringa is relatively soft. Because of this, is n' t used in heavy construction. The wood is light, but provides a good fuel for cooking. It yields approximately 4,600 kcal/kg (F/FR.F-D, 1992). The fiber from the bark used to make ropes. (Jahn, 1981)

The flowers

The Moringa flowers also produce a good honey and Honey clarifier in India and Sudan (Jahn, 1984)

The root

The root tastes similar to horse radish and is a popular food in East Africa (; Sattaur, 1983; Jahn, 1986).

Plant growth enhancer

Lab experimentation had shown that Moringa spray had a wide range of beneficial effects on plant crop. Effects of spray indicated accelerated growth of young plants. Plants were firmer, more resistant to pests and disease, longer life-span, heavier roots, stems and leaves, produced more fruit, larger fruit, increase in yield 20-35%f even a fraction of these results could be reproduced in the field, it could be a great help in increasing food supplies for millions of hungry people (Fahey 2005)

Medicinal Benefits

A number of natural compounds have been isolated from M. oleifera leaves including fully acetylated glycosides bearing thiocarbamates, carbamates or nitriles (Faizi et al., 1995; Murakami et al, 1998). Glycosides containing isothiocyanates, malonates and flavonoids also identified and isolated in the leaves of the Moringa plant.(Faiziet al, 1995; Bennett et al, 2003; Miean et al, 2001).Plant glycosides can be used as treatments for cancer or chronic conditions such as high cholesterol and atherosclerosis (Chumark et al, 2008; Ghasi et al, 2000). Plant flavonoids are important to the diet becauseof their effects on human nutrition. These phyto chemicals can modulate lipid peroxidation involved in atherogenesis, carcinogenesis and thrombosis and other known properties of free radical scavenging or inhibition of hydrolytic and oxidative enzymes (phospholipidase A2, cyclooxygenase, lipooxygenase), shows strong antioxidant and anti-inflammatory activity (Siddhuraju et al, 2003). Numerous studies have indicated that flavonoids also have anti-carcinogenic, anti-viral and anti-estrogenic activities (Havstee, 2002; Miean et al., 2001and Middletonet al., 2000). These identified bioactive compounds in the leaves of M. oleifera make this an excellentcandidate for nutritional and pharmaceutical supplementation.The World Health Organization (WHO) has been studying the use of M. oleifera for many decades as a low cost supplement enhancer in the poorest countries around the world (WHO Readers Forum, 1999). This organization has been promoting the use of this plant to help those countries suffering from malnutrition, which is one of the major causes of death worldwide. United Nations Food and Agriculture reported that one in twelve people worldwide is malnourished, including 160 million children under the age of five (United Nations Food and Agriculture Statistics, 2008).

Water Purifier

In parts of the world where clean drinking water is scarce, Moringa offers another crucial benefit such as the ability to purify water. Many countries use river water as their primary water source, but this water can contain harmful particles, bacteria and microorganisms. Water treatment plants are not available in many countries, but if there, is a Moringa tree nearby, river water users can still enjoy clean water seed from (Jahn and Dirar, 1979; Jahn, 1981). The Moringa tree harvested from the seedpod, crushed, and then put into vats of the water. Harmful particles bind to the seed and sink after an hour of treatment, and then clean water can remove from the top of the vat (Jahn 1986).

VII. CONCLUSION

These findings together demonstrate that moringa is an excellent multiuse plant Used to improve the health and nutrition of communities and appears to be a most promising candidate from which specific nutraceutical bioactive products could be developed. In Sudan moringa has a great potential to cultivate as an economically profitable crop and contribute in poverty elevation.

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Table (1): Uses and locations of the most important *Moringa* species

Use	*Moringaoleifera*	*Moringaperegrina*	*Moringastenopetala*	*Moringalongituba*	*Moringadrouhardii*	*Moringaovalifolia*
Vegetable	Asia, Africa, America *(leaves green pods. flowers roasted seeds)*		S. Ethiopia N. Kenya *(leaves)*			Namibia *(roots)*
Spice	Asia, Africa *(mainly roots)*		Kenya *(bark)*			
Oil (cooking, cosmetics, miscellaneous)	Asia, Madagascar, Africa, *(seeds)*	Near East			S. Madagascar	*Lab. test*, Angola
Water coagulant	Sudan *(trad., seeds)* Indonesia *(new project)*	*Lab. test*	*Lab. & field tests*	*Lab. & field tests*	*Lab. & field tests*	*Lab. Test*
Honey clarifier	Sudan *(trad., seeds)*					
Honey tree	*Recent* Indian *studies*					
Medicinal plant	Asia, Africa, Central America *(all plant organs)*	Near East as far as Sudan *(seeds)*	Kenya *(bark, root, leaves)*	Somalia *(root)*	S. Madagascar *(bark, root)*	Pounded seeds have also "horseradish smell"
Nematocide	Experiments in Philippines *(roots)*					
Fodder	India, Indonesia, *(leaves)*					Namibia *(roots: game; leaves and fruits: giraffes)*
Fences and windsheds	Asia, Africa, Central America		*New suggestion in* Sudan			
Supports for cultivation of climbers	Asia					
Firewood	Recent Indian project; Togo *(trad.)*					
Ornamental	Central & South America, USA, Africa	Saudi Arabia, Near East	Kenya		S. Madagascar	Namibia

Source: Jahn, 1986

The Emotion Perspective of an Ethnic Group of North India

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Abstract- Emotion is important because they avail the perceiver with information about another's behavioral readiness and information about the environment. Various theories suggest that emotion involve complex combination of physiological sensation, cognitive appraisal of situation, cultural beliefs and social environment. Feelings and behaviors are consistent with gender specific emotion beliefs. By using self administered questionnaire the authors investigated whether Bania-an ethnic group of North India differs in expressing their emotions in different situations. Also, they tried to find out whether strength of emotion varies according to the age groups.

Index Terms- Cognitive, Cultural beliefs, Emotion, Social environment.

I. Introduction

It almost seems as if there are as many definitions of emotion as there are writers on the subject. Attempts to arrive at a comprehensive definition of emotion (Kleinginna et al, 1981) should (1) say something about the way we feel when we are emotional; (2) mention psychological, or bodily basis of emotional feeling; (3) include the effects of emotion on perception, thinking and behavior; (4) point out the driving, or motivational, properties of certain emotions such as fear and anger; and (5) refers to the ways in which emotions are expressed in language, facial expressions, and gestures. While we have not given a concise definition of emotion because none is generally accepted, this lists gives the highlight of what is meant by the term (Morgan et al, 2000).

Successfully reading others emotions is important because they avail the perceiver with information about another's behavioral readiness and information about the environment for instance, emotional expressions signal upcoming behaviors (eg. Anger, I am going to fight you) or environmental conditionals (eg. Fear; Danger is nearby). As other facial expressions warn and ready perceivers for impending action, and because such actions are most likely to happen within one's culture, the emotions that are expressed by members of one's own culture (weisbuch and Ambady, 2008). It has been proposed that one's cultural background may influence the recognition of other's emotions.

II. Research Elaboration

Many studies have been done in this fields whose results are as follows:

- In-group advantage in emotion recognition Elfenbein et al 2002, conducted a study in which Black and white photographs of facial expression served as experimental stimuli. The set of photographs residing used included one man and one woman from each cultural group of similar age: anger, fear, disgust, happiness, neutral, sadness and surprise.

Each set of photographs was developed by researchers residing in the nation from which the posers negated the study was conducted in India (50 observers), Japan, and USA. The author created hemi facial composites for each photograph. They were 25 males and 25 females university students from each country, for a total of 150 participants.

Result:- Meta analysis studies involving face emotion recognition task across multiple cultures led to the conclusion that individuals are better at recognizing own culture expression relative to other culture expressions, pointing to a robust cultural specificity in emotion recognition. Also, judgment of the left hemisphere of facial expression shows a greater extent of cultural in group advantage than judgment of the right hemisphere of facial expression.

Masuda et al conducted a study on 39 American students (17 females and 22 males) and 36 Japanese students (15 females and 21 males) on perception of facial emotional. They used cartoon images as stimuli in the context of a group of 4 other people. Participants were asked to judge the emotions of the central person. In another study 22 Westerner (8 females and 14 males) and 27 Japanese students (14 females and 13 males) participant in experiment to know that whether emotion is individual.

Study I demonstrates when Americans are trying to figure out what a person is feeling, they focus on that particulars person, whereas Japanese consider the emotions of the other people in the situation i.e. Japanese emotions judgments of the central person would be affected by the emotion of other people in the situation but westerners emotions judgment were not. Study II revealed that Japanese looked at the surrounding people more than did westerners. i.e. westerners see them as inseparable from the feeling of the group

III. Neural Basis of recognizing emotion

Chio et al (2008) conducted an FMRI study with American and native Japanese participants. 12 native Japanese living in Japan (6 females and males) and 10 livings in USA (5 females and 5 males) between the ages of 18 and 25 years. Participants

were presented with Americans and Japanese faces expressing fear, anger, joy or nothing.

Behaviorally, Americans were more accurate at judging own-culture emotions relative to those of the other culture. Similarly, Japanese individual although not reliably more accurate, were quicker to judge own culture emotions relative to those of the other culture. This, thus conformed to Elfenbain et al (2002) conclusion of a cultural specificity was reflected by brain activity as well. Neuroimaging results revealed that own culture fearful faces elicited greater activity in the bilateral amygdala relative to fearful faces of the other culture.

Although the role of amygdala in responding to fear expressions is often interpreted as the direct detection of negative affect or threat, it has long been knows that the amygdala does not necessarily process valence per se, but instead driven flexibly by a stimulants motivational importance for instance the study conducted by Schiller et al (2009). On 19 right handed normal volunteers (12 males) between 18 and 31 yrs of age the researcher constructed 20 person profile using 120 person descriptive statements implying different personality traits. Participants were told that they would see information about different people and would be asked to give their impression of them.

Response in the amygdala and the posterior cingulate cortex (PCC) were stronger while encoding social information that was consistent, with subsequent evaluation. These findings provide evidence for encoding differences on the basis of subsequent evaluations, suggesting that the amygdala and PCC are important for forming first impressions. The amygdala responds to both positive and negative stimuli so long as the stimuli are subjectively valued and predictive of social evolution. These findings provide evidence for encoding differences on the basis of subsequent evaluations, suggesting that the amygdale an PCC are important for forming first impressions. The amygdala responds to both positive and negative stimuli so long as the stimuli are subjectively valued and predictive of social evaluation.

Anderson and Phelphs (2001) conducted attentional blink for affectively significant words in a patient with bilateral damage to amygdala and in 10 patients with unilateral lesions of the right or left amygdala. Observers were asked to report the identify of two green target words occurring amongst a stream of black destructor items during rapid several resula presentation (RSVP)

The finding showed that healthy observers demonstrate robust benefits for the perception of verbal stimuli of abrasive content compare with stimuli of neural content. In contrast, a patient with bilateral amygdala damage has no enhanced perception for such oversize stimulus events. Selective responses to own culture fearful faces (relative to other culture fearful faces) likely reflect the amygdala‘s enhancement of perception of motivationally significant stimuli.

IV. Theories About Gender and Emotion

Contemporary emotions theorists generally agree that emotions involve complex combinations of physiological sensations, cognitive appraisals of situations, cultural labels, and free or inhibited affective displays (Thoits, 1989 and Schechter and Singer, 1962). Hochschild’s normative theory about emotion (1975, 1981) predicts male-female differences in feelings and expressive behavior that are consistent with gender-specific emotion beliefs. Hochschild argues that cultural beliefs about emotion influence individuals’ feelings and expressions vis-a-vis feeling and expression norms that specify the emotions individuals should (and should not) feel and express in given situations. Feeling rules are cultural norms that specify the appropriate type, intensity, duration, and target of subjective feelings (or internal experience). Expression rules are cultural norms that regulate the type, intensity, duration, and target of emotional behavior (or affective displays). According to Hochschild, feeling and expression rules provide standards by which individuals judge their own and other’s emotions. When people’s feelings and expressions depart from cultural norms, they often engage in emotion management, expression management, or both in order to create a more appropriate emotional response. To the extent that our emotion culture includes feeling and expression norms—which specify that women should, and men should not, be emotional and emotionally expressive—we should find that women report that they experience and express emotions more often than men in general. Similarly, insofar as our emotion culture contains norms that discourage men from feeling and expressing sadness and women from feeling and expressing anger, we should also find that women report that they experience and express *sadness* more often than men and that men report that they experience and express *anger* more often than women.

In contrast to Hochschild’s normative theory, Kemper’s structural theory about emotion (1978, 1981, 1990, 1991) predicts a pattern of subjective feelings for men and women that departs from cultural beliefs about gender and emotion. According to Kemper, structural factors such as individuals’ social position vis-a` -vis others—rather than cultural derived emotion norms—influence their emotional responses to social situations.

Kemper argues that status and power are two fundamental dimensions of social relationships that elicit specific emotions during social interaction when relational power and status are maintained or changed. He claims that persons with more status and power in a relationship experience positive emotions such as happiness and security, whereas those with less power and status experience negative emotions such as fear, sadness, and anger. Although Kemper focuses on relational status and power between persons in interaction episodes, an implication of his theory is that persons with higher status and power in society experience more positive feelings, whereas persons with lower status and power experience more negative feelings(Simon and Nath, 2004).

V. Research on Gender and Emotion

In light of prevailing cultural beliefs about gender differences in emotion as well as the surge of scholarly interest in both gender and emotion, there is surprisingly little sociological research that compares men’s and women’s everyday feelings and expressive behavior. As Thoits (1989) and Smith-Lovin (1995) both note in their reviews of the field of emotion, the

sociology of affect is theoretically rich but limited in empirical evidence. There is some evidence of gender-specific norms about certain emotions.

In-depth qualitative studies have identified feeling and expression norms about male and female anger (Cancian and Gordon 1988; Shields and Koster 1989; Stearns and Stearns 1986) and romantic love (Cancian 1987; Simon, Eder, and Evans 1992; Swidler 1980). There is also some evidence of gender differences in feelings and expressive behavior that are consistent with cultural beliefs and norms about males' and females' emotions.

Thorne's (1993) and Eder's (1995) ethnographic studies of children and adolescents show that boys are given greater latitude than girls in expressing anger in playground and school contexts. Moreover, based on college students' descriptions of an emotional experience, Hochschild (1981) finds that women pay closer attention to feelings of love than do men (Peplau and Gordon 1985). Using a similar methodology, Thoits's (1989) study of college students reveals that, when faced with a stressful situation, women are more likely to express their feelings and cope with their emotions by seeking social support. However, while these studies provide rich and detailed information about gender and emotion in the United States with respect to a few specific emotions in small, highly select samples, they offer little insight into gender differences in a range of feelings and expressive behavior in the general population (Simon and Nath, 2004).

To date, most of the empirical research on gender and emotion has been conducted by psychologists, who focus on gender differences in emotion beliefs as well as on subjective feelings and expressive behavior among children, adolescents, and young adults. This research, which tends to be based on experimental methods, indicates that both males and females judge and subsequently label females as more emotional and emotionally expressive than males (e.g., Robinson and Johnson 1997). This research also shows that as early as preschool age, both males and females believe that sadness and fear are closely associated with females, whereas anger is closely associated with males (e.g., Birnbaum 1983). Studies further find that females are more expressive than males in response to a variety of experimental stimuli (Blier and Blier 1989; Brody 1997; Kring and Gordon 1998), although findings are mixed with regard to gender differences in experienced emotion. While some studies find that females report more feelings than males, others find no significant gender differences in experienced emotion (Brody, 1985 and Brody and Hall, 1993). On the basis of these and other findings, developmental psychologists have suggested that the expression of emotion may be more heavily socialized than the experience of emotion (Brody 1993; Fischer 2000; Kring and Gordon 1998). From a developmental perspective, these findings also suggest that males learn to conceal their feelings relatively early in life, whereas females learn to express their emotions more freely. These studies contribute to our understanding of gender and emotion among children, adolescents, and young adults in experimental settings; however, they tell us little about gender differences in everyday feelings and expressive behavior in the general population.

VI. Material and Methods

The present study has been done among Bania of Delhi of age groups of 11 – 41 + years. Both males and females have been included in the study. Total 60 samples have been analyzed.

Tool: - Self administered questionnaire and observation have been used in the study to collect the data. Questions from different situations of emotions were asked to the subjects to know about how do they react, feel or behave when they are in particular situation. The questionnaire contained questions about the frequency with which respondents experienced 15 different emotions, including a range of both positive and negative feelings.

VII. Result and Finding

Table 1. Frequency of Bania male and female describing feeling when someone compliments them.

Age Group(Years)	Gender	Options	Frequency(N)	Percentage (%)
11-20	male	little effect	5	62.5
		mildly pleased	1	12.5
		pleased	1	12.5
		very pleased	1	12.5
		Total	8	100.0
	female	little effect	3	42.9
		mildly pleased	1	14.3
		pleased	2	28.6
		very pleased	1	14.3
		Total	7	100.0
21-30	male	little effect	7	77.8
		pleased	1	11.1

		very pleased	1	11.1
		Total	9	100.0
	female	little effect	4	50.0
		mildly pleased	2	25.0
		pleased	1	12.5
		very pleased	1	12.5
		Total	8	100.0
31-40	male	little effect	2	28.6
		pleased	4	57.1
		ecstatic	1	14.3
		Total	7	100.0
	female	little effect	1	14.3
		mildly pleased	1	14.3
		pleased	2	28.6
		very pleased	2	28.6
		ecstatic	1	14.3
		Total	7	100.0
40+	male	little effect	2	28.6
		mildly pleased	3	42.9
		very pleased	2	28.6
		Total	7	100.0
	female	mildly pleased	3	42.9
		very pleased	3	42.9
		ecstatic	1	14.3
		Total	7	100.0

Table 1 show that 62.5% males in the age group of 11-20 years have no effect when someone compliments them. 42.9% females of the same age group have same effect. In the age group of 21-30 years, the percentage is 77.8% in case of males and 50% in case of females who have the similar responses as stated above but in the age groups of 31-40 and 40+ years both males and females are mildly or very pleased if someone compliments them. Table 1 show that percentages of females to express their feelings are higher than males.

Table 2. Frequency of Bania male and female describing feeling when they are at a fun party.

Age Group(Years)	Gender	Options	Frequency(N)	Percentage (%)
11-20	male	little effect	5	62.5
		little light hearted	1	12.5
		very lively	1	12.5
		so much lively	1	12.5
		Total	8	100.0
	female	lively	3	42.9
		very lively	3	42.9
		so much lively	1	14.3
		Total	7	100.0
21-30	male	little effect	4	44.4
		little light hearted	1	11.1
		lively	2	22.2

		so much lively	2	22.2
		Total	9	100.0
	female	little effect	3	37.5
		lively	2	25.0
		very lively	2	25.0
		so much lively	1	12.5
		Total	8	100.0
31-40	male	lively	6	85.7
		very lively	1	14.3
		Total	7	100.0
	female	lively	3	42.9
		very lively	2	28.6
		so much lively	2	28.6
		Total	7	100.0
40+	male	little effect	1	14.3
		little light hearted	3	42.9
		lively	2	28.6
		very lively	1	14.3
		Total	7	100.0
	female	little light hearted	3	42.9
		lively	2	28.6
		very lively	2	28.6
		Total	7	100.0

Table 2 show that in the age group of 11-20 years, 62.5% males have no effect when they are at a fun party while 42.9% females in the same age group feel lively or very lively. In the age group of 21-30 years percentages are higher for both males and females who have no effect at a fun party. In the age group of 31-40 years 85.7% males and 42.9% females feel lively at a fun party. In the age group of 40+ years, 42.9% males and similar percentage of females are light hearted at a fun party or in a celebration.

Table 3. Frequency of Bania male and female describing feeling when they accomplished something valuable.

Age Group(Years)	Gender	Options	Frequency(N)	Percentage (%)
11-20	male	Little effect	4	50.0
		Little satisfied	3	37.5
		very satisfied	1	12.5
		Total	8	100.0
	female	Little effect	2	28.6
		Little satisfied	1	14.3
		satisfied	3	42.9
		very satisfied	1	14.3
		Total	7	100.0
21-30	male	little effect	5	55.6
		satisfied	2	22.2
		very satisfied	1	11.1
		so satisfied	1	11.1
		Total	9	100.0

	female	Little effect	2	25.0
		satisfied	2	25.0
		very satisfied	3	37.5
		so satisfied	1	12.5
		Total	8	100.0
31-40	male	Little satisfied	1	14.3
		satisfied	5	71.4
		very satisfied	1	14.3
		Total	7	100.0
	female	Little effect	1	14.3
		satisfied	2	28.6
		very satisfied	2	28.6
		so satisfied	2	28.6
		Total	7	100.0
40+	male	Little effect	1	14.3
		Little satisfied	2	28.6
		satisfied	3	42.9
		so satisfied	1	14.3
		Total	7	100.0
	female	Little effect	1	14.3
		Little satisfied	1	14.3
		satisfied	1	14.3
		very satisfied	1	14.3
		so satisfied	3	42.9
		Total	7	100.0

Table 3 shows that in the age groups of 11-20 and 21-30 years 50% males have no effect when they accomplished something valuable. In both the age groups percentages are higher for females who feel satisfied when they achieve something valuable. In the age groups of 31-40 and 40+ years , 71.4% and 42.9% males feel satisfied. Similarly in the same age groups 28.6% and 42.9% females feel satisfied and so satisfied respectively on accomplishing something valuable.

In the age group of 11-20 years 37.5% males are slightly romantic and 28.6% females are romantic and so passionate when someone with whom they are involved prepare candlelight dinner for them. In rest of all the age groups percentages are higher for females who feel very passionate if their husbands or boyfriends make such plan. On the counterpart, males in the age group of 21-30 years have little effect, 31-40 and 40+ age groups feel romantic if their wives or girlfriends prepare candlelight dinner for them.

Table 4. Frequency of Bania male and female describing feeling when someone surprises with a gift.

Age Group(Years)	**Gender**	**Options**	**Frequency(N)**	**Percentage (%)**
11-20	male	little effect	2	25.0
		grateful	2	25.0
		very grateful	2	25.0
		buy gift in return	2	25.0
		Total	8	100.0
	female	little effect	1	14.3
		little grateful	2	28.6
		grateful	1	14.3

		very grateful	2	28.6
		buy gift in return	1	14.3
		Total	7	100.0
21-30	male	little effect	3	33.3
		little grateful	2	22.2
		grateful	2	22.2
		buy gift in return	2	22.2
		Total	9	100.0
	female	little effect	1	12.5
		grateful	1	12.5
		very grateful	1	12.5
		buy gift in return	5	62.5
		Total	8	100.0
31-40	male	little grateful	1	14.3
		grateful	3	42.9
		very grateful	2	28.6
		buy gift in return	1	14.3
		Total	7	100.0
	female	little grateful	1	14.3
		grateful	2	28.6
		very grateful	1	14.3
		buy gift in return	3	42.9
		Total	7	100.0
40+	male	little grateful	1	14.3
		grateful	2	28.6
		very grateful	2	28.6
		buy gift in return	2	28.6
		Total	7	100.0
	female	little grateful	1	14.3
		grateful	1	14.3
		very grateful	2	28.6
		buy gift in return	3	42.9
		Total	7	100.0

Table 4 shows that in the age groups of 11-20 and 21-30 years percentages are higher for males who have no effect if someone surprises them with gifts. But in age groups of 31-40 and 40+ years males are grateful and very grateful on receiving gifts. On the other hand, percentages are higher for females who feel not only grateful but also want to give them a return gift.

In all the age groups percentages are higher for males who are little frustrated when someone frustrates them. On the other hand, in the age group of 11-20 years percentages are higher for females who are little frustrated. But they become very frustrated and extremely tensed as the age increases.

In the age groups of 11-20, 31-40 and 40+ years percentages are higher for males who feel guilty if they say or do something which they should not have done. But in the age group of 21-30 years 44.4% males have little effect. Similarly in case of females , who are relatively at lesser in the scale from the males in the age groups of 11-20 and 21-30 years. In rest of the age groups percentages are higher for females who feel guilty and extremely guilty when they say or do something which they should not have done.

Table 5. Frequency of Bania male and female describing feeling when someone criticizes them.

Age Group(Years)	Gender	Options	Frequency(N)	Percentage (%)

11-20	male	no effect	6	75.0
		very upset	1	12.5
		Extremely upset	1	12.5
		Total	8	100.0
	female	no effect	3	42.9
		upset	2	28.6
		very upset	2	28.6
		Total	7	100.0
21-30	male	no effect	4	44.4
		a little effect	4	44.4
		upset	1	11.1
		Total	9	100.0
	female	no effect	3	37.5
		a little effect	1	12.5
		upset	2	25.0
		very upset	2	25.0
		Total	8	100.0
31-40	male	no effect	2	28.6
		a little effect	1	14.3
		upset	3	42.9
		v upset	1	14.3
		Total	7	100.0
	female	upset	5	71.4
		very upset	2	28.6
		Total	7	100.0
40+	male	no effect	2	28.6
		a little effect	2	28.6
		upset	1	14.3
		very upset	2	28.6
		Total	7	100.0
	female	a little effect	1	14.3
		upset	1	14.3
		very upset	3	42.9
		Extremely upset	2	28.6
		Total	7	100.0

Table 5 shows that in the age group of 11-20 and 21-30 years majority of both males and females have no effect or little effect when someone criticizes them. But in the age groups of 31-40 and 40+ years majority of males and females feel upset and very upset when someone criticizes them.

In the age group of 11-20 years 75% males have no effect when they have an embarrassing situation. But in the age groups of 21-30, 31-40 and 40+ years 44.4%, 85.7% and 28.6% males are embarrassed respectively. On the other hand among females percentages are higher for those who feel embarrassed when they face an embarrassing situation.

Among males in the age groups of 11-20, 21-30 and 31-40 years , majority of them responded that they have no effect on seeing sad movie. In the age group of 40+ years 57.1% males stated that they feel sad. Among females in the age groups of 11-20 and 21-30 years majority of them have no effect or little sad during watching sad movie. But in the age group of 31-40 and 40+ years percentages are higher for females who become extremely sad so that they feel like weeping during watching sad movie or TV serial.

Table 6. Frequency of Bania male and female describing feeling when someone they know is rude to them.

Age Group(Years)	Gender	Options	Frequency(N)	Percentage (%)
11-20	male	no effect	5	62.5
		hurt	1	12.5
		very hurt	1	12.5
		incredibly hurt	1	12.5
		Total	8	100.0
	female	no effect	2	28.6
		little hurt	1	14.3
		hurt	1	14.3
		very hurt	2	28.6
		incredibly hurt	1	14.3
		Total	7	100.0
21-30	male	no effect	5	55.6
		little hurt	1	11.1
		hurt	2	22.2
		very hurt	1	11.1
		Total	9	100.0
	female	no effect	4	50.0
		hurt	1	12.5
		very hurt	1	12.5
		incredibly hurt	2	25.0
		Total	8	100.0
31-40	male	little hurt	2	28.6
		hurt	2	28.6
		very hurt	3	42.9
		Total	7	100.0
	female	no effect	1	14.3
		hurt	3	42.9
		very hurt	1	14.3
		incredibly hurt	2	28.6
		Total	7	100.0
40+	male	no effect	1	14.3
		little hurt	2	28.6
		hurt	1	14.3
		very hurt	2	28.6
		incredibly hurt	1	14.3
		Total	7	100.0
	female	hurt	1	14.3
		very hurt	4	57.1
		incredibly hurt	2	28.6

Age Group(Years)	Gender	Options	Frequency(N)	Percentage (%)
11-20	male	no effect	5	62.5
		hurt	1	12.5
		very hurt	1	12.5
		incredibly hurt	1	12.5
		Total	8	100.0
	female	no effect	2	28.6
		little hurt	1	14.3
		hurt	1	14.3
		very hurt	2	28.6
		incredibly hurt	1	14.3
		Total	7	100.0
21-30	male	no effect	5	55.6
		little hurt	1	11.1
		hurt	2	22.2
		very hurt	1	11.1
		Total	9	100.0
	female	no effect	4	50.0
		hurt	1	12.5
		very hurt	1	12.5
		incredibly hurt	2	25.0
		Total	8	100.0
31-40	male	little hurt	2	28.6
		hurt	2	28.6
		very hurt	3	42.9
		Total	7	100.0
	female	no effect	1	14.3
		hurt	3	42.9
		very hurt	1	14.3
		incredibly hurt	2	28.6
		Total	7	100.0
40+	male	no effect	1	14.3
		little hurt	2	28.6
		hurt	1	14.3
		very hurt	2	28.6
		incredibly hurt	1	14.3
		Total	7	100.0
	female	hurt	1	14.3
		very hurt	4	57.1
		incredibly hurt	2	28.6
		Total	7	100.0

Table 6 shows that in the age groups of 11-20 and 21-30 years 62.5% and 56.6% males have no effect if someone they know is rude to them. In the age groups of 31-40 and 41+ years 42.9% and 28.6% males get very hurt. Among females, in the age groups of 11-20, 31-40 and 40+ years percentages are higher who get hurt if someone known is rude to them. But in the age group of 21-30 years 50% females have no effect.

Table 7. Frequency of Bania male and female describing feeling when they are involved in an important situation.

Age Group(Years)	Gender	Options	Frequency(N)	Percentage (%)
11-20	male	no effect	4	50.0
		slightly anxious	1	12.5
		anxious	1	12.5
		very anxious	1	12.5
		extremely anxious	1	12.5
		Total	8	100.0
	female	no effect	2	28.6
		anxious	3	42.9
		very anxious	1	14.3
		extremely anxious	1	14.3
		Total	7	100.0
21-30	male	no effect	3	33.3
		anxious	3	33.3
		very anxious	2	22.2
		extremely anxious	1	11.1
		Total	9	100.0
	female	no effect	1	12.5
		slightly anxious	1	12.5
		anxious	3	37.5
		very anxious	2	25.0
		extremely anxious	1	12.5
		Total	8	100.0
31-40	male	no effect	1	14.3
		slightly anxious	3	42.9
		very anxious	2	28.6
		extremely anxious	1	14.3
		Total	7	100.0
	female	slightly anxious	4	57.1
		anxious	3	42.9
		Total	7	100.0
40+	male	no effect	2	28.6
		slightly anxious	4	57.1
		very anxious	1	14.3
		Total	7	100.0
	female	slightly anxious	2	28.6
		anxious	3	42.9
		extremely anxious	2	28.6

Age Group(Years)	Gender	Options	Frequency(N)	Percentage (%)
11-20	male	no effect	4	50.0
		slightly anxious	1	12.5
		anxious	1	12.5
		very anxious	1	12.5
		extremely anxious	1	12.5
		Total	8	100.0
	female	no effect	2	28.6
		anxious	3	42.9
		very anxious	1	14.3
		extremely anxious	1	14.3
		Total	7	100.0
21-30	male	no effect	3	33.3
		anxious	3	33.3
		very anxious	2	22.2
		extremely anxious	1	11.1
		Total	9	100.0
	female	no effect	1	12.5
		slightly anxious	1	12.5
		anxious	3	37.5
		very anxious	2	25.0
		extremely anxious	1	12.5
		Total	8	100.0
31-40	male	no effect	1	14.3
		slightly anxious	3	42.9
		very anxious	2	28.6
		extremely anxious	1	14.3
		Total	7	100.0
	female	slightly anxious	4	57.1
		anxious	3	42.9
		Total	7	100.0
40+	male	no effect	2	28.6
		slightly anxious	4	57.1
		very anxious	1	14.3
		Total	7	100.0
	female	slightly anxious	2	28.6
		anxious	3	42.9
		extremely anxious	2	28.6
		Total	7	100.0

Table 7 shows that in all the age groups majority of males responded that they get slightly anxious in a situation in which they must do well such as important exam or job etc. On the other hand majority of females become anxious in all the age groups.

In the age group of 11-20 years 37.5% males get incredible angry, in the age groups of 21-30 and 31-40 years 44.4% and 57.1% males get mildly angry if they are in argument with someone and they find themselves difficult to remain composed respectively. Among females 42.9% get angry in the age group of 11-20 years, 37.5% get mildly angry in the age group of 21-30 years and 42.9% females in rest of the age groups get incredible angry when they have argument with someone.

Table 8 show that majority of males in the age groups 11-20 and 40+ years have opion that environment in which they live affect their emotion. 57.1% males in the age group of 31-40 years somewhat agree with the above mentioned statement. While in the age group of 21-30 years majority of males disagree with the statement that environment i.e. social environment affects emotion. Among females percentages are higher are higher in all the age groups who have opion that environment in which they live affects emotion.

VIII. Conclusion

The present study indicates that there is little differences between men's and women's feelings and expressive behavior and gender-linked cultural beliefs about emotion. Majority of Bania males and females report neutral feelings till the age of 30 year, our analyses also revealed that men do not report more frequent emotional experiences than women in general. We did, however, find differences in the frequency with which men and women report positive and negative emotions. Women report more frequent positive and negative feelings than men. Percentages of males and females, who reported their emotions increase after the age of 30 years. Further analyses revealed differences in the frequency with which men and women report specific emotions. Men report more frequent feelings of lively, satisfied, excitement and angry. Whereas women report more frequent feelings of anxiety, sadness and embarrassment after the age of 40 years.

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Multi Touch: An Optical Approach (Comparison of various techniques)

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Abstract- This paper will explain how the Multi-touch technology presents a wide range of new opportunities for interaction with graphical user interfaces, allowing expressive gestural control and fluid multi-user collaboration through relatively simple and inexpensive hardware and software configurations. We as the developers of the low cost multi-touch table draw our experience to provide the practical knowledge to build and deploy applications on the multi-touch surface. This will include the hardware and software requirements, comparison of various optical techniques and implementation of the multi-touch surface.

Index Terms- Multi-touch surfaces, Multi-touch technology, Optical Techniques

I. Introduction

As we all know, Multi-touch surfaces have found their way into the futuristic visions of human-computer interaction. It is a method of input on a touch screen that allows two or more fingers to be used on the screen at one time. In computing, **multi-touch** refers to a touch sensing surface's (track pad or touch screen) ability to recognize the presence of two or more points of contact with the surface

Now the question is why multi-touch? For the human computer interaction multi-touch has become a common interface thereby resulting in increase in productivity and one step ahead in creating an innovative interface.

Concept of multi-touch is not new to the world, it dated back to1970. Our multi-touch table is low cost and denotes a set of interaction techniques which allow computer users to control graphical applications with several fingers and consists of a touch screen (e.g., computer display, table and wall) or touchpad, as well as software that recognize multiple simultaneous touch points, as opposed to the other touch screens.

There are numerous ways to construct multi touch surfaces:-

- Resistance Based Touch Surfaces: It is made up of two electrical resistive layers separated by thin air space. The layers have electrodes which are facing each other, when a contact is made both layers are pressed together to give the precise location.
- Capacitance Based Touch Surfaces: It consists of an insulator coated with a transparent conductor such as Indium Tin Oxide. When we touch the surface it results in distortion of electrostatic field, measurable as change in capacitance used for sensing the touch.
- Surface Wave Touch Surfaces: In this technique transmitting and receiving piezoelectric transducers, for both the X- and Y axis, are mounted on a faceplate and ultra-sonic waves on a glass surface are created and then directed by reflectors.
- Optical Touch Surfaces: Optical Touch technology is an economical way to add touch capability to any display. It accommodates multiple touch points with a high degree of precision and user can touch any part of the screen with any material- finger, pen or a stylus.

There are several ways to make a multi-touch surface. But this research paper will focus on Optical techniques and implementation of the multi-touch table using the front diffused illumination technique along with the comparison of various optical techniques. The main focus will be on optical techniques since they are the easiest and most cost effective for the average person to create.

II. Optical Techniques

Optical Techniques provide the most cost effective way of implementing multi touch. Each technique utilizes 3 main components:

1. Infrared Camera (or other optical sensor)
2. Infrared light
3. Visual Feedback (projector or LCD)

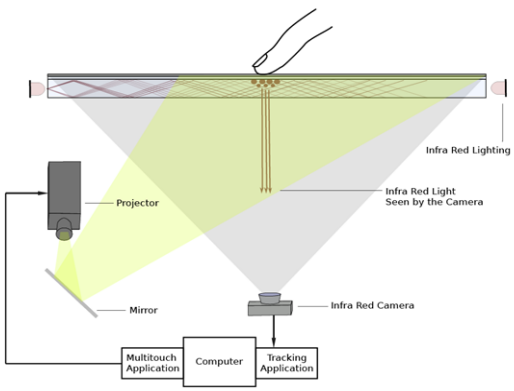

Fig 1: Illustration of basic structure of touch surface

An **infrared camera** is pointed at the touch surface to detect fingers/objects that touches the surface. **Infrared light** is used to so as to differentiate between visual images and the objects/fingers. Most systems have a **visual feedback** system

where an image from a **projector or LCD** is projected or placed below the touch surface. A tracking application uses the camera image to track and create touch coordinates.

Major optical techniques to construct touch surfaces:

- Frustrated Total Internal Reflection (i.e. Perceptive Pixels)
- Rear Diffused Illumination (i.e. Microsoft Surface)
- Front Diffused Illumination
- Diffused Surface Illumination
- Laser Light Plane
- LED Light Plane

1. FRUSTRATED TOTAL INTERNAL REFLECTION

FTIR uses the phenomenon of Total Internal Reflection in which infrared lights are placed adjacent to the edges (sides) of an acrylic panel. When the user touches the acrylic surface, the infrared light is "frustrated" which causes the light to escape internal reflection and scatter downwards which is scanned by an infrared camera (modified webcam).

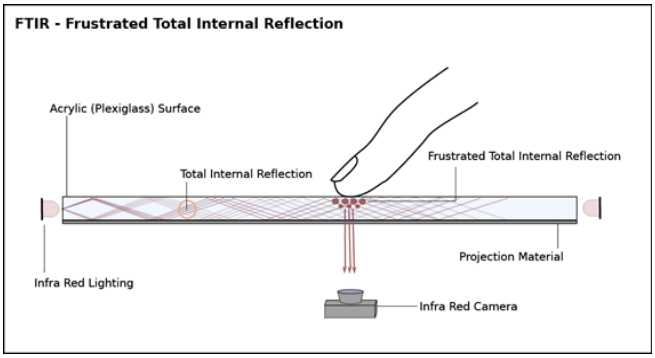

Fig 2: Illustration of Frustrated Total Internal Reflection

2. REAR DIFFUSED ILLUMINATION MULTITOUCH TECHNIQUE

In Rear DI, infrared light illuminates the screen from below the touch surface. A diffuser needs to be placed on top/bottom of the touch surface. When the user touches the surface, the infrared light hits his finger and is reflected downward and sensed by an infrared camera below the surface.

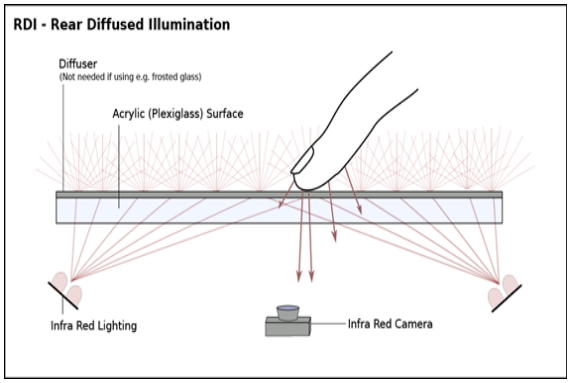

Fig 3: Illustration of Rear Diffused Illumination Technique

3. FRONT DIFFUSED ILLUMINATION MULTITOUCH TECHNIQUE

In Front DI, infrared light is shown from above the touch surface. A diffuser is placed on top/ bottom of the touch surface. When the user touches the surface, a shadow is formed under it and this is seen by an infrared camera placed below the surface.

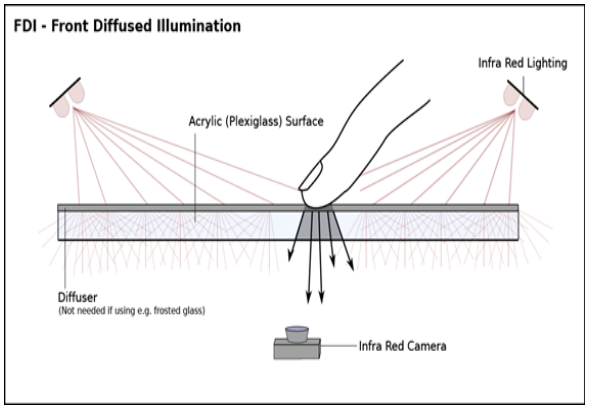

Fig 4: Demonstration of Front Diffused Illumination

4. DIFFUSED SURFACE ILLUMINATION

Like FTIR, in DSI infrared light is placed adjacent to the edges of an acrylic panel, directed towards the inside. It uses a special acrylic with small particles inside it, which act like hundreds of small mirrors for evenly distributing the light. On touching the diffuser, the light escapes out of the surface and thus the object/finger can be seen by the camera.

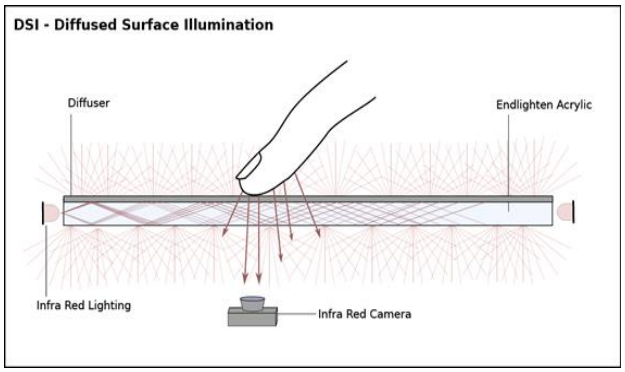

Fig 5: Illustration of DSI (Diffused Surface Illumination)

5. LASER LIGHT PLANE MULTITOUCH TECHNIQUE

In LLP, the surface is illuminated by the single or multiple lasers. The laser plane of light is about 1mm thick and proximally positioned to the touch surface. When a finger touches the light plane, the object lights up and viewed by an infrared camera below the surface.

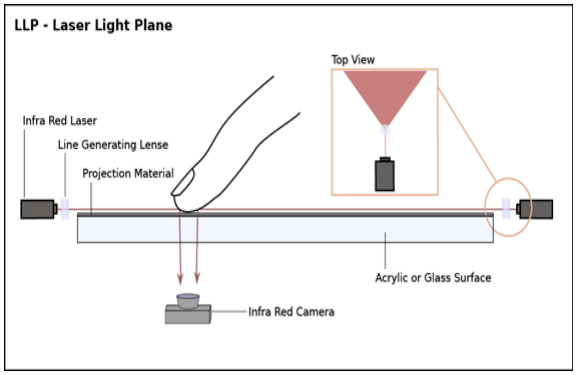

Fig 6: Demonstration of Laser Light Plane Technique

6. LED LIGHT PLANE MULTITOUCH TECHNIQUE

LED-LP, alike LLP forms a plane of light above the surface. In this narrow angle LEDs are used. The LEDs are placed just above the touch surface in order to create a plane of light. When a finger makes a contact with the light plane, it is seen by a infrared camera below the surface.

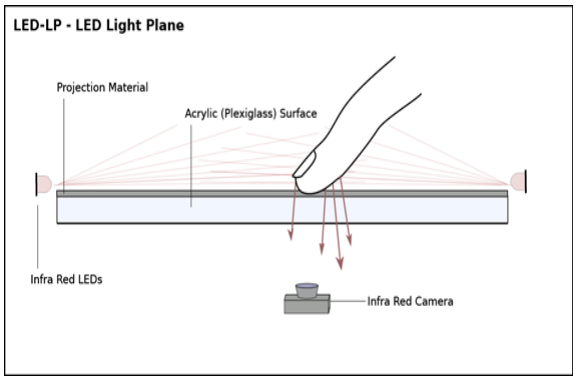

Fig 7: Illustration of LED Light Plane Technique

III. Comparison of Multi Touch Techniques

Even though FTIR is more popular than other techniques, there is no best technique. Each technique has its pros and cons as listed below.

Table 1: FTIR

Advantages	Disadvantages
In FTIR an enclosed box is not required	Soldering is required for making the LED frame
The Blobs have strong contrast	It Requires a compliant surface (silicone rubber) for proper application
It allows for varying the blob pressure on the surface.	It Cannot recognize the objects or fiducial markers
It can be used with something as pen tip since it uses a compliant surface.	It Cannot use a glass surface only the acrylic sheet is required for FTIR.

Table 2: Rear DI

Advantages	Disadvantages
Just a diffuser/projection surface on Top/bottom is used, no need for the compliant surface.	Even illumination is difficult to get
Any transparent material (glass or acrylic) can be used.	The lower contrast of blobs is harder to pick up by software.
LED frame not required.	More chances of detecting false blobs.
No soldering required.	Requires enclosed box.

Table 3: Front DI

Advantages	Disadvantages
Just a diffuser/projection surface on Top/bottom is used, no need for the compliant surface.	Objects and fiducials cannot be tracked.
Any transparent material (glass or acrylic) can be used.	It relies heavily on ambient lighting environment
LED frame not required	More chances of detecting false blobs.

Table 4: DSI

Advantages	Disadvantages
Compliant surface not required.	Cost of Enlighten Acrylic is more than regular acrylic
Easy to switch back and forth between DI (DSI) and FTIR	Lower contrast of blobs is harder to pick up by software compared to FTIR and LLP.
Objects, hovering, and fiducials can be detected.	
Hotspots are not formed.	

Table 5: LLP

Advantages	Disadvantages
Compliant surface not required.	Traditional objects and fiducials are not tracked by the setup.
Any transparent material (glass or acrylic) can be used.	It is not pressure sensitive as light intensity doesn't change with pressure.
LED frame not required.	Occlusion may occur if only 1 or 2 lasers are used ,where light striking one finger may block another finger from receiving light.
Slightly cheaper than other multi-touch Techniques	

Table 6: LED-LP

Advantages	Disadvantages
Compliant surface is not required.	Traditional objects and fiducials are not tracked by the setup.
Any transparent material (glass or acrylic) can be used.	It is not pressure sensitive as light intensity doesn't change with pressure.
Frame not required.	Occlusion may occur if only 1 or 2 lasers are used ,where light striking one finger may block another finger from receiving Light.

IV. Implementing a Low Cost Touch Table- Front Diffused Illumination Technique

1. Hardware Requirements

A. Infrared Illumination

Optical methods for multi touch require an infrared light source. Achieving the correct infrared illumination is challenging and requires a good knowledge of both the different methods of illuminating a surface and the various types of IR LEDs (5mm, 3mm, SMD (synchronous mirror delay)) that are available. All the IR-based set-ups employ light-emitting diodes (LEDs) as light sources. Commonly used types of IR LEDs include Osram SFH4250 (SMD) and Osram SFH485 (5 mm). Whether SMD devices or standard LEDs are more appropriate depends on various factors; say for example, if the LEDs are to be placed along the rim of an acrylic glass plate, SMDs can be used, as it is possible to simply attach them to the rim with instant glue.

B. Cameras

Optical Techniques rely on cameras to detect fingers touching the surface. For functional surface, a camera set-up must be capable of sensing light in the near-IR spectrum. Although challenging, the correct choice of camera is essential to gain high camera-signal quality required for a multi-touch surface. Camera sensors, capable of detecting IR light are required. When choosing a camera, sensors must be sensitive to wavelength of IR light. Web cameras contain an infrared filter to block ambient infrared light and must be removed. Although often it is either glued on to the lens or applied as a coating on the camera and must be scratched from the lens.

C. Projectors

Rear projection is generally used for displaying the actual image on the surface; but a number of factors must be considered when deciding upon an appropriate projector. Major factor is display resolution. The necessary projection resolution is strongly application dependent; however, a resolution of at least 1024_768 pixels (XGA) is usually sufficient.

D. Compliant Surfaces and Projection Screens

Multi touch set-up is comprised of a layer of acrylic or glass sheet augmented with a frame.

Compliant layer: A plain acrylic surface requires the user to apply significant pressure to get a responsive tracking. The use of a compliant surface overcomes this problem. Applying an additional layer on top of the acrylic or glass material greatly improves the sensitivity of the surface.

Projection layer: Depending on what material is used (silicone or latex), a different projection screen must be chosen. The main requirement is that an air gap should be achievable between the two layers when the screen chosen is pressed against the compliant surface.

2. Software Requirements

A. **OpenCV (Open Source Computer Vision Library)**

It is a library of programming functions aimed at real-time computer vision. It has interfaces for C++, C, Python and Java and supports a number of OS- Windows, Linux, MAC, and Android. OpenCV was designed to achieve computational efficiency and with a strong focus on real-time applications.

B. **CCV (Community Core Vision (CCV)**

CCV is a tracking application (Main Finger Tracking Application) – needed to track fingers and for calibration. It supports Windows, Linux, and Mac.

C. **TUIO (Tangible User Interface Object) Mouse Driver**

It is a protocol used for communicating the location (X and Y coordinates), size, and relative velocity of blobs.

D. **Flash Player Projector**

This is a flash player that runs outside of the browser. It is used to run multi-touch applications. Flash is used in advertisements and games.

4. Hardware Setup:

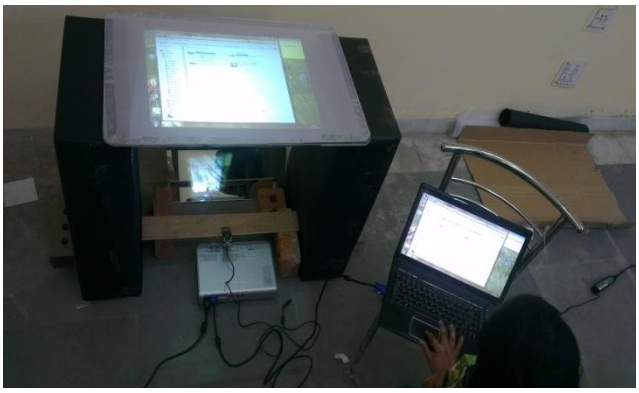

Fig 8: Multi Touch Panel

The **Glass (8 mm)** was placed and **Tracing Paper** was placed on the surface of the glass in order to display the projected image. An **infrared camera- (Intex Night Vision Camera)** was pointed at the touch surface to detect when fingers/objects that touches the surface. **Infrared light** – **ambient light** was used to distinguish between a visual image on the touch surface and the objects/fingers being tracked. **Panasonic Projector** was projected or placed below the touch surface. **Mirror** was placed at an angle of **45° (approx.)** in front of the projector to redirect the beamed image from the projector to the glass surface**.** The camera was also connected to the computer and a tracking application (CCV) used the camera image to track and create touch coordinates.

4. Camera Modification: Removal of IR Blocking Filter
Materials required:

1. Webcam
2. Screw Driver
3. Blade

Modification of a webcam was necessary to detect the IR blobs that were generated by our fingers. IR is not visible to the human eye, neither is it to most of the webcams. In order for a webcam to detect IR light, IR blocking filter has to be removed from it. In our case the IR-block filter was painted on the lens and could be scratched off. In most cases, it's impossible to remove the IR-block filter, without damaging the lens, as it is glued to the lens. In that case, it is wise to replace the default lens with a special lens that doesn't have an IR-block filter attached to it.

5. CCV Installation and Calibration:

Latest version of Community Core Vision was downloaded. The camera was attached to computer and the Community Core Vision (CCV) application was launched.Settings were changed to detect the fingers and calibration screen was opened by pressing "c" on keyboard. After calibrating test that the surface detects all your fingers.

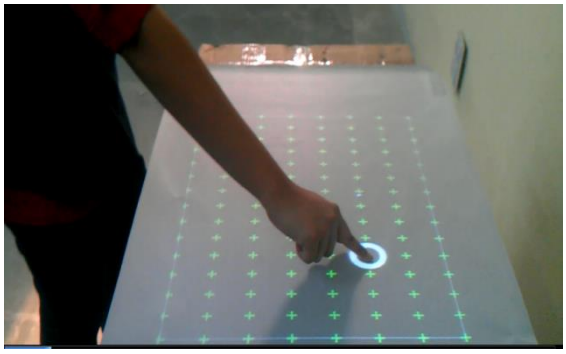

Fig 9: Calibration of the surface

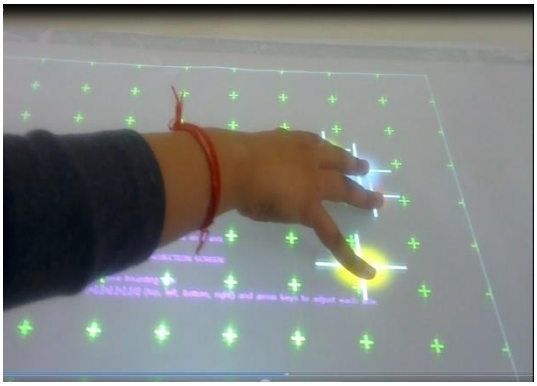

Fig 10: Testing for successful calibration

6. Working on multi touch

The laptop was connected to the projector and Windows-7 desktop screen was displayed. One can open different applications and can even type documents using touch instead of a keyboard.

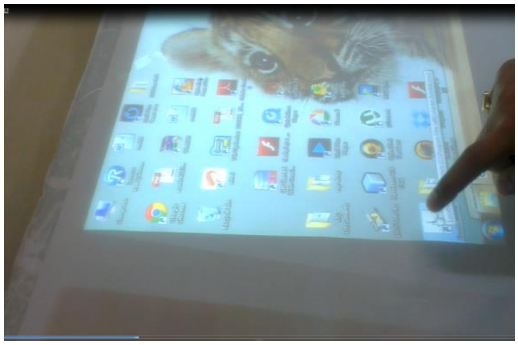

Fig 11: Clicking icons using multi touch

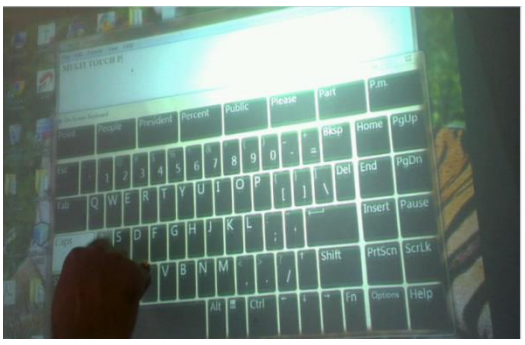

Fig 12: Typing Documents

Also Flash application was created and opened using Flash Projector. Tracking information from the CCV was send to the flash projector using one of the following:

- (CCV 1.1 or below), turn on **Send TUIO** in CCV
- (CCV 1.2+), turn on **Flash XML** in CCV

The screen was checked for input and location. If the system is not pointing to the correct location calibration has to be redone.

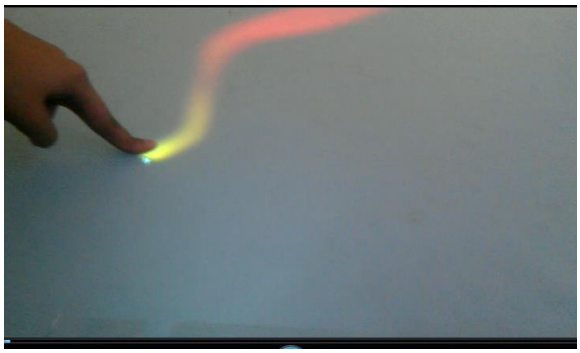

Fig 13: Fire Application

V. Conclusion

This report covers areas including but not limited to different methods and tools to build a multi-touch system and its working. The setup implemented by us is fairly simple. Literally every setup differs so there is a need to play around with most of the steps to find what works best for the setup. The basic concepts covered here will always apply. Similarly other techniques can be implemented practically. We also conclude that there is no best technique and each technique has its pros on cons. One can select which one to use depending upon the time, resources and conditions.

Multi-touch technology with computationally enhanced surfaces has attracted considerable attention in recent years as it greatly improves human-computer interaction. Optical approach for touch surface uses image processing to determine the locations velocity of the interacting object. Infrared illumination and simple setups mean that these systems can potentially be very robust. Implementation of frustrated total internal reflection (FTIR) and diffused illumination (DI), have enabled low-cost development of touch surfaces. Laser-light plane and diffused-screen illumination have their own advantages.

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DISCLAIMER: ALL AUTHORS CONTRIBUTED EQUALLY IN THE RESEARCH AND PUBLICATION OF THIS WORK.

Math Anxiety: The Poor Problem Solving Factor in School Mathematics

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Abstract- Math anxiety indicates psychologically a negative mind-set towards solving mathematical problems which impacts on students' learning practices and outcomes. In looking more closely at why a remarkable no of students may be struggling for improvement in mathematics in comparison to other subjects, it is timely to consider, the math anxiety factor. This literature review looks at the concept of math anxiety and solving mathematical problems highlighting math anxiety as an important factor of poor performance in terms of solving mathematical problems of school students in mathematics and how to assist in mitigating math anxiety.

Index Terms- problem solving, teaching and learning, math anxiety.

I. INTRODUCTION

From home to the work place, mathematics tools have become a part of our day-to-day life. In many circle of computer literacy, technological and mathematical competences are perceived as essential skills. Mathematics is truly the gateway of engineering and all other scientific and technological fields. The shortage of workers in technological fields is a matter of national concern. Thus, schools have responded to this need for mathematical competence by enhancing their course offering in mathematics. Also, as mathematics become part of daily life, the need for proper teaching and learning has become extremely essential. Because of this increase in needs, demands have also been placed in schools to educated students and make them "mathematically equipped". In today's high-tech world it is very important for young children to build confidence in mathematics for ever-increasing dynamic and competitive society (Furner and Berman-2003). At one time, Mathematics literacy might have been defined as knowing basic number facts and having proficiency with basic skills and procedures. But there is an increasing need for the students to understand and be able to use mathematics for the rapid change of the world. The way of describing mathematics literacy stated (Washington, D C-2001) how a person becomes mathematically literate:

Conceptual understanding: understanding mathematical concepts, operations and relations

Procedural fluency: skill in carrying out procedures flexibly, accurately, efficiently and appropriately

Strategic competence: the ability to formulate, represent, and solve mathematical problems.

Adaptive reasoning: the capacity of logical thought, reflection, explanation and justification.

Productive disposition: habitual inclination to see mathematics as sensible, useful and worthwhile combine with a belief in diligence and one's own efficiency.

Problem solving plays a significantly important role in mathematics teaching and learning. Trough problem solving students can enhance their thinking skills, apply procedures, deepen their conceptual understanding. Problem solving is like a vehicle by which students start thinking critically. During problem solving activities students involve with given information, formula, computations, figures, graphs etc. with synthesizing ideas, for which students must be involved with the entire learning system without any math anxiety. The study intends to highlight math anxiety as a factor of poor performance in problem solving and how to assist in mitigating this anxiety among school students.

Problem solving is the significantly important task and central part of teaching mathematics. Strategically handled proper effective teaching methodology can help the students to develop their problem solving skills. A crucial situation arises when one is in danger but does not know. A problem becomes critical due to lack of awareness of its existence. Teaching via problem solving has been emphasized more than teaching problem solving (Lester, Masingila, Mau, Lambdin, dos Santon and Raymond, 1994). Emphasising more on problem solving skill development many researchers have attempted to clarify what is meant by a problem-solving approach to teaching mathematics highlighting on teaching mathematical topics through problem-solving devices and relating with the real day to day life activities characterized by the teacher, which are helping students construct a deep understanding of mathematical concepts and procedures by involving them in doing mathematics: creating, conjecturing, exploring, testing, and verifying' (Lester et al., 1994, p.154). A problem-solving skill is like a vehicle for students to construct their own ideas, enhance logical thinking, transfer skills to unfamiliar situations and to take responsibility for their own development of learning. So, the problem solving component in mathematics teaching and learning has to be emphasized by the educators, parents and schools. Several researchers noted some important characteristics regarding problem-solving approach which include:

- mathematical dialogue and consensus between students (Van Zoest et al., 1994)
- To encourage students by making generalisations of rules and concepts (Evan and Lappin, 1994).
- teachers accepting right/wrong answers in a non-evaluative way (Cobb et al., 1991)

- teachers guiding, coaching, asking insightful questions and sharing in the process of solving problems (Lester et al., 1994)
- teachers knowing when it is appropriate to intervene, and when to step back and let the pupils make their own way (Lester et al., 1994)
- interactions between students/students and teacher/students (Van Zoest et al., 1994)
- teachers providing just enough information to establish background/intent of the problem, and students clarifying, interpreting, and attempting to construct one or more solution processes (Cobb et al., 1991)

The research on math anxiety has become more extensive as the researchers try to find why so many people intend to keep themselves away from mathematics. National Council of Teachers of Mathematics (NCTM, 1989) recognizes math anxiety as a problem and has specifically included in its assessment practices, as the teachers' job is to assess students' mathematical dispositions.

II. Concept "Math anxiety"

Mathematics anxiety can be defined as feelings of tension and anxiety that interfere with the manipulation of numbers and the solving mathematical problems in a open variety of societal life and academic situations. Math anxiety indicates psychologically a feelings of tension (Richardson & Suinn, 1972) which interferes in learning and performances. In D'Ailly & Bergering, (1992) it was mentioned as a fear and apprehension. Mathematics anxiety, considered a fear or phobia, produces 'a negative response specific to the learning, or doing, of mathematical activities that interferes with performance' (Whyte, 2009, p. 4). It is defined as low self confidence, a negative mind-set towards mathematics learning (Jain & Dowson, 2009), feeling threatened (Zohar, 1998), a factor of failing to reach potential (Perry, 2004) and a temporary reduction in working memory (Ashcraft & Kirk, 2001). Tobias and Weissbrod (1980) defined math anxiety as panic, helplessness, paralysis and mental disorganization that arises at the time of solving mathematical problems (Fiore, 1999, p-403). Tobias (1993) described math anxiety as a feeling of sudden death. In pursuance of these definitions and considerations by the researchers it can be considered to be a factor, which interferes with the manipulation of numbers and the solving of mathematical problems in academic, and social environments (Richardson & Suinn, 1972; Suinn, Taylor & Edwards, 1988). Thus, math anxiety as a psychological construct interferes in developing students' thinking skills it can be considered as a significantly important factor of poor problem solving skills of school students in mathematics.

There are some symptomatic characteristics of math anxiety helping in identifying children suffering from math anxiety, which are in the form of physical, psychological and behavioral (Plaisance, 2009; Jackson, 2008; Woodard, 2004):

Physical symptoms: It includes increased heart rate, clammy hands, upset stomach, light headedness.

Psychological symptoms: It includes inability to concentrate, feeling of helplessness, worry, and disgrace.

Behavioral Symptoms: It includes avoidance of mathematics classes, putting off the mathematics homework until the last minute and irregular study.

III. Co-relation with performance and achievement

Several researchers revealed in their studies that math anxiety is negatively correlated with the performance in mathematics, and if this issue is not dealt with properly, it could have a terrible effect in many areas of our education system. Fennema and Sherman (1976), using their math attitudes scales (MAS), found that math anxiety and mathematics ability concepts were highly correlated *(r* = -.89) in a sample of secondary school students. Mathematics anxiety has a highly negative relationship with mathematics performance and achievement, which has been described in Green, 1990; Hembree, 1990; Mevarech, Silber & Fine, 1991; Norwood, 1994; and Wigfield & Meece, 1988. All the mathematical features of a syllabus concerned may not be interested to some individuals. How certain features of mathematics, such as its precision, logic, and over emphasis to develop problem solving skills, make it particularly anxiety provoking for some individuals was discussed in Richardson and Woolfolk (1980). Math anxiety is an important factor in poor performing and achieving mathematics, Richardson & Suinn, 1972; Suinn, Edie, Nicoletti, & Spinelli,1972 documented the negative effects of math anxiety on math performance and achievement.

From the extensive research on co-relationship between math anxiety and students' performance in mathematics it can be noted that the math anxiety is a crucial factor for the students' poor performances in mathematics. If this factor is not taken into account sincerely and handled properly and effectively by the parents, educators, students themselves, schools and the policy making authorities concerned then it could have a terrible consequences for the entire education system, for instance, some major negative consequences of math anxiety are mathematics avoidance (Hembree, 1990), the effects of math anxiety are tied to those cognitive operations that rely on the resources of working memory (Mark H. Ashcraft1, 2002), distress (Tobias, 1978) and interference with conceptual thinking and memory processes (Skemp, 1986).

IV. Causes and prevention or mitigation

It is not easy task to determine the causes of math anxiety, where and how it begins and grows. There is some lack of agreement about the possible causes of mathematics anxiety in children (Newstead, 1995). In his studies the possible causes include teacher anxiety, societal, educational or environmental factors, innate characteristics of mathematics, failure and the influence of pre-school experiences of mathematics. Beginnings of anxiety can often be traced to negative classroom experiences seems particularly strong and well-documented (Tobias, 1978; Stodolsky, 1985). In assessment and evaluation system also there is a tendency to develop a negative attitude to the students' minds. Scarpello (2007) stated over reliance on high-stakes tests

has reinforced development of negative attitudes towards mathematics and increased students anxiety levels in mathematics. Teachers have reduced the diagnostic efficiency of mathematics tests by administering them in pressurised situation (Spark, 2011 ; Geist, 2010; Asheraft & Krause, 2007; Cavanaugh, 2007).

Without conceptual understanding the mathematical situations with computation skills it is harder to learn, for instance, memorised-by-writing rules and the manipulation of symbols with little are harder to learn mathematics than an integrated conceptual and understandable structure, and this can result in affective blocks for the children (Skemp, 1986). Traditional belief on results of any tests or examinations also may create a tendency towards math anxiety (Schoenfeld, 1988). Teachers' too much emphasis on memorising formulae, learning mathematics through drill and practice (Greenwood, 1984), applying traditional way- committing at home and vomiting in examination hall, can create math anxiety. Everyone is capable of learning, but may learn best in different ways. Therefore, lessons must be presented in a variety of ways, for instance, some teaching techniques could cause math anxiety Oberlin (1982): i) assigning the same work for everyone, ii) covering the text book problem by problem, iii) giving written work every day, iv) insisting that there is only one correct way to solve a problem, and v) assigning math problems as punishment for misbehavior. Thus, math anxiety can be considered as a function of teaching techniques used in the class rooms to convey mathematical skills which involve how a mathematical problem is solved mechanically, 'revision-memorise' teaching principles, which emphasises memorisation rather than conceptual understanding, procedural fluency and reasoning.

It can be expected that a friendly teaching approach which includes how to relate the mathematical problems with day to day life activities and process-oriented teaching method emphasising conceptual understanding rather than drill and practice will mitigate anxiety. It has also been suggested that encouraging students to discuss mathematical situations among themselves may have better consequences in terms of mitigation of anxiety (von Glasersfeld, 1991; Vacc, 1993). Also, to develop problem solving skills it is extremely important to take place discussion about various devices of solving problems among the students which prevents also math anxiety (Greenwood, 1984).
It is also possible that if the teachers start with a recapitulation of previous discussion and present the next in clearer and easier manner so that the students can handle the concepts comfortably with helping them to solve the mathematical problems rather than imposing them to create own solution steps, then it will strongly support the students' emotional need (Vinner, 1994). The teachers provide a leadership or guiding role in teaching and learning environment and therefore are highly influential as there exists no doubt of the role teachers play in what happens in their classrooms (Kulbir Singh Sidhu 1995). In the studies of Clute, (1984) and Norwood, (1994) it has been revealed that teachers' strong presentation in a transmission-type class room actually mitigate discomfort and math anxiety for those students who lack self efficacy in their own intuitions.

New concepts can be taught through play acting, cooperative groups, visual aids, hands on activities or information technology. Some important and effective suggestions provided in National Council of Teachers of Mathematics (NCTM,1989, 1995b) for teachers seeking to prevent math anxiety include:

- Accommodating for different learning styles
- Creating a variety of testing environments
- Designing positive experiences in math classes
- Refraining from tying self-esteem to success with math
- Emphasizing that everyone makes mistakes in mathematics
- Making math relevant
- Letting students have some input into their own evaluations
- Allowing for different social approaches to learning mathematics
- Emphasizing the importance of original, quality thinking rather than rote manipulation of formulas

According to the study, Hackworth (1992), the important activities, reported also in Furner, J, M. & Berman, B, T. (2003) which will assist in mitigating math anxiety:

i) discussing and writing about mathematics feelings ii) becoming acquainted with good math instruction as well as study techniques

iii) learning study methods that enable students to recognize different types of information learning

iv) being an active learner and creating problem-solving techniques

v) evaluating one's own learning

vi) developing calming, positive ways to deal with fear of math (e.g., visualization, positive messages, relaxation techniques, and "frustration breaks")

vii) building confidence in math through gradual, repeated success, .

In the study of Zemelman, S., Daniels, H., & Hyde, A. (1998) they found that the use of the following methods constituted best practices for teaching mathematics:
i) manipulatives (to make learning math concrete)
ii) cooperative group work
iii) discussion
iv) emphasizing the importance of questioning and making conjectures
v) justifying one's thinking
vi) writing as a way to explore mathematical thoughts and problem-solving methods
vii) the problem-solving approach to instruction
viii) and calculators, computers, and other technology.

The summery of the study is provided in figure-
1, 2 and 3 reflecting causes, consequences and commitment to mitigate math anxiety:

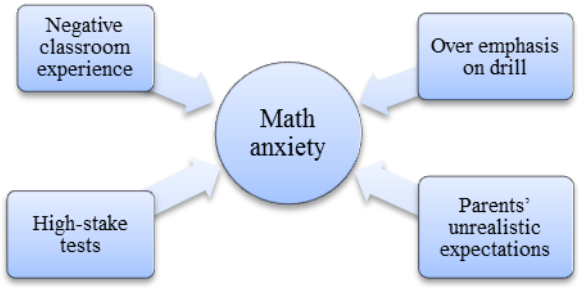

Figure-1: Causes

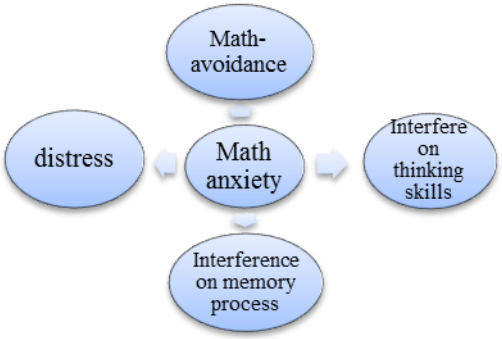

Figure-2: Consequences

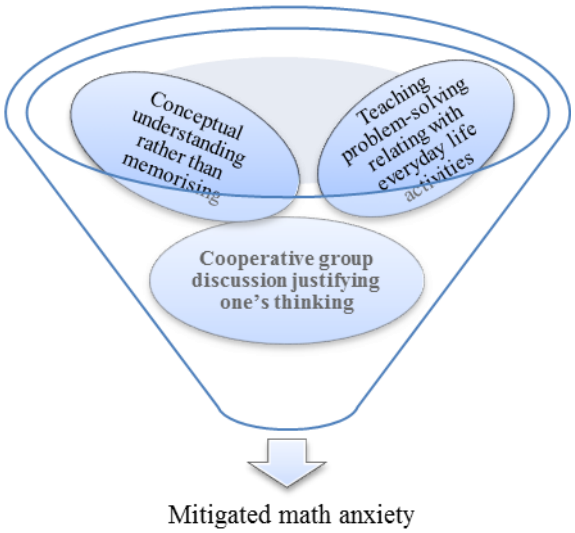

Figure-3: Commitment

V. CONCLUSION

Problem solving is the significantly important component of teaching and learning mathematics, it must be looked upon in a positive light to mitigate math anxiety, the psychological construct which interferes in developing students' thinking skills. It can be considered as a significantly important factor of poor problem solving skills of school students in mathematics. Therefore, the educators must re-define traditional teaching methodologies which often do not match students' learning styles and skills needed in society. Educators can play an instrumental role in fostering an environment of teaching and learning by presenting topics in a activity oriented manner to mitigate or prevent math anxiety. For instance, concept can be taught through mathematical modeling related to day to day life activities, forming math clubs for interaction among students on mathematical phenomena, audio-visual aids, hands on activities and technology. If once children see mathematics as fun, then it will be anxiety free for them, and, the enjoyment with mathematics could remain within them throughout the rest of their lives.

If this issue is not dealt with sincerely, it could have a terrible effect in many areas of our education system. There are some major consequences for math anxiety to be alert as mentioned earlier, mathematics avoidance, distress, interference with conceptual thinking and memory processes which may create deficiency of workers in the field of today's high-tech world.

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Analysis of Down-Wind Propeller Vehicle

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***Abstract* –** In the last decade many researches have been carried out on wind driven vehicle; a large number of academic publications have been presented. There have been many "Down Wind Faster Than The Wind (DWFTTW)" arguments based on energy flow. Wind driven vehicle systems travel faster than the wind along its direction. This paper deals with energy conversion mechanism of a vehicle driven by wind. The Downwind vehicle is designed with a propeller that pushes it along like an aircraft's propeller does. The propeller is connected to the wheels of the vehicle through a chain/belt drive, so that when the vehicle moves forward, the propeller spins in such a way to provide a thrust that will speed up the vehicle if there is energy available from a tailwind, which explains the energy conversion mechanism. Specifically, focus is on the horizontal axis propeller of a downwind vehicle to make it have a relative motion with the wind at a speed greater than the speed of the wind. The simulations of the analysis are carried out in JavaProp software.

***Index Terms* –** Aerofoil, DWFTTW, Momentum, Propeller, Thrust Force

I. INTRODUCTION

The energy crisis, environmental concerns and scarcity of conventional fuel have increased interest in green engineering. Using wind-power to produce energy in order to propel a vehicle is one such application of green engineering.

If a uniform wind is available over a level surface, is it possible to construct a man-carrying vehicle which by use of wind energy alone can accelerate in the wind direction from zero speed up to a speed larger than the wind speed? The work described in the paper has been carried out for the purpose of answering this problem.

The first downwind vehicle was created by Andrew Bauer [1] in 1969 which went downwind at a speed of 1.2 times the wind speed. In 1978, B. L. Blackford [2] applied the basic laws of mass, energy and momentum conservation to this novel wind-craft. Later in 2002, a wind test apparatus was introduced by Frank Bailey [3] which offered a means of testing a model of the downwind vehicle. The apparatus basically consisted of a wind tunnel erected over a towing tank so that velocity could be measured along the course of the model.

In 2003, a suggestion was made by Theo Schmidt [4] to use either an ogival or an un–cambered profile propeller in combination with a swiveling drive so that the thrust produced by it pushes the vehicle forward. In addition to it, Victor Korepanov [5] in 2004 explained the analysis behind achieving a speed of four times the wind speed. Also there has been an unofficial claim of going 4.2 times faster downwind, but there has been no authentic tests run of it.

John C. Wilson [6] in 2005 demonstrated that the appropriate principle to get forward thrust in a downwind vehicle is that the propulsion mechanism must be moving at a speed less than the wind speed. The same year, in July, Peter A. Sharp [7] proposed a simplified demonstration model of a Bauer air propeller vehicle that was relatively easy to construct.

In 2006, Jack Goodman [8] explained that the correct gearing of the propeller to the wheels of the car will speed up the car if energy is available from a tailwind. Again John C. Wilson [9] in 2007 conducted a study of the gear-ratio between the speed of the propeller and the speed of the wheels for a Bauer vehicle.

In 2009, Mark Drela [10], [11] formulated the Velocity as well as the Power analysis for the downwind vehicle. Later, the same year, a simple optimization method for both the wind turbine and propeller rotor, based on the Blade Element Momentum Theory was presented by Mac Gaunaa, Stig Øye and Robert Mikkelsen [12].

A comparison of the various aerofoil profiles was made by Shethal Thomas Kodiyattu [13] in 2010 while designing the propeller of a downwind vehicle. The latest explanation was given by S. Morris [14] in the same year, in June, which compared the propeller of a downwind vehicle to that of an aircraft.

The most successful and the latest wind-driven vehicle was the Blackbird [15] built by thin Air Designs Team in July 2010 which travelled downwind upto 2.8 times the wind speed, according to the NALSA speed regulations [16].

Nomenclature

V	Vehicle Speed
W	Wind Speed
F_t	Drag Force on Vehicle
F_p	Thrust Force on air Propeller
F_{net}	Net Thrust of the Vehicle
C_t	Coefficient of Thrust
C_p	Coefficient of Power
C_r	Coefficient of Rolling Resistance
P_t	shaft Power of Vehicle wheels
P_p	shaft Power of air Propeller
P_{net}	Net Power developed
A_d	Air Drag Area
A_p	Area of air Propeller disk
N	rpm of Propeller
D	Diameter of Propeller
D_{sp}	Diameter of Spinner or hub
ρ	Density of Air
η_g	Gearing/transmission Efficiency
η_p	Total Efficiency of air Propeller
η_{swirl}	Swirl Efficiency of air Propeller
ΔV	Change in Velocity of Vehicle
ΔW	Change in Velocity of air Propeller

II. RESEARCH AND ANALYSIS

A. Force and Power Analysis

Considering the forces acting on the vehicle as shown in the Figure 1,

$F_{net} = F_p - F_t$

Since we know that $P_p = P_t \eta_g$ or

$$F_p \frac{V-W}{\eta_p} = F_t V \eta_g$$

Therefore $F_{net} = F_t \cdot (\frac{V}{V-W} \cdot \eta_g \cdot \eta_p - 1)$

which is positive only as long as

$\frac{V}{V-W} \cdot \eta_g \cdot \eta_p > 1$ (requirement of DWFTTW)

Applying energy conversion principle, the net power developed by the vehicle could be written as the sum of power produced by propeller thrust and power lost by drag force on the vehicle and kinetic energy deposited in vehicle and air,

$P_{net} = P_t - P_p$

$= F_p W - F_{net} V - \frac{F_t \Delta V}{2} - \frac{F_p \Delta V}{2}$

In steady-state operation, P_{net} must be sufficiently positive to balance the remaining power losses in the system. A positive P_{net} indicates that the power delivered by wind energy could be used to accelerate the vehicle beyond the wind speed. The other losses not accounted for were power-transmission losses, profile-drag losses on the propeller and turbine blades, and swirl losses in the propeller and turbine slipstreams.

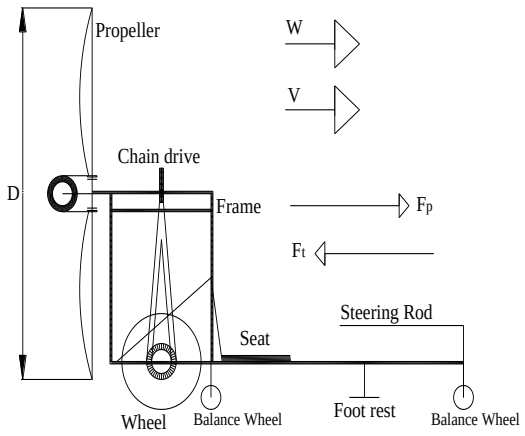

Figure 1: Side view of the air propeller vehicle

For optimization purposes, the following dimensionless parameters were introduced which characterize the operation of any Downwind vehicle.

Excess Thrust Ratio: $F = \frac{F_{net}}{F_p}$

Apparent Velocity Ratio: $Z = \frac{V-W}{W}$

Air Propeller Thrust Coefficient: $C_t = \frac{2F_p}{\rho V^2 A_p}$

Using the above relations, the equation could be rewritten as

$$\frac{W}{F_p} C_r + Z^2 \frac{1}{C_t} \frac{A_d}{A_p} = \left\{ 1 + \left[\frac{2\eta_g \eta_v}{z + \sqrt{z^2 + \frac{C_t}{\eta_{swirl}}}} - 1 \right]^{-1} \right\}^{-1}$$

Z could be replaced here by $\frac{V-W}{V}$ to obtain the $\frac{V}{W}$ ratio.

B. Analysis using JavaProp

For a quick and accurate analysis of the various forces acting on the propeller, JavaProp software was used.

The input parameters of the propeller required to be given in this software were Propeller name, Propeller Diameter, Spinner (or Hub) Diameter, Speed of rotation, axial inflow Speed, Number of blades and Thrust/power/torque required. For these input parameters as shown in Table I, the output obtained in JavaProp are as shown in Table II.

The basic equations used for the analysis were

$C_t = \frac{F_p}{\rho n^2 D^4}$

$C_P = \frac{P}{\rho n^3 D^5}$

Advance Ratio, $J = \frac{V}{nD}$

Efficiency, $\eta = \frac{V}{nD} \frac{C_t}{C_p}$

Table I: Input values in design tab of JavaProp

Propeller Name	NACA 6412
Propeller Diameter, D	4 m
Spinner Diameter, D_{sp}	0.15 m
Revolutions per minute (rpm)	120 rpm
Velocity, V	4 m/s
Number of Blades, B	2
Thrust	100 N

The output given by JavaProp included Advance Ratio, Efficiency, Thrust, Power, Coefficient of Thrust, Coefficient of Power and Pitch.

Table II: Output result in design tab of JavaProp

$\frac{V}{n\,D}$	0.5	$\frac{V}{\Omega R}$	0.159
Efficiency, η	70.47%	*loading*	medium
Thrust, T	100 N	C_t	0.08
Power, P	567.65 W	*Cp*	0.0568
β at 75% R	18°	*Pitch H*	3.05 m

Solving for $\frac{V}{W}$ using η_{swirl} = 0.95, $\frac{W}{F_p} C_r$ = 0.02, $\frac{A_d}{A_p}$ = 0.04 and with the result obtained from Table II, it was obtained that $\frac{V}{W}$ = **2.525** which indicates that the velocity of vehicle is 2.525 times faster than the velocity of wind. Correspondingly, $\frac{V}{W}$ Vs C_t graphs have been drawn for various values of η_{net} as shown in Figure 2 which confirms that $\frac{V}{W} > 1$.

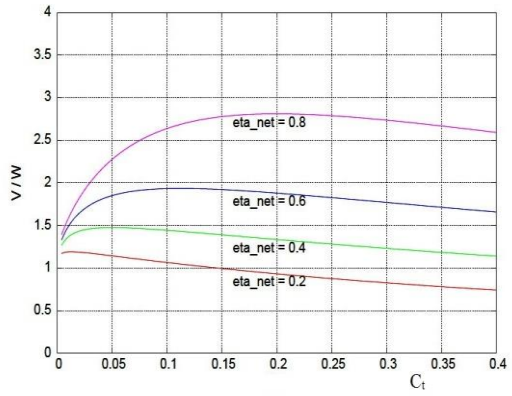

Figure 2: $\frac{V}{W}$ ratio vs. propeller thrust coefficient (C_t)

III. RESULTS OF SIMULATION

The results shown in this paper are for NACA 6412 aerofoil profile which was chosen after comparison between CLARK Y, NACA 6412, NACA 9412 and MH 114 profiles since it has a better L/D ratio, less camber and is comparatively easier to construct. The profile and geometry for NACA 6412 are shown in Table III and Figure 3.

Table III: Aerofoil profile characteristics

Propeller Profile Name	NACA 6412
Thickness	0.12c
Camber	0.06c
Trailing edge angle	14.2°
Lower flatness	0.812c
Leading edge radius	0.17c
Max CL	1.785
Max CL angle	12.0^{o}
Max L/D	60.34
Max L/D angle	4.0^{o}
Max L/D CL	1.268
Stall angle	4^{o}
Zero lift angle	-6.0^{o}

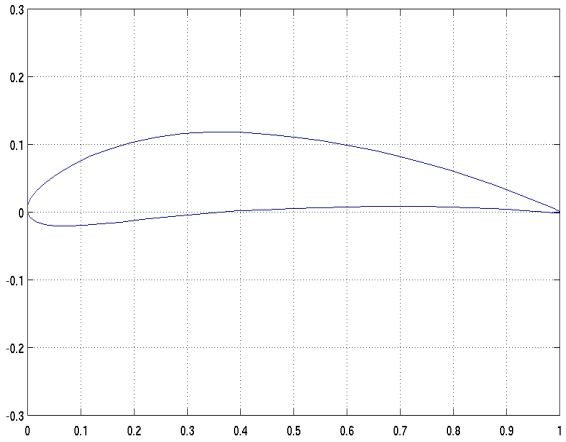

Figure 3: Geometry of NACA 6412 aerofoil profile

When analyzed using JavaProp for 4m and 8m diameters of the propeller, it was seen that the thrust and power coefficients decreased as the diameter was doubled (Figure 4) and the efficiency remained approximately the same but for a smaller advance ratio (Figure 5) of the propeller.

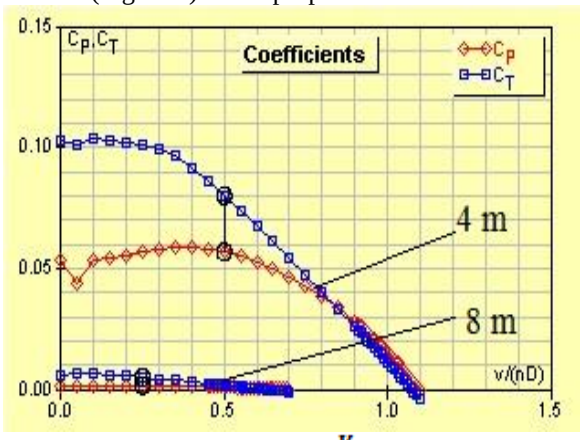

Figure 4: C_t, C_p vs. advance ratio $\frac{V}{nD}$ for NACA 6412 aerofoil for different diameters

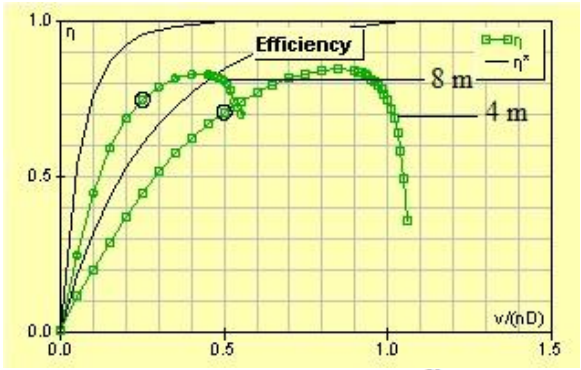

Figure 5: Efficiency (η) Vs advance ratio $\frac{V}{nD}$ for NACA 6412 aerofoil for different diameters

IV. CONCLUSION

It is theoretically possible to build a wind driven vehicle that can go in the downwind direction faster than the free stream wind speed (using a propeller in the air). There does not exist a definite upper limit for vehicles of this kind. As long as efficiencies are improved, the velocities would also increase unasymptotically. The calculations above show that it was possible to go downwind even 2.5 times the speed of wind in a wind propelled vehicle.

A variable pitch propeller is suggested so that by varying pitch angle we can maintain an optimal angle of attack (maximum lift to drag ratio) on the propeller blades as vehicle speed varies. Further analysis using computational software is suggested to understand the velocity and pressure changes that occur around the propeller in order to get a better design to travel faster downwind.

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Introducing Feature Extraction & Least square regression in Automatic Image Annotation

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Abstract- Automatic image annotation is the process of assigning keywords to digital images depending on the content information. Automatically assigning keywords to images is of great interest as it allows one to index, retrieve, and understand large collections of image data. Selection & implementation of proper feature extraction and weight calculation method plays important role in image annotation. Many techniques have been proposed for image annotation in the last decade that gives reasonable performance on standard datasets. This paper presents detailed literature about the Automatic Image Annotation and explains the feature extraction as well as weight calculation method.

Index Terms- Automatic image annotation, Literature, feature vector, concept modeling,Saliency.

I. INTRODUCTION

With the prevalence of digital imaging devices such as digital cameras and mobile phones, a large number of images are produced everyday. An emerging issue is how to efficiently and effectively search required image items from a huge image database. A promising approach is keyword-based image retrieval that allows the user to search for image items using keywords, but traditional manual image annotation is both laborious and subjective. Hence automatic image annotation has attracted considerable attention. Automatic Image annotation has been a topic of on-going research and several techniques have been proposed. [1,2,3,4,7,8,9,10]

Automatic image annotation consists of analyzing the visual content (colors, textures, shapes) of the images (or the objects existing in the images) in order to transform it into meaningful symbolic (or textual) information.. It is used in many applications domains such as entertainment, commerce education, biomedicine, military, and web image classification also saves lot of manual work of human. But AIA is difficult task because visual content to be analyzed depend on many factors such as shooting conditions, instances of objects, lighting condition, resolution of camera , and the background clutters.

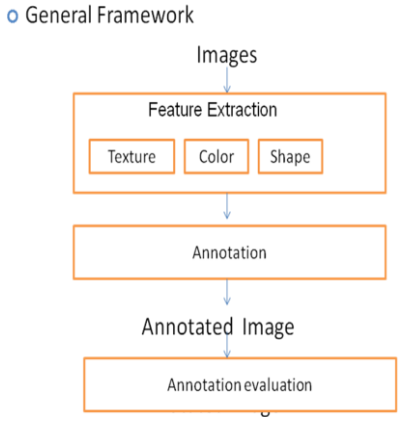

Fig. 1 General Framework Automatic Image annotation

State of the art automatic image annotation systems can be analyzed and grouped from various point of views. The remainder of paper is organized as follows. Review of the existing approaches is explain in section II. Proposed method is explain in III and we conclude our work in section IV.

II. RELATED WORK

A. Image Annotation

In Automatic Image Annotation three main groups are identified according to learning techniques and the application domain. classification model, probabilistic model and graph based models.

a) Classification model: Use discriminative model to classify images. Annotating images treats as classifying it to number of predefined groups called concepts or keyword. Since image is annotated by more than two words . So AIA problem is also considered as multi classification problem Classifier calculates the similarity of all trained classes and models unobserved instance to the class with highest similarity[6,7,9,12]. Both supervised[21,28] and unsupervised learning methods are used for classification. In supervised classification pre-trained classes are used whereas in unsupervised clustering used for classification.

b) Probabilistic model: Generative models are used to specify joint probability of image region feature and set of words [13,21,22] word to image relation and word to word relation are estimated. After estimating words are ranked according to probabilities. Probabilistic model are more flexible

in learning process but computational cost of parameter estimation is more.

c) Graph based model : Various graphical models are used for finding relation between word to image or word to word relation.[3,11,27] However time complexity and space complexity makes it difficult to apply on real word annotation.

Authors Reviewed many supervised learning models such as probabilistic classifier, artificial neural network, support vector machine, decision trees ,k-nearest neighbor, majority voting, bagging, boosting, stacked generalization. Also reviewed hybrid learning models which based on both clustering and classification[15].Most of these treat annotation as translation from image instances to keywords. The translation paradigm is typically based on some model of image and text co-occurrences.

B. Features

The essential requirement in image retrieval, indexing, classification is extracting efficient features from images. The color, texture and shape features are the most widely used visual features. There are many color spaces used RGB,HSV, LUV etc. Color feature is extracted from images or regions. Many color feature can be used such as color moments such as mean, standard deviation and skewness. Usually they are calculated for each color channels (components) separately .But all color information in image is not represented by moments.

The color histogram describes the color distribution of an image. The color coherence vector (CCV) incorporates spatial information into the basic color histogram. CCV is preferable than histrogram.

MPEG-7 also standardizes a number of color features including dominant color descriptor (DCD), color layout descriptor (CLD), color structure descriptor (CSD), and scalable color descriptor (SCD).

Texture is another important image feature.

Many methods have been proposed. In spatial approach, texture features are extracted by finding the local pixel structures in original image domain. This method is easy to understand but sensitive to noise.

In spectral texture feature extraction techniques, an image is transformed into frequency domain and then feature is calculated from the transformed image. For images or regions with sufficient size, spectral texture features best choice. However, for small images or regions, especially when the regions are irregular, spatial features desirable.

Shape feature is useful for extracting feature in many applications. Many region shape descriptors are commonly used in color image retrieval, including, area, moments, circularity, and eccentricity. In this work salient regions are consider for color feature extraction.

C. Saliency

Human eye is perceptually more sensitive to certain colors and intensities and objects with such features are considered more salient. Salient regions are most important point in image which attracts greater attention by visual system than other part of the image. These regions has distinctive features when compared with others in image. Eg. a polar bear is salient on dark rocks, but almost invisible in snow.

Recently, several saliency approaches came up that are based on computational and mathematical ideas and usually less biologically motivated. These approaches range from the computation of entropy[28],[29] over determining features that best discriminate between a target and a null hypothesis to learning the optimal feature combination with machine learning techniques.

In[30] proposed saliency detection method spectral residual. Spectral residual is the difference between original log spectrum and its mean-filtered version. The saliency map is obtained by applying inverse Fourier transform to the spectral residual. We compute the color histogram of saliency regions for the color space RGB..

III. PROPOSED WORK

A. Automatic Image Annotation

AIA is a process of assigning keywords to digital images automatically depending on the content information.

More formally, an image I with set of visual features Vi={V1,V2….Vn} and set of keywords W={w1,w2….wm},find automatically the keyword subset Wi which describe image I.

B. Concept

A training set S consisting of N images with n feature vectors. n feature vectors forms a feature matrix and A pair of similar and dissimilar images(L). The main purpose of this paper is to investigate the feature selection properties in the image annotation task. This image pair setting helps us to create a feature matrix that contains the same groups of features. Thus, we can directly do feature analysis on this matrix within the same framework.

Calculate the weight assign to each feature vector by using feature matrix and L, which is final step of training stage. Weight vector is used to find relevancy of keyword to the image. Sufficient training is necessary to have correct annotations to the image in testing stage.

Feature vector of input image is compared with feature matrix of the training images. Based on the weights calculated, most similar images are find out from L. Keywords from L are getting assigned to test image which are annotations.

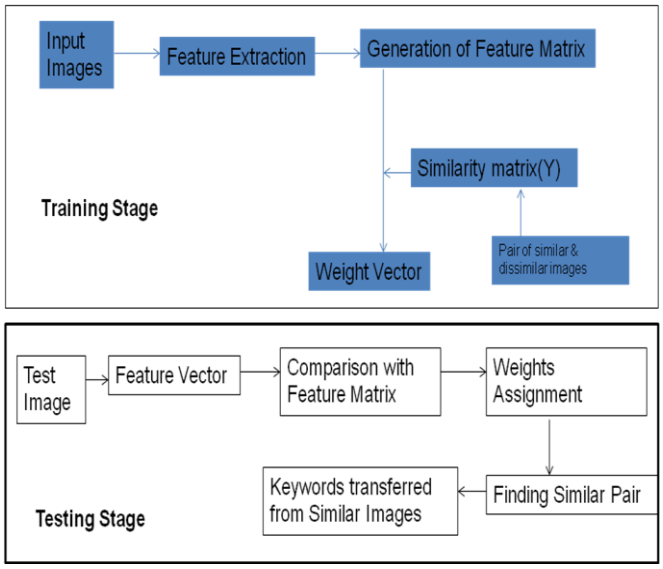

Fig 2.Proposed Architecture

C. *Feature Extraction*

[30] Paper argues that the spectrum residuals corresponds to image saliency. Given an input image, the log spectrum L(*f*)is computed from the down-sampled image with height (or width) equals 64 px. The selection of the input size is related to visual scale.

If the information contained in the L(*f*) is obtained previously,the information required to be processed is:

$$H(\mathrm{R}(f)) = H(\mathrm{L}(f))|\mathrm{A}(f)) \quad (1)$$

Where A(f) denotes the general shape of log spectra, which is given as prior information. R (f) denotes the statistical singularities that is particular to the input image, define R (f) as the spectral residual of an image. The average spectrum A(f) can be approximated by convoluting image

$$\mathrm{A}(f) = h_n(f) * \mathrm{L}(f) \quad (2)$$

Therefore the *spectral residual*R(f) can be obtained by:

$$\mathrm{R}(f) = \mathrm{L}(f) - h_n(f) * \mathrm{L}(f) \quad (3)$$

And saliency is obtained by

$$\mathrm{S}(x) = F^{-1}[\exp(\mathrm{R}(f) + \mathrm{P}(f)]$$

In order to obtain a better visual display the final saliency map was actually presented as

$$\mathrm{S}(x) = g(x) * F^{-1}[\exp(\mathrm{R}(f) + \mathrm{P}(f)]^2 \quad (4)$$

Where F and F^{-1} denote Fourier transform and inverse Fourier transform respectively; *g and* h_n are low pass filter. P(f) denotes phase spectrum of image which is assumed to be preserved during process.

D. *Weight Vector calculation*

Here weight of each feature vector is calculated by using feature matrix obtained using SR method and set of similar and dissimilar image pair. In this setting we consider any pair of images that share enough keywords to be positive training samples and any pair with no keywords in common to be negative example. In this work we obtained training samples from the designated training set of the Corel5K dataset. Image paires that had at least four common keywords were treated as positive sample for training and those with no common keywords were used as negative samples[31].

Weighted least square is an efficient method that makes good use of small data sets.It also shares the ability to provide different types of easily interpretable statistical intervals for estimation , prediction

The most popular loss function to calculate *w* in this regression problem is the least square estimate, which is also named as the minimizer of the residual sum of squared errors and is given as

$$w = \arg\min_{w \in Rp} ||Xw - Y||^2_2$$

This weight is used for testing stage in image annotation task.

Here we are using image pairs instead of one to-the-others mode, In this image annotation task, we have hundreds of keywords to model. Each image may have three to five keywords. Furthermore, different data set can have different sets of keywords. Thus, it is complicated to model exhaustively such relations using one-to-the-others mode. Image pair setting does not need to concern modelling these keywords directly. Instead, it just models the keyword similarity among images. Thus, it can easily assign keywords to testing images using existing annotations from training data.

IV. CONCLUSION

In this paper, detailed literature about various image annotation algorithms is mentioned. Proper feature extraction and weight calculation method plays important role in image annotation. This paper explains the method for the same. These methods are used as training stage in image annotation task. Future work is to work for testing stage and to implement this algorithm and to compare with various algorithms available.

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A Comprehensive Survey on Semantics for Data Compression

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***Abstract*-** Natural phenomena show that many creatures form large social groups and move in regular patterns. To reduce the data an efficient distributed mining algorithms are used to jointly identify a group of moving objects and discover their movement patterns in wireless sensor networks. In object tracking applications, many natural phenomena show that objects often exhibit some degree of regularity in their movements. To reduce the data volume, various algorithms have been proposed for data compression and data aggregation. In this paper we have surveyed various research papers of movement pattern mining, clustering, and data compression techniques.

Index Terms- Movement pattern Mining,Clustring, Data Compression.

I. INTRODUCTION

RECENT advances in location-acquisition technologies, such as global positioning systems (GPSs) and wireless sensor networks (WSNs), have fostered many novel applications like object tracking, environmental monitoring, and location-dependent service. These applications generate a large amount of location data, and thus, lead to transmission and storage challenges, especially in resource constrained environments like WSNs. To reduce the data volume, various algorithms have been proposed for data compression and data aggregation [1].

Discovering the group movement patterns is more difficult than finding the patterns of a single object or all objects, because we need to jointly identify a group of objects and discover their aggregated group movement patterns. The constrained resource of WSNs should also be considered in approaching the moving object clustering problem. [5]

In order to study the movement behavior of dynamic objects, it is important to take a closer look at movement itself. In other words, it is necessary to know what exactly the variables are that define movement, what constraints and external factors affect movement and most importantly to understand what types of movement patterns can be composed from these primitives of movement.

Generally, movement patterns include any recognizable spatial and temporal regularity or any interesting relationship in a set of movement data, whereas the proper definition (i.e. the instantiation) of "pattern interestingness" depends on the application domain.[2]

Clustering is the unsupervised classification of patterns (observations, data items, or feature vectors) into groups (clusters). The clustering problem has been addressed in many contexts and by researchers in many disciplines; this reflects its broad appeal and usefulness as one of the steps in exploratory data analysis. However, clustering is a difficult problem combinatorially, and differences in assumptions and contexts in different communities has made the transfer of useful generic concepts and methodologies slow to occur[3]

Process of reducing the amount of data needed for storage or transmission of a given piece of information typically by use of encoding techniques. Data compression is characterized as either lossy or lossless depending on whether some data is discarded or not, respectively.

Data compression can reduce the storage and energy consumption for resource-constrained applications.

In [1], Distributed source coding uses joint entropy to encode two nodes' data individually without sharing any data between them; however, it requires prior knowledge of cross correlations of sources.

The related works are classified into following three types.

Movement Pattern Mining
Clustering
Data compression.

II. MOVEMENT PATTERN MINING

In object tracking applications, many natural phenomenon show that moving objects often exhibit some degree of regularity in their movements. For example, the famous annual wildebeest migration demonstrates that the movement of creatures is temporally and spatially correlated. In addition, biologists have found that many creatures, such as elephants, zebra, whales, and birds, form large social groups when migrating to find food, or for breeding, wintering, or other unknown reasons. These characteristics indicate that the trajectory data of multiple objects may be correlated.Moreover, some research domains, such as the study of animals' social behavior and wildlife migration [16], [17], are more concerned with a group of animals' movement patterns than each individual's. This raises a new challenge of finding moving animals belonging to the same group and identifying their aggregated movement patterns.Many researchers model the temporal-and-spatial correlations of moving objects as sequential patterns in data mining, and various algorithms have been proposed to discover frequent movement patterns [18], [19], [20].

However, such works only consider the movement characteristics of a single object or all objects. Other works, such as [11] take the euclidean distance to measure the similarity of two entire trajectories, and then derive groups of mobile users based on their movement data. Since objects may be close together in some types of terrain, such as gorges, and widely distributed in less rugged areas, such as grassland, their group relationships are distinct in some areas and vague in others. Instead of applying global clustering on entire trajectories, examining partial trajectories of individual areas shows more opportunities of revealing the local group relationships of moving objects.

Following figure shows an example of tracking a group of moving objects

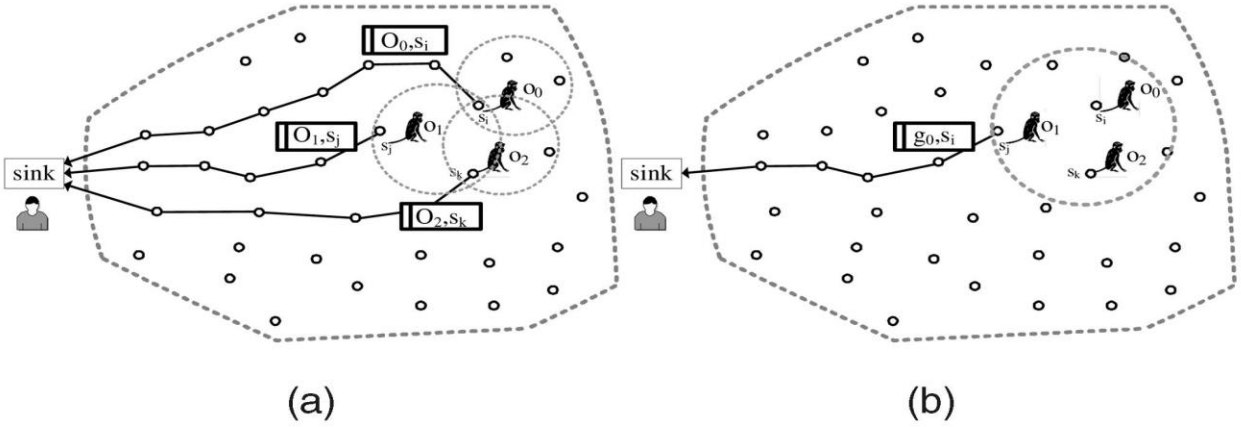

Fig. 12. An example of tracking a group of moving objects.(a) Monitoring multiple objects, respectively. (b) Monitoring multiple objects with group data aggregation.

Agrawal and Srikant first defined the sequential pattern mining problem and proposed an two algorithms AprioriSome and AprioriAll algorithm to find the frequent sequential patterns. AprioriSome and Apriori- All, have comparable performance, albeit AprioriSome performs a little better when the minimum number of objects that must support a sequential pattern is low.They show that the Accuracy of AprioriSome is better than AprioriAll algorithm[4].

Han et al. consider the pattern projection method in mining sequential patterns and proposed FreeSpan, which is an FP-growth-based algorithm.They reexamin pattern mining problem and proposed novel sequential pattern mining method called Freespan (i,e Frequent pattern- projected sequential pattern mining). The general idea of the method is to integrate the mining of frequent sequences with the frequent pattern and used projected sequence databases to confine the search and the growth of the subsequences fragments.FreeSpan mines the complete set of patterns but greatly reduces the effort of candidate subsequence generation.They show that FreeSpan examines a substantially smaller number of combination of subsequences and runs considerably faster than the Apriori based GSP algorithm.[5]

Yang and Hu developed a new match measure for imprecise trajectory data and proposed TrajPattern to mine sequential patterns. Many variations derived from sequential patterns are used in various applications.They propose the model of the trajectory patterns and a novel measure to represent the expected occurrences of a pattern in a set of imprecise trajectories. The concept of pattern groups is introduced to present the trajectory patterns in a concise manner. Since the Apriori property no longer holds on the trajectory patterns, a new min-max property is identified and a novel TrajPattern algorithm is devised based on the newly discovered property. Last but not least, they also calculate efficiency, and scalability TrajPattern algorithm.[6]

Chen et al. discover path traversal patterns in a Web environment, while Peng and Chen mine user moving patterns incrementally in a mobile computing system. However, sequential patterns and its variations like do not provide sufficient information for location prediction or clustering. First, they carry no time information between consecutive items, sothey cannot provide accurate information for location prediction when time is concerned. Second, they consider the characteristics of all objects, which make the meaningful movement characteristics of individual objects or a group of moving objects inconspicuous and ignored. Third, because a sequential pattern lacks information about its significance regarding to each individual trajectory, they are not fully representative to individual trajectories. [7][8]

Giannotti et al extract T-patterns from spatiotemporal data sets to provide concise descriptions of frequent movements. [9]Tseng and Lin used the TMPMine algorithm for discovering the temporal movement patterns. TMP-Tree for efficiently discovering the temporal movement patterns of objects in sensor networks. This is the first work on mining the movement patterns associated with time intervals in OTSNs. However, novel location prediction strategies that utilize the discovered temporal movement patterns so as to reduce the prediction errors for energy savings. Through empirical evaluation on various simulation conditions and real dataset, TMP-Mine and the prediction strategies are shown to deliver excellent performance in terms of scalability, accuracy and energy efficiency.[10]

III. CLUSTERING

Clustering is the unsupervised classification of patterns (observations, data items, or feature vectors) into groups (clusters). clustering denotes the grouping of a set of data items so that similar data items are in the same groups and different data items are placed in distinct groups. Clustering thus constitutes fundamental data analysis functionality that provides a summary of data distribution patterns and correlations in a dataset. Clustering is finding application in diverse areas such as image processing, data compression, pattern recognition, and market research, and many specific clustering techniques have been proposed for static datasets [21].The clustering problem has been addressed in many contexts and by researchers in many disciplines; this reflects its broad appeal and usefulness as one of the steps in exploratory data analysis. Recently, clustering based on objects' movement behavior has attracted more attention.

A straightforward approach to the clustering of a large set of continuously moving objects is to do so periodically. However, if the period is short, this approach is overly expensive, mainly because the effort expended on previous clustering are not leveraged. If the period is long, long durations of time exist with no clustering information available. Moreover, this brute-force approach effectively treats the objects as static object and does not take into account the information about their movement. For example, this has the implication that it is impossible to detect that some groups of data are moving together. Rather, clustering of continuously moving objects should take into account not just the objects' current positions, but also their anticipated movements.

Figure 2 illustrates the clustering effect where Connected black and the white points denote object positions at the current time and a near-future time.

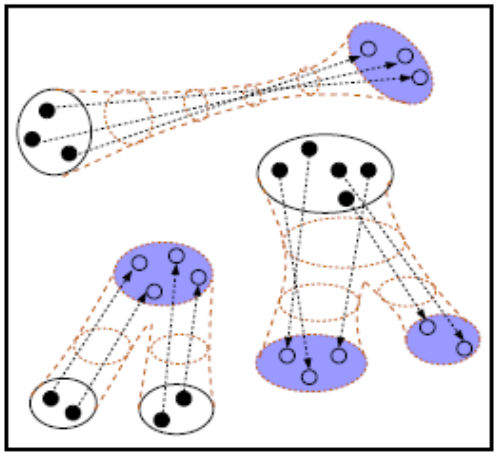

Fig 2: Clustering of moving object

Wang et al. present a new approach to derive groupings of mobile users based on their movement data. They assume that the user movement data are collected by logging location data emitted from mobile devices tracking users. We formally define group pattern as a group of users that are within a distance threshold from one another for at least a minimum duration. To mine group patterns, we first propose two algorithms, namely AGP and VG-growth. In our first set of experiments, it is shown when both the number of users and logging duration are large, AGP and VG-growth are inefficient for the mining group patterns of size two. We therefore propose a framework that summarizes user movement data before group pattern mining. In the second

series of experiments, They show that the methods using location summarization reduce the mining overheads for group patterns of size two significantly also they conclude that the cuboid based summarization methods give better performance when the summarized database size is small compared to the original movement database.[11]

Nanni and Pedreschi [12] proposed a density-based clustering algorithm, which makes use of an optimal time interval and the average euclidean distance between each point of two trajectories, to approach the trajectory clustering problem. However, the above works discover global group relationships based on the proportion of the time a group of users stay close together to the whole time duration or the average euclidean distance of the entire trajectories. Thus, they may not be able to reveal the local group relationships, which are required for many applications.

In addition, though computing the average Euclidean distance of two geometric trajectories is simple and useful, the geometric coordinates are expensive and not always available. Approaches, such as EDR, LCSS, and DTW, are widely used to compute the similarity of symbolic trajectory sequences [13], but the above dynamic programming approaches suffer from scalability problem [23]. To provide scalability, approximation or summarization techniques are used to represent original data. Guralnik and Karypis [23] project each sequence into a vector space of sequential patterns and use a vector-based K-means algorithm to cluster objects. However, the importance of a sequential pattern regarding individual sequences can be very different, which is not considered in this work.

IV. DATA COMPRESSION

Data compression can reduce the storage and energy consumption for resource-constrained applications.
Pradhan et al.[1] uses distributed source (Slepian-Wolf) coding technique for data compression which uses joint entropy to encode two nodes' data individually without sharing any data between them; however, it requires prior knowledge of cross correlations of sources.in this work they presented a new domain of collaborative information communication and processing through the framework on distributed source coding. This framework enables highly effective and efficient compression across a sensor network without the need to establish inter-node communication, using well-studied and fast error- correcting coding algorithms.

A. Scaglione and S.D. Servetto [14], used routing algorithm for data compression and also combine data compression with routing by exploiting cross correlations between sensor nodes to reduce the data size.

In [15], a tailed LZW has been proposed to address the memory constraint of a sensor device. Summarization of the original data by regression or linear modeling has been proposed for trajectory data compression.

V. CONCLUSION

With the growth of the data in resource constrined application there is need to compress the amount of data shared between the two nodes.In this survey paper, we briefly explored various techniques of movement pattern mining ,clustering and data compression suggested by authors. Anyway we believe that the most interesting research area deals with the discovering of semantics within movement pattern mining so to improve the results of data compression. Efforts in this direction are likely to be the most fruitful and compresses the much more location data quite effectively.

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Review Paper on Video Watermarking Techniques

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Abstract- The protection and illegal redistribution of digital media has become an important issue in the digital era. This is due to the popularity and accessibility of the Internet now a days by people. This results in recording, editing and replication of multimedia contents. Digital watermarking can be used to protect digital information against illegal manipulations and distributions. Digital watermarking technique is the process of embedding noise-tolerant signal such as audio or image data in the carrier signal. This technique provides a robust solution to the problem of intellectual property rights for online contents. This paper reviews different aspects and techniques of digital watermarking for protecting digital contents.
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Index Terms- Attacks, Content protection, Digital properties, DWT, SVD, Security, Watermarking techniques.

I. INTRODUCTION

Digital data are distributed across high-speed networks like the Internet and World Wide Web. This data is easily accessible for sharing. Due to this access possibility of tempering data and republishing it as own is increased. This leads the motivation of techniques providing security to this multimedia content. Digital watermarking is the technique used for this purpose. Various techniques of watermarking are used to insert data about ownership of contents, which help to keep the integrity of data.
A watermark is information about origin, ownership, copy control etc. This information is embedded in multimedia content with taking care imperceptibly and robustness.The watermark is embedded and extracted as per requirement. Video watermarking is different from image watermarking, because additional data are available here that allows information to be more redundantly and reliably embedded.

Digital video is a sequence or collection of consecutive still images. The amount of information that can be embedded in the video sequence is called payload. In reality video watermarking techniques need to meet other challenges than that in image watermarking schemes such as large volume of the inherently repeated sequence of data between frames.

The watermark embedding scheme can either embed the watermark into the host signal or to a transformed version of the host signal. Transform domain watermarking is a scheme that is used to transform image frequency domain in such a way to modify the transform coefficient. Some common transform domain watermarking for image data can be Discrete Cosine Transform (DCT) based [2, 3] or Discrete Wavelet Transform (DWT) based [4]. This scheme is very useful for taking advantage of perceptual criteria in the embedding process for designing watermark techniques. Spatial domain watermarking on the other hand has the capability of performing some transformation directly on image pixels. The use of perceptual models is also an important component in generating an effective and acceptable watermarking scheme for audio just as it is used in image watermarking [3, 4].

This paper is organized into five sections. The subsequent section explains the important aspects of video watermarking. Section II focuses the widespread techniques of video watermarking where various domains of video watermarking are explored and a robust algorithm in each domain is considered.

II. ASPECTS OF VIDEO WATERMARKING

Video sequencing is a collection of consecutive and equally time spaced still images. Apparently any image watermarking technique can be extended to watermark videos, but in reality video watermarking techniques needs to meet other challenges. Watermarked video sequences are very much susceptible to pirate attacks such as frame averaging, frame swapping, statistical analysis, digital-analog (AD/DA) conversion, and lossy compressions [1].

Watermarking systems can be characterized by a number of defining properties including embedding effectiveness, fidelity, data payload, blind or informed detection, false positive rate, capacity, robustness, perceptual transparency, security, cipher and watermark keys, modification and multiple watermark, cost, tamper resistance, unobtrusiveness, ready detection, unambiguous, sensitivity, and scalability. Some of them are common to more practical applications. In this section, such general properties will be listed and briefly discussed and focus will put on video watermarking. These properties are discussed due to their importance in watermarking applications.

A. Perceptual Transparency

Invisibility is the degree at which an embedded watermark remains unnoticeable when a user views the watermarked contents. However this requirement conflicts with other requirements such as tamper resistance and robustness, especially against lossy compression algorithms. To survive the next generation of compression algorithms, it will probably be necessary for a watermark to be noticeable to a trained observer which is asked to compare the original and the marked version of the video.

B. Robustness

Robustness is the resilience of an embedded watermark against removal by signal processing. The use of music, images and video signals in digital form, commonly involves many types of distortions, such as lossy compression. For watermarking to be useful, the mark should be detectable even after such distortions occurred. Robustness against signal distortion is better achieved if the watermark is placed in perceptually significant parts of the signal.

Due to large amounts of data and inherent redundancy between frames, video signals are highly vulnerable to pirate attacks, such as frame averaging, frame dropping, rotation, sharpening [1].

C. Capacity

Capacity is that amount of information that can be expressed by an embedded watermark. Depending on the application at hand, the watermarking algorithm should allow a predefined number of bits to be hidden.

III. REVIEW OF VIDEO WATERMARKING TECHNIQUES

Many digital watermarking schemes have been proposed in the literature for still images and videos. Most of them operate on uncompressed videos [8, 11, 13], while others embed watermarks directly into compressed videos [12, 13]. Video watermarking applications can be grouped as security related like Copy control [8], fingerprinting, ownership identification, authentication, tamper resistance etc. or value added applications like legacy system enhancement, database linking, video tagging, digital video broadcast monitoring [8], Media Bridge etc.

Existing video watermarking techniques are divided into different categories as shown in Figure 1. They can be divided into 3 main groups based on the domain that the watermark is embedded.

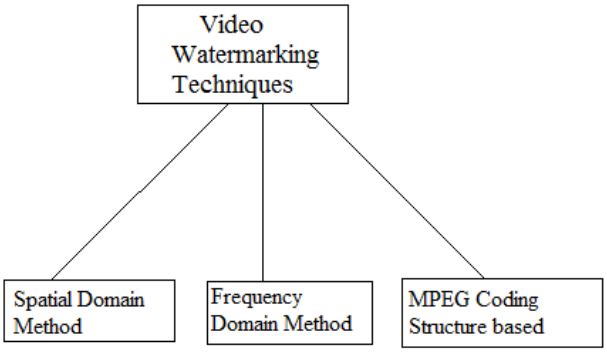

Figure 1. Classification map of existing digital video watermark techniques

A. Spatial Domain Watermarking

The spatial domain watermarking techniques embed the watermark by modifying the pixel values of the host image/video directly. Low computational complexities and simplicity are the main strengths of pixel domain methods. For better performance in real time these techniques are more attractive.

- *Least Significant Bit Modification [15]*

In this technique, the Least Significant Bit of each pixel is used to embed the watermark or the copyright information. In this technique cover image is used to store the watermark, in which we can embed a smaller object multiple times. The pixels are identified where embedding will be done using a pseudo-random number generator based on a given key.

LSB modification is suitable tool for stegnography as it is a simple and powerful tool for it.But it can not preserves robustness which is required in watermarking applications.

- *Correlation-Based Techniques*

The most straightforward way to add a watermark to an image in the spatial domain is to add a pseudo random noise pattern to the luminance values of its pixels. A Pseudo-random Noise (PN) pattern W (x, y) is added to the cover image I (x, y), according to the given below:

$$I\,w\,(x, y) = I\,(x, y) + k\,*W\,(x, y)$$

Where k denotes a gain factor and Iw the watermarked image. The robustness of the watermark is increased by increasing the value of k at the expense of the quality of the watermarked image. The same key is given as an input to retrieve the watermark, to the same pseudo-random noise generator algorithm, and the correlation between the noise pattern and possibly watermarked image is computed. If the correlation exceeds a certain threshold T, the watermark is detected, and a single bit is set [14]. This method can easily be extended to a multiple-bit watermark by dividing the image into blocks and performing the above procedure independently on each block.

B. Frequency Domain Watermarking

Most of watermarking techniques [6-8], the watermark will be embedded into the frequency domain instead of the spatial domain for the robustness of the watermarking mechanism. Discrete Cosine Transformation (DCT), Discrete Fourier Transformation (DFT) and Discrete Wavelet Transformation (DWT) are the three main methods of data transformation in this domain. The main strength offered by transforming domain techniques is that they can take advantage of special properties of alternate domains to address the limitations of pixel-based methods or to support additional features. Generally, transform domain methods require higher computational time.

- *Discrete Fourier Transform [14]*

This approach first extracts the brightness of the to-be-marked frame, computing its full-frame DFT and then taking the magnitude of the coefficients. The watermark is composed of two alphanumerical strings. The DFT coefficient is altered, then IDFT. Only the first frame of each GOP is watermarked, which was composed of twelve frames, leaving the other ones uncorrupted. It is good robustness to the usual image processing

as linear/non-linear filtering, sharpening, JPEG compression and resist to geometric transformations as scaling, rotation and cropping.

- *Discrete Cosine Transform [14]*

In Discrete Cosine Transform, an image is broken up into different frequency bands, to get middle frequency bands of an image where watermark can be embedded easily.

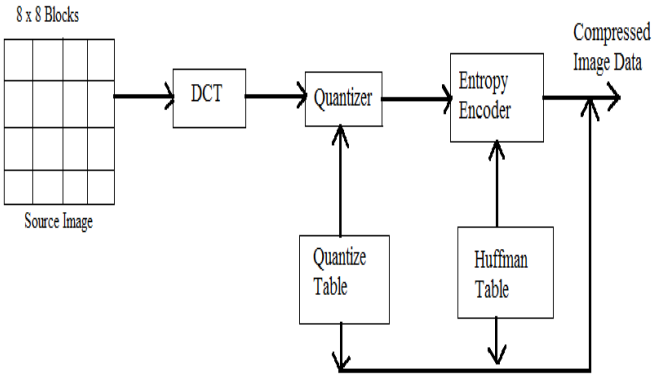

Figure 2. Process of Discrete Cosine Transform (DCT)

The following are steps carried out in the encoding procedure of DCT:

1. The image is broken into N*N blocks of pixels.

$$D(i,j) = \frac{1}{\sqrt{2N}} C(i) C(j) \sum_{x=0}^{N-1} \sum_{y=0}^{N-1} p(x,y) \cos\left[\frac{(2x+1) i\pi}{2N}\right] \cos\left[\frac{(2y+1) j\pi}{2N}\right]$$

P (x, y) is x, the element of an image represented by matrix p .N is size of block that DCT is done on. The equation calculates one entry (i, j) of the transformed image from pixel values of the original image matrix.

2. In matrix multiplication the DCT is applied to each block from left to right, top to bottom.
3. Each block's element is compressed through quantization means dividing by some specific value.Qunatization is achieved by dividing each element in transforming image matrix by the corresponding element in quantization matrix.
4. The array of compressed blocks which represent the image is stored in a reduced amount of space. It is carried out using zig-zag sequences.

- *Discrete Wavelet Transform*

The DWT decomposes an input image into four components namely LL, HL, LH and HH where the first letter corresponds to applying either a low pass frequency operation or high pass frequency operation to the rows, and the second letter refers to the filter applied to the columns.

The basic idea in the DWT for a one dimensional signal is the following. A signal is split into two parts, usually high frequencies and low frequencies. The edge components of the signal are largely confined to the high frequency part. The low frequency part is split again into two parts of high and low frequencies. This process is continued an arbitrary number of times, which is usually determined by the application at hand.

Let

$$H(\omega) = \sum_k h_k e^{-jk\omega};$$

And

$$G(\omega) = \sum_k h_k e^{-jk\omega};$$

be a lowpass and a highpass filter, respectively. A signal, x[n] can be decomposed recursively as

$$c_{j-1,k} = \sum_n h_{n-2k} c_{j,n}$$

$$d_{j-1,k} = \sum_n g_{n-2k} c_{j,n}$$

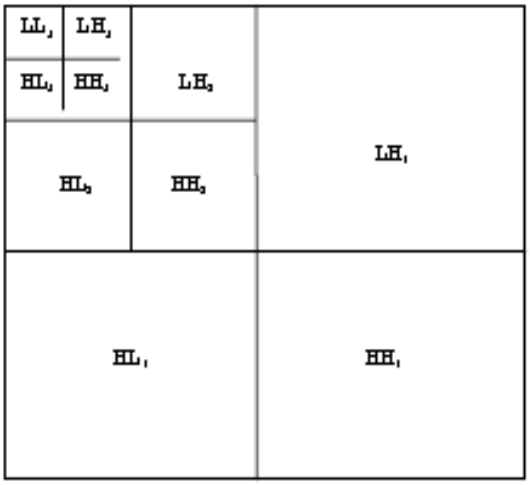

Figure 3. The model of DWT decomposition

for $j = J+1, J, \ldots, J_0$ where $c_{J+1,k} = x[k]$, $k \in Z$, $J+1$ is the high resolution level index, and J_0 is the low resolution level index.

In the encoding part of DWT while watermarking, we first decompose an image into several bands with a pyramid structure as shown in Fig. 3 and then add a pseudo-random sequence (Gaussian noise) to the

largest coefficients which are not located in the lowest resolution.

- *Contourlet Transform (CT)*

DWT offers multistage and time frequency localization of the image. However, it fails to represent the image effectively, if the image contains smooth contours in different directions. CT addresses this problem due to its inherent characteristics, viz., directionality and anisotropy.

Laplacian pyramid and the directional filter Bank [4] are combined in Contourlet transform. Figure 4 shows a multiscale and directional decomposition using a combination of a Laplacian pyramid (LP) and a directional filter bank (DFB) [4]. Bandpass images from the LP are fed into a DFB so that directional information can be captured. The scheme can be iterated on the coarse image. The combined result is a double iterated filter bank structure, named contourlet filter bank, which decomposes images into directional subbands at multiple scales.

The result of the process described above returns decomposition of the input video frames with different frequency bands. A watermark is also decomposed using CT and mapped with low pass coefficients of host video frames. Inverse Contourlet Transform (ICT) is applied to the modified sub bands to get final watermarked video.

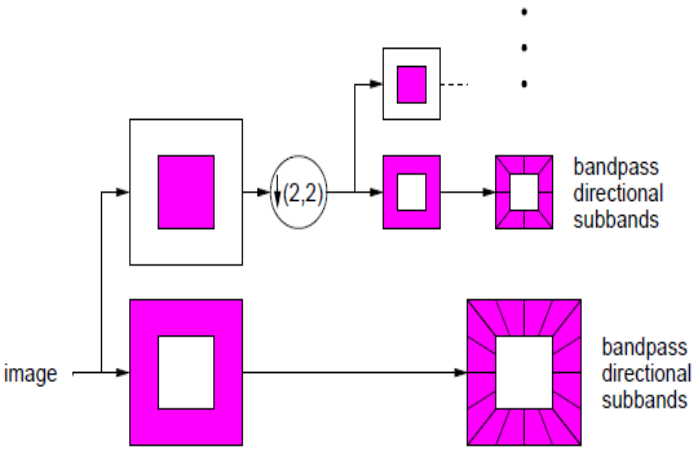

Figure 4. The contourlet filter bank

C. Watermarks Based on MPEG Coding Structures

Video watermarking techniques that use MPEG-1, -2 and -4 coding structures as primitive components are primarily motivated by the goal of integrating watermarking and compression to reduce overall real-time video processing complexity. Compression in block-based schemes like MPEG-2 is achieved by using forward and bi-directional motion prediction. There are a number of MPEG-2 and -4-based techniques that have been proposed, including approaches based on GOP modification, high frequency DCT coefficient manipulation, DCT block classification [1].

In [5], video object watermarking is based on the structure of MPEG-4. This technique can be easily added to the embedding and detection schemes without changing the watermarking algorithm. It modifies some predefined pairs of quantized DCT coefficient in the luminance block of pseudo-randomly selected MBs. It is based on spread-spectrum techniques. Dividing the image into blocks of equal size, them binary sequence is generated using a secret key, and then adds to the image.

In [6], watermarking procedure embeds copyright protection into video sequences which is object based transparent. Each watermark is created by shaping an author and video dependent pseudo-random sequence according to the perceptual masking characteristics of the video. As a result, the watermark adapts to each video to ensure invisibility and robustness. Furthermore, the noise like watermark is statistically undetectable to prevent unauthorized removal.

In [7], Mobasseri proposed direct sequence watermarking using m-frames This scheme applies a direct sequence spread spectrum model to the watermarking of the digital video. First, video signal is modeled as a sequence of bit planes arranged along the time axis. Watermarking of this sequence is a two layer operation. A controlling m-sequence first establishes a pseudorandom order in the bit plane stream for later watermarking.

D. Other Watermarking Techniques

The watermark can be either directly inserted in the raw video data or integrated during encoding process or implemented after compressing the video data. In this section we briefly discuss some of video watermarking techniques present in literature.

A novel collusion resistant (CR) video watermarking approach is proposed in [8]. This is a practical frame by frame video watermarking technique. Here a basic s × s watermark pattern is first created and this pattern is repeatedly embedded so that it is centered on a fixed number of selected points known as anchors in every video frame. The part of the video frame where the basic watermark is embedded is called the footprint [8]. Anchor points are calculated using a feature extraction algorithm. After generating these watermark frames within a given host frame, spatial masking is applied on it to ensue robustness and imperceptibility criteria. Then the scaled watermark is embedded in the host data using addition.

In Watermarking using CDMA modulation [9] one of the four least significant billions are replaced by watermark planes. The bitlanes to be replaced are selected according to a random periodic quaternary sequence. The watermark plane is generated using the 1D spread spectrum methodology.

In [10] a watermarking method using variable length code (VLC) swapping was proposed. This methodology was based on the observation that in the MPEG-2, H.26l VLC tables there are pairs of code words $(r, l) \rightarrow c_0$ and 1 $(r, l \pm 1) \rightarrow c_1$ such that *length*

$(c_0) = length\ (c_1)$, $lsb\ (c_0) \neq lsb\ (c_1)$. Such level-adjacent pairs are called label-carrying VLC (lc-VLC) [10]. A covert data bit U_i is embedded into a frame by extracting eligible lc-VLC, $c_i \in \{c_0\} \cup \{c_1\}$, and swapping a codeword, if necessary such that to ensure $lsb(c_i) = U_i$. This process does not use any random key based component as a result of that this method is not robust against attacks.

A perceptual watermarking (PW) method explicitly model masking properties of the HVS and utilizes these models to analyze video sequences of frames to embed watermark in the optimal way. The five main properties of the HSV namely, frequency sensitivity, luminance sensitivity, contrast masking, edge masking and temporal masking can be exploited by video watermarking techniques [11], [12].

IV. CONCLUSION

This study discuss a number of techniques for the watermarking of digital images, also focus on the limitations and promises of each. LSB substitution does not provide robustness hence it is not a very efficient approach for digital watermarking. LSB embedded watermarks can easily be extracted using techniques that do not visually degrade the image to the point of being noticeable. DCT domain watermarking proved to be highly challenging to JPEG compression as well as considerable amounts of random noise. The wavelet domain as well proved to be highly resistant to both compression and noise, with minimal amounts of visual degradation. The counters proposed to geometric distortion typically rely on discovering the exact rotation, or shifting used in the attack. Typically these techniques are computationally pricey, and unpredictable.

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Agent Based Network Sniffer Detection

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Abstract- A Sniffer is a program on the network traffic by grabbing information travelling over a network [1]. Many people assume computers connected to a switch are safe from sniffing here we go for Antisniffing [2]. Nothing could be further from the truth. Computers connected to switches are just from vulnerable to sniffing. Computers connected to switches are just as vulnerable to sniffing as those connected to a hub. Here in this paper we propose Mobile Agents to detect sniffers. Mobile agents perform a task by migrating and executing on several nodes connected to the network. Ignored to detect sniffers [3], the network administrator sends some special types of mobile agents in the network and collects information from different nodes. After analyzing this information the network administrator can identify the computer system running in promiscuous mode.

Index Terms- Computer Security, Mobile Agent, Sniffer, Sniffer Detection

I. INTRODUCTION

Packet sniffing is a technique of monitoring every packet that crosses the network. In theory, it's impossible to detect these sniffing tools because they are passive in nature, meaning that they only collect data. While they can be fully passive, some aren't therefore they can be detected. This paper discusses the different packet sniffing methods and explains how AntiSniff tries to detect these sniffing programs.

II. METHODOLOGY

A mobile agent[5] is a software agent that has the additional property that it is not bound to operate only in the system in which it started. A mobile agent is autonomous because it may decide itself where it will go, what it will do there, and how long it will exist for. However, its environment or other mobile agents may also influence it.

Although mobile agents do not provide a solution to any previously unsolvable problems, they do have advantages over other technologies. They can be used to benefit or to simplify different types of application areas. Some examples of these application areas include ecommerce, distributed information retrieval, telecommunication networks services, and monitoring and notification.

Adaptive Learning: Mobile agents[6] can learn from experiences and adapt themselves to the environment. They can monitor traffic in large networks and learn about the trouble spots in the network. Based on the experiences of the agent in the network the agent can choose better routes to reach the next host.

Autonomy: Mobile agents can take some decisions on its own. For example, mobile agents are free to choose the next host and when to migrate to the next host. These decisions are transparent to the user and the decisions are taken in the interest of the user.

Mobility: Mobile agents have the ability to move from one host to another in the network.

III. LITERATURE SURVEY

Packet Sniffing Tools: developers debugging communication protocol implementations, or anyone trying to learn how their networks work. Because attackers use sniffers for network reconnaissance and to intercept transmitted credentials and data, learning about the capabilities and limitations of packet sniffers is an important facet of understanding the security risks.

Tcpdump is the most widely used UNIX/Linux tool to record network traffic. It captures packets based on a wide range user-specified criteria, and can save the traffic in different formats. Tcpdump is commonly included in most Linux distributions.

Wireshark is the most widely used graphical application for network monitoring and analysis. It is open-source and runs on most popular computing platforms, including UNIX, Linux, and Windows..

IV. SYSTEM ARCHITECTURE

Here The architecture consists of the following components:

- An intrusion detection processor,
- Network Administrator,
- Mobile Agent Platform,
- Mobile Agent for detection.

Mobile Agent Platform is basically providing all the services like creation, interpretation, execution, transfer and termination for the mobile agents. This platform is responsible for accepting instructions given by network administrator, sending mobile agents to other nodes etc.

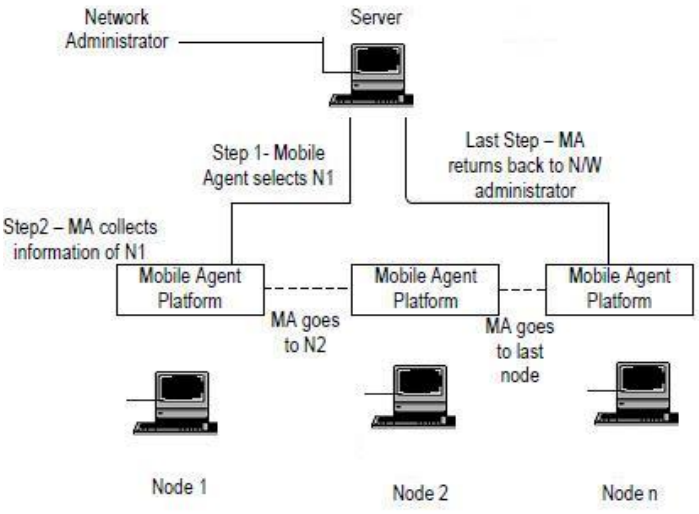

Fig.1 Architectural View

Mobile Agents are the main component of this architecture. They are specially designed to perform network analysis task. Whenever a mobile agent starts execution on a specified node, it monitors all the incoming and outgoing network traffic for that node. If it finds any abnormal incoming traffic (in the case of sniffer) or any other malicious activity, it immediately sends an alarm message to the network administrator for necessary action

V. DESIGN

The following steps can be used for the detection of sniffer in the network. For the mobile agent implementation, we use aglets2.5-alpha which is a java based tool and freely available on the internet. This part we present the architecture of our distributed IDS.
The architecture consists of the following components: (1) an intrusion detection processor, (2) a mobile agent platform, and (3) network administrator, (4) mobile agent for detection.

A. Data Flow Capture

As the main network node monitor, the Data Flow Capture network traffic which incoming monitors, it captures dump of data and sends to IDA for detecting intrusion, then, Data Analysis component to analyze and self-learning.

B. Intrusion Detection Agent (IDA)

IDA is the most important component of the system. It is responsible for monitoring network segments (subnets), and acts as a central intrusion detection agent and data processing unit. The unit is placed on a node that entry into intranet to monitor network traffic for all devices on the segment. And it is setup to send alert in time, so that, checking the errant packets using rule sets when it enter into the segment. Its main capabilities is detecting intrusion and judging whether the behaviour is abnormal, if it is abnormal, alerting to Administrator or make some decisions.

C. Sniffer Detection Techniques

1. Network administrator installs and configures Mobile Agent Platform on all the computers connected in the Local Area Network (LAN)
2. Now whenever the whole system starts, the network administrator activates some specially designed mobile agents.
3. Now these mobile agents travel in the network and select any random node for execution.
4. Mobile Agent collects all the information about network activities including network traffic for that node.
5. As we know if any node runs a Sniffer, then it collects all the packets moving in the network. So mobile agent sends an alarm message to the network administrator if it finds that the incoming network traffic is greater than a pre specified value.
6. After receiving this alarm message, network administrator can take necessary action.
7. If everything is normal then the mobile agent moves to another node and repeats the steps 4 and step 5.

So this whole process can detect the sniffer present in the network. For the mobile agent implementation, we use aglets2.5-alpha which is a java based tool and freely available on the internet

VI. CONCLUSION

Distributed computing involving several computers in a network can be achieved using message passing or remote procedure calls (RPC). The recently developed mobile agent technology adds a new dimension to distributed computing. Experts suggest that mobile agents will be used in many Internet applications in the years to come.

However there still exist many technical hurdles that need to be tackled, the most important of them being security. Only when security issues are properly addressed, will the mobile agent technology be widely accepted. However if intruder makes some changes in our mobile agent platform or mobile agent, then it may fail the whole process.

So in future, some more security measures should be taken for the guaranteed security. Mobile agent selects any node randomly and investigates that node, if it finds excessive incoming traffic on the network interface card then report to network administrator. So the sniffer can be detected.

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Optimizing Back-Propagation using PSO_Hill and PSO_A*

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Abstract- Back propagation algorithm (BPA) have the complexity, local minima problem so we are using Particle Swarm optimization (PSO) algorithms to reduce and optimize BPA. In this paper, two variants of Particle Swarm Optimization (PSO) PSO_Hill and PSO_A* is used as optimization algorithm. PSO_Hill and PSO_A* algorithms are analyzed and evaluated on the basis of their advantages, applied to feed forward neural network(FNN) for back propagation algorithm(BPA) which is a gredient desent technique. where BPA is used for non_linear problems. These non_linear problems are improved by a PSO_Hill and PSO_A* algorithms.

Index Terms- BPA, PSO_hill, PSO_A*, ANN

I. INTRODUCTION

Optimization is an active area of research as many algorithms that are introduced earlier are complex and not optimize so better optimization algorithms are needed. The objective of optimization algorithm is to find a solution to satisfying a set of constraints such that objective function is maximized or minimized.

Particle swarm optimization (PSO) is an alternative population-based evolutionary computation technique. It has been shown to be capable of optimizing hard mathematical problems in continuous or binary space. The particle swarm optimization (PSO) algorithm is based on the evolutionary computation technique [11-13]. PSO optimizes an objective function by conducting population-based search. The population consists of potential solutions, called particles, similar to birds in a flock. The particles are randomly initialized and then freely fly across the multi-dimensional search space. While flying, every particle updates its velocity and position based on its own best experience and that of the entire population. The updating policy will cause the particle swarm to move toward a region with a higher object value. Eventually, all the particles will gather around the point with the highest object value.

PSO processes the search scheme using populations of particles where each particle is equivalent to a candidate solution of a problem [9]. The particle moves according to an adjusted velocity, which is based on that particle's experience and the experience of its companions. For the D-dimensional function f(.), the i^{th} particle for the j^{th} iteration can be represented as

$X_j = X_1, X_2 \ldots\ldots\ldots\ldots X_n$

$W_j = W_1, W_2 \ldots\ldots\ldots W_n$

Assume that the best local position of the j^{th} particle at then nth iteration is represented as

$L\_best_j = L\_best_1, L\_best_2 \ldots\ldots\ldots\ldots L\_best_n$

The best position amongst all the particles, G_best, from the first iteration to the j^{th} iteration, where best is defined by some function of the swarm, is

$G\_best_j = G\_best_1, G\_best2 \ldots\ldots\ldots G\_best_n$

The original particle swarm optimization algorithm can be expressed as follows:

$P_j = X_j * W_j$

$V_j^t = f(p_j)$ means $V_j^t = 1/(1+\exp(-P_j))$

$V_j^{t+1} = V_j^t + r1 * c1(L\_best_j - X_j) + r2 * c2(G\_best_j - X_j)$

$X_j^{t+1} = X_j^t + V_i^{t+1}$

r1and r_2 are random variables acts as learning signals such that $0 \le r_1, r_2 \le 1$.

where W_j is the inertia weight at the j^{th} iteration. The weighting factors, C_1 and C_2 are used as constants. In this paper, the optimization and analysis of back propagation algorithm is done by particle swarm optimization, and two variant of particle swarm optimization PSO_Hill and PSO_A*and their algorithm, architecture are proposed.

This paper is organized as follows. In Section II, we explained original swarm techniques that are PSO and ACO. In Section III, we explained original Back-propagation network with algorithms. In section IV, we explained Related work in PSO. In Section V, we explained proposed work (PSO_Hill and PSO_A* variants with diagram and algorithm). In section VI, we explained Conclusion of paper.

II. SWARM TECHNIQUES

Swarm intelligence (SI) is the collective behavior of decentralized, self-organized systems, natural or artificial. The concept is employed in work on artificial intelligence. SI systems are typically made up of a population of simple agents interacting locally with one another and with their environment. The inspiration often comes from nature, especially biological systems. The agents follow very simple rules, and although there

is no centralized control structure dictating how individual agents should behave, local, and to a certain degree random, interactions between such agents lead to the emergence of "intelligent" global behavior, unknown to the individual agents. Natural examples of SI include ant colonies, bird flocking, animal herding, bacterial growth, and fish schooling. Two well known approaches of swarm intelligence are:

Ant Colony Optimization (ACO): Ant colony optimization (ACO) is a class of optimization algorithms modeled on the actions of an ant colony [2], [13]. ACO methods are useful in problems that need to find paths to goals. Artificial 'ants'—simulation agents—locate optimal solutions by moving through a parameter space representing all possible solutions. Natural ants lay down pheromones directing each other to resources while exploring their environment. The simulated 'ants' similarly record their positions and the quality of their solutions, so that in later simulation iterations more ants locate better solutions.

Particle Swarm Optimization (PSO): Particle Swarm Optimization (PSO) is a swarm-based intelligence algorithm [6] influenced by the social behavior of animals such as a flock of birds finding a food source or a school of fish protecting themselves from a predator. A particle in PSO is analogous to a bird or fish flying through a search (problem) space. The movement of each particle is co-coordinated by a velocity which has both magnitude and direction. Each particle position at any instance of time is influenced by its best position and the position of the best particle in a problem space. The performance of a particle is measured by a fitness value, which is problem specific. Each particle will have a fitness value, which will be evaluated by a fitness function to be optimized in each generation. Each particle knows its best position L_best and the best position so far among the entire group of particles G_best. The L_best of a particle is the best result (fitness value) so far reached

The particle swarm optimization (PSO) algorithm is based on the evolutionary computation technique. PSO optimizes an objective function by conducting population-based search. The population consists of potential solutions, called particles, similar to birds in a flock. The particles are randomly initialized and then freely fly across the multi-dimensional search space. While flying, every particle updates its velocity and position based on its own best experience and that of the entire population. The updating policy will cause the particle swarm to move toward a region with a higher object value. Eventually, all the particles will gather around the point with the highest object value. PSO attempts to simulate social behavior, which differs from the natural selection schemes of genetic algorithms.

PSO processes the search scheme using populations of particles, which corresponds to the use of individuals in genetic algorithms. Each particle is equivalent to a candidate solution of a problem. The particle moves according to an adjusted velocity, which is based on that particle's experience and the experience of its companions.

A) Conventional Particle Swarm Optimization

The particle swarm optimization algorithm was introduced by Kennedy and Eberhart in 1995 [6], [13]. The algorithm consists of a swarm of particles flying through the search space. Each individual i in the swarm contains parameters for position $\mathbf{x}_i$ and velocity $\mathbf{v}_i$, where $\mathbf{x}_i \in \mathbf{R}^n$, $\mathbf{v}_i \in \mathbf{R}^n$ while n is the dimension of the search space. The position of each particle represents a potential solution to the optimization problem. The dynamics of the swarm are governed by a set of rules that modify the velocity of each particle according to the experience of the particle and by adding a velocity vector to the current position, the position of each particle is modified. As the particles move around the space, different fitness values are given to the particles at different locations according to how the current positions of particles satisfy the objective. At each iteration, each particle keeps track of its local best position, L_best and depending on the social network structure of the swarm, the global best position, G_best.

B) Application of PSO

1. Neural Network Training
2. Telecommunications
3. Data Mining
4. Design and Combinatorial optimization
5. Power systems
6. Signal processing

III. Back-Propagation Algorithm

The back- propagation algorithm is used in layered feed-forward ANNs [14]. This means that the artificial neurons are organized in layers, and send their signals "forward", and then the errors are propagated backwards. The network receives inputs by neurons in the input layer, and the output of the network is given by the neurons on an output layer. There may be one or more intermediate hidden layers. The back-propagation algorithm uses supervised learning, which means that we provide the algorithm with examples of the inputs and outputs we want the network to compute, and then the error (difference between actual and expected results) is calculated. The idea of the back propagation algorithm is to reduce this error, until the ANN learns the training data. The training begins with random weights, and the goal is to adjust them so that the error will be minimal.

The basic back propagation algorithm consists of three steps.

1) The input pattern is presented to the input layer of the network. These inputs are propagated through the network until they reach the output units. This forward pass produces the actual or predicted output pattern.
2) The actual network outputs are subtracted from the desired outputs and an error signal is produced.
3) This error signal is then the basis for the back propagation step, whereby the errors are passed back through the neural network by computing the contribution of each hidden processing unit and deriving the corresponding adjustment needed to produce the correct output. The connection weights are then adjusted and the neural network has just learned from an experience.
4) Learning parameters are used to control the training process of a back propagation network.
5) The learn rate is used to specify whether the neural network is going to make major adjustments after each

learning trial or if it is only going to make minor adjustments.

6) Momentum is used to control possible oscillations in the weights, which could be caused by alternately signed error signals.

IV. RELATED WORK

Pillai , K. G. [1] explains a novel overlapping swarm intelligence algorithm is introduced to train the weights of an artificial neural network. Training a neural network is a difficult task that requires an effective search methodology to compute the weights along the edges of a network. The back propagation algorithm, a gradient based method, is frequently used to train multilayer feed-forward networks. On the other hand, training algorithms based on evolutionary computation have been used to train multilayer feed-forward networks in an attempt to overcome the limitations of gradient based algorithms with mixed results. This paper introduces an overlapping swarm intelligence technique to train multilayer feedforward networks. The results show that OSI method performs either on par with or better than the other methods tested.

M. Conforth and Y. Meng [2] propose a swarm intelligence based reinforcement learning (SWIRL) method to train artificial neural networks (ANN). Basically, two swarm intelligence based algorithms are combined together to train the ANN models. Ant Colony Optimization (ACO) is applied to select ANN topology, while Particle Swarm Optimization (PSO) is applied to adjust ANN connection weights. To evaluate the performance of the SWIRL model, it is applied to double pole problem and robot localization through reinforcement learning. Extensive simulation results successfully demonstrate that SWIRL offers performance that is competitive with modern neuro-evolutionary techniques, as well as its viability for real-world problems.

Marco Dorigo and Mauro Birattari [3] defines Swarm intelligence as the discipline that deals with natural and artificial systems composed of many individuals that coordinate using decentralized control and self-organization. In particular, the discipline focuses on the collective behaviors that result from the local interactions of the individuals with each other and with their environment. Examples of systems studied by swarm intelligence are colonies of ants and termites, schools of fish, flocks of birds, herds of land animals. Some human artifacts also fall into the domain of swarm intelligence, notably some multi-robot systems, and also certain computer programs that are written to tackle optimization and data analysis problems.

Y. Karpat and Tugrul Ozel [4] propose a concept of particle swarm optimization, which is a recently developed evolutionary algorithm, is used to optimize machining parameters in hard turning processes where multiple conflicting objectives are present .The relationships between machining parameters and the performance measures of interest are obtained by using experimental data and swarm intelligent neural network systems (SINNS). The results showed that particle swarm optimization is an effective method for solving multi-objective optimization problems, and an integrated system of neural networks and swarm intelligence can be used in solving complex machining optimization problems.

James kennedy and Russell Eberhart [5] propose a concept for the optimization of nonlinear functions using particle swarm methodology is introduced. The evolution of several paradigms is outlined, and an implementation of one of the paradigms is discussed. Benchmark testing of the paradigm is described, and applications, including nonlinear function optimization and neural network training, are proposed. The relationships between particle swarm optimization and both artificial life and genetic algorithms are described.

V. PROPOSED WORK

5.1) *Analysis of Different Algorithm*: This is done on the basis of analysis of their advantage to introduce proposed variants for optimization. In this we analysis the limitation of BPA and to optimize BPA we use PSO approach.

5.1.1) Disadvantages in BPA:

a) Local Minima
b) Low Speed
c) Higher cost of computation
d) Error Problem
e) Gloal Maxima but less as compared to Local
f) Less accuracy

5.1.2) PSO Advantages:

a) PSO can be applied to scientific research and Engineering.
b) High computational speed
c) Learning achieved from particle own experienced
d) Learning achieved from experience of cooperation between particles

5.2) *Proposed Variants for optimization*:

1) PSO_hill
2) PSO_A*

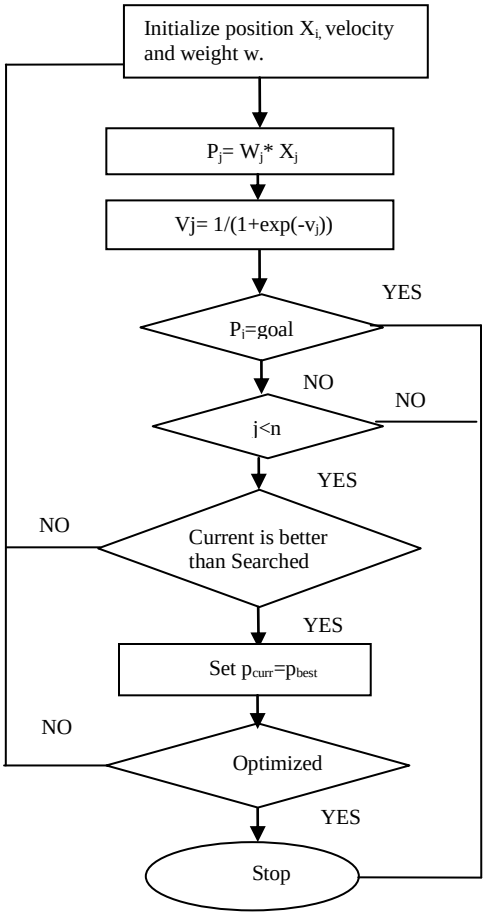
5.2.1) PSO_hill Advantge:

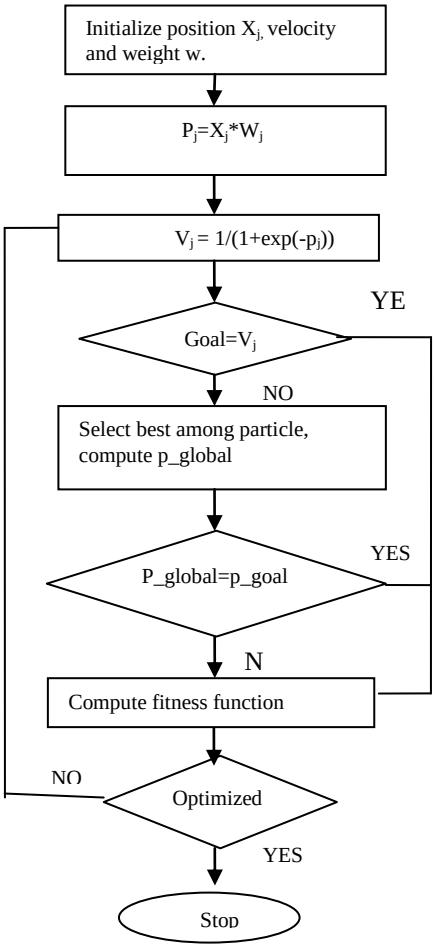
a) High computational speed
b) Strong ability in global search
c) Higher Accuracy
d) Learning achieved from particle own experienced
e) Learning achieved from experience of cooperation between particles.

5.2.2) PSO_Hill variables Notations:

X: Initial particle position.
W: Weight for each particle.
n: Total no. of iterations.
P_{curr}= current position of particles
P_{best} = particle best position

5.2.3) PSO_Hill Diagram:





5.2.4) PSO_Hill Algorithm:

1) Start with initializing particle position X, their velocity V, Weight W, p_loc→0, p_glob→0,n→0
2) If goal=v_j then terminate it; Otherwise
3) $P_j=X_j*W_j$
4) $v_j=1/(1+\exp(-p_j))$
5) If (j<n)
6) {
7) If(Evaluate a new position which is better than current position but not goal)
8) {
9) Curr_position=better_position;
10) Else
11) Keep current position & continue the search to find goal
12) }
13) Else
14) Return to step 2 to continue till it reaches to goal

5.3.1) PSO_A* Advantages:

I. Most heuristic solution
II. Optimized in terms of fitness function

5.3.2) PSO_A* Diagram:

5.2.3) PSO_A* Variables Notations:

Open_list: A list which have nodes that are generated but not expanded.

Closed_list: A list which have nodes that are expanded and its childrens are available to search program.

5.2.4) PSO_A* Algorithm:

1. Start with initial position of particles and place them on open node, p_loc→0,p_glob→0.
2. If (open_list=empty) stop and return as failure
3. Select p_loc particles n from open list that has smallest fitness function
4. {
5. if node n= goal node
6. return success
7. stop
8. }
9. Otherwise
10. Expand the successor particles of node n and as node n is explored so keep it on closed list
11. For each successor h
12. {
13. If h is not in open_list or closed_list
14. {
15. Attach a back pointer to n particle to backtrack it
16. {

17. Compute fitness of particle f*(h)
18. Else
19. Compute lowest or p_glob (g*(h))
20. }
21. Place on open list
22. Return to Step to till the goal

VI. CONCLUSION

In this paper, two variants of the particle swarm optimization scheme is presented. Two PSO Variants PSO_Hill and PSO_A* are proposed with their algorithm, architecture, advantages and disadvantages, which can be used to the optimize the BPA. In next paper, a third strategy is proposed PSO_Hill_A* on the basis of strength of two variants PSO_Hill and PSO_A* algorithm. The particle local best and global best positions help the variants to move towards the solution.

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A Study of Composite Material-“ Lithium-Cadmium-Copper Ferrite”- Curie Temperature

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Abstract- Ceramic magnets became commercially available in the 1950’s (Alnico magnets were the first true magnets to be introduced n the 1930’s).There are currently 27 grades of Ferrite Permanent Magnet available. Ferromagnetic, Paramagnetic, and Antiferromagnetic materials are all made up of intrinsic magnetic moments. Magnetic moments depend on temperature as thermal disorder destroys alignment between dipoles. It is at this temperature point, the Curie temperature where the materials change from Ferromagnetic order to Paramagnetic order or vice versa. Ferrite is ceramic material with magnetic properties, is useful in many types of electronic devices and building materials.. It is the component which gives steel and cast iron their magnetic properties, and is the classic example of a ferromagnetic material. Our observations is that the Curie temperature decreases with addition of Cu^{2+},Cd^{2+} ions tends to decrease the number of A-B interactions. This may be due to A-O-B distance. Due to thermal vibrations, the magnetic A-B interactions in ceramic material are broken at Curie temperature.

Index Terms- Ceramic material, , Ferromagnetic material, Curie temperature, A-B interactions.

I. INTRODUCTION

In physics and materials science, the Curie temperature (Tc), or Curie **point**, is the critical point of a second order phase transition where magnetic moments change their orientations. Ferromagnetic, Paramagnetic, Ferromagnetic and A ntiferromagntic materia have different structures of intrinsic magnetic moments that depend on temperature. It is at a material specific Curie temperature where they change properties.

Ferrite, a ceramic-like material with magnetic properties, is useful in many types of electronic devices and building materials. They are composed of iron oxide and one or more other metals in chemical combination. Ferrites are hard, brittle, generally gray or black, and typically have a crystal structure which has more than one type of site for the cations. Usually the magnetic moments of the metal ions on sites of one type are parallel to each other and anti-parallel to the moments on at least one site of another type. It is the component which gives steel and cast iron their magnetic properties, and is the classic example of a ferromagnetic material.

A large number of applications of ceramic materials depend on their electrical properties. Composite materials are formed from combination of two or more single phase compounds. Physical properties of composites are determined by the properties of their constituent phases and the interaction between them. Composite have some properties which give rise to sum of their constituents. The electrical properties of composite are quantitatively considered as sum properties of their individual electrical and ionic behavior. In industrial electronics Ferrites are play an important role as material.

II. EXPERIMENTAL SET UP

For Curie temperature measurement we require proper arrangement .which is shown in Fig.1.

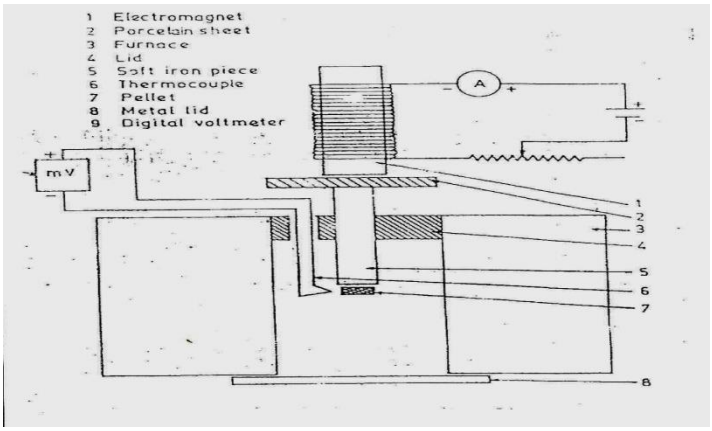


Fig.1 Experimental set up

Set up consist of

- Tabular furnace- For heating purpose.
- Soft iron rod-For holding sample in pellet formation.
- Dimmerstat –For changing voltage required for heating purpose.
- D.C.Supply-For maintaining supply given to pellet.
- Pellet sample

III. EXPERIMENTATION

3.1 Experimental methods

A transition temperature marking a change in the magnetic or ferroelectric properties of a substance, especially the change from ferromagnetism to paramagnetic. Also called Curie temperature.

Curie temperature is one of the important properties of microwave ferrite. There are few experimental methods to determine curie temperature T_C.

There are few methods for Curie temperature measurements as given below-

1. Loria method
2. Permeability method.

In this paper for measurement of Curie temperature we use famous method i.e .Loria method.

3.2 Preparation of Li-Cd – Cu Ferrite-

Ferrite arte homogeneous ceramic materials. Such material expressed by the molecular formula as MFe2O4 , where M is an oxide of divalent metal.

The process leading the formation of ferrite from their oxide component belongs to the class of solid state reaction. This solid state reaction is between two solids resulting in the formation of new solid.

The ferrite preparation methods have been classified in three category

1. Ceramic Method.
2. Pre- cursor Method.
3. Wet chemical Method.

For formation of sample we use ceramic method , and sintering process play an important role in ferrite sample preparation. as shown

Mechanism for Solid – State reaction of composition formation

Carbonates –

i. Li_2CO_3

ii. $CdCO_3$

Lithium ferrite, Cadmium ferrite and Copper ferrite formation by adding Fe_2O_3 in composition .

Li_2CO_3	$CdCO_3$	CuO
↓	↓	↓
decomposition	decomposition	
↓	↓	
$Li_2O + CO_2\uparrow$	$CdO + CO_2\uparrow$	
↓	↓	
$Li_2O + 5Fe_2O_3$	$CdO + Fe_2O_3$	$CuO + Fe_2O_3$
↓	↓	↓
presintering	presintering	presintering
↓	↓	↓
$Li_2O\ Fe_{10}O_{15}$	$Cd\ Fe_2O_4$	$Cu\ Fe_2O_4$
↓ 1/2		
$LiFe_5O_8$		

During Sintering Solid State reaction takes place as

$$LiFe_5O_8 + Cd\ Fe_2O_4 + Cu\ Fe_2O_4$$
↓
$$LiCdCuFe_9O_{16}$$
↓1/2
$$Li_{0.5}Cd_{0.5}Cu_{0.5}Fe_{4.5}O_8$$
↓
$$Li_{0.5}Fe_{0.5}Cd_{0.5}Cu_{0.5}Fe_4O_8$$
↓
$$[(Li_{0.5}Fe_{0.5})(Cd_{0.5}Cu_{0.5})]\ Fe_4O_8$$
↓
$$(Li_{0.5}Fe_{0.5})(Cd\ Cu)_{0.5}\ Fe_4O_8$$
↓1/2
$$(Li_{0.5}Fe_{0.5})_{0.5}(Cd\ Cu)_{0.25}\ Fe_2O_4$$

Put $0.5 = 1\text{-}X \Rightarrow X = 0.5$

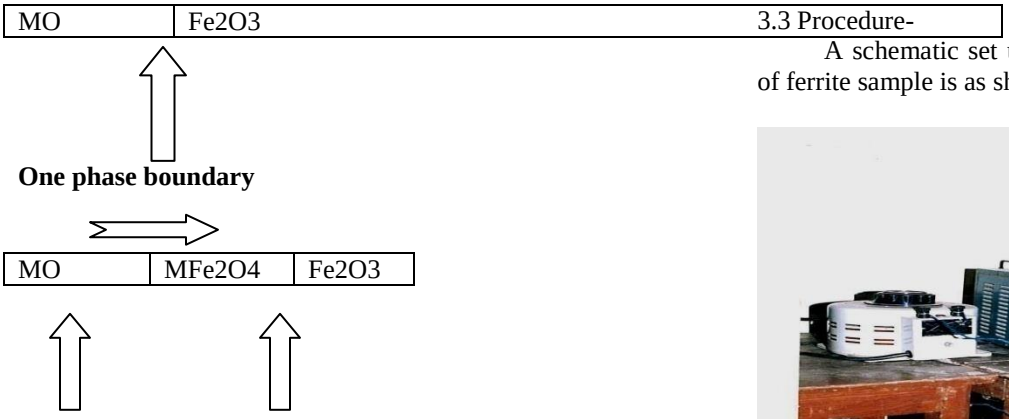
$0.25 = 1/2X = 0.5X$

$\therefore (Li_{0.5}Fe_{0.5})_{1\text{-}X}(Cd\ Cu)_{0.5X}\ Fe_2O_4$

$\therefore (Li_{0.5}Fe_{0.5})_{1.X}(Cd_{0.5X}Cu_{0.5X})\ Fe_2O_4$

Mechanism

Such mechanism has been discussed on the basis of the simple diffusion couple, involving divalent metal and iron oxide. To initiate the mechanism , homogenous mixture of the metal oxide(MO) and ferrite oxide (Fe2O3).is heated at an elevated temperature.



Transfer Mechanism

Such transfer mechanism can be discussed in three mode as Mode1, Mode2, Mode3.

Mode1-

In such mode only cations migrate in the opposite direction and oxygen ions remains stationary.

Mode2-

In such mode diffusion of cations is compensated by an associated flux of anions instead of another cation migration.

Mode3-

In such mode the diffusion of iron through ferrite layer in the reduced state Fe 2+

Sample preparation

The general formula of ferrite system is $(Li_{.5}Fe_{.5})1-x(Cd_{.5x}Cu_{.5x})Fe_2O_4$where x=0,0.1,0.2,0.3,0.4,0.5,0.6,0.8 and 1. The AR grade of oxides of Li_2O_3,$CdCo_3$,CuO,Fe_2O_3were used for the preparation of the sample.

By putting x=0,0.1,0.2,0.3,0.4,0.5,0.6,0.8 and 1 in the general formula. We have calculated weights of Li_2O_3,$CdCo_3$,CuO,Fe_2O_3.Then according to calculated weights carbonates and oxides were weighted on a sensitive mono- pan balance.

Table 1

Sample Notation	Composition
M0	X=0
M1	X=0.1
M2	X=0.2
M3	X=0.3
M4	X=0.4
M5	X=0.5
M6	X=0.6
M7	X=0.8
M8	X=1

3.3 Procedure-

A schematic set up of the curie temperature measurement of ferrite sample is as shown.

Fig.2 Photograph of set up

The electromagnet and soft iron rod were well insulated from each other using porcelain sheet.

The soft iron rod kept at the centre of the furnace got magnetized due to induced magnetation.

One end of chromel- alumel junction was kept in the vicinity of the sample in order to measure temperature of furnace and as result Curie temperature of a sample. A digital multivoltmeter was used for the temperature measurement equivalent to emf.

A sample of which T_C was to be measured , introduced into the furnace with the aid of metallic lid . the current in the heating coil was raised gradually to heat the sample upto its Curie point. When temperature reached at which ferrite looses its magnetation and drops under gravity,this temperature was measured by digital multivoltmeter.

IV. RESULT AND DISCUSSIONS

The data of Curie temperature obtained by Loria technique is presented in Table 2.

Table 2

Sample Notation	Composition	Curie temperature ^oC
M0	X=0	645 ^oC
M1	X=0.1	540 ^oC
M2	X=0.2	470 ^oC
M3	X=0.3	420 ^oC
M4	X=0.4	365 ^oC
M5	X=0.5	370 ^oC
M6	X=0.6	330 ^oC
M7	X=0.8	345 ^oC
M8	X=1	280 ^oC

We observed that the curie temperature decreases with addition of Cu^{2+},Cd^{2+} ions which tends to decrease the number of A-B interactions. This may be due to A-O-B distance.

Due to thermal vibrations, the magnetic A-B interactions in spinet ferrites are broken at Curie temperature.

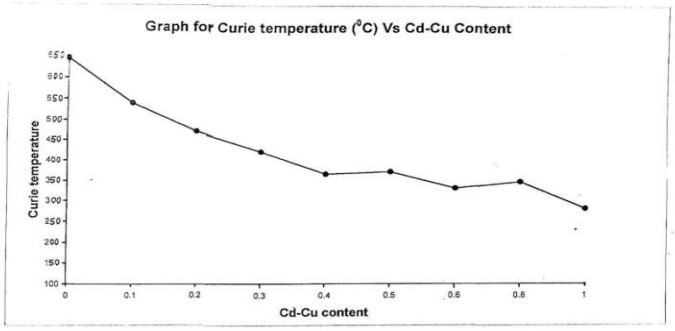


Fig.3 Variation of Curie temperature

Graphical variations of T_C with increasing Cd-Cu content is shown in Fig.3.This graph shows that all compositions are magnetic, which is verified also by Hysteresis loops. It is also observed that T_C decreases with deceases of Li concentration, which is attributed to decrease of A-B interaction force.

V. CONCLUSIONS

1. T_C value of Li ferrite agrees well with reported values [1].
2. The magnetic A-B interactions in spinet ferrites are broken at Curie temperature.
3. The addition of Cd^{2+} ions tends to decrease number of A-B interactions.

NOMENCLATURE

Cu^{2+}- Copper divalent metal ions.
Cd^{2+} -Cadmium divalent metal ions.
T_c - Curie temperature.
M- Oxide of divalent metal.
A-B interactions.
M0- Sample notation for X=0

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Cotton Dust Level in Textile Industries and Its Impact on Human

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***Abstract*-** In India, the textile industry contributes substantially to the foreign exchange earned by the country. The textile industry is providing employment opportunities to numerous people in the country. The emphasis on awareness about the environmental concern such as air, water and noise pollution during the processing from fibre to fabric is essential in the present circumstances. Information regarding cotton dust exposure impacts and the control strategies is lacking among textile employers and its management. The main aim of this paper is to provide, dust level in the textile industry and the available air quality standards are discussed to facilitate textile mill employers and management to establish cotton dust control strategies to save their workers from its harmful impacts. The study has been carried out in textile industries located in Tirupur. This study is based on the analysis and monitoring of air pollutants using respiratory dust sampler in work place.

***Index Terms*-** Asthma, Byssinosis, Chronic bronchitis, cotton dust, Pulmonary Diseases, Emphysema, respiratory dust sampler, Particulate matter

I. INTRODUCTION

Textile industry is the second largest industry in the world next to agriculture. In India, the textile industry contributes substantially to the foreign exchange earned by the country. The textile industry consists of a number of units engaged in spinning, weaving, dyeing, printing, finishing and a number of other processes that are required to convert fibre into a finished fabric or garment. The textile industry is providing employment to numerous people in the country. The emphasis on awareness about the environmental concern such as air, water and noise pollution during the processing from fibre to fabric is essential in the present circumstances. There were 1818 mills (non-SSI) in the country as on January 31, 2007 with a capacity of 35.37 million spindles, 4, 48,000 rotors and 69,000 looms. Information regarding cotton dust exposure impacts on workers and its control strategies is missing among textile employers, management and employees. [1]

Cotton dust is defined as dust present in the air during the handling or processing of cotton, which may contain a mixture of many substances including ground up plant matter, fiber, bacteria, fungi, soil, pesticides, non-cotton plant matter and other contaminants which may have accumulated with the cotton during the growing, harvesting and subsequent processing or storage periods.[1]

Any dust present during the handling and processing of cotton through the weaving or knitting of fabrics, and dust present in other operations or manufacturing processes using raw or waste cotton fibers and cotton fiber byproducts from textile mills are considered cotton dust within this definition. [1]

II. MATERIALS AND METHODS

The analysis was carried out in the textile industry in tirupur. The analysis was done in various sections in the mill such as cutting and ironing, knitting and stitching. The experiment was carried out by using an air sampler. The method of measurement is done by gravimetric method. With the analysis from each section, PM2.5 and PM10 were collected. And the duration of the collection of samples is with the time period of 8 hours. The sampling collected is done with process and without process except for knitting sections.

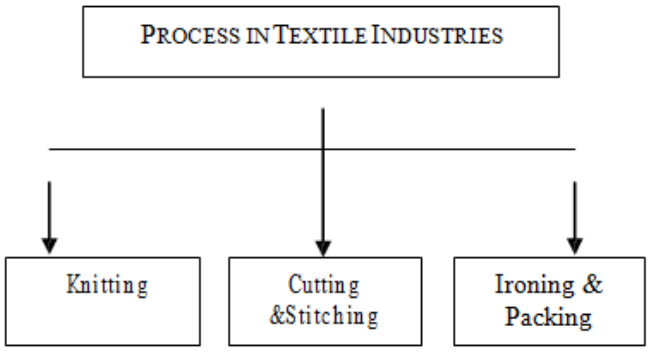


TABLE I
STANDARD METHODS USED FOR TARGET POLLUTANTS

Particulars	$PM_{2.5}$	PM_{10}
Equipments	Respirable dust sampler	Respirable dust sampler
Analytical method	Gravimetric	Gravimetric
Sampling period	8/24 hourly	8/24 hourly
Measuring principal	Gravimetric	Gravimetric
Sampling frequency	Four days continuous	Four days continuous

Sampling

The numbers of samples collected at each zone for a period of 8 hours and the meteorological parameters such as wind speed, wind direction, temperature and relative humidity are also noted.

METHODOLOGY

Method for sampling particulate matter

Total suspended particulate matter (TSPM) comprises of particles above 10 µ (non Respirable particulate matter) and particles below 10 µ (Respirable particulate matter).TSPM present in ambient air is measured by High volume sampling method by using a Respirable Dust Sampler with a cyclone attachment over a period of 8 hours by sucking a known quantity of air through glass fibre filters. The mass of concentration of SPM is computed by measuring the weight of collected matter in known volume of air sampled. [6]

RESPIRABLE DUST SAMPLER

III. CLASSIFICATION OF COTTON DUST

A.According to the type of dust

1. Inhalable Dust

It is a term used to describe dust that is hazardous when deposited anywhere in the respiratory tree including the mouth and nose. [2]

2. Thoracic Dust

It is defined as those materials that are hazardous when deposited anywhere within the lung airways and the gas exchange region. [2]

3. Respirable Dust

Respirable dust is defined as that fraction of the dust reaching alveolar region of the lungs. [2]

According to the size of the particle (Table I1)

TABLE II

Types	Size of particle (µm)
Trash	Above 500
Dust	50 -500
Micro dust	15 -50
Breathable	Below 15

IV. SAMPLING MONITORING

Sampling of the workplace must be done at least every six months to determine the amount of cotton dust in the environment. Measurements must be representative of all employees in the workplace. Sampling will be done in all work areas and on each shift.

Sampling involves measuring with a standard cotton dust monitor called a vertical elutriator (VE) or by an approved similar device. Air is drawn into the vertical elutriator at a specified speed, and particles of 15 microns or smaller is collected on a filter. The particles collected are measured to determine the amount of Respirable dust (dust that can get into the lungs) present is in the work area. It is important to realize that other "dusts," such as starch or oil mist are also collected on the filter and may contribute to the cotton dust levels.

Sampling is done for a period equal to at least three-quarters of the shift (for example, six hours of an eight-hour shift). While sampling is being done, other information is collected that may pertain to the generation of cotton dust. The percent of cotton fiber in the mix; the grade of the cotton and where it was grown; types of yarn being run; and the number and types of machines operating in each area may all affect the amount of cotton dust in the workplace.

Employees will be notified of the findings of cotton dust air sampling once the results have been received. When the results show that an employee is exposed to dust greater than the permissible exposure limit, control measures must be taken to reduce exposure to an acceptable level. Detailed exposure measurement records must be made, and they must be retained for 20 years. [4]

V. PROPERTIES

A.Analysed particulate matter characteristics

1) PM10

Particulate matter is the term for solid or liquid particles found in the air which is of 10 µ size. Some particles are large or dark enough to be seen as soot or smoke. Others are so small they can be detected only with an electron microscope. Because particles originate from a variety of sections their chemical and physical compositions vary widely. Particulate matter can be directly emitted or can be formed in the atmosphere when gaseous pollutants such as SO2 and NOx react to form fine particles.

2)PM2.5

The most important characteristic of particulate matter (PM) is the particle size. Particulate matter that is of 2.5 µ size is denoted as above. This property has the greatest impact on the behavior of particulate matter in control equipment, the atmosphere, and the respiratory tract. Particles of importance in air pollution control span a broad size range from extremely small (0.01 micrometer) to more than 1,000 micrometers. As a frame of reference, a human hair has a diameter of approximately 50 micrometers.

A particle size is usually expressed in terms of its aerodynamic diameter instead of its actual or physical diameter. The chemical composition of the particulate matter is also important. Absorption and heterogeneous nucleation of vapour phase pollutants onto existing particles can create toxic particulate matter. Other characteristics besides size and chemical composition should be considered when selecting an appropriate particulate control device for a gas stream. Other important characteristics of particulate matter in gas streams include stickiness, resistivity and explosiveness.

VI. HEALTH HAZARDS ASSOCIATED WITH COTTON DUST EXPOSURE

Workers exposed to cotton dust laden environment generally become patients of byssinosis.

A.Byssinosis

Byssinosis is a term taken from a Greek word meaning white thread. It is a breathing disorder that occurs in some individuals with exposure to raw cotton dust. Characteristically, workers exhibit shortness of breath and/or the feeling of chest tightness when returning to work after being in the mill for a day or more. There may be increased cough and phlegm production. [3]

Change in the levels of ESR, LDH3 and Histamine may be used as indicators to assess pulmonary dysfunction in the workers those are exposed to cotton dust. It was suggested that the low hemoglobin and poor immunity against diseases may also predispose the outcome pulmonary dysfunction at an earlier stage. Cotton dust extract induces the release of histamine from samples of human lung tissue in vitro. Therefore it is believed that histamine release is responsible for the major symptoms of byssinosis. [3]

Dr. Richard Schilling, a British physician developed a system of grading workers based on their breathing complaints on the first workday of the week. Schillings classification grades byssinosis according to how far it has progressed. [3]

A.1.Schillings classifications are as follows.

- Grade 0 = No complaints of breathing problems.
- Grade 1/2 = Chest tightness and/or shortness of breath sometimes on the first day of the workweek.
- Grade 1 = Chest tightness and/or shortness of breath always on the first day of the workweek.[3]
- Grade 2 = Chest tightness and/or shortness of breath on the first workday and on other days of the workweek.[3]
- Grade 3 = Chest tightness and/or shortness of breath on the first workday and other days as well as impairment of lung function.[3]

TABLE III
Permissible Exposure Limits (PEL) for Cotton Dust for Different Work Areas[3]

TABLE III [5]

DEPARTMENT	PEL (micrograms per cubic meter)
Opening	200
Packing	200
Carding	200
Combing	200
Roving	200
Spinning	200
Winding	200
Warping	200
Slashing	750
Weaving & Knitting	750
Waste house	500

B. Pulmonary Diseases

There is a group of lung diseases called *chronic obstructive pulmonary diseases*. The diseases in the group are major causes of illness and disability among workers.

The most common types of chronic obstructive pulmonary disease are:

_ Chronic bronchitis
_ Asthma
_ Emphysema [3]

1. Chronic bronchitis

Is a disorder characterized by a cough and sputum lasting for three or more months of the year and recurring year after year. [3]

2. Asthma

It is thought to be an allergic type of response that causes airways to swell and become narrow. There is increased mucous causing a wheezy, "whistly" sound to breathing. Usually both chronic bronchitis and asthma improve when the person is removed from the irritation causing this response. [3]

3. Emphysema

It is the destruction of the delicate walls between the tiny air sacks in the lungs. As the walls are destroyed, the air sacks enlarge and the lungs have less ability to supply oxygen to the bloodstream. In emphysema, there is no way to repair the destroyed air sacs. [3]

VII. HEALTH MONITORING

Employees must be monitored if they are to work in an environment containing cotton dust. The environment containing the cotton dust must also be monitored.

A.Medical Monitoring

In any workplace where cotton dust is present there must be a medical surveillance program for all employees exposed to cotton dust. Examinations must be done by or under the direction of a licensed physician. People administering the pul- monary function (breathing) tests must have attended a course approved by the National Institute for Occupational Safety and Health (NIOSH).

Medical examinations are to be provided to prospective employees prior to their initial assignment. As a minimum, the examinations should include:

•A medical history to identify any existing health problems or diseases that may affect breathing.

•A standardized respiratory questionnaire inquiring about such concerns as cough, chest tightness and smoking history.

•A pulmonary function (breathing) test including the forced vital capacity (FVC), the amount of air one can force out after taking a deep breath, and forced expiratory volume in 1 second (FEV1), the amount of air forced out during the first second of expiration.

Test results are compared to a set of predicted tables based on a person's age, height, sex and race. Generally, tests are considered to be within the normal range if they are 80 percent or greater of the predicted value. The initial determinations should be made prior to entering the workplace on the first day worked and after having no cotton dust exposure for at least 35 hours. The pulmonary function tests will be repeated during the shift, at least four hours, but not longer than 10 hours after the first test. These tests are then compared for changes. If there is a decrease of 5 percent or greater on the second after- exposure test, it may indicate a reaction to cotton dust. Each employee will be assigned a byssinosis grade based on his or her response to the respiratory questionnaire.

Follow-up examinations are required annually for all employees exposed to cotton dust. The examinations include an update of the medical history and standardized questionnaire and a repeat of the pulmonary function test preformed both before and after the exposure to cotton dust.

Examination is required every six months for employees who are below the expected normal value when compared to predicted values, or for employees who show a decrease in pulmonary function on the after-exposure test. If the physician feels significant changes have occurred from year to year, or if the worker has complaints about breathing, six-month test- ing will also be done.

Employees who are below 60 percent of the predicted value on their breathing test will be sent to a physician for an evaluation. Employees will be furnished written information on the results of their examination.

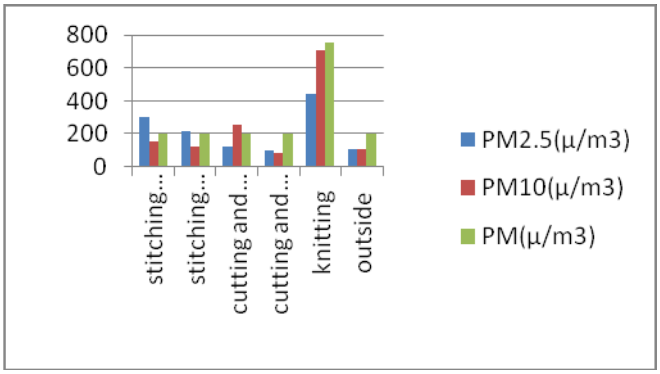
OSHA has not published specific criteria for hiring people to work in a cotton dust environment except to specify an ini- tial examination be conducted. Experience has shown that people who have a history of asthma or other respiratory diseases do not fare well in a cotton dust environment. Many medical specialists recommend that individuals who have a breathing test result below 80 percent of the predicted normal value should not be hired to work in a cotton dust environment. [4]

VIII. .RESULTS AND DISCUSSIONS

TABLE IV

DEPARTMENT	$PM_{2.5}$ (μ/m^3)	PM_{10} (μ/m^3)
Stitching with process	304.318	152.803
Stitching without process	215.303	120.606
Cutting and Ironing with process	126.287	255.075
Cutting and Ironing without process	101.666	86.515
Knitting	446.383	709.621
Outside	109.924	109.240

From this project we have observed that, the workers are exceeding the working limit (8hr to 11hr) and though the working hour is more than the limit for each worker, the value of the pollutant is not accurately exceeding the permissible exposure limit but almost closer whose effects, in case of a continuous exposure might lead to severe negative impacts.



The following recommendations are suggested for controlling the occupational lung diseases caused by cotton dust:

- Periodic health surveillance to be made essential.
- Proper treatment should be given to the affected worker.
- Effective dust control measure to be adopted
- Awareness to be given for the administrators and the Workers.
- Usage of personnel protective equipment should be Strongly advised.
- The management should strictly provide and follow Control technologies and protective strategies.
- The worker should be aware of the effects of cotton dust And exceeding the working limit.[5]

Dust Controls

Often employers can reduce dust levels by adjusting dust control equipment, such as ventilation systems, and by cleaning and repairing the equipment regularly.

An employer's dust control program must include, at a minimum, the following:

(1) Cleaning floors with a vacuum or any other method that cuts down the spreading of dust;

(2) Disposing of dust in such a way that as little dust scatters as possible;

(3) Using mechanical methods to stack, dump or otherwise handle cotton or cotton waste, when possible;

(4) Checking, cleaning, and repairing dust control equipment and ventilation systems.

Employees involved in cleaning must wear respirators. Compressed air may not be used to clean clothing and floors and may only be used to clean equipment if no other methods are possible. If these measures fail to reduce the cotton dust levels below the OSHA limits, employers must try additional engineering controls and work practices.

IX. DISCUSSION

The Project was done with an intention to study the hazardous effects over people working in the textile area. The project was carried out in a location where the above mentioned textile processes happen to be the mere source of income of over 50% of people in the surrounding area.

Hence the results suggest that a continuous exposure to these processes of the textile sector may result in tedious health hazards.

The results also suggest that Particulate matter easily reacts with Aerosols causing the atmosphere more acidic in nature. The effects may be severe some may lead to cardiac failure.

From the above graphical representation we observed that knitting process produces the maximum dust level compared to other process, so we suggest that the preventive measures in the knitting session particularly should be more effective than other sessions.

X. ACKNOWLEDGMENT

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Security for Windows Registry Using Carving

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Abstract- This paper describes the Windows registry which stores a lot of system information and can also be used as forensic evidence. Many researchers have worked to know how the information is stored in the registry, but carving the registry files from the raw disk is not described. Till now the researchers performed the researches on how the registry files are carved when each block is not fragmented using the internal structure of registry file. It is also based on the internal structure of registry files, but in this paper, the fragmentation is performed on the multiple HBIN blocks instead of two HBIN blocks. It also recovers the Windows registry files using back up when the file system is crashed or damaged. The carving technique is used which is more effective and accurate for Windows registry files.

Index Terms- computer forensic; registry; carving

I. INTRODUCTION

The Windows operating system is one of the most popular operating systems. It contains a lot of data in it and hence it is mostly attacked. As we know that the Windows registry is a database that stores configuration settings for System, users, applications, hardware and many other system parameters and as it has a lot of information contained in it, the registry can be an excellent source for potential evidential data in aiding forensic examiners on other aspects of forensic analysis. Although the Windows registry appears as a single hierarchy in registry editor such as regedit.exe, it is actually made up of a number of different binary files called hives on disk.

The Windows registry file has a lot of data which can be useful for the forensic examination hence forensic examiner shows a more interest in it. An important traditional field of computer forensic is data carving. In the data carving, the evidential files are extracted from raw images without using any file system metadata, such as the allocation information. In order to recover the data from registry files, carving is very important for computer forensic when the file system is crashed or damaged.

This paper uses the data carving technique to recover the Windows hive files. Section II briefly describes previous work on file carving and registry forensic. The internal structure of the hive will be illuminated in section III. In section IV, the process for carving the hive files from the raw disk is presented and some validation is offered to enhance. Section V describes some experimental results. In section VI, we outline future work, and in section VII, we conclude this paper.

II. RELATED WORK

Win Hex [1] supports manual extension to other files. It doesn't support the hive file carving in its file type definition database. It has a magic number which occurs in it and can be added to the database. This magic number is referred as "regf" and its length is specified by the user as the maximum. It is difficult to understand for the new user how the fragmentation is performed. Hence it is not more efficient.[2]. Morgan has provided an algorithm to recover deleted data from the Windows registry [3]. Dolan-Gavitt introduces a way on how to forensic analysis of the Windows registry in memory while the attack modifies the cached version of the registry without altering the on-disk version [4].Mark Russinovich described the hive internal structure for the first time [5]. Jolanta Thomassen has provided extensive information of how registry information is organized into data structures on disk [6]. These are very important to the hive file carving. It helps a lot to carry out this experiment.

III. REGISTRY INTERNAL STRUCTURE

The registry internal structure is important, in order to understand the carving as it is based on the internal structure of hive. The internal structure of Windows registry hives is shown in Fig. 1. Registry hive files start with a header, or base block, and continue with a series of hive bin blocks. The base block has a stable size of 4096 bytes and contains a magic number of "regf". This base block gives the information such as file name, its path, time stamp and other systems files. It can store the maximum length of 32 characters file name. It may be shortened by the Windows operating system.

As we know that the each base blocks has a series of hive bin (HBIN).HBIN consists of 4096 bytes, but sometimes may be any larger multiple of that size. These HBINs are linked together through the HBINs header. The size of a HBIN header is always 32 bytes, and the HBIN header contains the magic number "hbin". Each HBIN references the beginning of the next HBIN in addition to indicate its distance from the first HBIN. Information in the Registry is arranged in a tree-like system akin to folders and files.

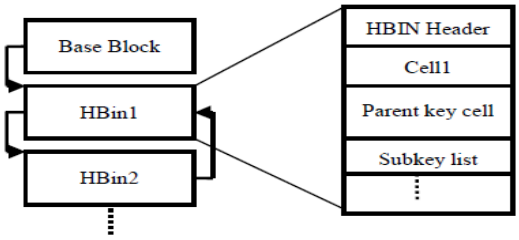

Figure 1. Registry internal structure

In the registry, the containers for information are called “keys”. Keys can have sub keys just as folders can have sub folders. Each HBIN has a series of variable-length cells and each cell’s total length is a multiple of 8 bytes, which is specified in the first four bytes of the cell. There are some types of cells in registry. They are: key cell, sub key lists, value lists, value cell, and security cell and class name. The key cell is the most important structure which contains a number of offset fields to other cells. It acts as a connecting link between two HBINs. This key cell is divided into some parts which contains the size of the cell, name, timestamp of the key, and offsets to parent key, sub key list, value list, security descriptor and class name. If a key does not have remaining cells, offsets are set to 0xffffffff. Key cells use sub key-lists to reference a set of other key cells. The name of data that is contained in a key is called a "value". This is something analogous to a file name. The actual data can have several formats and may be a string, a number, or a series of numbers. This cells, keys and values are arranged in tree structure as shown in fig 2.

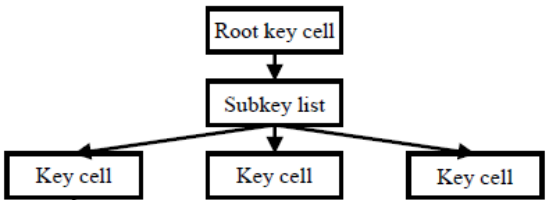

Figure 2. Registry tree

IV. CARVING THE HIVE FILES

The Registry can be seen as one unified “file system” which consist of number of blocks. And this block consist of number of HBINs, again in which a number of keys and values are present. In registry, there are 5 HKEYs, its abbreviation is “handle to a key”. They are HKEY_CLASSES_ROOT (HKCR),
HKEY_CURRENT_USER (HKCU),
HKEY_LOCAL_MACHINE (HKLM),
HKEY_USERS (HKU),
HKEY_CURRENT_CONFIG (HCU).

Only two of these are actually “real”. They are HKU and HKLM. The other three are sub keys of these. Each of these five hives is composed of keys, which contain values and sub keys. Values are the names of certain items within a key, which uniquely identify specific values pertaining to the operating system, or to applications that depend upon that value. Depending on the size of the partition, the hard disk cluster can vary in size, It means that when the hive files are fragmented, the internal structure of the base blocks are not destroyed, but sometimes the HBINs’ may be fragmented when their lengths are larger than the size of cluster. In typical Windows, all hives file are illuminated in following registry path: HKEY_LOCAL_MACHINE\SYSTEM\CurrentControlSet\Control\hivelist. 12 hives file can be found. The six user hives locate in user home directory, they are: the ntuser and UsrClass hives for the currently logged on system user, the Local Service user, and the Network Service user; and the rest store in %Systemroot%\system32\config path, they are: default; SAM; system; SECURITY; software; and the HARDWARE hive which is generated at boot and provides information on the hardware detected in the system.. Each system user contains two hives: ntuser and UsrClass. There are some hive files with the path of %Systemroot%\repair. These hive files were created when the Windows system was installed in the first time and their base blocks have neither timestamp nor file name. The structures of the base blocks and HBINs files are intact. The unallocated space which is not used on disk sometimes includes several registry data. These unavailable data are produced when the Windows system needs delete a hive; it drops the original space, but leaves the registry data. This situation always occurs when we delete a system user, the ntuser and UsrClass are deleted. There is another probability that these data are created by last Windows system and not covered by the current system. The structures of the base blocks and HBINs in above files are intact. From the above explanation, carving method for two blocks had been proved successfully. Using this carving method, our aim is to prove that the carving can be performed successfully on multiple blocks. The carving process shows how the blocks are fragmented.

Initially select the base blocks. Based on the magic number "regf" and the file name is included the hives file name; the cluster should be the base block of a hive file.

Secondly, check whether the hive file is fragmented. If it is fragmented then it jumps to the HBIN of that base blocks. The HBINs have the magic number "hbin", the offset to the first HBIN and the length. According to the hive internal structure, all HBINs follow after the base block, and the difference of the offset to the first HBIN between two conjoint HBIN blocks is the length of the previous HBIN. If this rule is broken, the fragment must happen. There is another fragmented case. If a HBIN’s next cluster not starts with the magic number "hbin" and the current found HBINs’ total size is not greater than the base block’s total size of all HBINs. Otherwise, all the HBIN blocks can be found and the found HBIN blocks length is equal to the base block’s total size of all HBINs.it continues for next base blocks if present.

Finally, find all the fragments, and reassemble them together. If the hive file is fragmented, the next fragment must start with the magic number "hbin", and its offset to the first HBIN can be made sure because the previous HBIN’ offset to the first HBIN and its size are known. If a found fragment already belong to another hive, we ignore this fragment and otherwise, try to next one while accords with our condition Try to find all the fragments until total size of all HBINs is the same as the base block’s total size of all HBINs and integrate those HBIN blocks together by the offsets to the first HBIN.

During this process it is important to backup the files of registry. There may be a possibility that the data may be lost .hence in order to recover the data, it is an essential factor. There is a high-level validation is mentioned to verify the all found fragments are the real ones. For each new carved hive, we try to build the registry tree from the root key, if the build fails, there must be some fragments don’t fit the hive. The available bases block should include the magic number, file name and timestamp. The timestamp has very important role in the carving. If some base blocks have the same file name, the timestamp can be used to select the latest base block.

There is another way to validate base blocks is checksum that can be stored at offset 0X1FC. In order to carve the hive of multi-user, first, we should carve SAM hive accurately and acquire the key with path \SAM\Domains\Account\Users\Names. The available system users' names are this key's sub keys.

V. EXPERIMENTAL RESULTS

In our experiment of carving the multiple registry files, first we selected the blocks for fragmentation and their HBINs. After selecting the blocks, validate the fragmented blocks. Later on all the fragmented blocks are merged. At last the registry tree is built to verify that all found fragments are the real ones or not. Because of fragmentation, the data may be scatter and the related data may be in any other blocks. During this process there is a possibility of data to be lost.

Carving technique in this experiment is maximum successful, regarding carving the file. The problem with this approach is that loss of data. There are more chances of loss of data as it is multiple files are carved .it works well when compared to other techniques but still we need to work on the back up files and recovery as well. Hence our technique is performed well for carving the multiple registry files than other.

VI. FUTURE WORK

Currently, the tool that we used for carving the registry hives file works well when multiple blocks are fragmented. In order to be more useful, in future we have to work on the recovery part of registry by using back up files so that data must not lost and can be recoverable when needed.

VII. CONCLUSIONS

In this paper, our aim is to propose a method to carve the registry hive files based on file's internal structure. The method is successful when the multiple HBIN hive blocks are fragmented. More importance is given to the recovery as well.

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Policies for Adaptability in Retrofitting Office Building

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Abstract- As sustainability becomes a central point in the conceptual stage and design process in most architectural practice, research on office buildings' impact on the environment is greatly increasing and gaining high attention around the globe. Although many designers claim their buildings to be sustainable, unless a comprehensive Life Cycle Assessment (LCA) study is carried-out to ascertain the real situation, it is difficult to ascertain the total environmental burden a particular building has on its surrounding and global environment. This paper demonstrates how design policies for adaptability is applied to choose more sustainable materials alternatives during the maintenance (retrofit) phase of the building with an estimated projected life cycle of 50 years. However, most designers believed that, their designs will never undergo any change, as it is designed to suit the demand of the client's requirements, without given much emphasis on future retrofitting of the building to carter for most especially the technology (smart) yet to be developed. The paper focus on the need to create an awareness of the new paradigm shift in building industry in Nigeria on sustainable development where green technology is the new idea that needs to be applied to design. A Review of; the World Green Building (WGB) movement with a focus on the progress of the Malaysian GBIM (Green Building Index Malaysia) and other progressive countries were compared with the present situation of Energy utilization and conservation in Nigeria. Emphasis on how to strategize the adaptation of developmental policies into the mainstreamed of Vision 2020:20 as propose by Nigerian Government.

Index Terms: Adaptability, Office-buildings, Retrofitting, Sustainability.

I. INTRODUCTION

There has been evidence of climate change in Nigeria and the consequences well elaborated (Okali, 2008), with Nigeria contributing 0.79 per cent (0.79%) of CO_2 into the atmosphere an estimation of 2.1 Metric tons/person this exclude the land use change according to world research institute (WRI 2005).We find ourselves to be among the most vulnerable to the effect of climate change, this call for local action to be taken. Massive public awareness and in increase in knowledge base are two key factors needed to be addressed. To reduce greenhouse gas (GHG) emission at least by 5% by 2012, in order to tackle global warming and climate change, new and old buildings construction require low-energy criteria (Chlela et al., 2009). As climate change negotiations advance, it is becoming increasingly clear that there is a greater need to shift the focus on developing countries, since the international negotiations have not adequately addressed other priorities for sustainable development, support for adaptation activities and technology transfer. This will be because affordable and accessible modern energy services along with poverty reduction are essential to achieve sustainable development "(Reddy and Assenza, 2009).

II. BACKGROUND

A. Source of energy in Nigeria.

The three main sources of energy in Nigeria are; Traditional, Conventional and Alternative (Thomas, 1980). The high energy demand by the world's population has made it mandatory for us to preserve and conserve energy as well as to source for alternative. The advent of international- style and our quest for development and lifestyle, saw an era were buildings were constructed neglecting original standards and requirements. Much energy was required to cool, light, ventilate and this required the use of fossil fuel which is non-renewable and contributes an amount of CO_2 to our atmosphere.

B. Energy utilization and conservation in Nigeria.

The method in which buildings are designed and built in Nigeria compliments conservation especially in the rural area; this is as a result of the limited resources. In the rural area the shapes, forms and functions are still very traditional and materials used are predominantly clay and mud. At the late end of colonization, traditional buildings were not preserved and subtly destroyed, especially in the metropolis. The discovery of oil and prosperity that comes with it saw a large influx of ideas and styles in building from international style to the postmodern style as well as high-tech. Most of these buildings did not consider the standard and requirement for building them in the various localities and a great deal of energy was required in buildings to provide; cooling, lighting and other gadget requirements. The energy sources, after all is exhaustible and coupled with the limitation in technological base, policy are needed to enhance the conservation of the energy source (Thomas 1980).

C. Effect of climate change in Nigeria.

Late onset of the rains around (1970 - 2000) was observed in Northern and Southern part of Nigeria, which is a sign of precipitation change. Extreme weather events: Droughts and floods were experienced in some part of the country. NEST (Nigeria Environmental study Team) has identified a framework at the local and National level that will be implemented, in order to address the issue of climate change.

One of the actions is massive public awareness (Okali, 2008). Climate change is a complex phenomenon and requires mutual action among; climate environment, economy, politics, institutional, social and technological processes. The hostile effect of climate has made Nigeria one of the vulnerable nations to desertification and soil erosion. It has an adverse impact on our; agriculture, energy, biodiversity and water resources.

D. Vision 2020:20 a mile stone for Nigeria energy sector

In 2005, Goldman Sachs looked at the economy of the world and made a prediction on the world economy potentials, he predicted that most economy will outplay the big economy if they continue on the path of growth rate, Nigeria will be the 12th largest economy by 2050, this made the Nigeria government plan for 2020 in lieu of 2050 prediction (Nwachukwu, 2009). For developmental growth the issue of energy efficiency was not established, except the issue of generation and deregulation of the downstream oil sector (Soludo, 2007)

III. METHOD

A. World green building movement and the green building index

The World Green Building (WGB) movement started in early 1990 due to world global warming and climate change. It is a union of councils from around the world that aims to accelerate global sustainable building practices. The world Green Building Council (GBC) has the sole authority to appoint and direct the formation of green building councils throughout the world, launched 1988 by David Gottfried, founder of US Green building council. Various affiliated bodies are; Green building council of Australia (GBICA), United State green building council (USGBC), New-Zealand, and South-Africa have adopted the Green Star rating tools with

modification to suite the unique conditions of their countries as well as the implementation stage for the completion of Malaysian green building council MGBC. Most of the countries mentioned above are developed nations except for Malaysia which is at the height of the fully implementation of the Green Building Index Malaysia (GBIM), what about other developing countries? (Jamaludin, et. al., 2010).

B. Malaysia quest for sustainable energy development and Green Building Index Malaysia (GBIM)

Malaysia has well-established set of plans that are laid out for implementation every five (5) years. The 2006 to 2010 plan addresses the issue of energy vividly, which all started in the 70's with the strategy of implementing energy policies which had a target of renewable energy providing 5% of electricity generation by the year 2005 (Jamaludin, et al., 2010).

Sustainable building program started with energy efficiency (EE) and renewable energy (RE) program beginning in the year 2000 with Low Energy Office (LEO) building demonstration project commissioned in 2003 and followed by national Malaysia Building Integrated photovoltaic (MBIPV) program started in 2005. The newly launched Green Building Index Malaysia (GBIM) in 2009 is another milestone for Malaysia towards green development after a long effort primarily initiated by the Malaysia energy policies and promotion on RE & EE.

LEO is focuses only on low energy which directly contributes to the reduction of CO_2 emissions. ZEO of PTM (commissioned in 2006) has further proved in achieving low energy demand by tremendous Building Energy Index (BEI) reduction as low as 30kWh/sq.m/year. ZEO demonstrates both RE & EE features successfully.

C. Office building Adaptability

Office buildings design has been very much influenced lately by the concept of open building system historically (Raji et al., 2012). From the very initial period of modern office building, purpose built buildings have often incorporated the future spatial requirements, when building needs to cater initially for current spatial needs of users. These initial designs were done without any serious emphasis on future adaptability. Hence, more creative and innovative perspective on designing office buildings which will accommodate adaptability is needed to enhance building performance, through design concept for adaptable office buildings. Adaptable building appears to be a vital response to rapid change, especially in terms of user demand for more space, as a result of organizational growth. As Graham (2006) stated, "A sustainable building is not a building that must last forever, but a building that can easily adapt to change"(Eguchi et al, 2009). Thus the creation of a more sustainable environment can be augmented by adaptable design strategies that produce a level of building flexibility, and which allow for a variety of changes to be accommodated. However, developing a better understanding of how buildings change over time is another issue argued by architects concerned with extending the life of buildings. The lengthening of the life of a building is believed to be useful to reduce the consumption of natural resources and the economic burden of building expenses for organization (Kendall et al, 2011). Therefore, adaptability is inevitable to the future office building of Nigeria.

According to Schmidt et al (2005), time as a design contingency relies on placing architecture in context, making it responsive to its temporal reality and biggest fear - change. Designers tend to ignore these temporal aspects focusing on an aesthetic fixation and functional performance, freezing out time in pursuit of a static idealized object of perfection. A reaction to this way of operating is the encouragement of a more dynamic and long-term understanding of the built environment. How then, does one design for time? (Eguchi et al, 2005).

Adaptability as a design characteristic embodies spatial, structural, and service strategies which allow the physical artifact a level of malleability in response to changing operational parameters over time. (Schmidt et al, 2009) This strategic shift reflects buildings, not as finished work removed from time, but as imperfect objects whose forms are in constant flux continuously evolving to fit functional, technological, and aesthetic metamorphosis in society. According Baldwin (2000), states that the capacity for buildings to respond to these changes are highly determined through design decisions early on resulting in the building's design structure – what it is, how it is constituted (Schmidt et al., 2009). Achieving adaptability then demands a shift away from the current emphasis on form and function in response to immediate priorities, towards a 'context' and 'time-based' view of design.

Adaptability as a design principle which brings to the forefront of this critical path- time, as Croxton (2003) points out, "If a building doesn't support change and reuse, you have only an illusion of sustainability."

D. Retrofit adaptability in office building

A basic interpretation of building adaptability is the *refitting* of a physical environment as the result of a new circumstance. According to (Friedman, 2002), states that the process can be brought about by conditions that are internal or external to the building. The fitting of a building can occur in some cases prior to occupancy, as a choice to the occupants or user. Retrofitting usually take place during the occupancy period, the user exercise the previously conceived and constructed options for adaptability in a particular unit so as to satisfy inevitable dynamic needs of time (Friedman, 2002).

E. *Fitting new technology*

Architects, designers and builders incorporate into building the technology of their time as they designed and built. In the ensuing years new advances are introduced and old technologies become obsolete (Friedman, 2002). Friedman further states that systems and utilities in office buildings are designed to be upgraded and adapt to the needs of the original and subsequent users. Subcomponents are the elements that are usually fitted into building once the structure has been erected. Recent advances in information technology have introduced additional and different kinds of subcomponents into buildings. They can be electrical or computer wiring, heating and ventilation ducts. The useful life of many such subcomponents is often shorter that the life of the building structure, which requires replacement when the part is obsolete (Friedman, 2002). Hence, designing for adaptability "retrofitting" would permit easy access and replacement when the subcomponents need repair or upgrade, and will contribute immensely in reducing the amount of waste being induced by conventional renovation in retrofitting or "reworking" and existing plan of the building to meet the challenges and demands of modern technology.

IV. RESULT & DISCUSSION

A. Adapting the green technology and initiative for future development in Nigeria

The emergence of two most serious threats that confront the world today are: the threat to the atmosphere and the biosphere, which are fundamentally urban and the solution is known and clearly understood (Low, et al. 2007). This require the emergence of a green building movement in Nigeria, and the world as a whole especially those countries who have not started the movement (Achyuthan and Balagopal, 2007).

B. Policies for developing countries a beneficial movement for Nigeria

Developing countries need to integrate both development and change climatic policies that have to be embedded in its development policy and the use of SDPAM (sustainable development policies and measures). This is of the global action on climate change.

C. Technology advancement

Artisan would have to be Information Technology (IT) literate, low carbon buildings, carbon - foot printing, health and safety would be of priority as well as a mobile computing and ICT. Architecture has to do with preservation of culture and heritage but principally sustainability. To achieve and deliver this, players involved in construction project must fully utilize all available resources and IT is one such resource and the functional role of IT can be best described in the role it plays in the various stages in construction Pre-tender, post-tender and post completion stages (Sommerville & Craig 2006, UNEP, 2009).

V. CONCLUSION

Sustainability development has clearly taken on a global dimension, even in recent years it has recently been acknowledged that there is a close mutual interaction between local and global process which requires; networks, knowledge and local milieu. To achieve this in a conventional building, it requires a lot of energy and most of the energy used in Nigeria are non-renewable. For the energy from the renewable source is not enough to serve the nation and therefore have adverse effect on our climate as a whole and much concern and effort is required to formulate the policies, strategies to implement programs to develop green building and its related technology.

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Design and Analysis of Low Power Pulse Triggered Flip-Flop

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Abstract- Practically, clocking system like flip-flop (FF) consumes large portion of total chip power. In this paper, a novel low-power pulse-triggered flip-flop (FF) design is presented. Pulse- triggered FF (P-FF) has been considered as a popular alternative to the conventional master –slave based FF in the applications of high speed. First, a simple two-transistor AND gate design is used to reduce the circuit complexity. Second, a conditional pulse-enhancement technique is devised to speed up the discharge along the critical path only when needed. As a result, transistor sizes in delay inverter and pulse-generation circuit can be reduced for power saving. The maximum power saving against rival designs is up to 39.4%.Compared with the conventional transmission gate-based FF design; the average leakage power consumption is also reduced by a factor of 3.52.

Index Terms *– Flip-flop, low power, pulse-triggered*

I. INTRODUCTION

Flip-flops (FFs) are the basic storage elements used extensively in all kinds of digital designs. In particular, digital designs now-a-days often adopt intensive pipelining techniques and employ many FF-rich modules. It is also estimated that the power consumption of the clock system, which consists of clock distribution networks and storage elements, is as high as 20%–45% of the total system power. Traditional master-slave flip-flops are made up of two stages, one master and one slave and they are characterized by their hard-edge property. Alternatively, pulse-triggered flip-flops reduce the two stages into one stage and are characterized by the soft edge property. Pulse-triggered FF (P-FF) has been considered a popular alternative to the conventional master–slave based FF in the applications of high-speed operations. Besides the speed advantage, its circuits simplicity are also beneficial to lowering the power consumption of the clock tree system. The circuit complexity of a P-FF is simplified since only one latch, as opposed to two used in conventional master–slave configuration, is needed. P-FFs also allow time borrowing across clock cycle boundaries and feature a zero or even negative setup time. Depending on the method of pulse generation, P-FF designs can be classified as implicit or explicit. In an implicit-type P-FF, the pulse generator is a built-in logic of the latch design and no explicit pulse signals are generated. In an explicit-type PFF, the designs of pulse generator and latch are separate. Implicit pulse generation is often considered to be more power efficient than explicit pulse generation. In this paper, we will present a novel low-power implicit-type P-FF design featuring a conditional pulse-enhancement scheme. Three additional transistors are employed to support this feature. In spite of a slight increase in total transistor count, transistors of the pulse generation logic benefit from significant size reductions and the overall layout area is even slightly reduced.

II. IMPLICIT-TYPE P-FF DESIGN WITH PULSE CONTROL SCHEME

A.Conventional Implicit-Type P-FF Designs

1)ip-DCO: ip-DCO is known as the implicit data close to output. It is an implicit type flip-flop. In this method the pulse is generated inside the flip-flop. A state-of-the-art P-FF design, named ip-DCO, is given in Fig.1. It contains an AND logic-based pulse generator and a semi-dynamic structured latch design. Semi-Dynamic Flip-Flop is a high performance flip-flop because of its small delay and simple topology. It is measured to be one of fastest flip-flops today.

Inverters I5 and I6 are used to latch data and inverters I7 and I8 are used to hold the internal node X. The pulse generator takes complementary and delay skewed clock signals to generate a transparent window equal in size to the delay by inverters I1-I3. Two practical problems exist in this design. First, during the rising edge, nMOS transistors N2 and N3 are turned on. If data remains high, node X will be discharged on every rising edge of the clock. This leads to a large switching power.

2)MHLFF: The modified hybrid latch flip-flop is known as MHLFF and this is an type of implicit type flip-flop. MHLFF shows an improved P-FF design in fig.2. It employs a static latch structure. A static latch can remember as long as gate power is supplied. It uses feed-back to remember, rather than depending on the charge on a capacitor. Node X is no longer precharged periodically by the clock signal. A weak pull-up transistor P1 controlled by the FF output signal Q is used to maintain the node level at high when Q is zero.

This design eliminates the unnecessary discharging problem at node. However, it encounters a longer Data-to-Q (D-to-Q) delay during “0” ,“1” transition because node is not pre-discharged. Larger transistors N3 and N4 are required to enhance the discharging capability. Another drawback of this design is that node becomes floating when output Q and input Data both equal to “1”. Extra DC power emerges if node X is drifted from an intact“1”.

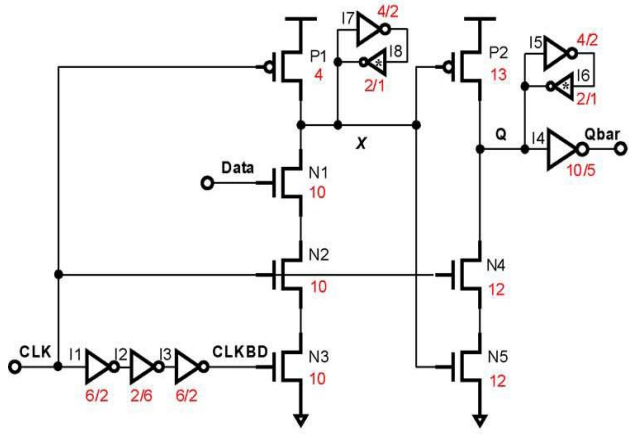


Figure.1. ip-DCO

3)SCCER: SCCER is known as the single ended conditional capturing energy recovery P-FF. It is a refined low power P-FF design using a conditional discharged technique. This technique is also used to present a new flip-flop Conditional Discharge flip-flop (CDFF). CDFF use a pulse generator which is suitable for double edge sampling. CDFF has two stages. First is responsible for capturing the Low-to-High transition and second stage captures the High-to-Low input transition. In this SCCER design, the keeper logic is replaced by a weak pull up transistor P1 in conjunction with an inverter I2 to reduce the load capacitance of node. The discharge path contains nMOS transistors N2 and N1 connected in series. In order to eliminate superfluous switching at node, an extra nMOS transistor N3 is employed. Since N3 is controlled by Q_fdbk, no discharge occurs if input data remains high.

The worst case timing of this design occurs when input data is "1" and node is discharged through four transistors in series, i.e., N1 through N4, while combating with the pull up transistor P1. A powerful pull-down circuitry is thus needed to ensure node can be properly discharged. This implies wider N1 and N2 transistors and a longer delay from the delay inverter I1 to widen the discharge pulse width.

III. PROPOSED P-FF DESIGN

Fig. 4 shows the proposed design. The proposed design, adopts two measures to overcome the problems associated with existing P-FF designs. The first one is reducing the number of nMOS transistors stacked in the discharging path. The second one is supporting a mechanism to conditionally enhance the pull down strength when the input data is "1." As opposed to the transistor stacking design in Fig. 1 and Fig. 3, transistor N2 is removed from the discharging path. Transistor N2, in conjunction with an additional transistor N3, forms a two-input pass transistor logic (PTL)-based AND gate to control the discharge of transistor N1. Pass transistors require lower switching energy to charge up a node, due to the reduced voltage swing.

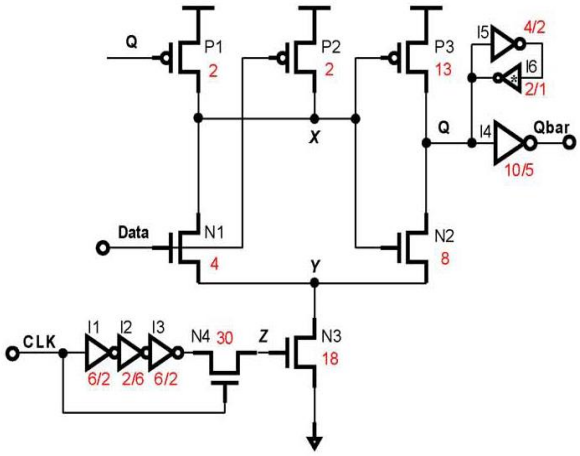


Figure.2. MHLFF

In SCCER design, the discharge control signal is driven by a single transistor. Parallel conduction of two nMOS transistors (N2and N3) speeds up the operations of pulse generation. Thus the number of stacked transistors along the discharging path is reduced. To enhance the discharging condition, transistor P3 is added. When the FF output Q changes from 0 to 1 the conditional pulse enhancement technique effectively takes place. Thus this leads to the better power performance compared to the indiscriminate pulse enhancement approach.

The post layout simulations on various P-FF were conducted to obtain the performance figure of the proposed design. These designs include three flip-flops namely ip-DCO, MHLFF and SCCER. And those designs are discussed above. The target technology is the UMC 90-nm CMOS process. The operating condition used in simulations is 500 MHz/1.0V.

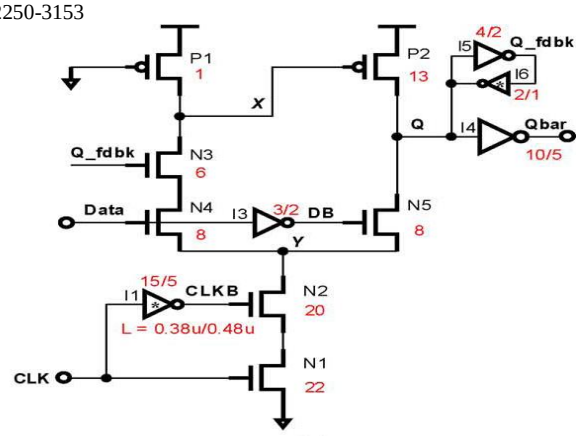
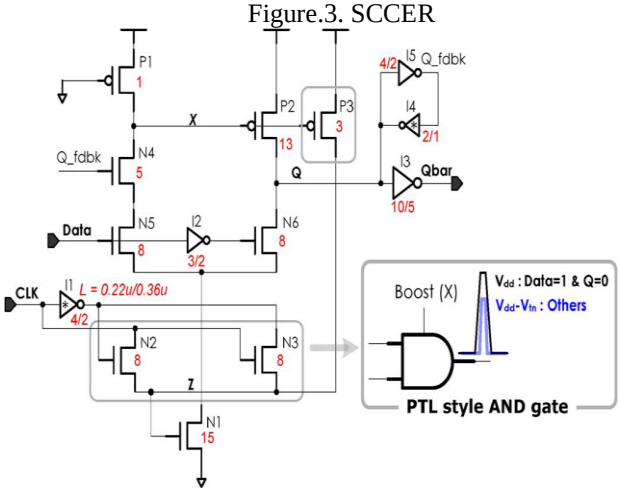


Figure.3. SCCER



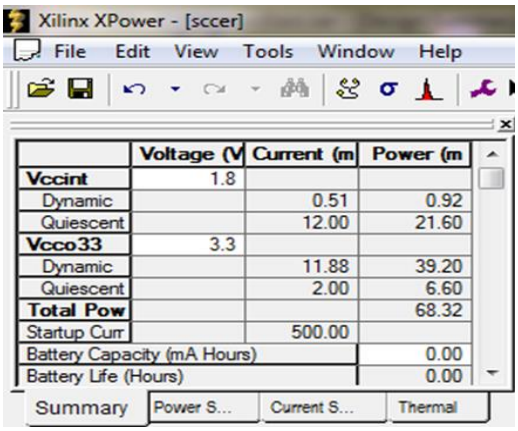


Figure.7. Power consumed by SCCER

Figure.4. Schematic of the proposed P-FF design with pulse control scheme

IV. Simulation Results

The simulation results of above designs are shown below in the Fig. 5 to Fig. 8.

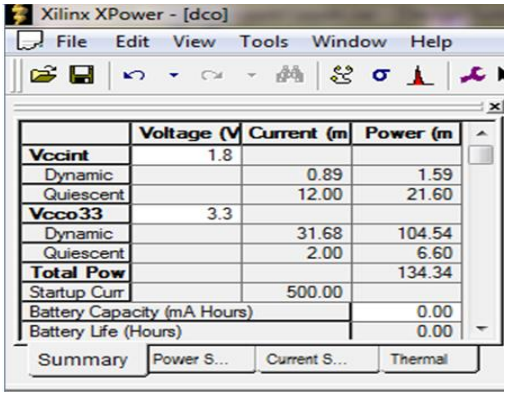


Figure.5. Power consumed by ip-DCO

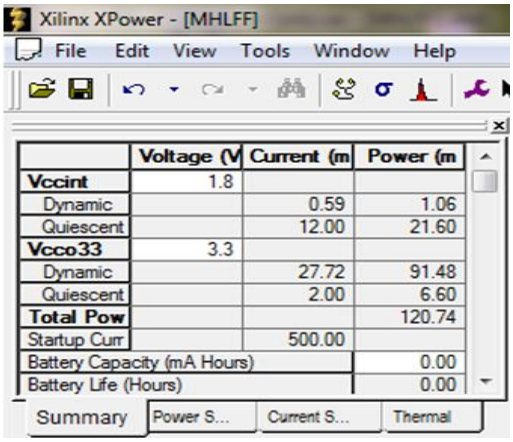


Figure.6. Power consumed by MHLFF

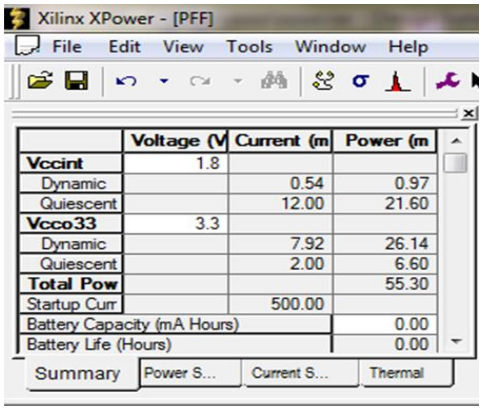


Figure.8. Power consumed by P-FF

The power consumed by the ip-DCO is 134.34mW, MHLFF is 120.74mW, SCCER is 68.32mW and P-FF is 55.3mW.By comparing the above results, we came to know that power consumption of P-FF design is low.

Table.I. Power Comparision

FF	TOTAL ESTIMATED POWER CONSUMPTION P(mW)
ip- DCO	134.34
MHLFF	120.74
SCCER	68.32
P-FF	55.3

V. Conclusion

In this paper, we devise a novel low-power pulse-triggered FF design. This was successfully done by reducing the number of transistors stacked along the discharging path by incorporating a PTL-based AND logic. The table 1 has been added to verify that the proposed design will be better compared to the existing design like ip-DCO, MHLLF, SCCER.

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Hierarchical Clustering For Cancer Discovery Using Range Check And Delta Check

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Abstract: Class discovery is one of the most important tasks in cancer classification using biomolecular data. To perform this, a multiple clustering approach called Hierarchical clustering is used. It uses one of the metrics called Manhattan Distance which measures the distance between the values of the data set and builds a hierarchy of clusters after analysing it. The clustering result enables to classify the cancer types and it is further evaluated by Range check and Delta check. The various test results are compared with the known initial range of values using Range check. Delta check is performed on the current test result and the immediate previous test result for better results. These techniques are used to improve the diagnosis of cancer.

Index Terms- Class Discovery, Hierarchical clustering, Range check, Delta check

I. INTRODUCTION

Cancer diagnosis and treatment involves discovering and classifying cancer types. Most of the previous works involve the single clustering algorithms. In Golub'swork [3], the self-organizing feature map and neighbourhood analysis were adopted to discover two types of human acute leukemia, which are acute myeloid leukemia (AML) and acute lymphoblastic leukemia (ALL). In Wigle's work [4], clustering approaches and statistical analysis were adopted to identify non-small cell lung cancer (NSCLC) from the normal cases. These works include certain limitations such as lack of robustness, stability and accuracy. But in our paper, we have adopted the concept called Hierarchical clustering which is one of the multiple clustering algorithms. This is the powerful method for improving both the robustness as well as the stability of unsupervised classification solutions.

Cancer classification using biomolecular data poses a major challenge in monitoring the levels of thousands of genes[1] [2], and this can be overcome by using the machine learning technique. For this purpose we use the WEKA tool which is the popular machine learning workbench. WEKA contains simple implementations of algorithms for classification, clustering, and association rule mining, along with graphical user interfaces and visualization utilities for data exploration and algorithm evaluation.

II. HIERARCHICAL CLUSTERING

Hierarchical methods for unsupervised and supervised data mining give multi-level description of data. It is relevant for many applications related to information extraction, retrieval navigation and organization [5]. Hierarchical clustering algorithm is of two types:

i) Agglomerative Hierarchical clustering algorithm or AGNES (agglomerative nesting) and

ii) Divisive Hierarchical clustering algorithm or DIANA (divisive analysis).

We focus on agglomerative probabilistic clustering. This algorithm works by grouping the data one by one on the basis of the nearest distance measure of all the pairwise distance between the data point. This way we go on grouping the data until one cluster is formed. Now on the basis of dendogram graph we can calculate how many numbers of clusters should be actually present.

Algorithmic steps for Agglomerative Hierarchical clustering

Let X = {x1, x2, x3, ..., xn} be the set of data points.

1) Begin with the disjoint clustering having level L(0) = 0 and sequence number m = 0.

2) Find the least distance pair of clusters in the current clustering, say pair (r), (s), according to d[(r),(s)] = min d[(i),(j)] where the minimum is over all pairs of clusters in the current clustering.

3) Increment the sequence number: m = m +1.Merge clusters (r) and (s) into a single cluster to form the next clustering m. Set the level of this clustering to L(m) = d[(r),(s)].

4) Update the distance matrix, D, by deleting the rows and columns corresponding to clusters (r) and (s) and adding a row and column corresponding to the newly formed cluster. The distance between the new cluster, denoted (r,s) and old cluster(k) is defined in this way: d[(k), (r,s)] = min (d[(k),(r)], d[(k),(s)]).

5) If all the data points are in one cluster then stop, else repeat from step 2).

Divisive Hierarchical clustering - It is just the reverse of Agglomerative Hierarchical approach.

The probabilistic scheme enables automatic detection of the final hierarchy level.

III. INTRODUCTION TO WEKA

WEKA may prove useful to others involved in the development of open-source machine learning software. It contains implementations of algorithms for classification, clustering, and association rule mining, along with graphical user interfaces and visualization utilities for data exploration and algorithm evaluation, which are described as the main features. WEKA uses the java language and therefore satisfies the promise of platform independence. Weka's standard file format is ARFF, any file which is used to be used in Weka should end with .arff format.

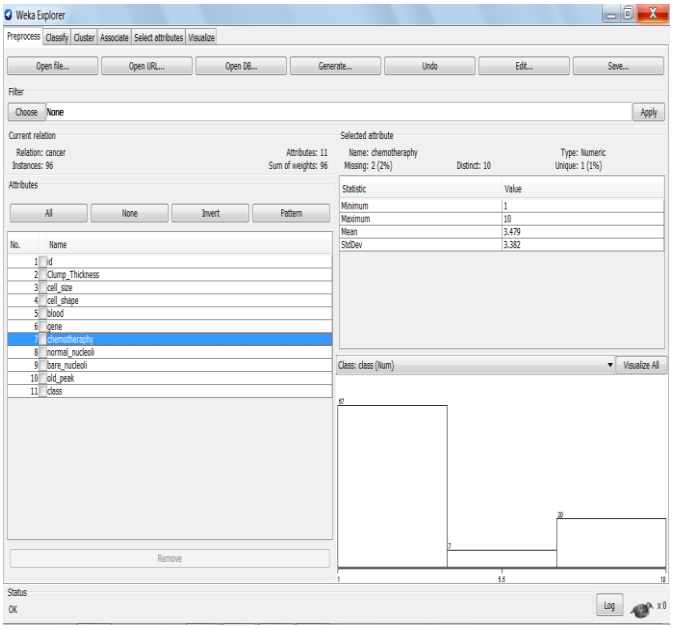
A. Sample cancer Data set

	id no	clump thic	cell size	cell shape	Blood cont	gene	chemother	Normal Nu	Bare Nucl	Old peak	Class
2	1000025	5	1	1	1	2	1	3	1	1	2
3	1002945	5	4	4	5	7	10	3	2	1	2
4	1015425	3	1	1	1	2	2	3	1	1	2
5	1016277	6	8	8	1	3	4	3	7	1	2
6	1017023	4	1	1	3	2	1	3	1	1	2
7	1017122	8	10	10	8	7	10	9	7	1	4
8	1018099	1	1	1	1	2	10	3	1	1	2
9	1050718	6	1	1	1	2	1	3	1	1	2
10	1054590	7	3	2	10	5	10	5	4	4	4
11	1054593	10	5	5	3	6	7	7	10	1	4
12	1056784	3	1	1	1	2	1	2	1	1	2
13	1057013	8	4	5	1	2	?	7	3	1	4
14	1059552	1	1	1	1	2	1	3	1	1	2
15	1065726	5	2	3	4	2	7	3	6	1	4
16	1066373	3	2	1	1	1	1	2	1	1	2
17	1066979	5	1	1	1	2	1	2	1	1	2
18	1067444	2	1	1	1	2	1	2	1	1	2
19	1070935	3	1	1	1	1	1	2	1	1	2
20	1071760	2	1	1	1	2	1	3	1	1	2
21	1072179	10	7	7	3	8	5	7	4	3	4
22	1074610	2	1	1	2	2	1	3	1	1	2

Cancer data set with missing values

The sample Cancer data set contains some of the missing values and these can be replaced by choosing the unsupervised filter in WEKA and select Replace Missing Values and click apply. After this all the missing values are replaced with an appropriate value either by taking mean or any nearer value. This will reduce the effort and the risk of replacing the values manually.

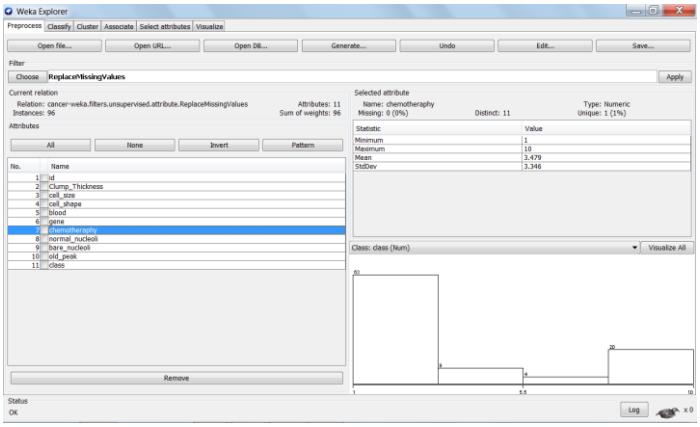
B. Output before replacing the missing values



Here in this data set there is 2% of missing value in the data chemotherapy and these missing values can be replaced by using filters. The output can be visualized at the bottom corner. As well as all of the data outputs can also be visualized by selecting visualize all.

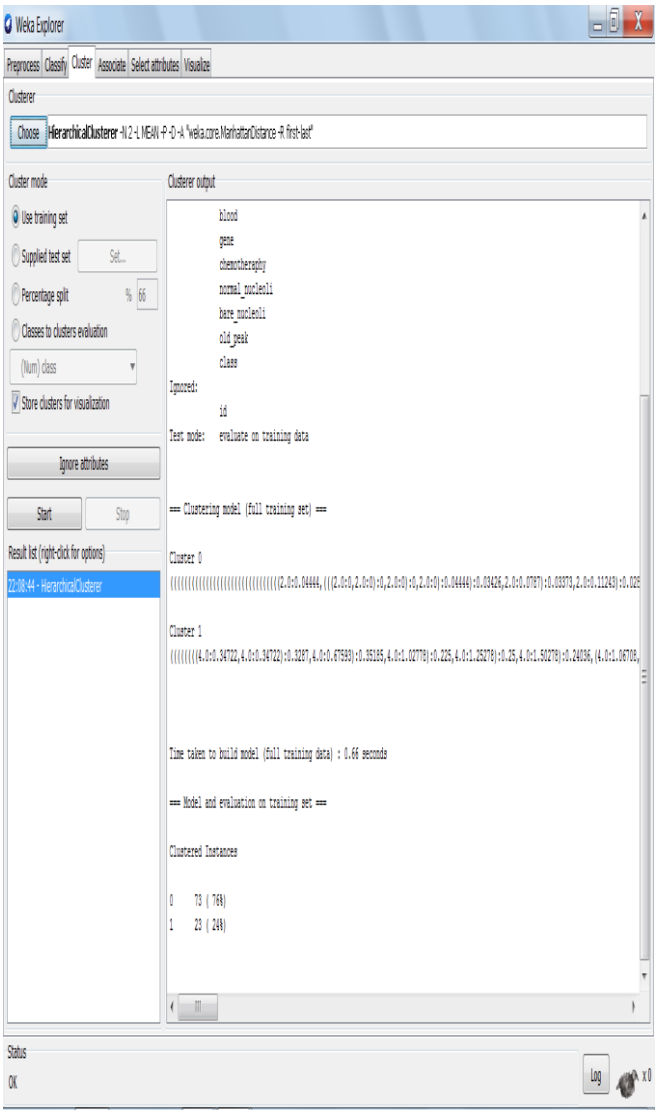
C. Output after replacing the missing values

Here the missing values are replaced by selecting the unsupervised filter. And since the Weka tool is an automated machine learning the missing values will be replaced automatically by selecting the filter.



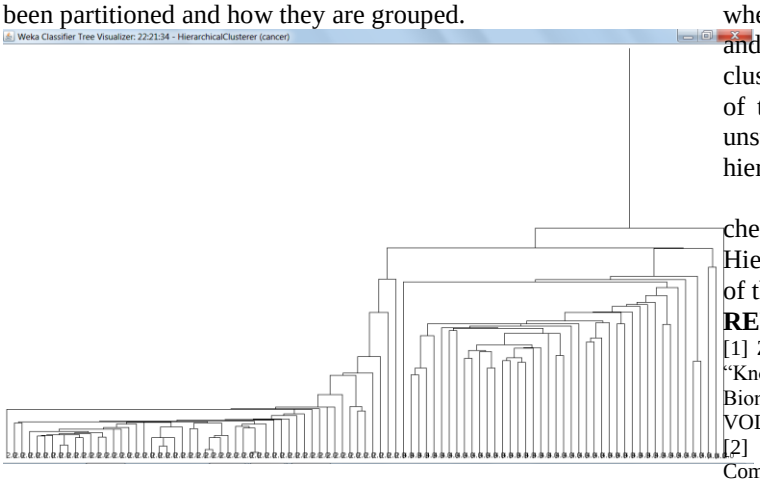
IV. CLUSTER OUTPUT

After choosing the Hierarchical clustering method in the weka tool select one of the class and number of clusters. If any attributes has to be ignored it can also be selected and ignored. Then click on start button below, the output window will show the mean, standard deviation, instances, attributes, and percentage split of clusters [6] [7].



V. HIERARCHICAL TREE VIEW

The cluster is formed by grouping the data and the dendogram grath (i.e., the hierarchical tree view for the cancer data set is shown). This also shows the number of clusters has been partitioned and how they are grouped.



VI. RANGE CHECK AND DELTA CHECK

Range check and Delta check are the techniques used to validate the results generated by the Hierarchical clustering approach.

The automated system normally uses a range check technique as a quality control measure. In **range check**, the results of the various tests are compared with known normal range of values for the tests and the normal results are accepted and stored. The results that are not within the normal range of values are identified as panic values and a smear of the blood is prepared from those samples and it is sent for an equivalent manual procedure to ensure quality of the system

Delta check is performed on the current and the immediate previous test result for better results and in order reduce the number of manual reviews.

VII. CONCLUSION

In this paper we have used one of the multiple clustering algorithms named Hierarchical clustering algorithm which replaces the existing use of the single clustering algorithm where certain limitations such as lack of robustness, stability, and accuracy follows. To avoid these limitations multiple clustering algorithm is used in this paper. The implementation of the algorithm was carried out by WEKA tool where the unsupervised data can be managed by using filters and the hierarchical tree view of the data set can be evaluated easily.

And further the techniques such as Range check and Delta check are used to validate the results generated by the Hierarchical clustering approach, in order to maintain accuracy of the result.

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Notes

[i] The Gods like Visnu and Siva in Hindu religion are here called as 'great' gods, because they obtain such place in the Hindu community.
[ii] Ramachandra Shastri Pandey (ed.), *Mahabharata, Anusasanaparva*, 95.18.96.8, Gita Press, Gorakhpur, 1998.
[iii] *ibid. Vanaparva*, 116.
[iv] N D Kambale, *The cult of Yellamma or Renuka*, dspace.vidyanidhi.org.

[v] Mang is a tribal community in Maharastra region, Edgar Thurston, *Castes and tribes of Southern India*, Asia Educational Service, New Delhi, 2001, pp.159, 222, 238,292, 386.
[vi] S., Charsley, www.simoncharsley.com, Madiga and Dalit: The *Yellamma* Cult.
[vii] It is also told that more than these practices, offering of blood, alcohol etc., were also dedicated to the goddesses during the time when Kapalika were priests of this temple.
[viii] Jangamas are Virasaivite order men wandering place to place. They had the place of Guru among ht Virasaiva people.
[ix] *Yellamma* was a daughter of a rich man staying near a hill. As she grew into youth, she got affected with skin disease. Once went to place where there were the Padukas of Ekayya and Jogayya are installed. As she did penance there, her skin disease got completely cured. This myth is popular among the Virasaivas. From this time, *Yellamma* was also worshipped by the devotees who arrived to the holy place of Ekayya and Joggayya. Here the local myth told now is changed like this. Renuka got skin disease because she viewed *Gandharva* sporting in the water and due to this she lost her chastity and got skin disease. Wandering here and there, she saw Aswini Devatas the twin physicians and they after seeing her, got back her to charm and they as usual the story of beheading and all follows.
[x] Sthanika means governor or tax collector of a place. Historians tell than the Tulu Brahmins are also called as Sthanikas. It is also historically proved that these persons were appointed as the chief managers of the *Yellamma* shrine also to claim the benefits from that temple. Nagendra singh, *Divine Prostitution*, APH Publishig Corporation, New Delhi, 1997, pp, 95, 96.
[xi] Nicholas J. Bradford, Transgenderism and the Cult of *Yellamma*: Heat, Sex, and Sickness in South Indian Ritual, Journal of Anthropological Research, Vol. 39, New Mexico, 1983, pp.307- 332.
[xii] Mohan Giri, *Kanya Exposition of Little Angels*, Gyan Publishing House, New Delhi, 1999, pp. 35-38.
[xiii] *ibid*, p.38.

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