NSRIT

AUTONOMOUS

ANSWER KEY & SCHEME OF EVALUATION

Fourth Semester

ACADEMIC REGULATION 2020

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NSRIT

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	Program	Common To All			Academic Year	2021 - 2022
Course Code	20HSX03	Test Duration	3 Hrs.	Max. Marks	70	Semester	IV
Course	MANAGERIAL ECONOMICS AND FINANCIAL ANALYSIS						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What is the definition of managerial economics?	20HSX03.1	L1
2	Define Angle of incidence.	20HSX03.2	L1
3	Write the proforma of journal entry.	20HSX03.1	L1
4	What is Pay Back Period?	20HSX03.4	L1
5	Write the formula of current ratio.	20HSX03.1	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	How do you link managerial economics with other disciplines/subjects?	6M	20HSX03.1	L2
6 (b)	Define elasticity of demand and explain different types of elasticity of demand.	6M	20HSX03.1	L2
	OR			
7 (a)	What is demand? Explain the different types of demand.	6M	20HSX03.1	L2
7 (b)	What do you mean by demand forecasting? Explain various demand forecasting methods.	6M	20HSX03.1	L2
8(a)	Explain the production function with one variable.	6M	20HSX03.2	L2
8 (b)	What is break even analysis? How do you determine breakeven point? Illustrate.	6M	20HSX03.2	L2
	OR			
9 (a)	Explain Cobb Douglas Production function.	4M	20HSX03.2	L2
9 (b)	You are required to calculate i) Margin of Safety ii) Total sales iii) Variable cost iv) Fixed costs Rs. 12,000, Profit Rs. 1,000, Break-Even Sales Rs.60,000	8M	20HSX03.2	L3
10 (a)	(see table below)	6M	20HSX03.3	L3

Particulars	Rs.	Particulars	Rs.
Opening stock	1,250	Plant and machinery	6,230
Sale	11,800	Returns outwards	1,380
Depreciation	667	Cash in hand	942
Commission (Cr).	211	Salaries	750
Insurance	380	Debtors	1,905
Carriage inwards	300	Discount (Dr.)	328
Furniture	670	Bills receivable	2,730
Printing charges	481	Wages	1,589
Carriage outwards	200	Return inwards	1,659
Capital	9,228	Bank overdraft	4,000
Creditors	1,780	Purchases	8,679
Bills payable	541	Bad debts	180

	The above balances extracted from the books of mythri & company, prepare a trading, profit and loss account and a balance sheet. The value of stock on 31st December, 1990 was Rs. 3,700.			
10 (b)	Write about short note on trading account and profit and loss account.	6M	20HSX03.3	L2
	OR			
11 (a)	Explain the concepts of journal and ledger accounts with performa.	6M	20HSX03.3	L2
11 (b)	Journalizing the following transactions: Jan 1 Started business with cash Rs.100,000 2 Deposited Rs.75,000 to bank 3 Purchased furniture Rs.20,000 and paid by cheque (Through Bank) 5 Paid shop rent Rs.2,500 cash 7 withdrew from bank for personal use Rs.1,000 8 Sold goods on credit Rs.6,000 to Jithesh 10 Received interest from bank Rs.600	6M	20HSX03.3	L3
12 (a)	What do you mean by payback period method? Explain the merits and demerits of payback period method.	6M	20HSX03.4	L2
12 (b)	Solve the payback period of the following projects each requiring a cash outlay of Rs 1,00,000 each. (see table below)	6M	20HSX03.3	L3

Year	Cash Inflows Rs.	
	Project A	Project B
1	30,000	30,000
2	30,000	40,000
3	30,000	20,000
4	30,000	25,000
5	30,000	5,000

OR

13 (a)	What is capital? Explain the types and significance.	6M	20HSX03.4	L2
13 (b)	Find the net present value at the rate of 10% per annum from the following data related to CNC machines 1 and 2. The estimated cash flows after taxes for each machine are as given below. (see table below)	6M	20HSX03.4	L2

Year	CNC Machine 1	CNC Machine 2
1	1,50,000	2,00,000
2	3,00,000	3,00,000
3	1,50,000	2,50,000
4	____	1,50,000
Total	6,00,000	6,00,000

Investment is 3,00,000 in each project

<table>
<tr><td>14 (a)</td><td>From the following information, solve
i. Debt-Equity ratio
ii. Current ratio
iii. Quick ratio
<table>
<tr><td>Liabilities</td><td>Rs.</td><td>Assests</td><td>Rs.</td></tr>
<tr><td>Debentures</td><td>1,40,000</td><td>Bank balance</td><td>30,000</td></tr>
<tr><td>Long term Loans</td><td>70,000</td><td>Sundry Debtors</td><td>70,000</td></tr>
<tr><td>General reserve</td><td>40,000</td><td></td><td></td></tr>
<tr><td>Creditors</td><td>66,000</td><td></td><td></td></tr>
<tr><td>Bills payable</td><td>14,000</td><td></td><td></td></tr>
<tr><td>Share capital</td><td>1,20,000</td><td></td><td></td></tr>
</table></td><td>6M</td><td>20HSX03.5</td><td>L3</td></tr>
<tr><td>14 (b)</td><td>What do you mean by accounting ratios? How are they useful?</td><td>6M</td><td>20HSX03.5</td><td>L2</td></tr>
<tr><td colspan="5">OR</td></tr>
<tr><td>15 (a)</td><td>Solve interest coverage ratio from the following information
<table>
<tr><td>Particulars</td><td>Rs.</td></tr>
<tr><td>Net profit after deducting interest and taxes</td><td>6,00,000</td></tr>
<tr><td>12% Debentures of the face value of</td><td>15,00,000</td></tr>
<tr><td>Amount provided towards taxation</td><td>1,20,000</td></tr>
</table></td><td>6M</td><td>20HSX03.5</td><td>L3</td></tr>
<tr><td>15 (b)</td><td>Explain the various profitability ratios and explain the meaning and method calculation of these methods.</td><td>6M</td><td>20HSX03.5</td><td>L2</td></tr>
</table>

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

Managerial Economics and Financial Analysis

1. **What is the definition of managerial economics?**

Answer :

Managerial economics is defined as the branch of economics which deals with the application of various concepts, theories, and methodologies of economics to solve practical problems in business management.

2. **Define Angle of incidence.**

Answer : **The angle which is created by cost and sales line** is called the angle of incidence. This angle is formed from the starting of a break-even point. The angle of incidence shows the rate at which a company is making profits. The simple rule is that the bigger the angle of incidence higher is the rate of profit.

3. **Write the proforma of journal entry.**

Answer: Journal is a first book of recording an account which all transactions recorded in chronological(Date wise) order.

Date	Particulars	Ledger folio	Debit Amount	Credit Amount

4. **What is Pay Back Period?**

Answer :

Payback period is defined as **the number of years required to recover the original cash investment**. In other words, it is the period of time at the end of which a machine, facility, or other investment has produced sufficient net revenue to recover its investment costs.

5. **Write the formula of current ratio.**

Answer :

The current ratio, also known as the working capital ratio, measures the capability of a business to meet its short-term obligations that are due within a year. The ratio considers the weight of total current assets versus total current liabilities. It indicates the financial health of a company and how it can maximize the liquidity of its current assets to settle debt and payables. The current ratio formula (below) can be used to easily measure a company's liquidity.

Current Ratio = Current Assets / Current Liabilities

6.a. How do you link managerial economics with other disciplines/subjects?

Answer:

Managerial Economics and Theory of Decision Making:

The theory of decision making is relatively a new subject that has a significance for managerial economics. In the process of management such as planning, organising, leading and controlling, decision making is always essential. Decision making is an integral part of today's business management. A manager faces a number of problems connected with his/her business such as production, inventory, cost, marketing, pricing, investment and personnel.

Economist are interested in the efficient use of scarce resources hence they are naturally interested in business decision problems and they apply economics in management of business problems. Hence managerial economics is economics applied in decision making.

Managerial Economics and Operations Research:

Mathematicians, statisticians, engineers and others join together and developed models and analytical tools which have grown into a specialised subject known as operation research. The basic purpose of the approach is to develop a scientific model of the system which may be utilised for policy making.

The development of techniques and concepts such as Linear Programming, Dynamic Programming, Input-output Analysis, Inventory Theory, Information Theory, Probability Theory, Queuing Theory, Game Theory, Decision Theory and Symbolic Logic.

Managerial Economics and Statistics:

Statistics is important to managerial economics. It provides the basis for the empirical testing of theory. It provides the individual firm with measures of appropriate functional relationship involved in decision making. Statistics is a very useful science for business executives because a business runs on estimates and probabilities.

Statistics supplies many tools to managerial economics. Suppose forecasting has to be done. For this purpose, trend projections are used. Similarly, multiple regression technique is used. In managerial economics, measures of central tendency like the mean, median, mode, and measures of dispersion, correlation, regression, least square, estimators are widely used.

Statistical tools are widely used in the solution of managerial problems. For eg. sampling is very useful in data collection. Managerial economics makes use of correlation and multiple regression in business problems involving some kind of cause and effect relationship.

Managerial Economics and Accounting:

Managerial economics is closely related to accounting. It is recording the financial operation of a business firm. A business is started with the main aim of earning profit. Capital is invested / employed for purchasing properties such as building, furniture, etc and for meeting the current expenses of the business.

Goods are bought and sold for cash as well as credit. Cash is paid to credit sellers. It is received from credit buyers. Expenses are met and incomes derived. This goes on the daily routine work of the business. The buying of goods, sale of goods, payment of cash, receipt of cash and similar dealings are called business transactions.

The business transactions are varied and multifarious. This has given rise to the necessity of recording business transaction in books. They are written in a set of books in a systematic manner so as to facilitate proper study of their results.

There are three classes of accounts:

(i) Personal account,

(ii) Property accounts, and

(iii) Nominal accounts.

Management accounting provides the accounting data for taking business decisions. The accounting techniques are very essential for the success of the firm because profit maximisation is the major objective of the firm.

Managerial Economics and Mathematics:

Mathematics is another important subject closely related to managerial economics. For the derivation and exposition of economic analysis, we require a set of mathematical tools. Mathematics has helped in the development of economic theories and now mathematical economics has become a very important branch of economics.

6.b. Define elasticity of demand and explain different types of elasticity of demand.

Answer:

Elasticity is a concept in economics that talks about the effect of change in one economic variable on the other.

<u>ELASTICITY OF DEMAND</u>

Elasticity of Demand, or Demand Elasticity, is the **measure of change in quantity demanded of a product in response to a change in any of the market variables, like price, income etc.** It measures the shift in demand when other economic factors change.

TYPES OF ELASTICITY OF DEMAND:

On the basis of different factors affecting the quantity demanded for a product, elasticity of demand is categorized into mainly three categories: **Price Elasticity of Demand (PED), Cross Elasticity of Demand (XED), and Income Elasticity of Demand (YED).**

1. <u>Price Elasticity of Demand (PED)</u>

Any change in the price of a commodity, whether it's a decrease or increase, affects the quantity demanded for a product. For example, when there is a rise in the prices of ceiling fans, the quantity demanded goes down.

This measure of responsiveness of quantity demanded when there is a change in price is termed as the Price Elasticity of Demand (PED).

The mathematical formula given to calculate the Price Elasticity of Demand is:

PED = % Change in Quantity Demanded % / Change in Price

2. <u>Income Elasticity of Demand (YED)</u>

The Income Elasticity of Demand, also represented by YED, refers to the sensitivity of quantity demanded for a certain good to a change in real income (the income earned by an individual after accounting for inflation) of the consumers who buy this good, keeping all other things constant.

The formula given to calculate the Income Elasticity of Demand is given as:

YED = % Change in Quantity Demanded% / Change in Income

3. <u>Cross Elasticity of Demand (XED)</u>

In a market where there is an oligopoly, multiple players compete. Thus, the quantity demanded for a product does not only depend on itself but rather, there is an effect even when prices of other goods change.

Cross Elasticity of Demand, also represented as XED, is an economic concept that measures the sensitiveness of quantity demanded of one good (X) when there is a change in the price of another good (Y), and that's why it is also referred to as Cross-Price Elasticity of Demand.

The formula given to calculate the Cross Elasticity of Demand is given as:

XED = (% Change in Quantity Demanded for one good (X)%) / (Change in Price of another Good (Y))

4. <u>Advertisement Elasticity of Demand (AED)</u>

The Advertisement Elasticity of Demand, also represented by AED, refers to the sensitivity of quantity demanded for a certain good to a change in advertisement of the consumers who buy this good, keeping all other things constant.

The formula given to calculate the Advertisement Elasticity of Demand is given as:

AED = % Change in Quantity Demanded% / Change in Advertisement

7.a. What is demand? Explain the different types of demand.
Answer :

MEANING OF DEMAND:

Demand is the number of goods that the customers are ready and able to buy at several prices during a given time frame. The association between price and quantity demanded is also known as demand curve. Preferences and choices, which are the basics of demand, can be depicted as the functions of costs, odds, benefits, and other variables.

The amount of a good that the customer picks up modestly relies on the cost of the commodity, the cost of other commodities, the customer's earnings, and his or her tastes and proclivity. The amount of a commodity that a customer is ready to purchase and is able to manage and afford, provided that the prices of goods, and customer's tastes and preferences are known, is referred to as demand for the commodity.

In our daily life, we often see that a consumer's preferences for products change according to their preferences, income, and the prices of the goods or the prices of the other goods.

Here, the demand of a product can be defined as the quantity of a product that a consumer is eager to purchase, can afford at a given price, and is according to his/her preferences and tastes. Whenever there is a change in any of those variables, the demand and supply of the product starts changing.

TYPES OF DEMAND:

Market or individual demand: Here, the individual demand is defined as the demand for products or services by an individual consumer. The market demand can be defined as a demand for a product made by a bunch of consumers who buy that product. Therefore, it is a collective demand of each individual's demand.

Derived demand: The derived demand is defined when the goods manufactured are related to the demand for other products . For example, the demand for silk yam is the result of the demand for silk cloth. However, the direct demand for goods can be defined when the demand for a product is independent. For example, there is an autonomous demand for cotton cloth.

Price demand: The price demand refers to the number of goods or services an individual is eager to buy at a given price.

Income demand: The income demand means the eagerness of a person to buy a definite quantity at a given income level.

Cross demand: This is one of the important types of demand where the demand of a product is not subjected to its own price but the price of other similar products is known as the cross demand

7.b. What do you mean by demand forecasting? Explain various demand forecasting methods.

Answer :

Demand forecasting is known as the process of making future estimations in relation to customer demand over a specific period. Generally, demand forecasting will consider historical data and other analytical information to produce the most accurate predictions.

Demand forecasting plays an important role for businesses in different industries, particularly in reducing risk in business activities. However, it is known to be a challenge that companies face due to the intricacies of analysis, specifically quantitative analysis. Yet, understanding customer needs is an indispensable part of any industry, so that business plans can be implemented more efficiently and can more appropriately respond to market needs. If businesses begin to master the concept of demand forecasting, it can result in several benefits. These include, but are not limited to, waste reduction, more optimal allocation of resources and potentially dramatic increases in sales and revenue.

Demand Forecasting Methods:

1] Survey of Buyer's Choice

When the demand needs to be forecasted in the short run, say a year, then the most feasible method is to ask the customers directly that what are they intending to buy in the forthcoming time period. Thus, under this method, potential customers are directly interviewed. This survey can be done in any of the following ways:

2] Collective Opinion Method

Under this method, the salesperson of a firm predicts the estimated future sales in their region. The individual estimates are aggregated to calculate the total estimated future sales. These estimates are reviewed in the light of factors like future changes in the selling price, product designs, changes in competition, advertisement campaigns, the purchasing power of the consumers, employment opportunities, population, etc.

3] Barometric Method

This method is based on the past demands of the product and tries to project the past into the future. The economic indicators are used to predict the future trends of the business. Based on future trends, the demand for the product is forecasted. An index of economic indicators is formed. There are three types of economic indicators, viz. leading indicators, lagging indicators, and coincidental indicators.

4] Market Experiment Method

Another one of the methods of demand forecasting is the market experiment method. Under this method, the demand is forecasted by conducting market studies and experiments on consumer behavior under actual but controlled, market conditions.

5] Expert Opinion Method

Usually, market experts have explicit knowledge about the factors affecting demand. Their opinion can help in demand forecasting. The Delphi technique, developed by Olaf Helmer is one such method.

6] Statistical Methods

The statistical method is one of the important methods of demand forecasting. Statistical methods are scientific, reliable and free from biases. The major statistical methods used for demand forecasting are:

a. **Trend Projection Method:** This method is useful where the organization has a sufficient amount of accumulated past data of the sales. This date is arranged chronologically to obtain a time series. Thus, the time series depicts the past trend and on the basis of it, the future market trend can be predicted. It is assumed that the past trend will continue in the future. Thus, on the basis of the predicted future trend, the demand for a product or service is forecasted.

b. **Regression Analysis:** This method establishes a relationship between the dependent variable and the independent variables. In our case, the quantity demanded is the dependent variable and income, the price of goods, the price of related goods, the price of substitute goods, etc. are independent variables. The regression equation is derived assuming the relationship to be linear. Regression Equation: $Y = a + bX$. Where Y is the forecasted demand for a product or service.

8. a. Explain the production function with one variable.

Answer:

"The production function as that function which decide the maximum amount of output that can be Produced with a given set of inputs."

Laws of production:

The production function can be studied in 3 ways.

→ Production function with one variable input & law of retu[rns]

→ production function with two variable input & law of returi[ns]

→ Production function where input factors are multiple cor[responding]

law of returns to scale.

- How to produce ?
  - Short run
    - one variable input
  - Long run
    - Two variable inputs [Iso quant Analysis]
    - Multi variable inpu[ts] [Returns to scale]

Production function with one variable inputs & Laws of returns

→ "The short run production function is otherwise called Single variable production function."

→ This is also called "Law of Return", "Law of Variabl[e] Proportions or" The law of Diminishing returns."

→ The law of production deals with the relation ship between additional inputs and additional outputs.

The law of returns states that when atleast one factor of production is varied and all other facto are fixed, the total output in the initial stages will increases at an increasing rate, and after reaching certain level of output, the total output will increase at declining rate.

If variable factor inputs are added furthur to the fixed factor input, the total output may decline.

This law is universal nature and it proved to be true in agriculture and Industry also.

The law states the relation ship between variable factors and output. How does output changes in that there are 3 stages.

This can be Explain with the help of the following table.

units of labour	Total product (TP)	Marginal product (MP)	Average product (AP)	Stages
1	5	5	5	I stag
2	12	7	6	
3	18	6	6	
4	20	2	5	II stag
5	20	0	4	
6	18	−2	3	III sta
7	14	−4	2	

1. Total Product:- It refers to the total amount of outpu

2. Average Product:- It refers to the product of each lab
If ~~de~~ we divide the total product by no. of labour.

$$A_p = \frac{\text{Total product}}{\text{No. of labour}}$$

3. Marginal product:- It refers to the additional produ obtained from the use of an additio labour

→ In the short run, It is assumed that capital is fixed factor input and labour is variable input. It is also assumed that technology is given and it is not going to change.

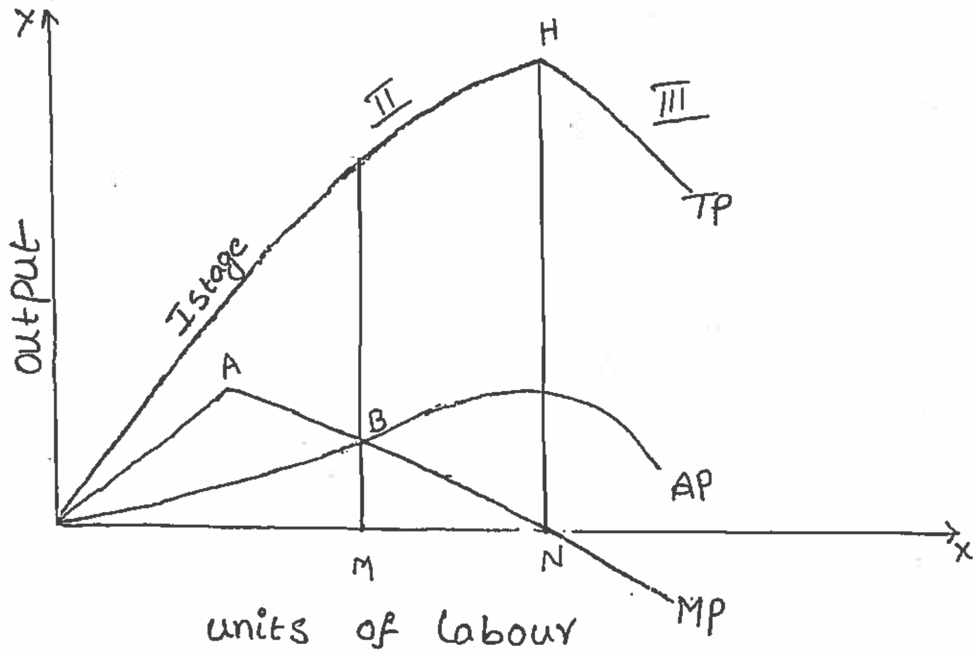
→ under such circumstances, the firm starts production with a fixed amount of capital and uses more and more units of labour.

→ At the third labourer, the production is continue to increase and (MP) marginal product and Average product (Ap) are equal. So this is the first stage, i.e it is called Increasing return to variable factors.

→ In the second stage, the total output is increased, after that the Marginal product (mp) and Average product(are gradually decreased and the marginal product (Mp) is reached zero at the 5th labour. So, this is called IInd stage i.e "diminishing return to variable factor."

In the third stage, The marginal product (MP) is goes into Negative (minus). The total product (TP) is also started to decrease. So it is called "Negative returns to variable factors."

Diagramatic representation of law:—



units of labour
(variable factor)

In the above diagram the variable factors labour is shown on the x-axis and the output is shown in y-axis.

MP is the marginal product curve.

AP is the Average product curve

TP is the total product curve.

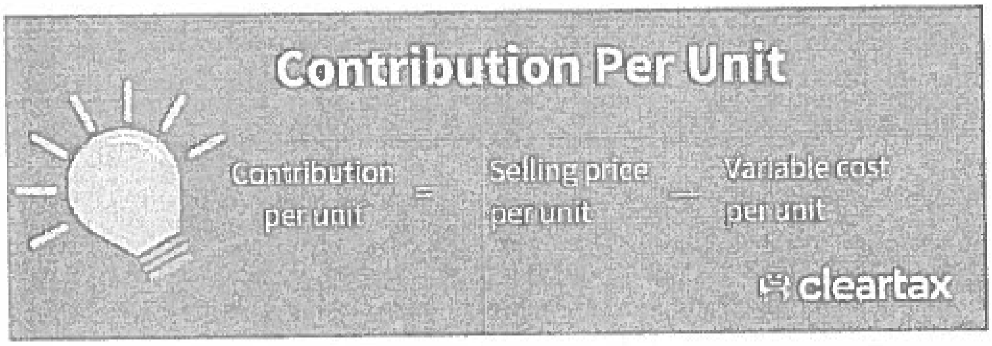
8.b. What is break even analysis? How do you determine breakeven point? Illustrate.

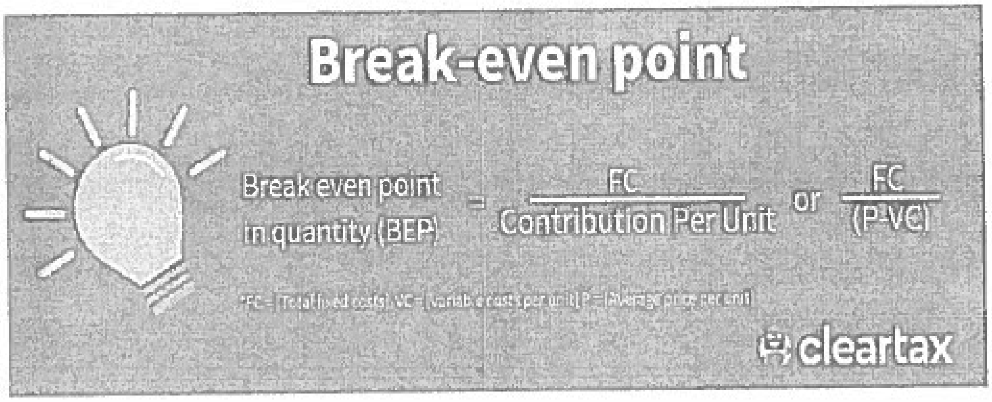
Answer:

A break-even analysis is a financial calculation that weighs the costs of a new business, service or product against the unit sell price to determine the point at which you will break even. In other words, it reveals the point at which you will have sold enough units to cover all of your costs. At that point, you will have neither lost money nor made a profit

CALCULATION OF BREAK-EVEN ANALYSIS

The basic formula for break-even analysis is derived by dividing the total fixed costs of production by the contribution per unit (price per unit less the variable costs).





9.a. Explain Cobb Douglas Production function.

Answer:

In economics and econometrics, the **Cobb–Douglas production function** is a particular functional form of the production function, widely used to represent the technological relationship between the amounts of two or more inputs (particularly physical capital and labor) and the amount of output that can be produced by those inputs. The Cobb–Douglas form was developed and tested against statistical evidence by Charles Cobb and Paul Douglas between 1927–1947; according to Douglas, the functional form itself was developed earlier by Philip Wick steed.

In its most standard form for production of a single good with two factors, the function is

$$P = b\, L^{a} C^{1-a}$$

Where:

- P = total production (the real value of all goods produced in a year or 365.25 days)
- L = labor input (person-hours worked in a year or 365.25 days)
- C = capital input (a measure of all machinery, equipment, and buildings; the value of capital input divided by the price of capital)
- A = total factor productivity
- α and β are the output elasticity's of capital and labor, respectively. These values are constants determined by available technology.

9.b. You are required to calculate

i) Margin of Safety
ii) Total sales
iii) Variable cost

Fixed costs Rs. 12,000, Profit Rs. 1,000, Break-Even Sales Rs.60,000

Answer:

[illegible]

10.a.

Particulars	Rs.	Particulars	Rs.
Opening stock	1,250	Plant and machinery	6,230
Sale	11,800	Returns outwards	1,380
Depreciation	667	Cash in hand	942
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The above balances extracted from the books of mythri & company, prepare a trading, profit and loss account and a balance sheet. The value of stock on 31st December, 1990 was Rs. 3,700.
Answer :

10.b. Write about short note on trading account and profit and loss account.

TRADING AND PROFIT AND LOSS ACCOUNT FORMAT

Trading and Profit and Loss Account format is represented separately as follows:

Format for Trading Account

Trading account for the year ended..........

Particulars			Particulars		
To opening stock		xxx	By Sales	xxxx	
To purchases	xxxx		Less returns	xx	
Less returns	xxx			--------	xxxx
	------------	xxxx	By closing stock		xxx
To Direct expenses:			By gross loss (if loss)		xxx
Carriage inward		xxx			
Freight		xxx			
Octroi		xxx			
Dock dues		xxx			
Excise duty		xxx			
Royalty		xxx			
Motive power		xx			
Coal, gas, water		xxx			
Factory expenses		xxx			
To Gross Profit (if profit)		xxx			
		xxxxx			xxxxx

Format for Profit and Loss Account

Profit & Loss Account

(For the year ended...)

Dr. Particulars	Amount	Particulars	Cr. Amount
To Gross loss b/d	Xxx	By Gross Profit b/d	Xxx
To Salaries	Xxx	By Discount Received	Xxx
To Office rent, rates and taxes	Xxx	By Commission Received	Xxx
To Printing & stationery	Xxx	By Bank Interest	Xxx
To Telephone expenses	Xxx	By Rent received	Xxx
To Postage & telegram	Xxx	By Dividend on shares	Xxx
To Discount Allowed	Xxx	By Interest earned on debentures	Xxx
To Insurance	Xxx	By Profit on sale of asset	Xxx
To Audit Fees	Xxx	By Net loss	Xxx
To Electricity charges	Xxx		
To Repairs & renewals	Xxx		
To Depreciation	Xxx		
To Advertisement	Xxx		
To Carriage Outwards	Xxx		
To Bad Debts	Xxx		
To Provision for Bad debts	Xxx		
To Selling commission	Xxx		
To Bank Charges	Xxx		
To Interest on loans	Xxx		
To Loss on sale of asset	Xxx		
To Net Profit	Xxx		
	xxx		xxx

11.a. Explain the concepts of journal and ledger accounts with performa.

WHAT IS A JOURNAL?

A journal is a subsidiary book of account that records monetary transactions according to accounting standards. These transactions get recorded in chronological order, and it gives details about the accounts that are affected by each transaction. It is known as the first step of the accounting process.

Date	Particulars	Ledger Folio	Debit Amount	Credit amount

WHAT IS A LEDGER?

A Ledger is a principal book of account, and its primary purpose is to transfer transactions from a journal and then classify it into separate accounts. Ledger is also known as the book of final entry as it helps businesses prepare accounting statements like the Trial Balance.

Debit					**Credit**
Date	Particulars	Amount	Date	Particulars	Amount

11.b. Journalizing the following transactions:

Jan 1 Started business with cash Rs.100,000
2 Deposited Rs.75,000 to bank
3 Purchased furniture Rs.20,000 and paid by cheque (Through Bank)
5 Paid shop rent Rs.2,500 cash
7 withdrew from bank for personal use Rs.1,000
8 Sold goods on credit Rs.6,000 to Jithesh
10 Received interest from bank Rs.600

Journal entries in the books of

Journal entries

Date	Particulars	L.f	Debit amount (Rs)	Credit amount (Rs)
Jan 1	Cash A/c -- Dr To Capital A/c (Being Business Commenced)		1,00,000	1,00000
Jan 2	Bank A/c -- Dr To Cash A/c (Being Cash deposited into Bank)		75,000	75,000
Jan 3rd	Furniture A/c Dr To Bank A/c (Being furniture purchased Paid by cheque)		20,000	20,000
Jan 5th	Shop rent A/c Dr To Cash (Being rent Paid)		2,500	2,500
Jan 7th	Drawings A/c -- Dr To Bank A/c (Being Cash withdrawn from bank for personal use		1,000	1,000
Jan 8th	Jithesh A/c Dr To Sales (Being goods Sold on Credit)		6,000	6,000
Jan 10th	Interest A/c Dr To Bank A/c		600	600

12.a. What do you mean by payback period method? Explain the merits and demerits of payback period method.

ANSWER:

PAYBACK PERIOD METHOD

This method is also known as pay out, pay off or recoupment period method. Under this method, the original investment of the project should be received back out of the implementation of the project as early as possible. It means that the company gets additional earnings or savings if the project is implemented. Thus, it measures the period of time for the original cost of a project to be recovered from the additional earnings or savings of the project itself. When the total cash inflows from investments equals the total outlay, then the period is the pay back period of that project. Cash inflows should be calculated to find the pay back period. The term cash inflows refers to annual net earnings (profit) before depreciation but after taxes.

Accept or Reject Criteria of payback period

A project may be accepted or rejected on the basis of the per-determined (standard) pay back period if only one independent project is to be evaluated. The standard pay back period is determined by the management in terms of maximum period during which initial investment must be recovered.

Payback period = (Initial Investment or Original Cost of the Asset / Cash Inflows)

If cash inflows uneven, cumulative cash inflow statement is prepared and the following formula is used.

P = PYFR + (BA / CIYER)

MERITS OR ADVANTAGES OF PAYBACK PERIOD METHOD

The chief merits of the payback period are briefly presented below.

1. It is very simple to understand and easy to calculate.
2. It requires less cost, time and labour when compared to other methods of capital budgeting.
3. This method reduces or avoids the loss through obsolescence since shorter payback period is preferred to longer payback period.
4. This method is mostly suitable to a company which has less amount of cash in hand and a company whose liquidity position is very weak.
5. It gives much importance to the speedy recovery of investment in capital assets.

DEMERITS / LIMITATIONS / DISADVANTAGES OF PAYBACK PERIOD

The payback period method has some limitations. They are given below:

1. A slight change made in the labour cost or cost of maintenance, there is a much change in its earnings and affects the payback period.
2. This method ignores the short term solvency or liquidity of the business concern.
3. It ignores capital wastage and economic life by restricting consideration to the project's gross earnings.
4. The time value of money is not considered in the payback period method.
5. It overlooks the cost of capital which is a main factor in sound capital budgeting decision. This method does not consider the cash inflows arising after the payback period.
6. This could be misleading in capital budgeting decisions.
7. This method fails to measure the productivity of capital expenditure plan because it does not attempt to measure the return on investment.
8. This method does not consider full earnings or full savings of the capital expenditure plan i.e. savings or earnings available during whole economic life of the project.
9. This method also fails to assign proper weightage to the unevenness of rate of profit of various projects.
10. It may be difficult to determine minimum acceptable payback period. Generally, it is a subjective decision.
11. This method treats the each asset individually in isolation with other assets. But, in practice, it is not feasible.

12.b. Solve the payback period of the following projects each requiring a cash outlay of Rs 1,00,000 each.

Year	Cash Inflows Rs.	
	Project A	Project B
1	30,000	30,000
2	30,000	40,000
3	30,000	20,000
4	30,000	25,000
5	30,000	5,000

Answer:

Initial Investment = 1,00,000.

12 b) Pay back Period

Year	Project A	Cumulative Cash flows
1	30,000	30,000
2	30,000	60,000
3	30,000	90,000
4	30,000	1,20,000
5	30,000	1,50,000

Pay back period lies between 3rd & 4th year.

$3 + \frac{1,20,000 - 90,000}{30,000}$

$\Rightarrow 3 + 0.3 = 3.3$ yrs.

Project B Year	Cash flows	Cumulative Cash flows
1	30,000	30,000
2	40,000	70,000
3	20,000	90,000
4	25,000	1,15,000
5	5,000	1,20,000

Pay back lies between 3rd & 4th year.

$3 + \frac{1,00,000 - 90,000}{25,000 \text{ (Difference)}}$

$3 + 0.4 \Rightarrow 3.4$ years.

the Pay back Period of Project A = 3.3 years
Project B = 3.4 years

13.a. What is capital? Explain the types and significance.

WHAT IS CAPITAL?

Nic Barnhart of Pareto Labs defines capital as simply, "Money that is used to make more money." This definition can apply to individuals in the greater economy and to companies. In the world of business, the term *capital* means anything a business owns that contributes to building wealth.

Sources of capital include:

- **Financial assets** that can be liquidated like cash, cash equivalents, and marketable securities.
- **Tangible assets** such as the machines and facilities used to make a product.
- **Human capital**; i.e. the people that work to produce goods and services.
- **Brand capital**; i.e. the perceived value of a brand recognition.

TOP 4 TYPES OF CAPITAL FOR BUSINESS

There are four common ways that businesses gather capital, whether it is to fund the company to launch or to help the company through a growth period. Working capital and debt and equity capital are sources of capital for any business, but trading capital is only found in companies in the financial space.

1. Working capital

Working capital—the difference between a company's assets and liabilities—measures a company's ability to produce cash to pay for its short term financial obligations, also known as liquidity.

Working capital = Current assets – Current liabilities

Positive working capital means the value of a company's current assets is more than its current liabilities Negative working capital, on the other hand, means that current liabilities outweigh current assets. For the company, this could lead to financial issues with creditors, growth, or production.

2. Debt capital

Debt capital is acquired by borrowing from financial institutions, banks, friends and family, credit cards, federal loan programs, and venture capital, or by issuing bonds. Just like an individual needs established credit history to borrow, so do businesses.

Debt capital has to be paid off on a regular basis (with interest) but unlike an individual's debt, it is seen as more of an essential part of building a business instead of a financial burden.

3. Equity capital

Equity capital is any capital raised through selling shares with a key difference being whether those shares are sold privately or publicly:

- **Private:** Shares of stock in a company within a private group of investors.
- **Public:** Shares of stock in a company that are listed on the stock exchange (think: IPO).

The money an investor pays for shares of stock in a company becomes equity capital for the business.

4. Trading capital

Trading capital applies exclusively to the financial industry where brokerage companies need enough capital to support their investment strategies. Trading capital supports the many daily trades that brokerage companies need to make to generate a profit and the large-scale trades made by the biggest brokerage firms. Sometimes it is granted to individual traders and sometimes to the firm as a whole.

13b)

Year	Machine 1	P.V @ 10%	Present value of future Cash flows
1	1,50,000	0.909	1,36,350
2	3,00,000	0.826	2,47,800
3	1,50,000	0.751	112650
4	—	0.683	-
			4,96,800

NPV = P.V. of future Cash flows − initial Investment.

4,96,800 − 3,00,000

1,96,800.

Year	Machine 2 Cash flows	P.V @ 10%	Present Value of Cash flows
1	2,00,000	0.909	18,180
2	3,00,000	0.826	2,47,800
3	2,50,000	0.751	1,87,750
4	1,50,000	0.683	1,02,450
			5,56,180.

NPV = 5,56,180 − 3,00,000

2,56,180

Hence Accept Machine 2 as it provides higher NPV.

14 a) Debt equity

$\frac{\text{Debt}}{\text{Equity}} \Rightarrow \frac{1,40,000 + 70,000 + \cancel{1,40000} + 66,000 + 14,000 + \cancel{1,20,000}}{} = 2,90,000.$

$\frac{2,90,000}{1,20,000 \text{ [share capital]}}$

~~2.416~~ // or 2.42

Debt {Debentures + long term loans + Creditors + B/P}

$$\text{Current ratio} = \frac{\text{Current Assets}}{\text{Current liabilities}} = \frac{\text{Bank Balance + debtors}}{\text{Creditors + B/P}}$$

$$\Rightarrow \frac{1,00,000}{66,000 + 14,000}$$

$$80,000.$$

$$\frac{\cancel{1,00,000}^{5}}{\cancel{80,000}_{4}} \Rightarrow \frac{5}{4}$$

$$\Rightarrow \underline{\underline{1.25}}$$

Quick ratio $= \boxed{\dfrac{\text{Quick } \cancel{\text{liabilities}}}{\text{Current Assets}}} =$

$$\frac{\text{Quick Assets}}{\text{Quick liabilities}} = \frac{1,00,000}{80,000} \Rightarrow \underline{\underline{1.25}}$$

(159) Interest Coverage Ratio =

$$\frac{\text{Net Profit before Interest \& taxes}}{\text{Interest charges.}}$$

Net Profit after interest & taxes = 6,00,000.

Add interest taxes = 1,20,000

7,20,000.

$$\text{Interest charges} = 15,00,000 \times \frac{12}{100}$$

$$1,80,000.$$

$$\text{I.C.R} \Rightarrow \frac{7,20,000}{1,80,000} = \underline{\underline{5,40,000}}$$

14.b. What do you mean by accounting ratios? How are they useful?

Answer:

Accounting ratios are an important tool for analysing financial statements. It is a comparison of two or more financial data that is used to analyse a company's financial statements. These depict a connection between two or more accounting numbers obtained from financial statements. It is a useful tool for shareholders, creditors, and other stakeholders to understand a company's profitability, strength, and financial health. This is also known as financial ratios, which are used to track corporate performance and make key business choices.

TYPES OF ACCOUNTING RATIOS

- Liquidity Ratios
- Profitability Ratios
- Solvency Ratios
- Activity or Efficiency Ratio

Liquidity Ratios

The liquidity ratio is used to determine whether or not a company has enough cash on hand to pay down its short-term debts. A high liquidity ratio indicates that the corporation will be able to pay its creditors. It is allowed to have a liquid ratio of 2 or more.

Ratio	Formula	Objective
Current Ratio: Current assets include cash, inventory, accounts receivable or debtors, interest receivable, etc Current liabilities include accounts payable or creditors, income tax payable and any other current liabilities	{(Current Assets)/(Current Liabilities)}	This is the most widely used liquidity ratio for comparing a company's current assets to its current liabilities. The current ratio can be used to determine whether or not a company will be able to pay its debts in the next twelve months.
Quick Ratio: Quick assets excludes assets such as inventory and prepaid expenses which are difficult to liquidate quickly.	{(Quick Assets)/(Current Liabilities)}	Acid test is another name for Quick Ratio. The quick ratio is a more cautious approach to determining a company's short-term solvency. It solely comprises the company's quick assets, which are its most liquid assets.
Cash Ratio: The cash ratio is a ratio that compares a company's total cash and cash equivalents to its current liabilities. This metric represents a company's ability to meet short-term debt obligations with its most liquid assets.	{(Cash + Marketable securities)/(Current Liabilities)}	This ratio converts current assets into an account that is immediately available to a company in order to pay its liabilities. Any company with a Cash Ratio of one or greater is regarded as financially sound.

Profitability Ratio

Profitability ratios are a group of financial indicators that are used to evaluate a company's ability to create earnings over time in relation to its revenue, operational costs, assets, or shareholders' equity. The evaluation is done by utilising financial information from a certain point in time. Efficiency ratios assess how successfully a corporation uses its assets internally to generate income. These efficiency ratios can be compared to profitability ratios (as opposed to after-cost profits).

Higher ratio outcomes are often more beneficial. However, such a ratio outcome must be compared to

- The results of similar companies
- The company's own previous performance
- Industry average

Ratio	Formula	Objective
Gross Profit Margin: Revenue is the income from sale of goods or services Cost of goods sold, as the name suggests, is the cost that a company incurs to produce the goods that it sold. COGS includes raw materials, processing cost, labour, and other production expenses.	{(Revenue – Cost of Goods Sold (COGS))/(Revenue)} Gross Profit = Revenue – Cost of Goods Sold	Using the Gross Profit Ratio, any company can compare its performance to that of its competitors or to that of its own historical performance. The gross profit ratio expresses the proportion of factory costs to sales revenue.A higher gross profit margin shows that a company's operations are more efficient. The Gross Profit Margin compares a company's gross profit to its sales revenue. This margin shows how much money a company makes after all of the costs of producing goods and services have been deducted.
Operating Margin: Operating income is also known as EBIT earning before interest and taxes. EBIT, or operational earnings = Revenue minus cost of goods sold (COGS) and normal selling, general, and administrative costs of running a firm, excluding interest and taxes,	{(Gross Profits- Operating Expense)/(Revenue)}	The operating margin quantifies how much profit a company generates on a dollar of sales after paying for variable expenses. Such variable expenses include production expenses, wages and raw materials, but before paying interest or taxes. Higher ratios indicate that a company's operations are efficient and that it is good at converting revenues into profits. Unlike Gross Profit Ratio, this includes more expenses and is thus used to more efficiently determine a company's profitability.
Profit Margin	{(Revenue – Operating expense + non-operating	Any business can determine the amount of profit gained from its entire generated

	income-Interest Expense-Income taxes)/(Revenue)}	revenue using the Profit Margin ratio. A company's overall profitability may be easily assessed and compared to that of its competitors.
Earnings per Share (EPS): The net profit of a corporation is divided by the number of common shares it has outstanding to calculate earnings per share (EPS) Usually, weighted average number of outstanding shares is considered to calculate. This is because the company issues shares during the year. Moreover, Diluted EPS covers options, convertible securities and warrants outstanding which affects outstanding shares.	{(Net Income – Preferred Dividend)/(Weighted Average Outstanding Shares)}	EPS is a widely used indicator for measuring corporate value since it shows how much money a firm produces for each share of its stock.Investors will pay more for a company's shares if they believe the company's profits are higher than its share price, so a higher EPS signals more value. The higher a company's earnings per share (EPS), the more profitable it is deemed to be.

Solvency Ratios

A solvency ratio is a crucial metric used by prospective business lenders to assess an organisation's capacity to satisfy long-term debt obligations. A solvency ratio is a measure of a company's financial health that determines if its cash flow is sufficient to cover its long-term liabilities. An unfavourable ratio can suggest that a corporation is at risk of defaulting on its debt obligations. Solvency ratios are frequently utilised by prospective lenders and bond investors when evaluating a company's creditworthiness. Although both solvency and liquidity ratios are used to assess a company's financial health, solvency ratios have a longer-term outlook than liquidity ratios.

Ratio	Formula	Objective
Debt Equity Ratio or D/E ratio: Debt includes long term and short term debt obligations Equity includes the shareholder's capital i.e. value of outstanding shares plus reserves	{(Total Debt)/(Total Equity)}	The D/E ratio is similar to the debt-to-assets ratio in that it shows how debt is used to fund a company. The higher the ratio, the more debt a business has on its books, and the greater the risk of default. The debt-to-equity ratio examines how much of the debt can be covered by equity in the event of a liquidation. This is also known as the gearing ratio. It is used by creditors and investors to assess a company's financial leverage.
Debt to Asset Ratio: Debt includes long term and short term debt obligations	{(Total Debt)/(Total Asset)}	The debt-to-assets ratio compares the overall debt of a corporation to its total assets. It calculates a firm's leverage and shows how much of the company is

Total assets is the total assets for the period as reflected in the balance sheet.		funded by debt vs assets. It also measures the company's ability to repay debt with available assets. A higher ratio, particularly one above 1.0, suggests that a corporation is heavily reliant on debt and may struggle to satisfy its obligations.
Debt Ratio	{(Total Liabilities)/(Total Asset)}	A debt ratio is a measurement of a company's indebtedness in terms of total debt to total assets.The debt ratio varies greatly by industry, with capital-intensive enterprises having substantially greater debt ratios than others.A debt ratio more than 1.0, or 100 percent, shows that a company's debt exceeds its assets. On the other hand, a debt ratio less than 100 percent implies that the company's assets exceed its debt.
Interest Coverage Ratio: EBIT, or operational earnings = Revenue minus cost of goods sold (COGS) and normal selling, general, and administrative costs of running a firm, excluding interest and taxes,	{(Earnings before interest and taxes (EBIT))/(Interest Expense)}	The <u>interest coverage ratio</u> determines how many times a company's available earnings can cover its existing interest payments. In other words, it calculates a company's margin of safety for paying interest on its debt over a specific time period. It is preferable to have a larger ratio. If the ratio falls below 1.5, it may suggest that a corporation will have trouble paying its obligations' interest.

Activity or Efficiency Ratio

Activity ratio determines the efficiency by which a company is utilizing its assets to generate revenue and cash or bank balance. In other words, it calculates a company's margin of safety for paying interest on its debt over a specific time period.

Analysts can use activity ratios to assess a company's inventory management, which is critical to its operational flexibility and overall financial health. An activity ratio is a financial indicator that investors and research analysts use to determine how well a firm uses its assets to create revenue and cash.

Activity ratios can be used to compare two organizations in the same industry, or they can be used to track the financial health of a single company over time.

Ratio	Formula	Objective
Accounts Receivable Ratio	{(Annual Sales Credit) / (Accounts Receivable)}	The ability of a business to collect money from its clients is determined by the accounts receivable turnover ratio. For a given period, total credit sales are divided by the average accounts receivable balance. A low ratio indicates a problem with the collection procedure. A high receivables turnover ratio may suggest that a company's accounts receivable collection is effective and that the company has a

		large number of high-quality customers who pay their bills on time.Inefficient collection, poor credit policies, or clients that are not financially viable or creditworthy could all contribute to a low receivables turnover percentage.
Inventory Turnover Ratio	{(Cost of Goods Sold) / (Average Inventory)}	The inventory turnover ratio determines how frequently the inventory balance is sold over the course of a financial year. The average inventory for a given period is divided by the cost of items sold. Higher estimations indicate that a company's inventory can be moved with relative ease.
Asset Turnover Ratio	{(Net Revenue)/(Assets)}	The assets turnover ratio is a metric that assesses how effectively a company utilises its assets to make a sale. Total revenues are divided by total assets to determine how well a company uses its resources. Smaller ratios could suggest that a corporation is having difficulty moving its goods.

Course Instructor

NSRIT

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	Program	CE		Academic Year	2021 - 2022
Course Code	20CE402	Test Duration	3 Hrs.	Max. Marks 70	Semester	IV
Course	HYDRAULICS & HYDRAULIC MACHINERY					

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What is the condition for Reynold's Number for the case of laminar flow in open channels	20CE405.1	L1
2	What are the different dimensionless numbers?	20CE405.2	L2
3	A jet of water strikes with a velocity of 40 m/s a flat plate inclined at 300 with the axis of the jet. If the cross sectional area of the jet is 25 cm² determine the force exerted by the jet on the plate.	20CE405.3	L2
4	Classify different types of turbines according to discharge.	20CE405.4	L2
5	What are various components of reciprocating pump?	20CE405.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Describe the different types of flow in open channels	12M	20CE405.1	L2
	OR			
7 (a)	Determine the economical cross-section for an open channel of trapezoidal section with side slopes of 1 vertical to 2 horizontal, to carry 10 m³ /s, the bed slope being 1/2000. Assume Manning coefficient as 0.022.	8M	20CE405.1	L2
7 (b)	Differentiate between uniform and non uniform flow.	4M	20CE405.1	L2
8 (a)	What are similarities between model and prototype?	4M	20CE405.2	L3
8 (b)	A spillway model is constructed on a scale of 1:25. Calculate (i) the prototype discharge corresponding to model discharge of 0.12 m³/sec (ii) the velocity in model corresponding to prototype velocity of 3.5 m/s.	8M	20CE405.2	L3
	OR			
9 (a)	Write in detail about Geometric, Kinematic and Dynamic Similarities.	6M	20CE405.2	L3
9 (b)	What do you mean by dimensionless numbers? Name any four dimensionless numbers. Define and explain Reynolds's number and Froude's number.	6M	20CE405.2	L3
10 (a)	A jet of water 50 mm in diameter and moving with a velocity of 26 m/s is impinging normally on a plate. Determine the pressure on the plate when it is fixed and when it is moving with a velocity of 10 m/s in the direction of the jet. Also determine the work done per second by the jet.	~~10M~~ 7	20CE405.3	L3
10 (b)	A jet of water strikes with a velocity of 40 m/s a flat plate inclined at 300 with the axis of the jet. If the cross sectional area of the jet is 25 cm² determine the force exerted by the jet on the plate.	5M	20CE405.3	L3
	OR			
11 (a)	A jet of water of 60 mm diameter strikes a curved vane at its centre with a velocity of 18 m/s. The curved vane is moving with a velocity of 6 m/s in the direction of the jet. The jet is deflected through an angle of 165°. Assuming the plate to be smooth find: (i) Thrust on the plate in the direction of jet, (ii) Power of the jet, and (iii) Efficiency of the jet.	6M	20CE405.3	L3
11 (b)	Explain about Angular momentum principle.	6M	20CE405.3	L3
12	Write in detail about a hydropower installation	12M	20CE405.4	L2
	OR			
13 (a)	A pelton wheel has to be designed for the following data: power to be developed =6000 kW, Net head available = 400 m, speed = 550 rpm, Ratio of jet diameter to the wheel diameter = 1/10 and overall efficiency = 85%. Find the number of jets, diameter of jet, diameter of the wheel and quantity of water required.	8M	20CE405.4	L2
13 (b)	Write short note on Francis turbine.	4M	20CE405.4	L2

14 (a)	Define centrifugal pump and explain the working procedure of a single-stage centrifugal pump with neat sketch.	7M	20CE405.5	L3
14 (b)	A centrifugal pump rotating at 1080 rpm delivers 168 liters/s of water against a head of 30 m. The pump is installed at a place where atmospheric pressure is 1×10^5 Pa(abs.) and vapour pressure of water is 2 kPa (abs.). The head loss is suction pipe is equivalent to 0.2 m of water. Calculate minimum NPSH.	5M	20CE405.5	L3
	OR			
15 (a)	Write in detail about Kaplan turbine.	8M	20CE405.5	L3
15 (b)	Write about cavitation in pumps.	4M	20CE405.5	L3

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM, ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

PART-A

1. What is the condition for Reynold's Number for the case of laminar flow in open channels?

   Ans)

   - If the Reynolds number Re is less than 500 or 600, then the flow is called laminar flow.
   - If the Reynolds number is more than 2000, then the flow is said to be turbulent.

2. What are the different dimensionless numbers?

   Ans) Different dimensionless numbers are:

   - Reynolds Number.
   - Froude Number.
   - Weber Number.
   - Mach Number.
   - Euler's Number

3. A jet of water strikes with a velocity of 40 m/s a flat plate inclined at 30^0 with the axis of the jet. If the cross-sectional area of the jet is 25 cm^2 determine the force exerted by the jet on the plate.

   Ans) $F_n = \rho a V^2 \sin\theta$

   $$F_n = 1000 \times (25/100^{\wedge}2) \times 40^2 \sin 30^0 = 2000\ N = 2kN$$

4. Classify different types of turbines according to discharge.

   Ans) Types of turbines according to discharge

   - Pelton turbine
   - Francis turbine
   - Kaplan turbine

5. What are various components of reciprocating pump?

   Ans) The main components of reciprocating pump are as follows:

   - Suction Pipe.
   - Suction Valve.
   - Delivery Pipe.
   - Delivery Valve.
   - Cylinder.
   - Piston and Piston Rod.
   - Crank and Connecting Rod.
   - Strainer.

PART-B

6. Describe the different types of flow in open channels

Ans)

- The flow of liquid through a channel with a free surface is defined as open channel flow. This free surface of the liquid is subjected to atmospheric pressure.
- The flow in an open channel takes place due to gravity that is achieved by providing a bed slope.
- The flow of liquid through the open channel can be of several types like steady and unsteady flow, laminar or turbulent flow or uniform or non-uniform flow and finally sub-critical, critical and supercritical flow.

Types of Flow in Open Channels

1. Steady and Unsteady Flow
   - In an open channel flow, if the flow parameters such as depth of flow, the velocity of flow and the rate of flow at a particular point on the fluid do not change with respect to time, then it is called as steady flow.
   - And is at any point on the open channel flow, the flow parameters like depth of flow, the velocity of flow and rate of flow do change their value with respect to time, then it is called as an unsteady flow.
2. Uniform Flow and Non-Uniform Flow
   - The flow in the channel is said to be uniform, if, for a given length of the channel, the velocity of flow, the depth of flow remains constant
   - In a Non-uniform flow, the flow parameters like velocity, depth of flow, etc do not remain constant for a given length of the channel.
   - The Non-uniform flow can be again defined as Rapidly varying flow (R.V.F) and Gradually Varied Flow (G.V.F). In the case of R.V.F, the depth of flow rapidly changes over a smaller length of the channel. It rises up suddenly for a short length and settles back. While in a G.V.F, the depth of flow changes gradually over a longer length of the channel.
3. Laminar Flow and Turbulent Flow
   - Laminar and turbulent flow in open channel flow is defined based on the Reynolds Number, Re.
   - If the Reynolds number Re is less than 500 or 600, then the flow is called laminar flow.
   - If the Reynolds number is more than 2000, then the flow is said to be turbulent.
   - A flow that has Reynolds number between 500 and 2000 is said to be in the transition state
4. Critical, Sub-Critical and Super - Critical Flow
   - The open channel flow is categorized as critical or sub-critical or super-critical based on the Froude number F_e. Froude number is given by the relation:
   - Open channel flow is Sub-critical if the Froude number is less than 1. Sub-Critical open channel flow is also defined as a tranquil or streaming flow.
   - An open channel flow with a Froude number equal to one is a critical flow.
   - And super-critical flow in open channel has a Froude number greater than 1. A supercritical flow is also termed as rapid flow or torrential flow or shooting flow

7. a) Determine the economical cross-section for an open channel of trapezoidal section with side slopes of 1 vertical to 2 horizontal, to carry 10 m^3/s, the bed slope being 1/2000. Assume Manning coefficient as 0.022.

Ans) Given data, Q =10 m^3/s; s = 1/2000; n = 0.022

Area of trapezoidal section is

For most economical trapezoidal section

Using manning's formula, discharge

Economical cross-section for an open channel of trapezoidal section is b =1.0 m & y = 2.0 m

7. b) Differentiate between uniform and non-uniform flow.

Ans) Types of fluid flow:

According to different considerations fluid flows may be classified in several ways as indicated below:

1. Steady flow and Unsteady flow.
2. Uniform flow and non-uniform flow.
3. One-dimensional flow
4. Two-dimensional flow
5. Three-dimensional flow
6. Rotational flow and Irrotational flow
7. Laminar flow and turbulent flow.

Difference between uniform and non-uniform flow

Uniform Flow

- When the velocity of flow of fluid does not change, both in magnitude and direction, from point to point in the flowing fluid, for any given instant of time, the flow is said to be uniform.
- In the mathematical form a uniform flow may therefore be expressed as

$$\left(\frac{\partial V}{\partial s}\right) = 0$$

Non-uniform Flow

- If the velocity of flow of fluid changes from point to point in the flowing fluid at any instant, the flow is said to be non-uniform.
- In the mathematical form a non-uniform flow may be expressed as

$$\left(\frac{\partial V}{\partial s}\right) \neq 0$$

8. a) What are similarities between model and prototype?

Ans)

- Three types of similarities must exists between the model and prototype.
- They are:
    - Geometric similarity
    - Kinemetic similarity
    - Dynamic similarity

8. b) A spillway model is constructed on a scale of 1:25. Calculate (i) the prototype discharge corresponding to model discharge of 0.12 m³/sec (ii) the velocity in model corresponding to prototype velocity of 3.5 m/s.

Ans)

8b) Given data,

Scale ratio of length $L_r = 25$ ($\because$ Scale 1:25)

Discharge in model $Q_m = 0.12\, m^3/s$

Velocity in prototype $V_p = 3.5\, m/s$

i) Prototype discharge (Q_p):

Using discharge ratio $\frac{Q_p}{Q_m} = L_r^{2.5}$

$$Q_p = Q_m L_r^{2.5} = 0.12 \times (25)^{2.5}$$
$$= 375\, m^3/s$$

ii) Velocity in model (V_m):

Using velocity ratio, $\frac{V_p}{V_m} = \sqrt{L_r}$

$$V_m = \frac{V_p}{\sqrt{L_r}} = \frac{3.5}{\sqrt{25}} = 0.7\, m/s$$

9. a) Write in detail about Geometric, Kinematic and Dynamic Similarities.

Ans) **Geometric Similarity:**

- The geometric similarity is said to exist between the model and the prototype.
- The ratio of all corresponding linear dimension in the model and prototype are equal.
- For geometric similarity between model and prototype we must have the relation

$$\frac{L_P}{L_m} = \frac{b_P}{b_m} = \frac{D_P}{D_m} = L_r$$

L_m = length of model L_p = length of prototype

b_m = Breadth of model b_p = breadth of prototype

D_m = Diameter of model D_p = Diameter of prototype

L_r = scale ratio

For area's ratio and volume's ratio the relation should be given below:-

Area ratio $\frac{A_P}{A_m} = \frac{L_P \times b_P}{L_m \times b_m} = L_x \times L_r = L_r^2$

Volume ratio $\frac{forall_P}{V_m} = (\frac{L_P}{L_m})^3 = (\frac{b_P}{b_m})^3 = (\frac{D_P}{D_m})^3$

Kinematic Similarity:

- Kinematic similarity means the similar of motion between model and prototype. Thus, kinematic similarity is said to exist between the model and the prototype if the ratios of the velocity and acceleration at the corresponding points in in the prototype are the same.
- Since velocity and acceleration are vector quantities; hence not only the ratio of magnitude of velocity and acceleration at the corresponding points in the model and prototype also should be parallel.
- For kinematic similarity, we must have

$$\frac{V_{p1}}{V_{m1}} = \frac{V_{p2}}{V_{m2}} = V_r(velocityratio)$$

For acceleration

$$\frac{ap_1}{am_1} = \frac{ap_2}{am_2} = a_r$$

Dynamic similarity:

- Dynamic similarity means the similar of forces between the model and prototype.
- Thus, dynamic similarity is said to exist between the model and the prototype if the ratios of the corresponding forces acting at the corresponding points are equal.
- Also, the directions of the corresponding forces at the corresponding points should be same.
- For dynamic similarity we have

$$\frac{(Fi)_p}{(Fi)_m} = \frac{(Fv)_p}{(Fv)_m} = \frac{(Fg)_P}{(Fg)_m} = F_r[Forceratio]$$

9. b) What do you mean by dimensionless numbers? Name any four dimensionless numbers. Define and explain Reynolds's number and Froude's number.

Ans)

- A dimensionless number is obtained by dividing the inertia force by viscous force or gravity force or pressure force or surface tension force or elastic force.
- As it is a ratio of one force to the other force, it has no dimensions, i.e. dimensionless.
- Some important dimensionless numbers which are used in model analysis of hydraulic structures and machines are given below:

1. Reynold's number
2. Froude's number
3. Weber number
4. Euler number
5. Mach number

Reynold Number (R_e)

In fluid mechanics, the Reynolds number R_e is a dimensionless number that gives a measure of the ratio of inertial forces to viscous forces and consequently quantifies the relative importance of these two types of forces for given flow conditions.

In the flow situations where the viscous forces plays an important, Reynolds number is taken as the criterion of dynamic similarity. Examples are as follows:

- Incompressible flow through small diameter pipes,
- Objects moving completely under water,
- Air movement under low velocity around airplanes and automobiles
- Open channel flow.

Froude Number (F_r)

Froude number (F_r)) is the ratio of the square root of the inertia force to the square root of the force due to gravity.

In flow situations where gravitational force is more important, Froude number governs the dynamic similarity. Other forces are comparatively small and negligible. Examples are:

- Flow through open channels
- Flow of liquid jets from orifices
- Flow over notches and weirs
- Flow over the spillway of a dam

10. a) A jet of water 50 mm in diameter and moving with a velocity of 26 m/s is impinging normally on a plate. Determine the pressure on the plate when it is fixed and when it is moving with a velocity of 10 m/s in the direction of the jet. Also determine the work done per second by the jet.

Ans) Given data,

Diameter of the jet, $d = 50$ mm $= 50 \times 10^{-3}$ m

Velocity of the jet , $V = 26$ m/s

Velocity of the plate, $u = 10$ m/s

i) Force exerted by the jet on the plate, when it is fixed

ii) Force exerted by the jet on the flat moving plate,

502.65 N

iii) Work done by the jet on the plate per second,

N-m/s

10. b) A jet of water strikes with a velocity of 40 m/s a flat plate inclined at 30^0 with the axis of the jet. If the cross-sectional area of the jet is 25 cm^2 determine the force exerted by the jet on the plate.

Ans) Given data,

cross-sectional area of the jet, $a = 25\ cm^2 = 2.5 \times 10^{-3}\ m^2$

Velocity of the jet, V = 40 m/s

Inclination of plate = 30^0

Force exerted by the jet on the plate is :

11. a) A jet of water of 60 mm diameter strikes a curved vane at its centre with a velocity of 18 m/s. The curved vane is moving with a velocity of 6 m/s in the direction of the jet. The jet is deflected through an angle of 165°. Assuming the plate to be smooth find: (i) Thrust on the plate in the direction of jet, (ii) Power of the jet, and (iii) Efficiency of the jet.

Ans) Given data,

Diameter of the jet, d = 60 mm = 0.06 m

Absolute velocity of the jet, V = 18 m/s

Angle of deflection of the jet, $(180-\theta) = 165^0$, $\theta = 180 - 165 = 15^0$

Velocity of the plate, u = 6 m/s

(i) Force exerted by the jet on the plate in the direction of jet,

$F_x = \rho a(V-u)^2[1 + Cos\theta]$

$= 1000 \times \Pi/4 \times (0.06)^2 \times (18 - 6)^2 (1 + cos15) = 800.43\ N$

(ii) Power of the jet

Workdone by the jet on the plate per second,

$W = F_x \times u = 800.43 \times 6 = 4802.58\ Nm/s$

Power of the jet, P = Workdone by the jet per second

= 4802.58 Watts = 4.802 kW

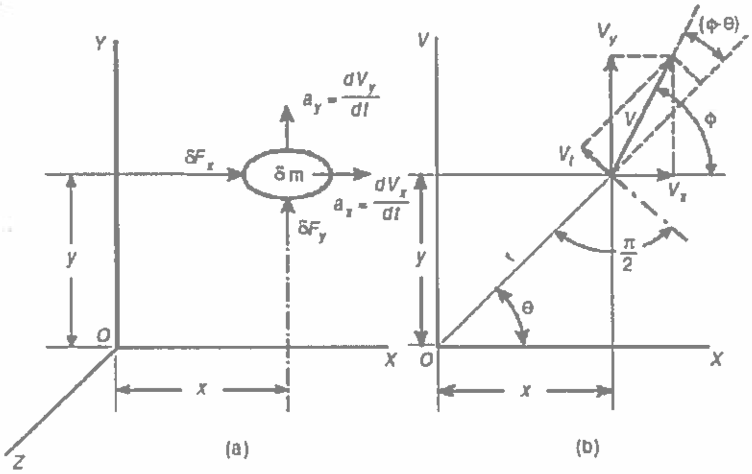
(iii) Efficiency of the jet, $\eta = \dfrac{\text{Work done per second}}{\text{Energy supplied per sec}}$

$$= \frac{4802.58}{\frac{1}{2} \times 1000 \times \frac{\Pi}{4} \times (0.06)^2 \times 18^3}$$

$$\eta = 0.5825 = 58.25\ \%$$

11. b) Explain about Angular momentum principle.

Ans) The angular momentum principle states that the torque exerted on any body is equal to the rate of change of angular momentum. The torque is defined as the moment of the force and the angular momentum is defined as the moment of momentum; the moments being taken about the axis of rotation.



Fluid mass subjected to torque–definition sketch

Consider a fluid mass δm which is rotating about the z-axis as shown in Fig. (a). Let V_x and V_y be its velocity components in x any y directions respectively.

If $a_x = \frac{dV_x}{dt}$ and $a_y = \frac{dV_y}{dt}$ represent the acceleration components of the fluid mass, one obtains

$$\delta F_x = \frac{dV_x}{dt}\delta m, \ \delta F_y = \frac{dV_y}{dt}\delta m$$

where δF_x and δF_y are the components of external forces causing the acceleration.

The moment of the external forces about z-axis (counter-clockwise being considered positive) or the torque δT_z, is then obtained as

$$\delta T_z = (x\,\delta F_y - y\,\delta F_x)$$

$$= \left(x\frac{dV_y}{dt} - y\frac{dV_x}{dt}\right)\delta m$$

By the rules of differentiation

$$\frac{d}{dt}(xV_y - yV_x) = \frac{dx}{dt}V_y - \frac{dy}{dt}V_x + x\frac{dV_y}{dt} - y\frac{dV_x}{dt}$$

$$= V_xV_y - V_yV_x + x\frac{dV_y}{dt} - y\frac{dV_x}{dt}$$

Therefore, since $(V_x V_y - V_yV_x) = 0$ and δm is constant,

$$\delta T_z = \frac{d}{dt}\left(xV_y - yV_x\right)\delta m$$

$$= \frac{d}{dt}[(xV_y - yV_x)\delta m]$$

The quantities $(\delta m V_y)x$ and $(\delta m V_x)y$ represent the "moments of momentum" or "angular momentum" about the z-axis. Therefore the right hand side of the above expression represents the rate of change of angular momentum about z-axis, and this is equal to the torque.

In the above derivation, since z-axis is arbitrarily chosen, a torque equation for the x or y axis may also be similarly obtained.

Hence it may be stated that the resultant external torque about any axis is equal to the rate of change of angular momentum about that axis. This is the *law of moment of momentum* (or law of angular momentum).

It is usually convenient to express $(xV_y - yV_x)$ in terms of V_t and r, where V_t is the tangential velocity and r is the radial distance as defined in Fig. 8.6 (b).

From Fig. 8.6 (b) since,

$$x = r\cos\theta;\ y = r\sin\theta$$

$$V_x = V\cos\phi; V_y = V\sin\phi$$

Hence

$$(xV_y - yV_x) = r\cos\theta\,(V\sin\phi) - r\sin\theta\,(V\cos\phi)$$

$$= rV\sin(\phi - \theta)$$

$$= rV_t$$

Thus by substituting in Eq. 8.27

$$\delta T_z = \frac{d(rV_t\delta m)}{dt}$$

Applying Eq. 8.27 or 8.28 to each of the several small fluid masses of a system and summing all the resulting equations, the resultant external torque T_z for a steady flow system is obtained as

$$\Sigma(\delta\Delta T_z) = \frac{\Sigma\, d(rV_t\delta m)}{dt}$$

or

$$T_z = \rho Q(r_2V_{t_2} - r_1V_{t_1})$$

in which r_2 and V_{t_2} are the radial distance and tangential velocity at section 2 and r_1 and V_{t_1} are the same quantities at section 1 of the control volume.

By rewriting Eq. in the form

$$T_z - \rho Qr_2V_{t_2} + \rho Qr_1V_{t_2} = 0$$

- It can be shown that the moment of the momentum flux across an area about any axis equals the moment of all the external forces applied at the centre of the area about the same axis.
- Further it may be seen from above equation that if the external forces that act on the fluid mass exert no net moment about a fixed axis (i.e., Tz = 0), the moment of momentum of the fluid mass with respect to that axis remains constant.
- This principle is known as the law of conservation of moment of momentum or the law of conservation of angular momentum.

12. Write in detail about a hydropower installation

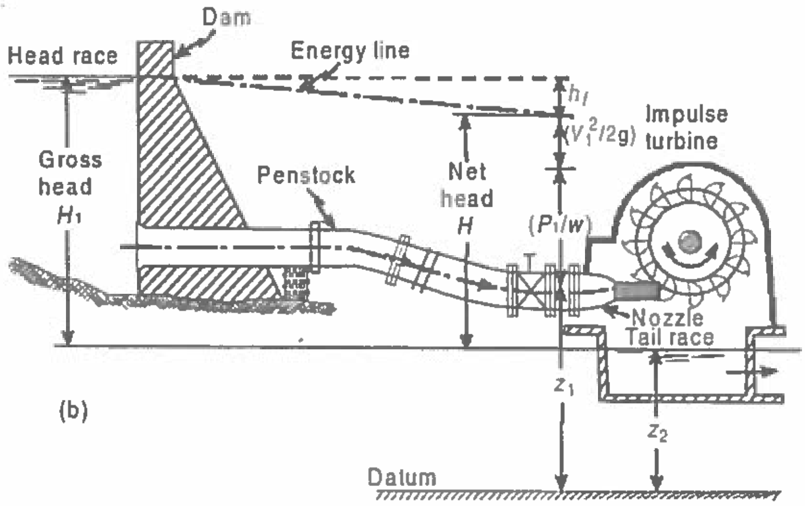
Ans)

- Hydraulic (or water) turbines are the machines which use the energy of water (hydro-power) and convert it into mechanical energy.

- The mechanical energy developed by a turbine is used in running an electric generator which is directly coupled to the shaft of the turbine.
- The electric generator thus develops electric power, which is known as hydroelectric power.

ELEMENTS OF HYDROELECTRIC POWER PLANTS

- One of the essential requirements of the hydroelectric power generation is the availability of a continuous source of water with a large amount of hydraulic energy.
- Such a source of water may be made available if a natural lake or a reservoir may be found at a higher elevation or an artificial reservoir may be formed by constructing a dam across a river.
- The following figure shows a general layout of a hydroelectric power plant, in which an artificial storage reservoir formed by constructing a dam has been shown.



General layout of a hydro-electric power plant

- The water surface in the storage reservoir is known as head race level or simply head race.
- Water from the storage reservoir is carried through penstock or canals to the power house. Penstocks are the pipes of large diameter, usually made of steel, wood or reinforced concrete, which carry water under pressure from the storage reservoir to the turbine.
- In some installations smaller reservoirs known as forebays are also provided. A forebay is essentially a storage reservoir at the head of the penstocks. The purpose of a forebay is to temporarily store water when it is not required by the turbine and supply the same when required.
- The water passing through the turbine is discharged to the tail race. The tail race is the channel which carries water (known as tail water) away from the power house after it has passed through the turbine.
- It may be a natural stream channel or a specially excavated channel entering the natural
- stream at some point downstream from the power house. The water surface in the tail race channel is known as tail race level or simply tail race.

13. a) A pelton wheel has to be designed for the following data: power to be developed =6000 kW, Net head available = 400 m, speed = 550 rpm, Ratio of jet diameter to the

wheel diameter = 1/10 and overall efficiency = 85%. Find the number of jets, diameter of jet, diameter of the wheel and quantity of water required.

Ans)

13a) Given data, Shaft power $SP = 6000\,kW$; Head $H = 400\,m$

Speed $N = 550\,rpm$; efficiency $\eta_o = 85\%$

ratio of jet dia to wheel dia $= \frac{d}{D} = \frac{1}{10}$

Taking Coefficient of velocity $K_{v_1} = C_v = 0.985$

Speed ratio $K_{u_1} = 0.45$

Velocity of jet $V_1 = C_v\sqrt{2gH} = 0.985\sqrt{2\times 9.81\times 400} = 87.26\,m/s$

Velocity of wheel $u = u_1 = u_2 = $ speed ratio $\times \sqrt{2gH}$

$= 0.45\sqrt{2\times 9.81\times 400} = 39.87\,m/s$

But $u = \frac{\pi DN}{60} \Rightarrow D = \frac{u\times 60}{\pi N} = \frac{39.87\times 60}{\pi\times 550} = 1.38\,m$

given $\frac{d}{D} = \frac{1}{10} \Rightarrow d = \frac{D}{10} = \frac{1.38}{10} = 0.138\,m$

Discharge of one jet, $q =$ Area of one jet $\times$ velocity of jet

$= \frac{\pi}{4}\times(0.138)^2\times 87.26 = 1.31\,m^3/s$

$$\eta_o = \frac{SP}{WP} = \frac{\text{Shaft Power}}{\text{Water Power}} = \frac{SP}{\left(\frac{\rho g\times Q\times H}{1000}\right)}$$

$$0.85 = \frac{6000}{\frac{1000\times 9.81\times Q\times 400}{1000}} \Rightarrow Q = 1.80\,m^3/s$$

$$\text{Number of jets} = \frac{\text{Total discharge}}{\text{Discharge of one jet}} = \frac{Q}{q} = \frac{1.80}{1.31}$$

$= 1.37$

$\simeq 2$ jets

13. b) Write short note on Francis turbine.

Ans)

- Francis Turbine is a combination of both impulse and reaction turbine, where the blades rotate using both reaction and impulse force of water flowing through them producing electricity more efficiently.
- Francis turbine is used for the production of electricity in hydro power stations.

The major components of Francis turbine are

- Spiral Casing
- Stay Vanes
- Guide Vanes
- Runner Blades
- Draft Tube

Working of Francis turbine:

- The water is allowed to enter the spiral casing of the turbine, which lead the water through the stay vanes and guide vanes.
- The spiral case is kept in decreasing diameter so as to maintain the flow pressure.

- The stay vanes being stationary at their place, removes the swirls from the water, which are generated due to flow through spiral casing and tries it to make the flow of water more linear to be deflected by adjustable guide vanes.
- The angle of guide vanes decides the angle of attack of water at the runner blades thus make sure the output of the turbine.
- The runner blades are stationary and can-not pitch or change their angle so it's all about the guide vanes which controls the power output of a turbine.
- The performance and efficiency of the turbine is dependent on the design of the runner blades.
- In a Francis turbine, runner blades are divided into 2 parts.
- The lower half is made in the shape of small bucket so that it uses the impulse action of water to rotate the turbine.
- The upper part of the blades use the reaction force of water flowing through it.
- Thus runner blades make use of both pressure energy and kinetic energy of water and rotates the runner in most efficient way.
- The water coming out of runner blades would lack both the kinetic energy and pressure energy, so we use the draft tube to recover the pressure as it advances towards tail race, but still we cannot recover the pressure to that extent that we can stop air to enter into the runner housing thus causing cavitation

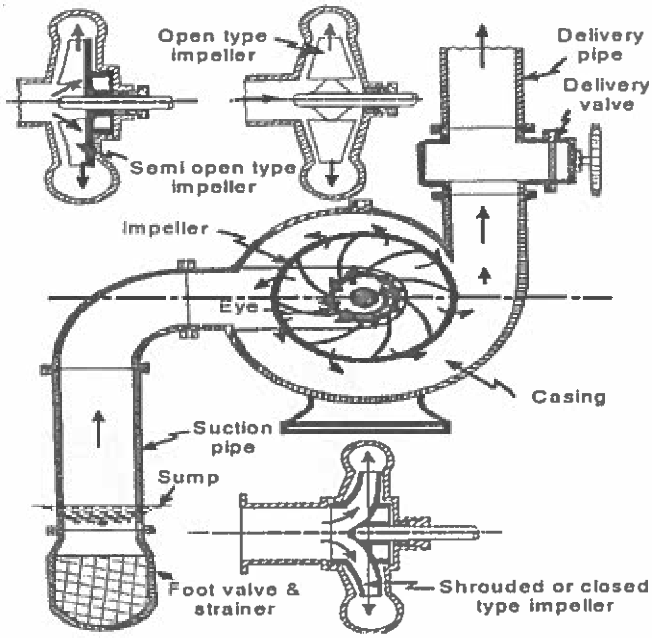
14. a) Define centrifugal pump and explain the working procedure of a single-stage centrifugal pump with neat sketch.

Ans) Centrifugal Pump

A centrifugal pump is a mechanical device designed to move a fluid by means of the transfer of rotational energy from one or more driven rotors, called impellers. Fluid enters the rapidly rotating impeller along its axis and is cast out by centrifugal force along its circumference through the impeller's vane tips

Working procedure of a single-stage centrifugal pump:

- A single-stage centrifugal pump consists of one impeller rotating on a shaft within a pump casing which is designed to produce fluid flow when driven by a motor.
- These pumps are excellent for applications with high flow rates and also, low-pressure purposes.
- Single-stage pumps are usually used in pumping services like high-flow and total dynamic head (TDH) from low to moderate ranges.



Working

- The process liquid enters the suction nozzle and then into eye (center) of a revolving device known as an impeller.
- When the impeller rotates, it spins the liquid sitting in the cavities between the vanes in an outward direction and provides centrifugal acceleration.
- As the liquid leaves the eye of the impeller a low-pressure area is created causing more liquid to flow toward the inlet.
- Because the impeller blades are curved, the fluid is pushed in a tangential and radial direction by the centrifugal force.
- This force acting inside the pump is same as the one that keeps water inside a bucket that is rotating at the end of a string.

14. b) A centrifugal pump rotating at 1080 rpm delivers 168 liters/s of water against a head of 30 m. The pump is installed at a place where atmospheric pressure is $1x10^5$ Pa(abs.) and vapour pressure of water is 2 kPa (abs.). The head loss is suction pipe is equivalent to 0.2 m of water. Calculate minimum NPSH.

Ans) Given data, N=1080 rpm, Q=168 litre/s=0.168 m^3/s, H_m =30 m, p_a= 10^5 Pa

p_v= 3 kPa, h_{fs} = 0.2 m of water

Minimum NPSH:

According to definition of thomas cavitation factor,

NPSH will be minimum when is minimum

minimum value of for no cavitation is , Hence =

= $1.03*10^{-3}$ * $N_s^{4/3}$, N_s = specific speed of pump = N. $Q^{0.5}$ / $H_m^{3/4}$

using values, we get = 1.03 * 10^{-3} * $1080^{4/3}$ * $0.168^{2/3}$ / 30 = 0.1158

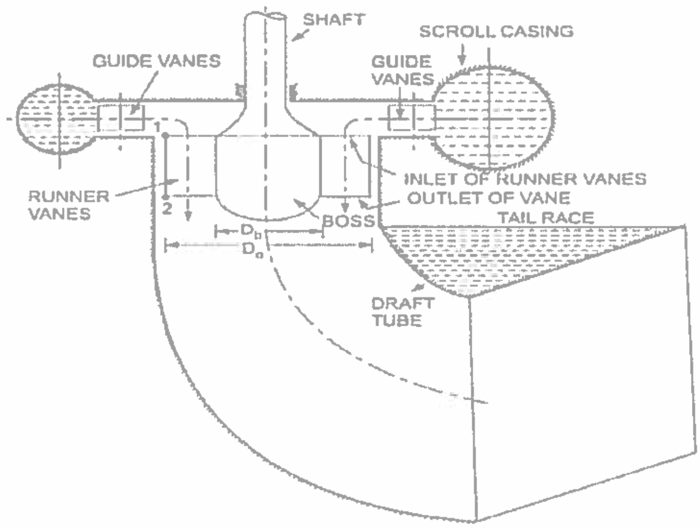
substituting the value of ,

$(NPSH)_{min}$ = H_m * 0.1158 = 30* 0.1158 = 3.474 m

15. a) Write in detail about Kaplan turbine.

Ans) **KAPLAN TURBINE**

- A Kaplan turbine is a type of propeller turbine which was developed by the Austrian engineer V. Kaplan (1876–1934).
- It is an axial flow turbine, which is suitable for relatively low heads, and hence requires a large quantity of water to develop large amount of power.
- It is also a reaction type of turbine and hence it operates in an entirely closed conduit from the head race to the tail race.



Main components of Kaplan turbine.

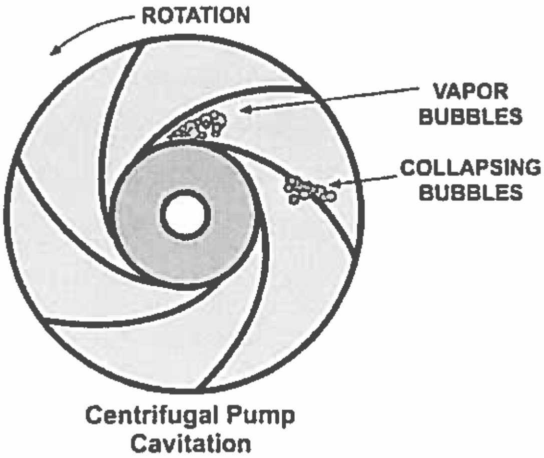
- The main components of a Kaplan turbine such as scroll casing, stay ring, arrangement of guide vanes, and the draft tube are similar to those of a Francis turbine.
- Between the guide vanes and the runner, the water in a Kaplan (or propeller) turbine turns through a right-angle into the axial direction and then passes through the runner.
- The runner of a Kaplan (or propeller) turbine has four or six blades and it closely resembles a ship's propeller.
- The blades (or vanes) attached to a hub or boss are so shaped that water flows axially through the runner.
- Ordinarily the runner blades of a propeller turbine are fixed, but the Kaplan turbine runner blades can be turned about their own axis, so that their angle of inclination may be adjusted while the turbine is in motion.
- This adjustment of the runner blades is usually carried out automatically by means of a servomotor operating inside the hollow coupling of turbine and generator shaft.
- When both guide-vane angle and runner-blade angle may thus be varied, a high efficiency can be maintained over a wide range of operating conditions.
- In other words, even at part load, when a lower discharge is flowing through the runner, a high efficiency can be attained in the case of a Kaplan turbine.

15. b) Write about cavitation in pumps.

Ans) *Cavitation*

- Cavitation occurs when the liquid in a pump turns to a vapor at low pressure.
- It occurs because there is not enough pressure at the suction end of the pump, or insufficient Net Positive Suction Head available (NPSHa). When cavitation takes place, air bubbles are created at low pressure
- If the pressure at the suction side of the pump drops below the vapour pressure of the liquid then the cavitation may occur.

- The cavitation in a pump can be noted by a sudden drop in efficiency and head.
- Cavitation includes the creation and breakdown of vapor bubbles in the liquid due to the variation in pressure values. The overall performance of the pump would be affected by the cavitation.



Suction cavitation in centrifugal pumps

- To find out if the pump is affected by cavitation or not, the following signs can be helpful:
  - Increase and decrease in discharge pressure values
  - Inconsistent power use
  - The reduction in the efficiency
  - Distinct crackling sounds
- By reducing the length of the pumps to 4 meters a head of the water level, the effect of cavitation disappears.

NSRIT

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	Program	ECE			Academic Year	2021 - 2022
Course Code	20EC403	Test Duration	3 Hrs.	Max. Marks	70	Semester	IV
Course	**Pulse and Digital Circuits**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Under what Condition high pass RC Circuit act as Differentiator?	20EC403.1	L1
2	Define the following for a transistor switch i) Rise Time and ii) Fall Time.	20EC403.2	L1
3	What type of triggering is used in Monostable Multivibrator?	20EC403.3	L1
4	What are the Time Base Generators?	20EC403.4	L1
5	Distinguish between Sampling Gates and Logic Gates.	20EC403.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain the response of High-Pass RC circuit for square wave input.	6M	20EC403.1	L2
6 (b)	A Step Generator of 50 Ω impedance applies a 10V step of 2.2ns rise time to a series combination of a capacitance C and Resistance 50Ω. There appears across R a pulse of amplitude 1V Find The Capacitance C.	6M	20EC403.1	L2
	OR			
7 (a)	With the help of a neat circuit diagram, explain the working of a two level diode clipper.	8M	20EC403.1	L2
7 (b)	If A Square Wave of 5kHz is applied to an RC High Pass Circuits and the resultant waveform is tilted from 15V to 10V Find the Lower 3dB frequency of the High Pass Circuit.	4M	20EC403.1	L3
8 (a)	Explain the design of the transistor switch.	8M	20EC403.2	L2
8 (b)	For a common emitter circuit, V_{cc} is 15V, R_c is 1.5kΩ and Ib is 0.3mA. Determine the value of h_{fe} for saturation to occur and If R_c is changed to 500Ω will the transistor be saturated?	4M	20EC403.2	L4
	OR			
9 (a)	Explain the working of Collector Coupled Bistable Multivibrator with the help of neat diagram.	6M	20EC403.2	L2
9 (b)	A Silicon Transistors with h_{fe} equal to 20 are available. If V_{cc}=V_{bb}=10V Design the Bistable Multivibrator.	6M	20EC405.2	L4
10 (a)	Derive an expression for the frequency of oscillation of an astable multivibrator.	12M	20EC403.3	L3
	OR			
11 (a)	Derive the expression for gate width of a monostable multivibrator.	6M	20EC403.3	L3
11 (b)	Design a collector coupled one shot with a gate width of 3ms, using npn transistors.	6M	20EC403.3	L2
12 (a)	Draw the Circuit of miller integrator and explain how it improves the linearity of the sweep waveform.	6M	20EC403.4	L2
12 (b)	Explain the basic principles behind Bootstrap time base generator.	6M	20EC403.4	L2
	OR			
13 (a)	Derive the relation between slope, transmission and displacement errors.	6M	20EC403.4	L4
13 (b)	Design a relaxation oscillator to have 2kHz output frequency, using specifications Ip=2µA,Iv=1mA,Veb(sat)=3V and intrinsic stand off ratio is 0.68 to 0.82	6M	20EC403.4	L4

14 (a)	Compare DTL, TTL, ECL, and RTL Logic families.	6M	20EC403.5	L4
14 (b)	Define Propagation Delay, Fan-in, Fan-out, Noise Margin, Speed Power Product and Power Dissipation.	6M	20EC403.5	L1
	OR			
15 (a)	Write advantages and disadvantages of unidirectional diode sampling gates and application of sampling gates.	6M	20EC403.5	L2
15 (b)	Explain the operation of Four Diode sampling gate.	6M	20EC403.5	L2

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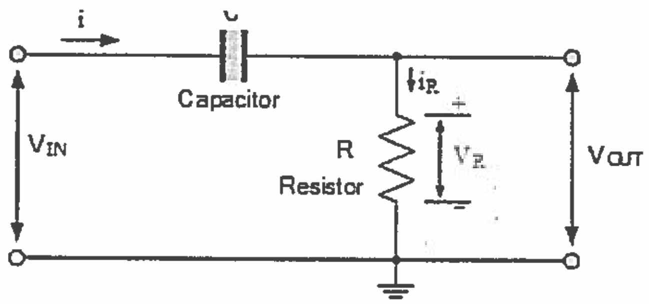
PULSE AND DIGITAL CIRCUITS
SCHEME OF VALUATION

1. For an RC differentiator circuit, the input signal is applied to one side of the capacitor with the output taken across the resistor, then V_{OUT} equals V_R. As the capacitor is a frequency dependant element, the amount of charge that is established across the plates is equal to the time domain integral of the current. That is it takes a certain amount of time for the capacitor to fully charge as the capacitor cannot charge instantaneously only charge exponentially.



Capacitor Current

$$i_{(t)} = \frac{dQ}{dt} = \frac{d(C \times dV_C)}{dt} = C\frac{dV_C}{dt} = C\frac{dV_{IN}}{dt}$$

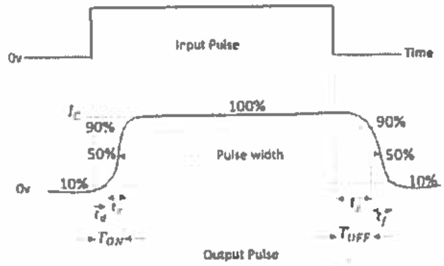
Therefore the capacitor current can be written as:

$$\boxed{i_{C(t)} = C\frac{dV_{IN(t)}}{dt}}$$

2. **Rise time(t_r)** – The time taken for the collector current to reach from 10% of its initial value to 90% of its final value is called as the **Rise Time.**

Fall time (t_f) – The time taken for the collector current to reach from 90% of its maximum value to 10% of its initial value is called as the **Fall Time.**

PULSE AND DIGITAL CIRCUITS
SCHEME OF VALUATION



3.

i) It has only one stable state. The other state is unstable referred as quasi- stable state.
ii) It is also known as one-short multivibrator or univibrator.
iii) When an external trigger pulse is applied to the circuit, the circuit goes into the quasi-stable state from its normal stable state.
iv) After some time interval, the circuit automatically returns to its stable state.
v) The circuit does not require any external pulse to change from quasi- stable state.
vi) The time interval for which the circuit remains in the quasi-stable state is determined by the circuit components and can be designed as per the requirement.

4.

There are two types of Time base Generators. They are –

- **Voltage Time Base Generators** – A time base generator that provides an output voltage waveform that varies linearly with time is called as a Voltage Time base Generator.
- **Current Time Base Generator** – A time base generator that provides an output current waveform that varies linearly with time is called as a Current Time base Generator.

5. A **logic gate** is a computer circuit with several inputs but only one output that can be activated by particular combinations of inputs, a diagram that shows the major gates can be found here.

A **sampling gate**, on the other hand is a circuit that produces an output only when first activated by a preliminary pulse. So if you have a current going through a wire and through a sampling gate, your output would be 0, unless you program the sampling gate to let the current through.

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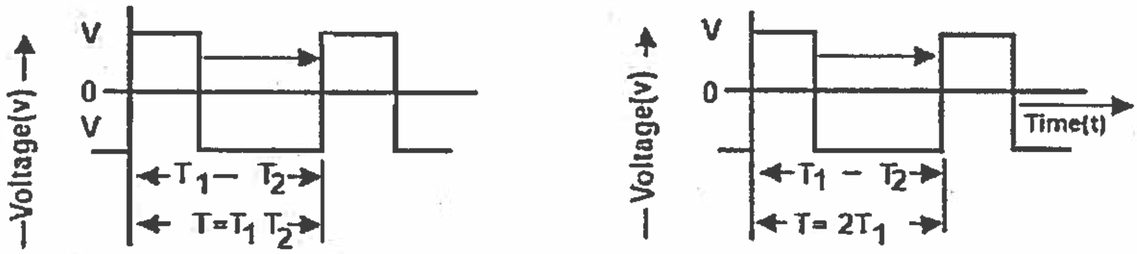
PULSE AND DIGITAL CIRCUITS
SCHEME OF VALUATION

6 (a).

A waveform which maintains itself at one constant voltage level V1 for a time T1 and at another constant level V2 for time T2 and is repetitive with a period T = T1 + T2 as shown in Fig.(a) is called a square waveform. The square waveform is used in digital electronic circuits, radars and as synchronizing pulses in television.



(a) Square waveform (b)Symmetrical square waveform

Expression for the percentage tilt

We will derive an expression for the percentage tilt when the time constant *RC* of the circuit is very large compared to the period of the input waveform, i.e. *RC* » *T*. For a symmetrical square wave with zero average value

$$V_1 = -V_2, \text{i.e } V_1 = |V_2|, V_1' = -V_2' \; i.e \; V_1' = |V_2'|, and\ T_1 = T_2 = \frac{T}{2}$$

The output waveform for RC>>T is shown in Figure 1.35. Here

$$V_1' = V_1 e^{-T/2RC} \text{ and } V_2' = V_2 e^{-T/2RC}$$

$$V_1 - V_2' = V$$

$$V_1 - V_2 e^{-T/2RC} = V_1 + V_1 e^{-T/2RC} = V$$

$$\Rightarrow V_1 = \frac{V}{1+e^{-T/2RC}} \text{ or } V = V_1(1+e^{-T/2RC})$$

$$\%\text{tilt, P}= \frac{V_1-V_1'}{\frac{V}{2}}\times 100\% = \frac{V_1-V_1e^{-T/2RC}}{V_1(1+e^{-T/2RC})} \times 200\% = \frac{1-e^{-T/2RC}}{1+e^{-T/2RC}} \times 200\%$$

PULSE AND DIGITAL CIRCUITS
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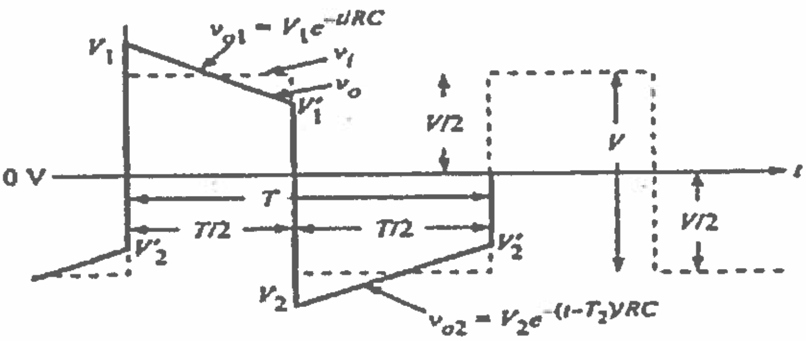


Figure Linear tilt of a symmetrical square wave when $RC >> T$.

When the time constant is very large, i.e $\frac{T}{RC} \ll 1$

$$P=\frac{1-\left[1+\left(-\frac{T}{2RC}\right)+\left(\frac{T}{2RC}\right)^2\frac{1}{2!}+\cdots\right]}{1+1+\left(-\frac{T}{2RC}\right)+\left(\frac{T}{2RC}\right)^2\frac{1}{2!}+\cdots}\times 200\%$$

$$=\frac{\frac{T}{2RC}}{2}\times 200\%$$

$$=\frac{T}{2RC}\times 100\%$$

$$=\frac{\pi f_1}{f}\times 100\%$$

Where $f_1=\frac{1}{2\pi RC}$ is the lower cut-off frequency of the high-pass circuit.

PULSE AND DIGITAL CIRCUITS
SCHEME OF VALUATION

6(b)

O/p response for square wave input

$$V_0(t) = V_{in} e^{-t/RC}$$

$$\ln\left(\frac{V_0}{V_i}\right) = -\frac{T}{RC}$$

$$T = +RC \ln\left(\frac{V_i}{V_0}\right)$$

$T = 2.2ns$

$V_{in} = 10V$ $\quad V_0 = 1V.$

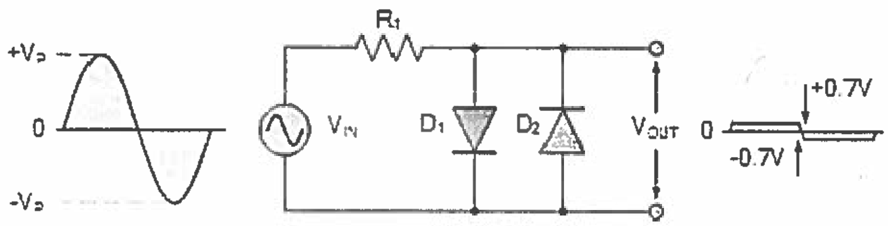
$R = 50\Omega$

$$C = \frac{T}{R} e^{-\left(\frac{V_i}{V_0}\right)}$$

$$\boxed{C = \frac{2.2ns}{50} e^{-(10)}}$$

$$\ln(10) \quad C = \frac{2.2ns}{50}$$

7(a) Clipping of Both Half Cycles



PULSE AND DIGITAL CIRCUITS
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If we connected two diodes in inverse parallel as shown, then both the positive and negative half cycles would be clipped as diode D_1 clips the positive half cycle of the sinusoidal input waveform while diode D_2 clips the negative half cycle. Then diode clipping circuits can be used to clip the positive half cycle, the negative half cycle or both.

For ideal diodes the output waveform above would be zero. However, due to the forward bias voltage drop across the diodes the actual clipping point occurs at +0.7 volts and –0.7 volts respectively. But we can increase this ±0.7V threshold to any value we want up to the maximum value, (V_{PEAK}) of the sinusoidal waveform either by connecting together more diodes in series creating multiples of 0.7 volts, or by adding a voltage bias to the diodes.

7(b) 5 KHz frequency square wave applied to RC Highpass circuit

$$\% \text{tilt } P = \frac{V_1 - V_1'}{\frac{V}{2}} \times 100\%$$

$$P = \frac{\pi f_1}{f} \times 100\%$$

$$\frac{V_1 - V_1'}{V/2} \times 100 = \frac{\pi f_1}{f} \times 100$$

$$\boxed{f_1 = \frac{2f}{\pi V}(V_1 - V_1')}$$

Given $f = 5\,KHz$

$V_1 = 15\,V$ & $V_1' = 10\,V$

$f_1 \rightarrow$ lower cutoff frequency

$$V = V_1(1 + e^{-T/2RC})$$

$$V \simeq V_1 = 15\,V$$

$$f_1 = \frac{2(5K)}{\pi(15)}(5) = 1\,KHz$$

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PULSE AND DIGITAL CIRCUITS
SCHEME OF VALUATION

8(A) Design of transistor as a switch

Transistor can operate in three regions

a. Cut-off region
b. Active region
c. Saturation region

Cut-Off region: When collector junction is in reverse bias and emitter junction of the transistor is in forward bias then we can say transistor is operated in active region and it acts as amplifier.
Active region: When both collector and emitter junctions of the transistor are reverse biased then we can say transistor is operated in cut-off region and it acts as open switch.
Saturation region: When both collector and emitter junctions of the transistor are forward bias then we can say transistor is in the saturation region and it acts as a closed switch.

i. The time interval between the instant of application input pulse and output (collector) currant to attain 10 percent of its maximum value is termed as the delay time t_d.
ii. Rise time, t_r is defined as the time required for the output current I_C to go from 10% to 90% of its maximum value.
iii. The sum of delay time t_d, and rise time t_r, is called the turn-ON time t_{ON}, i.e.

$$t_{ON} = t_d + t_r$$

iii. TURN- OFF time is made up of a storage time t_s, and a fall time t_f
i.e $t_{off} = t_s + t_f$
iv. Storage time, t_s is defined as the time interval between the end of the input pulse (trailing edge) and when the collector current falls to 90% of its maximum value.

OR

Storage time, t_s is equal to the sum of time taken in removing excess charge stored and the time taken by collector transition capacitance to discharge to 90% of its maximum but major portion of the time is taken in removing excess charge storage. The time duration of the output pulse measured between two 50% levels of rising and falling waveform is known as the **pulse width**.
For a fast-switching transistor, turn-on time t_{ON} and turn-off t_{off} time must be of the order of nano seconds.

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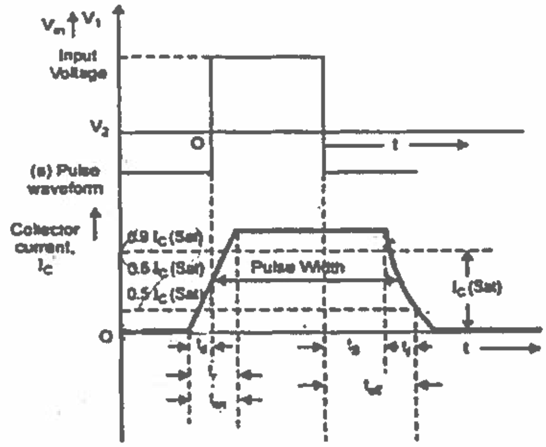


Fall time: The time required for the collector to drop from 90% to 10% of the saturation current is defined as a fall time t_f.

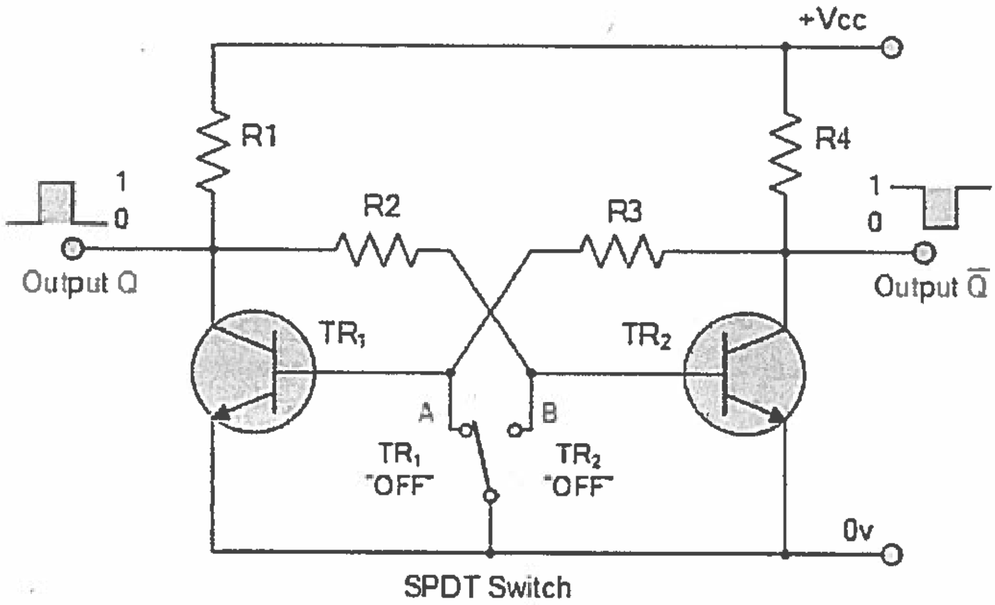
9(a) A bistable is the one of the type of multivibrator device similar to the monostable, we discus about monostable in previous article but the difference this time is that BOTH states are stable.

It has two stable states and maintains the state until the external trigger is not applied. This means the output will be shifted from one stage to another stage by applying trigger pulse. It required two external pulses to return original state. As it has two stable stages they are known as latches and flip flop.

The bistable multivibrator has two state non-regenerative devices. The circuit configuration is a cross coupling of two transistors one is in ON and OFF switching. That means one transistor is in cut-off region and other transistor is in saturation region. The bistable circuit is capable in either stable state without trigger pulse.

To change the stage from one stable state to another state it required external trigger pulse; when we applied two trigger pulses to bistable circuit it return to original position. It is also known as flip flop or latch circuit. The circuit diagram is shown below.

PULSE AND DIGITAL CIRCUITS
SCHEME OF VALUATION



BISTABLE MULTIVIBRATOR OPERATION

The circuit diagram of bistable is shown above is stable in both state. In this circuit shows two transistors, one transistor is in cut off region and other in saturation region. Let's suppose the base of transistor TR_1 is connected to the ground which is shown in figure, and it is cut off region producing output at Q. That would mean that transistor TR_2 is saturation region. The base of transistor TR_2 is connected to Vcc with the series combination of R_1 & R_2. As transistor TR_2 is "ON" there will be zero output at Q, the opposite or inverse of Q.

Now if we applied a trigger pulse at point "B". the transistor TR_2 will switch "OFF" and transistor TR_1 will switch "ON" through the combination of resistors R3 and R4 resulting in an output at Q and zero output at Q the reverse of above. Then we can say that the stable state exist when TR_1 is ON and TR_2 is OFF, switching position A. and other stable state is exist, TR_2 is ON and TR_1 is OFF, switching position B.

PULSE AND DIGITAL CIRCUITS
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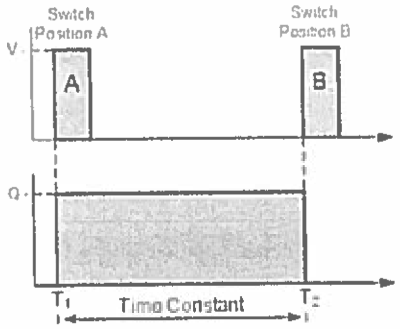

9(b)

Given
$h_{fe} = 20$, $V_{cc} = V_{bb} = 10\,V$
for Bistable Multivibrators

$$R_C = \frac{V_{CC} - V_{(sat)}}{I_{C_2}} = \frac{10 - 0.1V}{5mA} = \frac{9.9}{5mA}$$

$$I_C = \beta I_B$$

$$\beta = \frac{I_C}{I_B} = \frac{5mA}{I_B} = 20$$

$$I_B = \frac{\cancel{5}mA}{\cancel{20}\,4} = \frac{1}{4000}\,Amp$$

$$= \frac{0.25}{1000}$$

$$I_B = 0.25mA$$

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SCHEME OF VALUATION

10(a)
A voltage-controlled oscillator (VCO) is one in which the frequency of oscillations varies as a function of voltage. The same circuit is also called a voltage-to-frequency converter (VFC) because a given voltage gives rise to a specific frequency. An astable multivibrator is used as voltage-controlled oscillator [see Fig. 2 (a)].

When Q1 is OFF and Q2 is ON, C1 charges. When Q1 is ON, the charge on C1 decays with a time constant $\tau 1 = RC1$ as shown in Fig. 7.2(b). As a result, the voltage at the base of Q2, VB2 varies with time, [see Fig. 7.2(c)].

$vB2(t) = vf - (vf - vi)e-t/\tau 1$

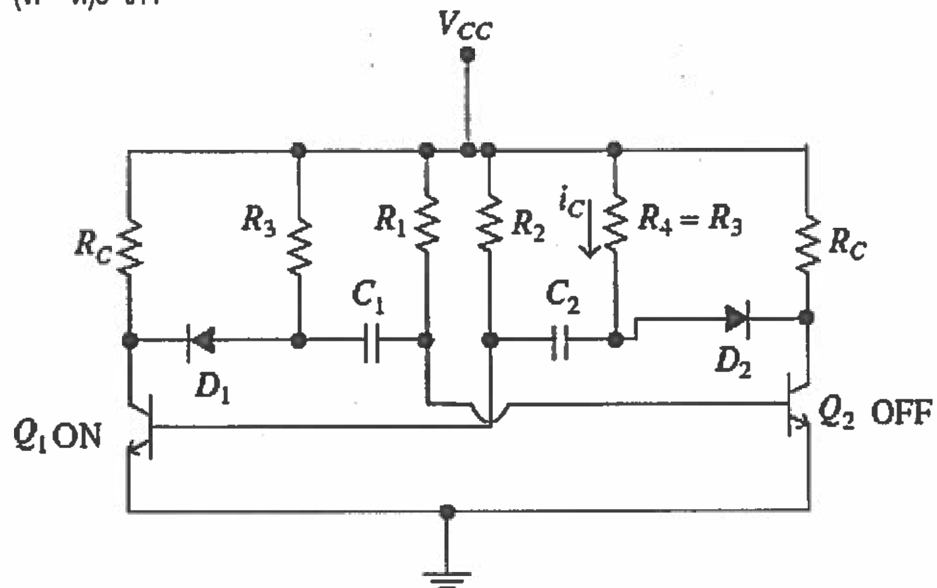

FIGURE 1 An astable multivibrator that generates pulses with vertical edges

PULSE AND DIGITAL CIRCUITS
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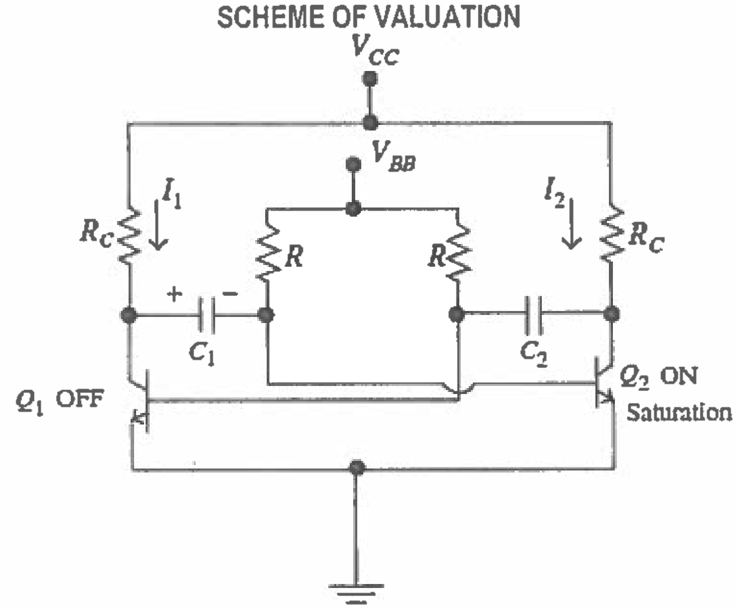

FIGURE 2(a) A voltage-to-frequency converter

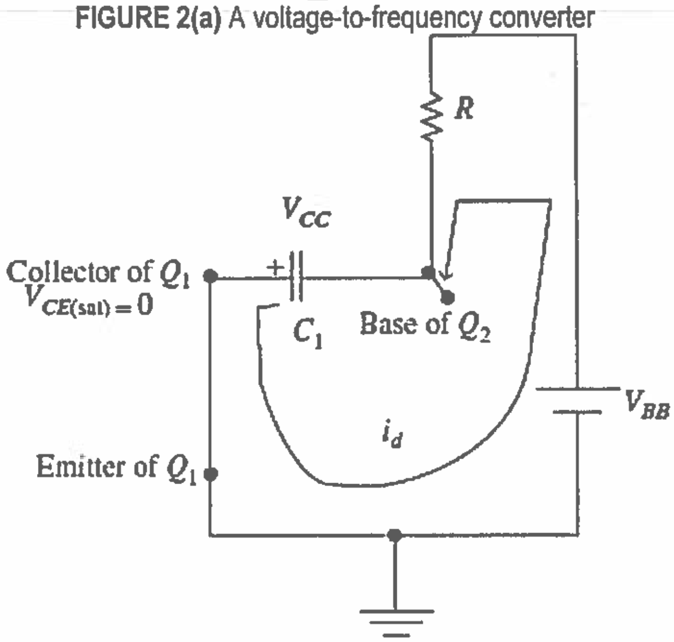

FIGURE 2(b) The discharge of condenser C_1 through R

PULSE AND DIGITAL CIRCUITS
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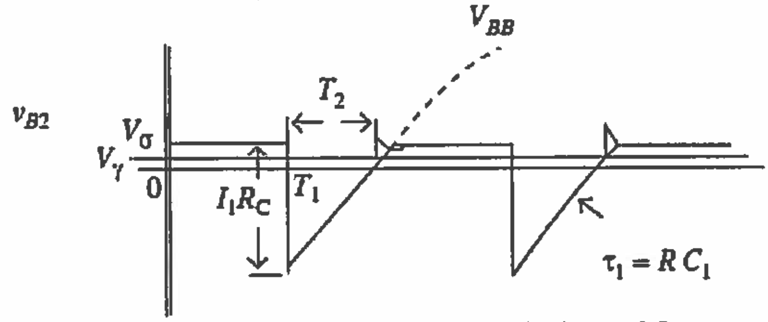


FIGURE 2(c) The voltage variation at the base of Q_2

vf = VBB   vi = Vσ − I1RC = Vσ − VCC + VCE(sat)
Since I1RC = VCC − VCE(sat);
vB2 = VBB − [VBB − Vσ + VCC − VCE(sat)]e−T2/τ1
If the junction voltages are small,
0 = VBB − (VBB + VCC)e−T2/τ1
For a symmetric circuit, $T_1 = T_2 = T/2$ and $\tau_1 = \tau_2 = \tau$

$$T_1 = T_2 = \frac{T}{2} = \tau \ln\left(\frac{V_{BB} + V_{CC}}{V_{BB}}\right)$$

Consequently, for a symmetric astable multivibrator:

$$T = 2\tau \ln\left(1 + \frac{V_{CC}}{V_{BB}}\right)$$

And, f = 1/T. As the frequency of the multivibrator can be varied by simply varying VBB, this circuit is called a voltage-controlled oscillator or voltage-to-frequency converter.

11(a)

We know that the voltage across the capacitor C rises exponentially. Hence the equation for the capacitor voltage VC can be written as

$$VC = VCC\,(1 - e^{-t/RC})$$

When the capacitor voltage is 2/3 VCC, then

$$2/3\ VCC = VCC\,(1 - e^{-t/RC})$$
$$2/3 = 1 - e^{-t/RC}$$
$$e^{-t/RC} = 1/3$$
$$-t/RC = \ln(1/3)$$
$$-t/RC = -1.098$$
$$t = 1.098\ RC$$
$$\therefore t \approx 1.1\ RC$$

PULSE AND DIGITAL CIRCUITS
SCHEME OF VALUATION

The pulse width of the output rectangular pulse is W = 1.1 RC.

11(b)

The collector-coupled monostable multivibrator is shown in Fig. 1. When compared to an astable multivibrator, it is evident that the output from the second collector to the first base is through a resistance R_1. Hence, this circuit has one stable state and one quasi-stable state.

As a negative voltage is connected to the base of the first device, it is possible that Q_1 may be OFF. In the stable state, let Q_1 be OFF and Q_2 be ON and in saturation. Therefore:

VC1 = VCC VC2 = VCE(sat) VB2 = VBE(sat) = Vσ

The capacitor, C now tries to charge to VCC through RC of Q1 and a small input resistance of Q2, as shown in Fig. 8.2. As t→∞, this voltage reaches VCC. To change the state of the devices, a trigger is applied at an appropriate point in the circuit.

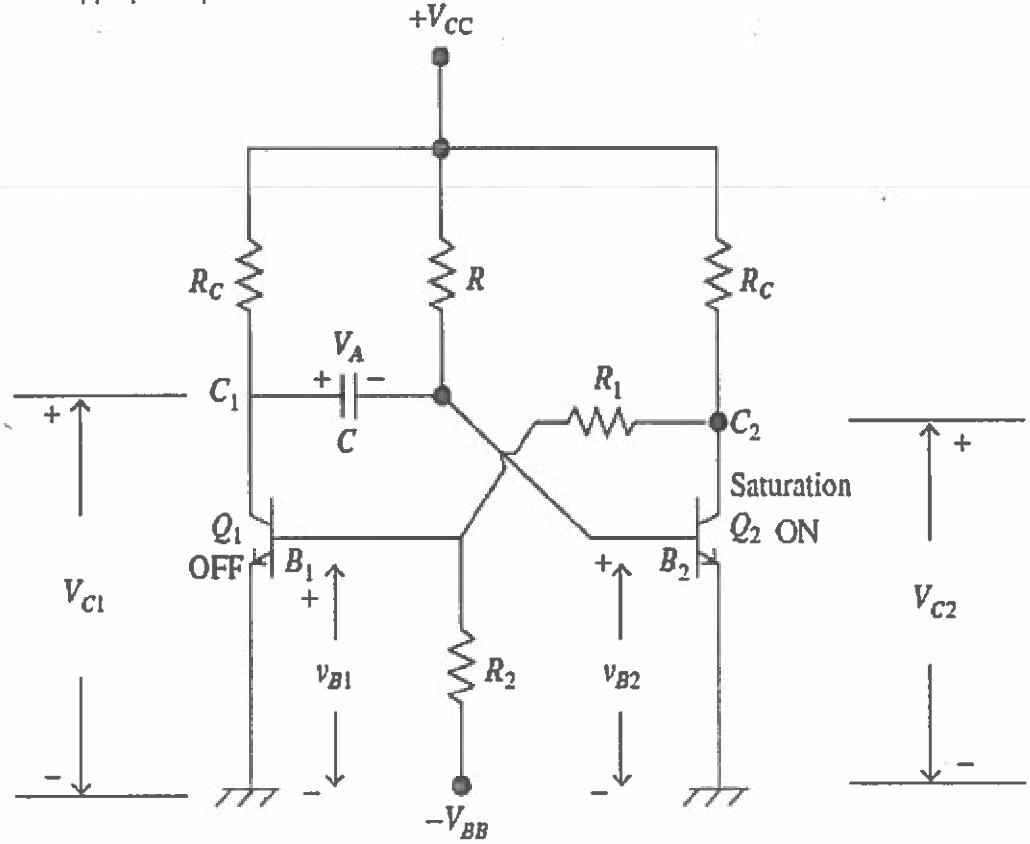


FIGURE 1 The collector-coupled monostable multivibrator

PULSE AND DIGITAL CIRCUITS
SCHEME OF VALUATION

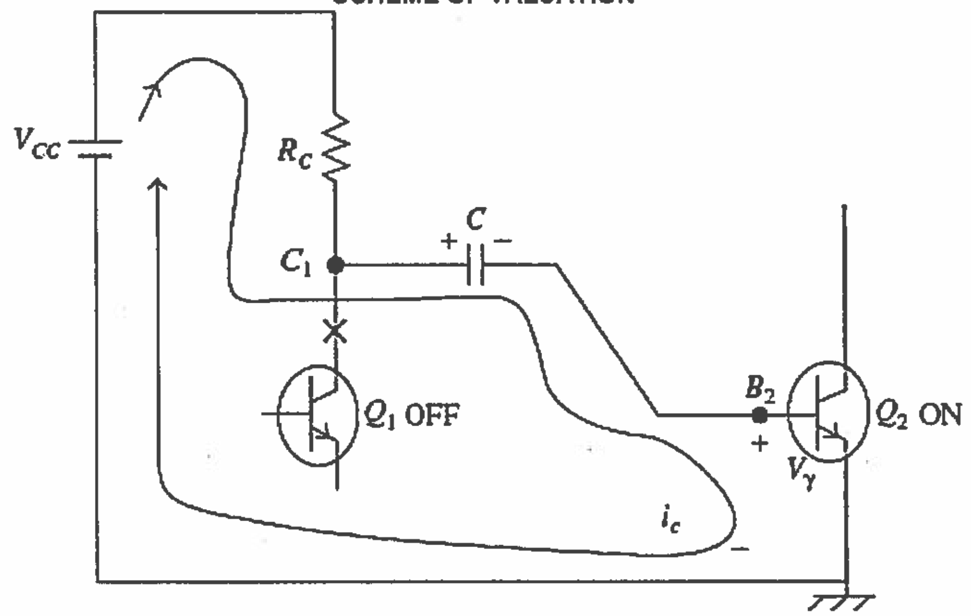


FIGURE 8.2 The charging of capacitor C

12(a) Let us consider the working of the triggered transistor Miller's sweep generator, as shown in Fig. 12.12(a).

1. In the quiescent state (before the application of the trigger): The circuit conditions are adjusted such that when the input is not present Q1 is ON and in saturation. Therefore, the voltage at C1(collector of Q1)is VCE(sat) ≈ 0. Transistor Q2 is OFF since VBE2 ≈ 0. The voltage at C2 (collector of Q2)is VCC, vo = VCC. The voltage across the capacitor Cs is VCC.
2. When trigger is applied at t = 0. When the input signal goes negative, Q1 is OFF and the voltage at the collector of Q1 rises; Q2 is ON and the voltage at its collector is required to decrease abruptly to VCE(sat). Due to the capacitor, the voltage falls almost linearly. The capacitor Cs charges through RC1 and the small resistance RCS (saturation resistance) that exists between the collector and emittér terminals of Q2, which is driven into saturation as shown in Fig. 12.12(b). Hence, the output voltage decreases linearly from VCC to VCE(sat) in Ts and hence, is a negative-going ramp as shown in Fig. 12.12(d). Depending on the time constant employed, Ts may be less than or equal to Tg.
3. At the end of the trigger: Again at the end of the input pulse, at t = Tg, Q1 goes ON, Q2 goes OFF and the capacitor discharges through RC2 and the output again reaches VCC, as shown in Fig. 12.12(c). The waveforms are shown in Fig. 12.12(d).
4. Calculation of Ts:

From Fig. 12.12(b), the charging current of Cs:

PULSE AND DIGITAL CIRCUITS
SCHEME OF VALUATION

$$i_C \simeq \frac{V_{CC}}{R_{C1}} \qquad v_o(t) = \frac{i_C}{C_s}t = \frac{V_{CC}}{R_{C1}} \times \frac{t}{C_s}$$

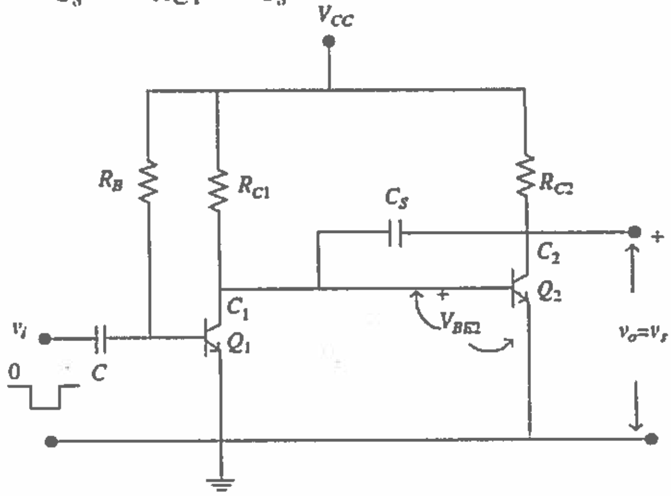


FIGURE 12.12(a) A transistor Miller sweep generator

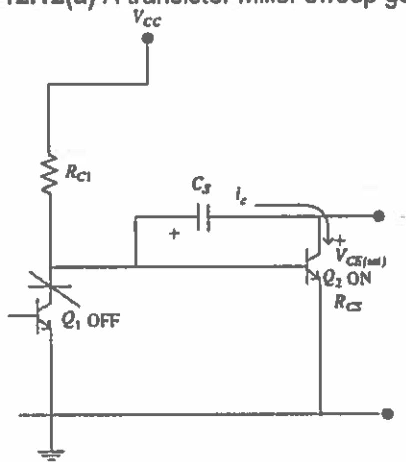


FIGURE 12.12(b) The circuit of Fig. 12.12(a) when Q1 is OFF and Q2 is ON

PULSE AND DIGITAL CIRCUITS
SCHEME OF VALUATION

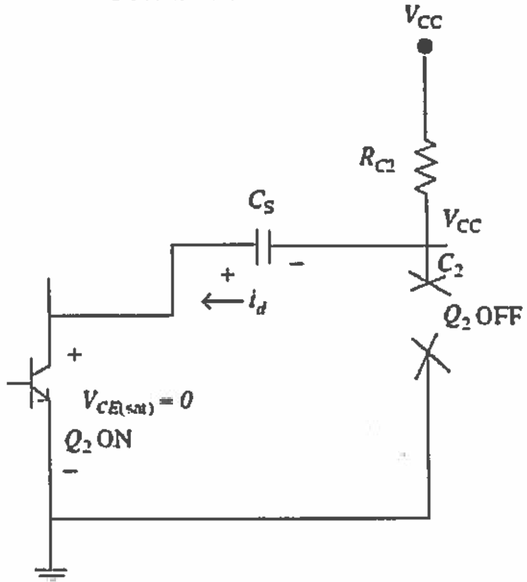
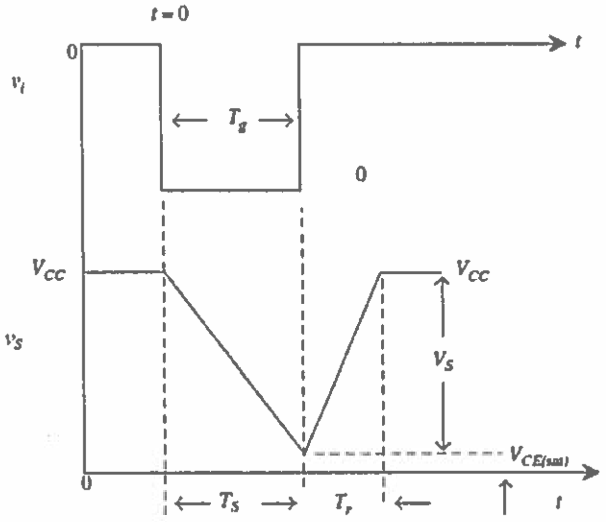
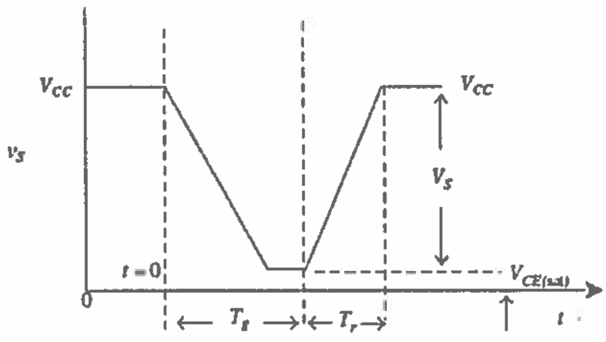


FIGURE 12.12(c) The discharge of C_s when Q_1 is ON and Q_2 is OFF



(i) When v_s reaches V_{CC} at T_g



(ii) When v_s reaches V_{CC} before T_g

FIGURE 12.12(d) The waveforms of Miller's sweep transistor

At t = Ts, vo(t) = Vs
Therefore,

$$V_s = \frac{V_{CC} T_s}{R_{C1} C_s} \qquad (12.36)$$

$$T_s = \frac{V_s}{V_{CC}} \times R_{C1} C_s$$

If Vs = VCC, Ts = RC1Cs.

PULSE AND DIGITAL CIRCUITS
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5. Calculation of Tr: From Fig.12.12(c), the discharging current:The change in voltage during Tr is once again Vs.

Therefore,

$$V_s = \frac{V_{CC} T_r}{R_{C2} C_s} \qquad (12.37)$$

If Vs = VCC, then, Tr = R,C2Cs,

12(b)

Consider the bootstrap sweep generator shown in Fig. 12.14(a) in which the auxiliary generator is replaced by an amplifier with gain 1, which obviously is an emitter follower. If initially the capacitor is uncharged and if S is closed at t = 0, then the voltage across C and Ri, i.e., vi = 0, Ri is replaced by a short circuit. As vi = 0, Avi = 0 and is also replaced by a short circuit. Hence, at t = 0, the circuit of Fig. 12.14(a) reduces to that in Fig.12.14(b).From Fig.12.14 (b):

From Fig.12.14 (b):

$$v_o = -V \times \frac{R_o}{R + R_o} \qquad (12.43)$$

And as Ro of the emitter follower is very small: vo ≈ 0. As t → ∞, C is fully charged and is open circuited and the resultant circuit is shown in Fig.12.14(c).From Fig. 12.14(c)

$$v_o(t \to \infty) = \frac{V(AR_i - R_O)}{R_o + R + R_i(1 - A)}$$

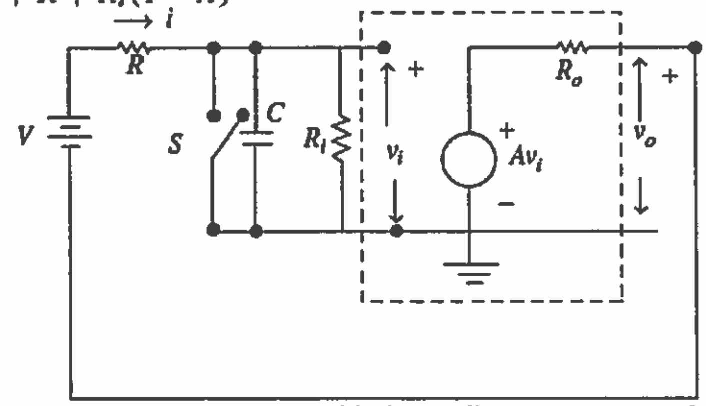

FIGURE 12.14(a) A bootstrap sweep generator with the auxilary generator replaced by an amplifier

PULSE AND DIGITAL CIRCUITS
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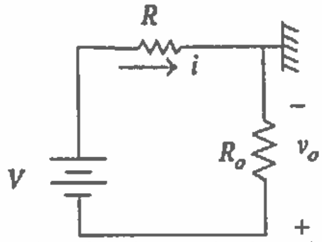


FIGURE 12.14(b) The circuit to calculate the output at t = 0

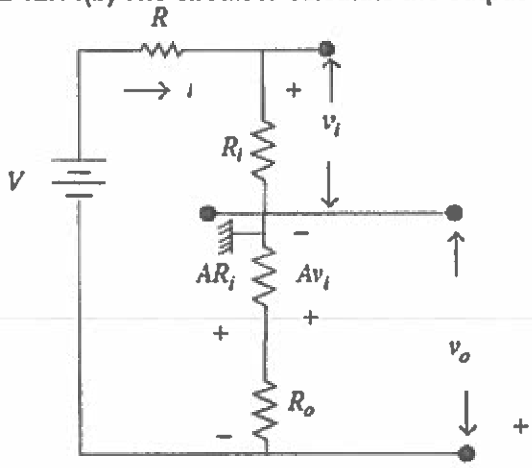


FIGURE 12.14(c) The circuit to calculate the output as t → ∞

Dividing by Ri:

$$v_o(t \to \infty) = \frac{V\left(A - \frac{R_o}{R_i}\right)}{(1-A) + \frac{R}{R_i} + \frac{R_o}{R_i}}$$

Here, Ro is the output resistance of the emitter follower, which is small and Ri is its input resistance, which is large. Therefore, Ro/Ri is negligible and A ≈ 1

$$v_o(t \to \infty) = \frac{V}{(1-A) + \frac{R}{R_i} + \frac{R_o}{R_i}} \approx \frac{V}{(1-A) + R/R_i} \qquad (12.44)$$

Eq. (12.44) gives the peak-to-peak excursion of the output swing. Therefore,

$$e_{s(\text{Bootstrap})} = \frac{V_s\left[(1-A) + \frac{R}{R_i}\right]}{V} \approx \frac{V_s}{V}\left(1 - A + \frac{R}{R_i}\right) \cong e_s\frac{R}{R_i} \qquad (12.45)$$

PULSE AND DIGITAL CIRCUITS
SCHEME OF VALUATION

since A ≈ 1. If R = Ri, es(Bootstrap) = esThis means that the bootstrap circuit will not provide any improvement in linearity if the input resistance of the amplifier is small. For the output of the sweep generator to be linear, Ri >> R.

13(a)

In a simple voltage sweep generator, a capacitor (C) is allowed to charge to a voltage (V) through a resistance (R) with the time constant, RC, deciding the rate of charge of the condenser. However, as the capacitor charges exponentially, the resultant sweep voltage so generated may tend to be exponential or in other words, not necessarily linear. Thus, there is a need make arrangements to linearize an exponential sweep. In such arrangements, a constant current is used to charge the capacitor. The three types of voltage sweep generators considered in this chapter include exponential sweep generators, Miller's sweep generators, and bootstrap sweep generators. A simple exponential sweep generator and its output are shown in Fig. 12.2(a) and (b), respectively. Initially, at t = 0, let the capacitor be uncharged. If now the switch S is open, then the capacitor tries to charge to the supply voltage V. At t = Ts (sweep duration), when the voltage across the capacitor is Vs, if the switch is suddenly closed, the voltage across the capacitor, ideally, is expected to abruptly fall to zero. However, if the resistance offered by the switch is ideally not zero, there is a finite time delay before the signal reaches its initial value. This time delay is called the fly-back time, restoration time or retrace time (Tr), as shown in Fig.12.2(c).
Normally, Tr << Ts, so that T ≈ Ts. The voltage variation of the sweep voltage vs is given as:
vs = vf − (vf − vi)e−t/τ
Here, vf = V and vi = 0. Therefore, vs = V − (V − 0)e−t/τ

$$v_S = V(1 - e^{-t/\tau}) \qquad (12.4)$$

We assume that after an interval Ts, when vs = Vs, the switch closes. Then the charge on the capacitor discharges with a negligible time constant and the voltage abruptly falls to zero at t = TS. From Eq. (12.4), we have:

$$\frac{dv_S}{dt} = -Ve^{-t/\tau}\left(-\frac{1}{\tau}\right) = \frac{V}{\tau}e^{-t/\tau}$$

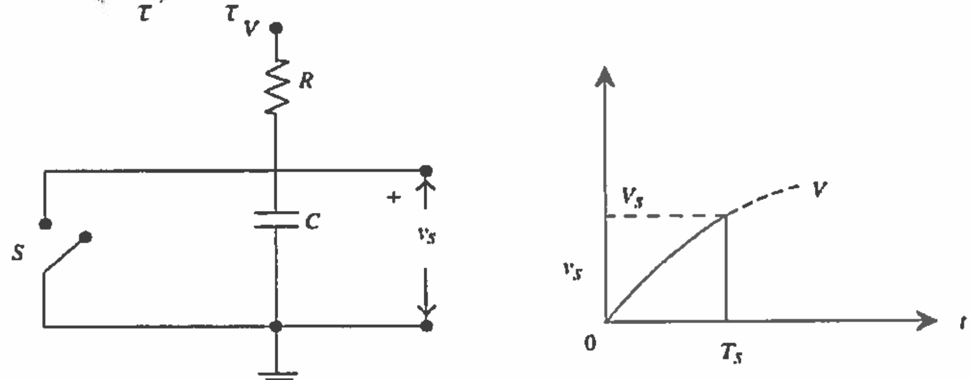


FIGURE 12.2(a) A simple exponential sweep generator; and (b) output of the sweep generator

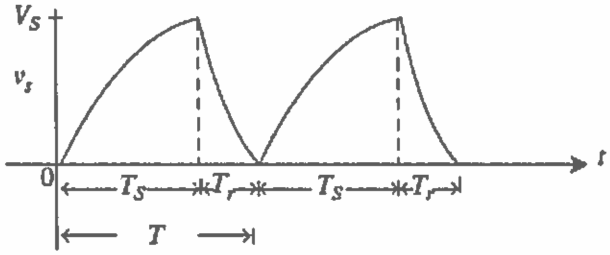


FIGURE 12.2(c) The waveform that depicts the sweep time and restoration time

The initial slope is:

$$\left.\frac{dv_s}{dt}\right|_{t=0} = \frac{V}{\tau}$$

The final slope is:

$$\left.\frac{dv_s}{dt}\right|_{t=T_S} = \frac{V}{\tau}e^{-T_S/\tau}$$

Therefore,

$$e_s = \frac{\frac{V}{\tau} - \frac{V}{\tau}e^{-T_S/\tau}}{\frac{V}{\tau}} = \left[1 - e^{-T_s/\tau}\right] \qquad (12.5)$$

From Eq. (12.4), at t = TS, vs = Vs:

Hence,

$$V_s = V\left(1 - e^{-T_S/\tau}\right) \qquad (12.6)$$

$$1 - e^{-T_S/\tau} = \frac{V_s}{V} \qquad (12.7)$$

Substituting Eq. (12.7) in Eq. (12.5) we get:

$$e_s = \frac{V_s}{V} \qquad (12.8)$$

From Eq. (12.8), it is evident that es is small when V >>Vs, i.e., linearity improves only if the supply voltage (V) is large when compared to Vs, the sweep amplitude. Therefore, the disadvantage of a simple exponential sweep is that a linear sweep is generated only when the sweep amplitude is much smaller than the applied supply voltage, V. For example if Vs = 20 V, V = 100 V

$$e_s = \frac{V_s}{V} = \frac{20}{100} \times 100\% = 20\%$$

And if Vs = 20 V, V = 1000 V

$$e_s = \frac{V_s}{V} = \frac{20}{1000} \times 100\% = 2\%$$

The above illustration explains that, for the same sweep amplitude, the smaller is the supply voltage the larger is the slope error. If the supply voltage is increased, the slope error decreases, which means linearity improves.

If t/τ << 1

$$e^{-t/\tau} = 1 - \frac{t}{\tau} + \frac{t^2}{2\tau^2} - \frac{t^3}{6\tau^3} + \cdots$$

We have from Eq.12.4:

$$v_s = V\left(1 - e^{-t/\tau}\right) = V\left[1 - 1 + \frac{t}{\tau} - \frac{t^2}{2\tau^2} + \frac{t^3}{6\tau^3}\cdots\right]$$

$$= \frac{Vt}{\tau}\left[1 - \frac{t}{2\tau} + \frac{t^2}{6\tau^2}\right] \qquad (12.9)$$

Since vs = at t = Ts, for a linear sweep, then to the first approximation.

$$V_s' = \frac{VT_s}{\tau} \qquad (12.10)$$

As this is a linear sweep:

$$e_s = \frac{V_s'}{V} = \frac{T_s}{\tau} \qquad (12.11)$$

Hence, for es to be small, τ >> Ts, i.e., the time constant employed in the circuit should be much larger than the sweep duration. If the actual sweep is non-linear, consider the first two terms given in Eq. (12.9):

$$v_s = \frac{Vt}{\tau}\left(1 - \frac{t}{2\tau}\right) \qquad (12.12)$$

Therefore, at: t = Ts:

$$V_s = \frac{VT_s}{\tau}\left(1 - \frac{T_s}{2\tau}\right) \qquad (12.13)$$

This is a non-linear sweep. Hence, the transmission error et is:

$$e_t = \frac{V_s' - V_s}{V_s'}$$

Where V_s' is the amplitude of the linear sweep and Vs is the amplitude of the non-linear sweep.

$$\therefore \; e_t = \frac{\dfrac{VT_s}{\tau} - \dfrac{VT_s}{\tau}\left(1 - \dfrac{T_s}{2\tau}\right)}{\dfrac{VT_s}{\tau}}$$

$$e_t = \frac{T_s}{2\tau} \qquad (12.14)$$

From Eq. (12.11) we have:

$$e_s = \frac{T_s}{\tau}$$

If we relate es and et:

$$e_t = \frac{T_s}{2\tau} = \frac{e_s}{2} \qquad (12.15)$$

Displacement error, ed is:

$$e_d = \frac{(v_s' - v_s)_{\max}}{V_s}$$

From Eq. (12.12)

$$v_s = \frac{Vt}{\tau}\left(1 - \frac{t}{2\tau}\right) \qquad v_s' = \frac{Vt}{\tau} \qquad (v_s' - v_s) = \frac{Vt}{\tau} \times \frac{t}{2\tau}$$

The deviation is maximum at t = (Ts/2) Therefore,

$$(v_s' - v_s)_{\max} = \frac{VT_s}{2\tau} \times \frac{T_s}{4\tau}$$

At t = TS = Vs

Therefore,

$$V_s = \frac{VT_s}{\tau}$$

Therefore,

$$e_d = \frac{(v_s' - v_s)_{\max}}{V_s} = \frac{\dfrac{VT_s}{2\tau} \times \dfrac{T_s}{4\tau}}{\dfrac{VT_s}{\tau}} = \frac{T_s}{8\tau}$$

$$e_d = \frac{1}{8}e_s \qquad (12.16)$$

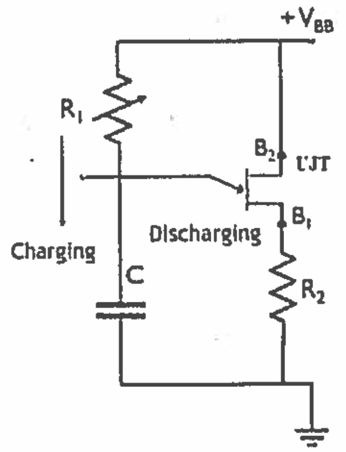
From Eqs. (12.15) and (12.16), the interrelationship between the three types of errors is given as:

$$e_d = \frac{1}{8}e_s = \frac{1}{4}e_t \qquad (12.17)$$

PULSE AND DIGITAL CIRCUITS
SCHEME OF VALUATION

13(b)

The basic UJT relaxation oscillator circuit diagram is shown below: This circuit can be built with a unijunction transistor (UJT) & a capacitor. Here the charging and discharging of this capacitor can be done through the resistors like R1 & R2 where the R1 resistor charges the capacitor and the R2 resistor discharges the capacitor. Here, VBB in the circuit is an external voltage supply.



The frequency of the UJT relaxation oscillator can be determined by the Resistor & capacitor. So, the frequency of this oscillator can be determined through the following equation.

$$F = 1/\left(RC \ln(1/(1-\eta)\right)$$

In the above equation,

η' = Intrinsic standoff ratio

ln = stand for natural logarithm.

The oscillation frequency for this oscillator can be given by F = 1/R1C. It is extremely significant to identify that the R1 resistor should include values that must be in a suitable range to oscillate to the circuit.

PULSE AND DIGITAL CIRCUITS
SCHEME OF VALUATION

The resistor values can be obtained by using these formulas

R1 max = (Vs-Vp)/Ip

14(a)

SL NO.	RTL Logic	DTL Logic	TTL Logic
1	Built with Resistor and Transistor	Built with Diode and Transistor	Built with Transistors
2	Slow Response	Better than RTL Logic	Much better than RTL and DTL
3	High Power Loss	Low Power Loss	Low Power Loss
4	Very simple in construction and operation	Simple in construction and operation	Complex in construction and operation
5	RTL Logic used in old computers	DTL logic used in basic digital circuits, switching circuits	All the modern digital circuits, Integrated Circuits are mostly built with TTL Logic

PULSE AND DIGITAL CIRCUITS
SCHEME OF VALUATION

14(b)

Propogation delay :

- The output of a logic gate does not change its state instantaneously in response to the change in the state of its input.
- There is a time delay between these two events ,which is called as the propogation delay . Thus propogation delay is defined as time delay between the instant application of an input pulse and the instant of occurrence of the corresponding output pulse

Fan Out:

- Fan out is defined as the maximum number of the inputs of the same IC family that a gate can drive without falling outside the specified voltage limits.
- Higher the fan out higher is the current supplying capacity of the gate.

Power dissipation:

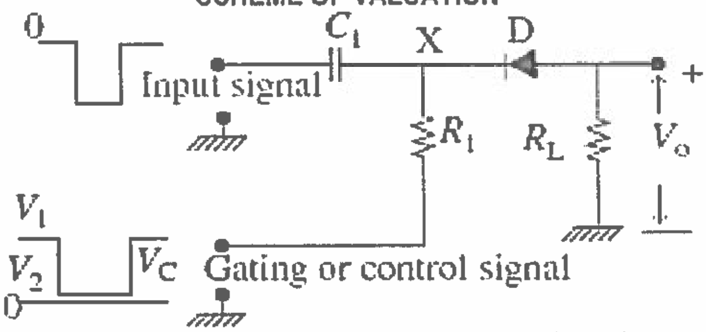
- Power dissipation is the amount of power drawn by the IC due to the current flowing current through the IC as a result of the applied voltage.

Noise Margin:

- Noise immunity is defined as the ability of logic circuit to tolerate the noise without causing any unwanted changes in the output.
- A quantitative measure of noise immunity is called as noise margin.

15(a) A **unidirectional gate** can transmit either positive or negative pulses (or signals) to the output. It means that this gate transmits pulses of only one polarity to the output. The signal to be transmitted to the output is the input signal. This input signal is transmitted to the output only when the control signal enables the gate circuit. Therefore, we discuss two types of unidirectional diode gates, namely, unidirectional diode gates that transmit positive pulses and unidirectional diode gates that transmit negative pulses.
Bidirectional sampling gates transmit both positive and negative signals. These gates can be derived using diodes, BJTs, FETs, etc. We are going to consider some variations of the bidirectional gates.

PULSE AND DIGITAL CIRCUITS
SCHEME OF VALUATION



The unidirectional diode gate to transmit negative pulses

Applications of Sampling Gates

- Sampling scopes.
- Multiplexers.
- Sample and hold circuits.
- Digital to Analog Converters.
- Chopped Stabilizer Amplifiers.

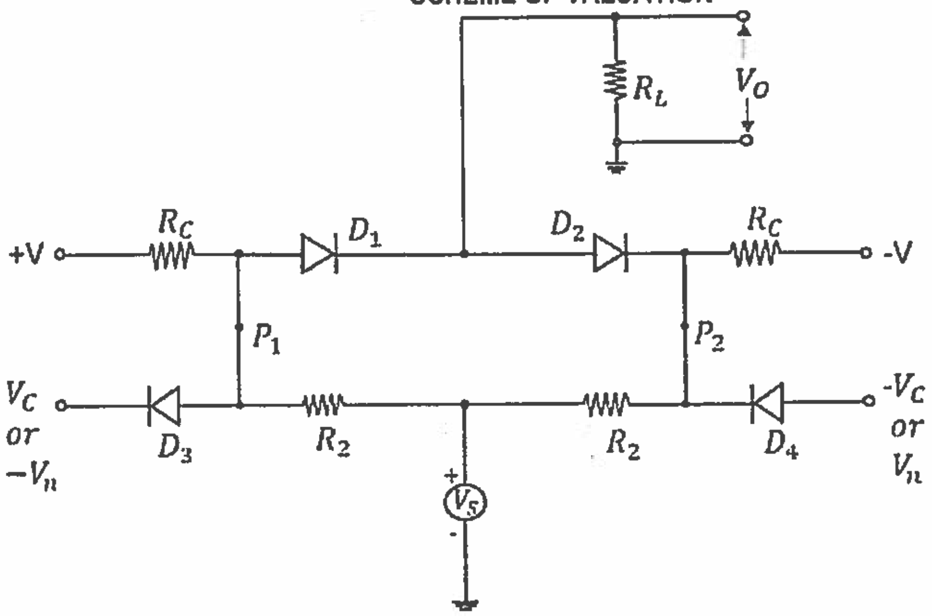
15(b)

Bidirectional sampling gate circuit is made using diodes also. A two diode bidirectional sampling gate is the basic one in this model. But it has few disadvantages such as

- It has low gain
- It is sensitive to the imbalances of control voltage
- $V_{n\,(min)}$ may be excessive
- Diode capacitance leakage is present

A four diode bidirectional sampling gate was developed, improving these features. A two bidirectional sampling gate circuit was improved adding two more diodes and two balanced voltages +v or –v to make the circuit of a four diode bidirectional sampling gate as shown in the figure.

PULSE AND DIGITAL CIRCUITS
SCHEME OF VALUATION



The control voltages V_C and $-V_C$ reverse bias the diodes D_3 and D_4 respectively. The voltages +v and –v forward bias the diodes D_1 and D_2 respectively. The signal source is coupled to the load through the resistors R_2 and the conducting diodes D_1 and D_2. As the diodes D_3 and D_4 are reverse biased, they are open and disconnect the control signals from gate. So, an imbalance in control signals will not affect the output.

When the control voltages applied are V_n and $-V_n$, then the diodes D_3 and D_4 conduct. The points P_2 and P_1 are clamped to these voltages, which make the diodes D_1 and D_2 revere biased. Now, the output is zero.

During transmission, the diodes D_3 and D_4 are OFF. The gain A of the circuit is given by

A=RCRC+R2×RLRL+(Rs/2)A=RCRC+R2×RLRL+(Rs/2)

Hence the choice of application of control voltages enables or disables the transmission. The signals of either polarities are transmitted depending upon the gating inputs.

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

II B.Tech II Semester Regular End Examination

Python Programming (20CS403)

(Common to CSE/Mechanical)

Part A (Short Answer Question 5X2 =10 Marks)

1. Represent Python Program Development Cycle [2M]

Program Development Cycle figure should mention Customer request, analysis, design, implementation, integration and maintenance – 2 M

There are several approaches to software development of which one version is known as the waterfall model.

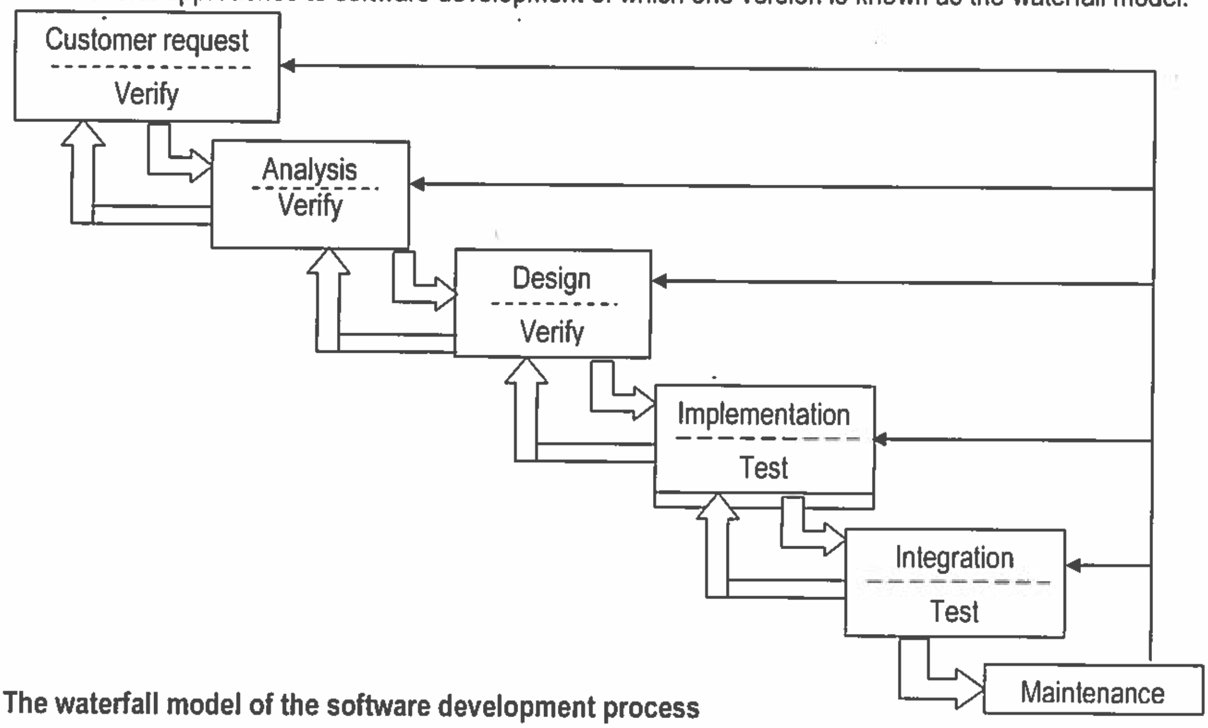


Fig. The waterfall model of the software development process

2. Develop a code to print "NSRIT" 5 times [2M]

Use either for or while loop to print NSRIT 5 times—2M

```
for count in range(5) :          (or)     c = 1
    print("NSRIT")                        while c<=5 :
                                              print("NSRIT")
                                              c = c + 1
```

3. List any four machine learning libraries that can be installed using PIP. [2M]

Mention any four machine learning libraries like Numpy, Pandas, Scikit-Learn, Matplotlib, Keras, etc —2M

4. Differentiate between class and object [2M]

Write atleast any two important differences between class and object which includes definition—2M

Class	Object
1. A template for creating or instantiating objects within a program or blue print of an object	1. An instance of a class is a object
2. A class is a logical entity	2. An object is a physical entity
3. A class does not get any memory when it is created.	3. Object gets memory when they are created
4. A class is declared once	4. Multiple objects are created using a class

5. Write the function of Matplotlib and GNUplot [2M]

Function of Matplotlib—1M and GNUplot—1M

Matplotlib is an amazing visualization library in Python for 2D plots of arrays. Matplotlib is a multi-platform data visualization library built on NumPy arrays and designed to work with the broader SciPy stack.

GNUplot is a Python package that interfaces to gnuplot, the popular open-source plotting program. It allows you to use gnuplot from within Python to plot arrays of data from memory, data files, or mathematical functions.

Part B (Long Answer Question 5X12 =60 Marks)

6 (a) Explain any three keywords with an example [6M]

Mention any three keywords in Python —3M Give example for each keyword—3M

Python has special reserved words, or keywords, that have specific meanings and purposes and can't be used for anything but those specific purposes.

Keywords: break, continue, else, if, in, pass, elif, for, True, while, None, def, class, as, and, or etc

The pass statement is used as a placeholder for future code. When the pass statement is executed, nothing happens, but you avoid getting an error when empty code is not allowed. Empty code is not allowed in loops, function definitions, class definitions, or in if statements.

Example:
```
def myfunction():
        pass
```
or
```
if a < 0 :
        pass
```

The Python break statement stops the loop in which the statement is placed. A Python continue statement skips a single iteration in a loop. Both break and continue statements can be used in a for or a while loop.

Write a program which prints the number from 1 to 9 and but not 6.

Example:
```
for k in range(1, 10) :
        if k == 6 :
            continue
        else :
            print(k, end = " ")
```
Output : 1 2 3 4 5 7 8 9

Write a program which loops from 1 to 20 and exits when number becomes a two digit value

Example:
```
for k in range(1, 21) :
        if k >9 :
            break
        else :
            print(k, end = " ")
```
Output : 1 2 3 4 5 6 7 8 9

6 (b) (i) Develop a code using rawinput () function to read the input from keyboard. [3M]

(ii) Develop a code to output multiple variables using "+" operator. [3M]

Sample Code using raw_input( ) in Python —3M

There are two versions of input functions used in Python. The **input( )** is used to read data from a standard input in Python 3 and **raw_input( )** is used to read data from a standard input in Python 2.

Python 2 included the built-in function raw_input(), to get to prompt command-line users for a user input string. This function prompts the user to type in some text at the command line and returns that text as a string.

```
fruit = raw_input( )
apples
>>> print(fruit)
apples
num = raw_input()
18
>>> num
'18'
>>> type(num)
<type   'str'>
```

(ii) Develop a code to output multiple variables using "+" operator. [3M]

Sample Code to output multiple variables in Python —2M and using + operator to display them—1M

```
x = 5
y =32
z = "Hello"
msg = "Good Morning"
print(z+" "+msg)  # prints Hello Good Morning
print(" Value of x = "+str(x)+ " and y value = "+str(y)) # prints Value of x =5 and y value = 32
```

7 (a) Develop a python code to perform arithmetic operations.
Note:(i) Use input ( ) function to get user input
(ii) format ( ) function to format the string and print the result with statements. [6M]

Using input( ) function to get user input and Arithmetic operations such as +,-,*,/,% and ** operators —4M

Using format( ) function to format the string and print result with statements—2M

```
a = int(input(" Enter  first  number: "))
b = int(input(" Enter  second  number: " ))
print(" Arithmetic operations are add, subtract, multiply, divide and exponent ")
print(" { } + { } = ".format(a,b),a + b)
print(" { } - { } = ".format(a,b),a - b)
print(" { } x { } = ".format(a,b),a * b)
print(" { } / { } = ".format(a,b),a // b)
print(" Floating Quotient { } / { } = ".format(a,b),a / b)
print(" Remainder of { } mod { } = ".format(a,b),a % b)
print(" Exponent result  of { } ^ { } = ".format(a,b),a ** b)
```

7 (b) Explain the logical operators with an example. [6M]

Logical operators in Python are **and, or, not** —4M + Simple example for logical operators—2M

Logical operators are used to combine conditional statements:
The logical operators used in Python are **and, or , not.**
To check multiple conditions at the same time we use logical operators in Python.

Operator	Description	Example
and	Returns **True** if both the statements are True, otherwise it returns **False**	**A, B,C = 10, 20, 30** **if A < B and A < C:** print("Smallest number =",A) # prints Smallest number = 10 as A<B and A<C
or	Returns **True** if either of the statements are True, otherwise it returns **False**	A = 40 if A >5 or A<31: print("Valid range") # prints Valid range and condition is True # as value of A 40 > 5 though A<31 fails
not	Reverse the result, returns False if the result is true	**x = 5** print(not(x > 3 and x < 10)) # returns False because not is used to reverse the result

8 (a) Develop the python code to find perimeter of square [6M]

Code to print perimeter of square with input taken from user—6M

```
# Print perimeter of a square
a  = int(input(" Enter side of a square : "))
perimeter = 4 * a # Here a is side of a square
print(" Perimeter of a square = ",perimeter)
```

8 (b) Develop the python code to print the numbers in the following pattern [6M]

```
2
4  6
8  10  12
14  16  18  20
```

Code to print numbers in the pattern as given above using for or while loop—6M

```
even = 2
for col in range(4):
   for x in range(col+1):
        print(even, " ",end=" ")
        even = even + 2
  print( )
```

[or]

```
even,count,outer = 1,1,1
while outer<=4:
         while inner< =outer:
                  inner = 1
                  even = count  * 2
                  print(x," ",end=" ")
                  inner = inner + 1
                  count = count + 1
         print( )
         outer = outer + 1
```

9 (a) Develop the python code to input any alphabet and check whether it is vowel or not [6M]

Python code to find if entered input alphabet matches vowels ('a','e','i','o','u')—6M

```
letter = input("Enter any alphabet: ")
vowels = ['a', 'e', 'i', 'o', 'u']
a = letter.lower()
print("Given input alphabet : ",a)
if a in vowels:
  print(" Vowel ")
else:
  print(" Not a vowel ")
```

9 (b) Distinguish between the list and tuples in terms of methods, iteration and memory consumption [6M]

List and Tuples (Definition + mutability property-2M, representation and iteration-1M, methods-2M, memory consumption-1M)—2+1+2+1→6M

Parameter	Lists	Tuples
Definition	Lists tore one or more objects or values in a specific order.	A Tuple is a collection of Python objects separated by commas.
Representation	Literal syntax of lists is shown by square brackets [ ]	Literal syntax of tuples is shown by parentheses ()
Mutable nature	Lists are mutable which means the elements can be changed or modified after its creation	Tuples are immutable which means the elements cannot be changed or modified after its creation

Mutable/Immutable Example	a = ['Car', 'Bike', 'Oven', 'TV']
	#will print Car, Bike, Oven, TV
	a[1]=20
	a[3]=5.6
	# It will print Car, 20, Oven, 5.6
	a.append(32)
	will add 32 at end of the list
	#prints Car, 20, Oven, 5.6, 32
	del a[0] will remove first element which is Car
	print(a) will output 20 , Oven 5.6
		z = ('Vizag', 'Chennai', 'Delhi', 'Hyd',)
		#will print Vizag, Chennai, Delhi, Hyd
		# z[0]=20 will print tuple object does not
		# support item assignment
		No items or elements can be modified as shown above and no element can either be added.

Iteration	List iteration is slower	Tuple iteration is faster compared to lists
Size comparision	Lists are mutable hence it has variable length, It occupies more size in memory as compared to tuples	As tuples are immutable they have fixed length It occupies less memory size as compared to lists
Where it should be used	Lists are not used as key in a dictionary because list can't handle __hash__() and have mutable nature.	Tuple can also be used as key in dictionary due to their hashable and immutable nature
Dynamic nature	Lists are dynamic as they can grow or shrink in size	Tuples are not dynamic as they cannot grow or shrink in size and hence they are fixed

10 (a) Develop the python code to find maximum and minimum between two numbers using functions [6M]

Use either user defined function to find maximum and minimum between 2 numbers—max( )-3M+min( )-3M

Or

Use built in methods max and min function of a list and find maximum and minimum from list- **(3+3=6M)**

```
# Print maximum and minimum between two numbers using functions
def max_min(a,b):
  if a >b:
    max = a
  else:
    max = b
  if a <b:
    min = a
  else:
    min = b
  return max,min

a = int(input(" Enter first value: "))
b = int(input(" Enter second value: "))
max,min=max_min(a,b)
print("Maximum is ",max)
print("Minimum is ",min)
```

10 (b) Explain any three functions of module with an example. [6M]

Module definition and three functions of modules—3M, Example of module—3M

Modules:

A file containing a set of functions you want to include in your application is said to be a module.

A module can define functions, classes and variables.

Modules are used to break down large programs into small manageable and organized files.

Modules provide reusability of code.

A module allows you to logically organize your Python code. Grouping related code into a module makes the code easier to understand and use

A module is loaded only once, regardless of the number of times it is imported. This prevents the module execution from happening over and over again if multiple imports occur.

A file containing Python code, for example sample.py, is called a module, and its module name would be sample.

We can use a module by using import statement.

A module can also include runnable code.

Example

Here's an example of a simple module, **support.py**

```
def print_branch( name ):
  print "Hello  ", name
  return
# include module support as follows
```

import support

Now you can call defined function of that module as follows

support.**print_branch("Mech students")** # prints Hello **Mech students**

support.**print_branch("CSE students")** # prints Hello **CSE students**

11 (a) Develop the python code accepts roll number and returns whether the student is present or absent. [4M]

Code to find whether student is present or absent from a list using in —4M

```
check_rno = int(input("Enter the student roll number"))
rollnumber = [1,3,23,55,67,2,44,67,68,32,25,10,9,8,12]
print("Roll numbers in list is ")
print(rollnumber)
if check_rno in rollnumber :
   print("\n Roll number "+str(check_rno)+ " is present ")
else :
   print(" \n Roll number "+str(check_rno)+ " is absent ")
```

11 (b) Interpret the Math module with an example. [8M]

Math module purpose—2M and explain atleast 5 functions in math module with examples—6M

Math module:

The math module is a standard module in Python and is always available. To use mathematical functions under this module, you have to import the module using import math.

When working with some kind of financial or scientific projects it becomes necessary to implement mathematical calculations in the project. Python provides the math module to deal with such calculations. Math module provides functions to deal with both basic operations such as addition(+), subtraction(-), multiplication(*), division(/) and advance operations like trigonometric, logarithmic, exponential functions.

```
# Program to demonstrate the math module functions-sqrt(), ceil, euler's number, pi & absolute value using fabs
# import the math module
import math
# print the square root of 4 ; print(math.sqrt(4)) # prints 2.0
# pi is depicted as either 22/7 or 3.14. math.pi provides a more precise value for the pi
print(" Pi value is : ",math.pi) # prints Pi value is  3.141592653589793.
# pow method returns b**e where b is base and e is exponent
print (" The value of 3**4 is : ",end=" ")   ; print (pow(3,4)) #  returns 81
print ("The ceil of 2.3 is : ", end=" ") ; print (math.ceil(a)) # prints The ceil of 2.3 is : 3
print ("The floor of 2.3 is : ", end=" "); print (math.floorl(a))    # prints The floor of 2.3 is : 2
print("The euler's number is "math.e)    # prints 2.71828182846
x = -10 ; radius=math.fabs(x)     # returns positive value of x so radius value is 10
print("Area of circle is ",math.pi*radius*radius)
```

print("Factorial of 4 is ",math.factorial(4)) # prints Factorial of 4 is 24

12 (a) Interpret the different functions of file with an example [6M]

File purpose—2M+Different functions of files like open( ),close( ),read( ),write( ) and readline( ) with example—4M
Files are named locations on disk to store related information. They are used to permanently store data in a non-volatile memory(hard disk).
Before performing any operations such as read from or write to a file, first we need to open the file. When we are done, it needs to be closed so that the resources that are tied with the file are freed.
Hence, in Python, a file operation takes place in the following order:
1. Open a file
2. Read or write (perform operation)
3. Close the file

Opening files in Python
Python has a built-in open() function to open a file. This function returns a file object, also called a handle, as it is used to read or modify the file accordingly.
f1 = open("test.txt") # open file in current directory
The mode is specified while opening a file- whether we want to read r, write w or append a to the file.
f2 = open("E:/Python38/sampledata.txt","r") # specifying full path for sampledata.txt in read mode

Writing to Files in Python
In order to write into a file in Python, we need to open it in write w, append a or exclusive creation x mode
Writing a string or sequence of bytes (for binary files) is done using the write() method. This method returns the number of characters written to the file.

Reading Files in Python
To read a file in Python, we must open the file in reading mode (**"r"**)

We can use the read(size) method to read in the size number of data. If the size parameter is not specified, it reads and returns up to the end of the file.
Alternatively, we can use the readline() method to read individual lines of a file. This method reads a file till the newline, including the newline character.
Lastly, the readlines( ) method returns a list of remaining lines of the entire file. All these reading methods return empty values when the end of file (EOF) is reached.

```
file = open("test.txt","w")    # open test.txt file in current directory with write mode
file.write("This file\n\n")
file.write("contains two lines\n")
# It will create a new file named test.txt in the current directory if it does not exist. If it does exist, it is overwritten.
f = open('test.txt','r')
f.read(4)    # read the first 4 data
f.read(4)    # read the next 4 data
f.readline( )
f.readline( )
f.readline( )
print("Using readlines method for printing below : ")
f.readlines( )
```

Output:

'This'

' is '

'This is my first file\n'

'This file\n'

'contains two lines\n'

Using readlines method for printing below :

['This is my first file\n', 'This file\n', 'contains three lines\n']

12 (b) Explain how to create a constructor in Python? Give an example. [6M]

Constructor definition and its purpose—1M+ Creating constructor—2M+Example—3M

A constructor is a special type of method (function) which is used to initialize the instance members of the class.

In Python, the method the **__init__()** simulates the constructor of the class. This method is called when the class is instantiated. It accepts the **self**-keyword as a first argument which allows accessing the attributes or method of the class. We can pass any number of arguments at the time of creating the class object, depending upon the **__init__()** definition.

```
class Employee:
  def __init__(self, name, empid) :
     self.empid = empid
     self.name = name
   def display(self):
     print("Employee Id: { } \nName: { }" .format(self.empid, self.name))

emp1 = Employee("Vasu", 101)
# accessing display() method to print employee 1 information
emp1.display( )
```

13 (a) Develop the python code to depict multiple inheritance. [6M]

Multiple inheritance definition with explanation—2M+ Any example of multiple inheritance with code—4M

A class can be derived from more than one base class in Python. This is called multiple inheritance.
In multiple inheritance, the features of all the base classes are inherited into the derived class.

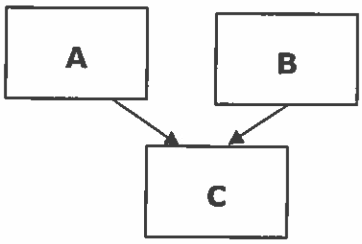


Fig. Multiple inheritance

In the above figure class C is inherited both base classes A and B

Syntax:

```
class Base1:
    Body of the class
class Base2:
   Body of the class
class Derived(Base1, Base2) :
   Body of the class
```

Multiple Inheritance Example:

```
class Father:
        def show1(self) :
```

```python
            print("Base class Father show1 function")
class Mother:
        def show2(self) :
            print("Base class Mother show2 function")

# Derived class child defined here as follows:
class Child(Father,Mother):
        def show3(self) :
            print("Derived class Child show3 function")
object =  Child( )
object. show1( )
object. show2( )
object. show3( )
```

13(b) Explain the operator overloading in Python with example. [6M]

Operator overloading in Python concept—2M+ Example of operator overloading—4M

The operator overloading in Python means provide extended meaning beyond their predefined operational meaning. Such as, we use the "+" operator for adding two integers as well as joining two strings or merging two lists. We can achieve this as the "+" operator is overloaded by the "int" class and "str" class.

```python
# Python Program illustrate how to overload an binary + operator
class A:
   def __init__(self, a):
      self.a = a
    # adding two objects
   def __add__(self, z):
      return self.a + z.a
ob1 = A(10)
ob2 = A(20)
ob3 = A("NSRIT-")
ob4 = A("Sontyam")
print(ob1 + ob2)
print(ob3 + ob4)
Output:
30
NSRIT-Sontyam
```

14 Explain the following terms:
(i) Types of Variables in Scratch
(ii) Use of Variable in Scratch [12M]

Types of variables in Scratch—Global, local and cloud-6M+Use of variables in Scratch—6M

Scratch is a high-level visual programming language tool that interacts with users through diagrams and blocks that have the basics of a program inbuilt in it.

A variable is a changeable value recorded in Scratch's memory. Variables can only hold one value at a time, unlike lists. Variables are like containers that can hold a number (numerical variables) or a word (alphanumerical variables).In Scratch there are three different types of variables :

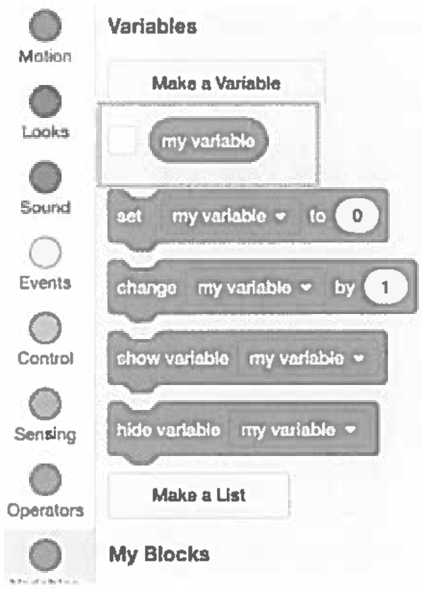
(1) Global (2) Local (3) Cloud

Global: It is the default variable. It means that it can be changed or accessed from any sprite in the project or stage, regardless of which sprite it was created on. Scratch allows the user or programmer to select a global variable by showing an option " choose for all sprites", or "choose and for this sprite only?". If the user or programmer chooses "choose for all sprites", then that variable becomes global as it can be accessed by anyone whoever needs it. All the global variables are stored in RAM and they are the default for those files in which they are created.

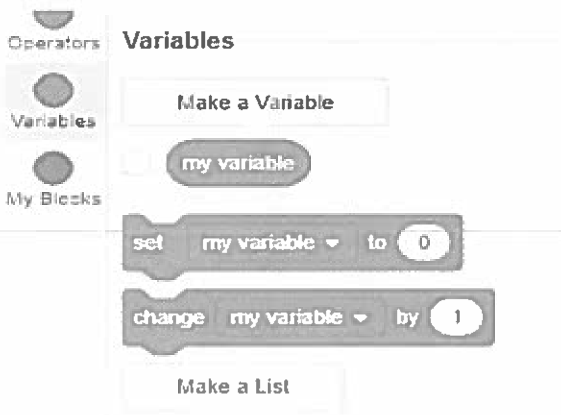
Local: It is one that can only be changed or accessed from the sprite on which it was created. Scratch allows the user or programmer to select a local variable by showing an option " choose for all sprites", or "choose and for this sprite only?". If the user or programmer chooses "choose for this sprite only", then that variable becomes local and only the current sprite has the access to it.

Cloud: It is a variable that allows users to store variables on the server of the scratch. Cloud variables have the cloud-like symbol in front of the variable name and they update themselves very quickly.

In Scratch, you can create variables in two different ways:
1. **Using build-in variable:** Scratch has an inbuilt variable named "my variable", so the users can directly use that.



2. **User-define variable:** In case the user wishes to make his own variable, with a different name, then click on the "make a variable" button in the variable palette.
Click on **Variables** in the Code tab, then click on **Make a Variable**



Type in the name of your variable. You can choose whether you would like your variable to be available to all sprites or to only this sprite. Press **OK.** After clicking the "make a variable block" a form will appear on the screen.

After clicking " **OK** ", the image that follows depict that variable has been created.

15. Explain the following terms:
(i) tkinter module in Python GUI.
(ii) Explain any 6 functions in NumPy with example. [12M]

tkinter module in Python GUI—6M

tkinter module in Python:
tkinter is the standard GUI library for Python. Creating a GUI application using Tkinter is an easy task. Foundational element of a Tkinter GUI is the "window". Windows are containers in which all other GUI elements live. These other GUI elements such as text boxes labels and buttons are known as widgets. Widgets are contained inside of windows.
Perform the following steps :-

1) Import the Tkinter module.

2) Create the GUI application main window.

3) Add one or more of the above-mentioned widgets to the GUI application.

4) Enter the main event loop to take action against each event triggered by the user.

There are two main methods used which the user needs to remember while creating the Python application with GUI.
1) Tk(screenName=None, baseName=None, className='Tk', useTk=1)

2) mainloop()
To create a main window, tkinter offers a method **Tk**

Tk(screenName=None, baseName=None, className='Tk', useTk=1).

To change the name of the window, you can change the className to the desired one.

The basic code used to create the main window of the application is

import Tkinter

top = Tkinter.Tk( )where top is the name of the main window object

There is a method known by the name mainloop() is used when your application is ready to run.

mainloop() is an infinite loop used to run the application, wait for an event to occur and process the event as long as the window is not closed.

Code to add widgets will go here...

top.mainloop( )

Tkinter Widgets

The term "Widgets" is a generic term that refers to the building blocks that make up an application in a graphical user interface. Tkinter provides various controls, such as buttons, labels and text boxes used in a GUI application. These controls are commonly called widgets.

Let us list out the core widgets with their categories:

Container : Under this category, the widgets that lies are frame, labelframe, toplevel, and paned window.

Buttons : Under the category of Buttons, there are buttons, radiobuttons, checkbuttons (checkbox), and menubuttons.

Text Widgets : Under the category of text widgets, there are labels, messages, text.

Entry Widgets : Under this category, the widgets are scale, scrollbar, Listbox, slider, spinbox, entry (single-line), optionmenu, text (multiline), and Canvas (vector and pixel graphics).

#Tkinter Basic Example

import tkinter as tk

win = tk.Tk( )

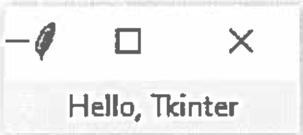
greeting = tk.Label(text="Hello, Tkinter") # Create a label with words or text Hello Tkinter

The window you created earlier doesn't change. You just created a Label widget, but you haven't added it to the window yet.We can use the Label widget's pack( ) method:

greeting.pack( ) # Put the label into the window

win.mainloop( )# Start the event loop

win.mainloop( ) tells Python to run the Tkinter event loop. This method listens for events, such as button clicks or keypresses, and blocks any code that comes after it from running until you close the window where you called the method.



15 (ii) Write any six functions in NumPy with examples-Each function in Numpy with example-1M[1X6=6M]

NumPy is a Python package for scientific computing that provides high-performance multidimensional array objects Numpy (Numerical Python) is an open-source core Python library for scientific computations. To use any package or library in our code, it needs to be made accessible. So we use import statement.

import numpy as np # we can use any name as alias for numpy like n or a or anything need not be np only

Before using any Numpy function

Six functions of Numpy are as follows:

1) **numpy.array() :** We can create a NumPy ndarray object by using the array() function. The array object in NumPy is called ndarray. It is basically a table of elements which are all of the same type and indexed by a tuple of positive integers.

Syntax: numpy.array(object, dtype=None, copy=True, order='K', subok=False, ndmin=0)

2) **linspace( )** function returns evenly spaced numbers over a specified interval defined by first two arguments of the function start and stop.

Syntax:linspace(start,stop,num=50,endpoint=True,retstep=False,dtype=None,axis=0)

3) **around(arr,decimals=0,out=None) :** This mathematical function helps user to evenly round array elements to the given number of decimals. Here out is the output resulted array which is optional parameter.

4) **arange:** It creates an array by using the evenly spaced values over the given interval.

Syntax:numpy.arange(start, stop, step, datatype)

5) **sum** function is used to compute the sum of all elements. It is also possible to add rows and column elements of an array. The output will be in the form of an array object.

Syntax:

numpy.sum(arr, axis=None, dtype=None, out=None, keepdims=<no value>, initial=<no value>)

6) zeros( ): It is used to get a new array of given shape and type filled with zeros

Syntax:

numpy.zeros(arr,dtype=None, order='K',subok=True)

```
# Program explaining linspace( ),around(),arange(),array() sum(), zeros(), max() etc

import  numpy as np
a = np.array( [ [10,9,4],[4,3,2] ] )
print(" Number of dimensions = ",a.ndim) # prints Number of dimesions = 2
print(" Shape of array = ",a.shape) # prints Shape of array =( 2,3)
print(" Size of array = ",a.size) # Prints Size of array as 6 (since total 6 elements are there)
print(" Sum of all array elements = ",a.sum( )) # prints sum of all array elements = 32
print(" Maximum element in array = ",a.max( )) # prints Maximum element in array = 10
input = [.53, 1.54, .71]
print("Input array :")
print(input)
# Displays Input array : [0.53, 1.54, 0.71]
output = np.around(input)
print(" Rounded values ")
# Displays Rounded values : [1, 2, 1]
M = np.linspace(0,5,6)
# An array with 5 values between 0 and 5
# [ 0,   0.55555556        1.11111111        1.6666667         2.2222222 ]
print(" A sequential array with steps of 4 : ")
M = np.arange(0,20,4)
# [0    4      8      12     16 ]
x = np.zeros((2,3))
print("Array initialized with zeros \n",x)
# [ [0,0,0]
#   [0,0,0] ]
```

NSRIT

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	**Program**	Mechanical Engg. / CSE	**Academic Year**	2021 - 2022
Course Code	20CS403	**Test Duration**	3 Hrs. Max. Marks 70	**Semester**	IV
Course	**Python Programming**				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Represent Python Program Development Cycle.	20CS403.1	L1
2	Develop a code to print "NSRIT" 5 times.	20CS403.2	L2
3	List any four machine learning libraries that can be installed using PIP.	20CS403.3	L1
4	Distinguish between class and object.	20CS403.4	L2
5	Write the function of Matplotlib and GNUplot.	20CS403.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain any three keywords with an example.	6M	20CS403.1	L2
6 (b)	(i) Develop a code using rawinput () function to read the input from keyboard. (ii) Develop a code to output multiple variables using "+" operator.	6M	20CS403.1	L3
	OR			
7 (a)	Develop a python code to perform arithmetic operations. Note:(i) Use input () function to get user input (ii) format () function to format the string and print the result with statements.	6M	20CS403.1	L3
7 (b)	Explain the logical operators with an example.	6M	20CS403.1	L2
8 (a)	Develop the python code to find perimeter of square.	6M	20CS403.2	L3
8 (b)	Develop the python code to print the numbers in following pattern. 2 4 6 8 10 12 14 16 18 20	6M	20CS403.2	L3
	OR			
9 (a)	Develop the python code to input any alphabet and check whether it is vowel or not.	6M	20CS403.2	L3
9 (b)	Distinguish between the list and tuples in terms of methods, iteration and memory consumption.	6M	20CS403.2	L2
10 (a)	Develop the python code to find maximum and minimum between two numbers using functions	6M	20CS403.3	L3
10 (b)	Explain any three functions of module with an example.	6M	20CS403.3	L2
	OR			
11 (a)	Develop the python code accepts roll number and returns whether the student is present or absent.	4M	20CS403.3	L3
11 (b)	Interpret the Math module with an example.	8M	20CS403.3	L2

12 (a)	Interpret the different functions of file with an example.	6M	20CS403.4	L2
12 (b)	Explain how to create a constructor in Python? Give an example.	6M	20CS403.4	L2
	OR			
13(a)	Develop the python code to depict multiple inheritance.	6M	20CS403.4	L3
13(b)	Explain the operator overloading in Python with example.	6M	20CS403.4	L2
14	Explain the following terms: (i)Types of Variables in Scratch (ii)Use of Variable in Scratch	12M	20CS403.5	L2
	OR			
15	Explain the following terms: (i) tkinter module in Python GUI. (ii) Explain any 6 functions in NumPy with example.	12M	20CS403.5	L2

NSRIT

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	**Program**	EEE/CSM/CSD	**Academic Year**	2021 - 2022
Course Code	20BSX15	**Test Duration**	3 Hrs. **Max. Marks** 70	**Semester**	IV
Course	**Probability and Statistics**				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Course Outcomes	DoK
1	Find the median of the marks of students in a class 60, 72, 96, 28, 35, 10, 40, 09, 85, 25	20BSX15.1	L2
2	Out of 800 families with 5 children each, how many families would be expected to have 3 boys? Assume equal probabilities for boys and girls	20BSX15.2	L3
3	Define the Sampling distribution of a statistic.	20BSX15.3	L2
4	Write the test statistic to test to the difference of two means in small samples.	20BSX15.4	L1
5	What is the difference between positive and negative correlation?	20BSX15.5	L2

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Course Outcomes	DoK
6 (a)	Calculate the Arithmetic mean and Standard deviation of the following continuous frequency distribution	8	20BSX15.1	L2
6 (b)	Find the Coefficient of Variation of the following data 3,8,6,10,12,9,11,10,12,7	4	20BSX15.1	L2

Table for 6 (a):

Class Interval	20-30	30-40	40-50	50-60	60-70	70-80	80-90
Frequency	3	61	132	153	140	51	2

OR

No.	Question	Marks	Course Outcomes	DoK
7	Calculate the Karl Pearson's coefficient of Skewness for the following data	12	20BSX15.1	L2

Variable	0-10	10-20	20-30	30-40	40-50	50-60	60-70
Frequency	5	6	11	21	35	30	22

8 (a) State and Prove Baye's Theorem. — Marks: 6 — 20BSX15.2 — L2

8(b) The probabilities of X, Y, Z becoming managers are $\frac{4}{9}, \frac{2}{9}$ and $\frac{1}{3}$ respectively. The probabilities that the bonus scheme will be introduced if X,Y and Z becomes managers are $\frac{3}{10}, \frac{1}{2}$ and $\frac{4}{5}$ respectively.
What is the probability that
(i) The bonus scheme will be introduced
(ii) If the bonus scheme has been introduced, what is the probability that the manager appointed was Y?
— Marks: 6 — 20BSX15.2 — L3

OR

9 (a) A continuous random variable X has the distribution function

$$F(x)=\begin{cases} 0, & if\ x \le 1 \\ k(x-1)^4 & if\ 1 < x \le 3 \\ 1 & if\ x > 3 \end{cases}$$

Determine i) f(x) ii) k iii) Mean(X)
— Marks: 6 — 20BSX15.2 — L3

9(b) If a random variable X has a poisson distribution such that P(1)=P(2), find i) Mean of the distribution ii) P(4) iii) P(X≥1) iv)P(1<X<4)
— Marks: 6 — 20BSX15.2 — L3

Q.No	Question	Marks	CO	BL
10	A Population consists of five members 2,3,6,8 and 11.Consider all possible samples of size two each can be drawn with replacement from the population find (a) Population mean (b) Standard deviation of the population (c) The mean of the sampling distribution of means (d) The Standard deviation of the sampling distribution of means	12	20BSX15.3	L3

OR

11 (a)	Define the following terms i) Population ii) Sample iii) Parameter iv) Statistic v) Standard Error of a Statistic	5	20BSX15.3	L1
11(b)	Measurements of the weights of a random sample of 200 ball bearings made by a certain machine during one week showed a mean of 0.824 and a standard deviation of 0.042. Find the maximum error at 95% confidence level. Also find the 95% confidence limits for the true mean.	7	20BSX15.3	L3

12 Two horses A and B were tested according to the time (in seconds) to run a particular track with the following results. — 12 — 20BSX15.4 — L3

Horse A	28	30	32	33	33	29	34
Horse B	29	30	30	24	27	29	

Test whether the two horses have same running capacity. (Table Value of t = 2.2)

OR

13 (a) Define i) Critical region ii) Level of Significance in hypothesis testing. — 4 — 20BSX15.4 — L1

13(b) The following table shows the number of air accidents of each day of a week. Test whether these accidents are uniformly distributed over a week — 8 — 20BSX15.4 — L3

Day	Sun	Mon	Tue	Wed	Thu	Fri	Sat
Accidents	147	125	160	118	149	128	150

(Chi-Square at 6df = 12.59

14 Obtain the regression lines of Y on X and X on Y from the following table and estimate the blood pressure when the age is 45 years. — 12 — 20BSX15.5 — L3

Age in years(X)	56	42	72	36	63	47	55	49	38	42	68	60
Blood pressure (Y)	147	125	160	118	149	128	150	145	115	140	152	155

OR

15(a) Write the normal equations to fit an straight line using the principle of least squares. — 3 — 20BSX15.5 — L2

15(b) Fit a second degree parabola to the following data using the principle of least squares — 9 — 20BSX15.5 — L3

X	1.0	1.5	2.0	2.5	3.0	3.5	4.0
Y	1.1	1.3	1.6	2.0	2.7	3.4	4.1

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N S RAJU INSTITUTE OF TECHNOLOGY
(AUTONOMOUS)
SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION (20BSX15)

— Prepared by
Dr UVSubbaRao
Professor in Statistics

Note:

* There are various methods in calculating standard deviation, mean, mode assign marks to any method when the value (answer) coincides.
* Question 8(b) can also be solved directly by using conditional probability
* Questions (12) can also be solved by using variances; consider that one also by using F distribution
* Question (14) can also be solved by using deviation method please consider
* Assign marks to the definitions sampling distribution, critical region, level of significance if they explained by means of Examples.
* table values are given at the end of the problem if required

Scheme of P&S — 20BSX15

— Dr. U.V.Subbbe
Professor

(1) Median of 60, 72, 96, 28, 35, 10, 40, 09, 85, 25

In ascending order 9, 10, 25, 28, 35, 40, 60, 72, 85, 96 → (1M)

$n = 10$ even

Median = average of $\frac{n}{2}$ and $\frac{n}{2}+1$ terms

$$= \frac{35+40}{2} = \frac{75}{2} = 37.5 \rightarrow (1M)$$

(2) Here $N = 800$; $n = 5$; $p = 1/2$ $q = 1 - p = 1 - 1/2 = 1/2$

$P(X=x) = {}^{n}C_{x}\, p^{x} q^{n-x}$ $x = 0 \ldots n$

$P(X=3) = {}^{5}C_{3} (1/2)^3 (1/2)^2 = 0.3125 \rightarrow (1M)$

expected number of families $= NP(x) =$
$800 \times 0.3125 = 250$ (1M)

(3) A sampling distribution is a statistic that is arrived out through repeated sampling from a larger population. It describes a range of possible out comes of a statistic such as mean or mode of some variable, as truly exists from the population (2M)

(4) Test statistic for difference of means in small samples

$$t = \frac{\bar{x} - \bar{y}}{S\sqrt{\frac{1}{n_1} + \frac{1}{n_2}}} \sim t_{n_1+n_2-2} \text{ df} \rightarrow (1M)$$

where pooled variance $S = \frac{n_1 s_1^2 + n_2 s_2^2}{n_1+n_2-2}$ (1M)

(5) In a bivariate distribution the change of the two variables is in the same direction it is called positive correlation

X	Y		X	Y
↑	↑		↓	↓

(1+1M)

Ex: X: Demand
Y: Supply

If the deviation is in the opposite direction it is called negative correlation

X: Price
Y: Demand

X	Y		X	Y
↑	↓		↓	↑

6 a)

C.I	fr	midvalue (x)	$d = \frac{X-A}{h} = \frac{X-55}{10}$; A = 55; h = 10	fd	fd^2
20-30	3	25	-3	-9	27
30-40	61	35	-2	-122	244
40-50	132	45	-1	-132	132
50-60	153	55(A)	0	0	0
60-70	140	65	1	140	140
70-80	51	75	2	102	204
80-90	2	85	3	6	18
	N = 542			$\Sigma f_i d_i = -15$	$\Sigma f_i d_i^2 = 765$

→ (2M)

Mean $\bar{x} = A + \frac{\Sigma f_i d_i}{N} \times h$

$= 55 + \left(\frac{-15}{542}\right) \times 10 = \underline{\underline{54.7232}}$ → (2M)

Population Variance $\sigma^2 = \left(\frac{\Sigma f_i d_i^2 - \frac{(\Sigma f_i d_i)^2}{N}}{N}\right) h^2$ → (2M)

$= \left(\frac{765 - \frac{(-15)^2}{542}}{542}\right) \times 100$

$= 141.0673$

$SD = +\sqrt{\text{Variance}} = \sqrt{141.0673} = \underline{\underline{11.877}}$ → (2M)

Note: * Assign marks if they solve it by alternative procedures

* As it is not mentioned as population and sample; Consider it if they solve it for sample

* Assign marks for the values which are very close to answers

6 (b) Coefficient of variation

3, 8, 6, 10, 12, 9, 11, 10, 12, 7

3 — 1
8 — 1
6 — 1
10 — 2
12 — 2
9 — 1
11 — 1
7 — 1

Mean $(\bar{x}) = \frac{\Sigma x_i}{n}$

$= \frac{3+8+6+10+12+9+11+10+12+7}{10}$

$= \frac{88}{10} = \underline{\underline{8.8}}$ (2M)

Variance $(\sigma^2) = \frac{1}{N} \Sigma (x_i - \bar{x})^2$

$= \frac{1}{N} \Sigma x_i^2 - (\bar{x})^2$

$= \frac{1}{10}(848) - (8.8)^2$

$= 7.36$

SD $(\sigma) = +\sqrt{\text{Variance}} = \sqrt{7.36} = \underline{\underline{2.7129}}$ (2M)

C.V $= \frac{\sigma}{\bar{x}} \times 100 = \frac{2.7129}{8.8} \times 100$

$= \underline{\underline{30.83\%}}$

* Assign marks if they solve it for sample.

(7) Karl Pearsons Coefficient of skewness

$S_k = \frac{\text{Mean} - \text{Mode}}{SD}$ → (1M)

Mean $= \frac{\Sigma f_i x_i}{N} = \underline{\underline{44.4615}}$

Mode $= l + \frac{C\Delta_1}{\Delta_1 + \Delta_2} = \underline{\underline{47.3684}}$ (→ 2M)

(3)

Variance $(\sigma^2) = \frac{1}{N} \Sigma f_i (x_i - \bar{x})^2 = \underline{\underline{243.5562}}$ (→ 4M)

Population SD (σ): $\sqrt{243.55} = \underline{\underline{15.6063}}$ (3M)

$$Sk = \frac{44.4615 - 47.3684}{15.6063} = \underline{\underline{-0.1862}} \quad (2M)$$

Negatively skewed.

8) a) <u>Bayes theorem</u>:

Let $E_1, E_2 \dots E_n$ are n mutually exclusive events in the sample space S such that if A is any event $P(A) > 0$ and $A \subset \bigcup_{i=1}^{n} E_i$ then $P(E_i) \neq 0$

$$P(E_i|A) = \frac{P(A|E_i)\,P(E_i)}{\sum_{i=1}^{n} P(A|E_i)\,P(E_i)} \quad \forall i = 1 \dots n \longrightarrow (3M)$$

<u>Proof</u>:

Since $A \subset \bigcup_{i=1}^{n} E_i$

$$A = A \cap \left(\bigcup_{i=1}^{n} E_i\right) = \bigcup_{i=1}^{n} (A \cap E_i)$$

↳ by distributive law

Since $(A \cap E_i) \subset E_i$ ($\forall i = 1 \dots n$) are mutually exclusive we have by addition theorem

$$P(A) = P\left(\bigcup_{i=1}^{n} (A \cap E_i)\right) = \sum_{i=1}^{n} P(A \cap E_i) = \sum_{i=1}^{n} P(E_i)\,P(A|E_i)$$

by Compound Probability theorem

$$P(A \cap E_i) = P(A)\,P(E_i|A)$$

$$P(E_i|A) = \frac{P(A \cap E_i)}{P(A)} = \frac{P(E_i)\,P(A|E_i)}{\sum_{i=1}^{n} P(E_i)\,P(A|E_i)} \quad \forall i = 1 \dots n \longrightarrow (3M)$$

* $P(E_1) \dots P(E_n)$ are called Prior Probabilities
* $P(A|E_i)$ $\forall i = 1 \dots n$ are called likelihoods
* $P(E_i|A)$ $\forall i = 1 \dots n$ are called Posterior Probabilities (4)

<u>Note</u> * There is another version of bayes theorem

8) b) Probability of x becoming manager

$P(x) = 4/9$

Probability of y becoming manager $P(y) = 2/9$

Probability of z becoming manager $P(z) = 1/3 = 3/9$

Let B denotes the event that Bonus scheme is introduced

then $P(B/x) = 3/10$; $P(B/y) = 1/2$; $P(B/z) = \frac{4}{5}$

i) Probability that the bonus scheme will be introduced (by Total Probability theorem)

$$P(B) = P(B/x)P(x) + P(B/y)P(y) + P(B/z)P(z)$$

$$= \left(\frac{3}{10} \times \frac{4}{9}\right) + \left(\frac{1}{2} \times \frac{2}{9}\right) + \left(\frac{4}{5} \times \frac{3}{9}\right)$$

$$= 0.133 + 0.111 + 0.2666 \longrightarrow (2m)$$

$$= 0.5106$$

ii) If the bonus scheme is introduced Probability that y was appointed as manager

by bayes theorem.

$$P(y/B) = \frac{P(B/y)\,P(y)}{P(B/x)P(x) + P(B/y)P(y) + P(B/z)P(z)}$$

$$= \frac{\left(\frac{1}{2} \times \frac{2}{9}\right)}{0.5106} = \frac{0.1111}{0.5106}$$

$$= \underline{\underline{0.21758}} \longrightarrow (2m)$$

5

4) a) $F_X(x) = \begin{cases} 0 & ; x \le 1 \\ K(x-1)^4 & ; 1 < x \le 3 \\ 1 & ; x > 3 \end{cases}$

$Pdf = \frac{d}{dx}(F_X(x))$

$f(x) = \begin{cases} 4K(x-1)^3 & ; 1 \le x \le 3 \\ 0 & ; \text{elsewhere} \end{cases}$

$\int_{-\infty}^{+\infty} f_X(x)dx = 1 \Rightarrow \int_1^3 4K(x-1)^3 dx = 1$

on simplification $\boxed{K = 1/16} \longrightarrow (2M)$

$f_X(x) = \begin{cases} 0 ; & x \le 1 \\ \frac{1}{4}(x-1)^3 ; & 1 < x \le 3 \\ 0 ; & x > 3 \end{cases} \longrightarrow (2M)$

$\text{Mean}(\mu) = E(X) = \int_{-\infty}^{+\infty} x f(x)dx = \int_{\cancel{-\infty}\,1}^{3\,\cancel{+\infty}} x\left(\frac{1}{4}(x-1)^3\right)dx$

on simplification

$\boxed{\mu = \frac{13}{5} = 2.6} \longrightarrow (2M)$

b) X follows Poisson

$P(x) = \frac{e^{-\lambda}\lambda^x}{x!} \quad x = 0 - -\infty$

$P(1) = P(2)$

$\frac{e^{-\lambda}\lambda^1}{1!} = \frac{e^{-\lambda}\lambda^2}{2!} \Rightarrow \lambda^2 - 2\lambda = 0$

$\lambda(\lambda - 2) = 0$

$\lambda = 0$ or $\lambda = 2$ but $\lambda > 0$

we select $\lambda = 2 \longrightarrow (2M)$

i, Mean $\underline{\underline{\lambda = 2}}$

ii, $P(X=4) = \frac{e^{-\lambda}\lambda^x}{x!} = \frac{e^{-2}2^4}{4!} = \underline{\underline{0.09022}} \rightarrow (1M)$

$P(X \ge 1) = 1 - P(X<1)$
$= 1 - P(X=0) = 1 - \frac{e^{-2}2^0}{0!} = \underline{\underline{0.8647}}$ (1M)

⑥

$P(1<X<4) = P(2) + P(3) \cancel{+P(4)}$ (2M)
$= 0.27067 + 0.18045 \cancel{+0.09022} = \underline{\underline{0.4511}}$

10) a) Population mean

$$M = \frac{2+3+6+8+11}{5} = \frac{30}{5} = 6 \qquad (4M)$$

b) Variance of the population

$$\sigma^2 = \frac{1}{n}\Sigma(x_i-\bar{x})^2$$

$$= \frac{(2-6)^2+(3-6)^2+(6-6)^2+(8-6)^2+(11-6)^2}{5}$$

$$= 10.8$$

$$\sigma = \sqrt{10.8} = \underline{\underline{3.29}} \qquad (4M)$$

c) No of possible samples with replacement

$$= N^n = 5^2 = 25$$

(2,2) (2,3) (2,6) (2,8) (2,11)
(3,2) (3,3) (3,6) (3,8) (3,11)
(6,2) (6,3) (6,6) (6,8) (6,11)
(8,2) (8,3) (8,6) (8,8) (8,11)
(11,2) (11,3) (11,6) (11,8) (11,11)

Sample means

2	2.5	4	5	6.5
2.5	3	4.5	5.5	7
4	4.5	6	7	8.5
5	5.5	7	8	9.5
6.5	7	8.5	9.5	11

sampling distribution of mean

$$M_{\bar{x}} = \frac{150}{25} = 6 \qquad (2M)$$

$$M_{\bar{x}} = M$$

d) The S.D of sampling distribution of means

$$\sigma_{\bar{x}}^2 = \frac{(2-6)^2 + \cdots + (11-6)^2}{25} = \frac{135}{25} = 5.40$$

$$\sigma_{\bar{x}} = \sqrt{5.40} = 2.32$$

⑦

in sampling with replacement

$$\sigma_{\bar{x}}^2 = \frac{\sigma^2}{n} = \frac{3.29}{2} = \underline{\underline{1.645}} \qquad (2M)$$

$$\sigma_{\bar{x}} = \sigma/\sqrt{n} = \frac{3.29}{\sqrt{2}} = \underline{2.32}$$

(8M)

11) a) <u>Population</u>: The group of individuals under study in any statistical investigation. It is also known as universe (1M)

<u>Sample</u>: Finite subset of the population (1M)

<u>Parameter</u>: Population constants
Ex: μ, σ, σ^2 → (1M)

<u>Statistic</u>: Sample constants
Ex: $\bar{x}, s, s^2 \ldots$ → (1M)

<u>standard error</u>: standard deviation of the sampling distribution of a statistic → (1M)

<u>Ex</u>: $SE(\bar{x}) = \sigma/\sqrt{n}$

11) b) mean of the sample $\bar{x} = 0.824$

$z_{\alpha/2}$ at 95% $= 1.96$

$\sigma = SD = 0.042$

$n =$ sample size $= 200$

→ (2M)

Confidence interval → (3M)

$$\left(\bar{x} - z_{\alpha/2}\frac{\sigma}{\sqrt{n}},\ \bar{x} + z_{\alpha/2}\sigma/\sqrt{n}\right)$$

$$= (0.824 - 0.0059,\ 0.824 + 0.0059)$$

$$= \underline{\underline{(0.8181,\ 0.8299)}}$$

Maximum error of the true mean

$$E = z_{\alpha/2}\sigma/\sqrt{n} = \frac{(1.96)(0.042)}{\sqrt{200}}$$ → (2M)

$$= \underline{\underline{0.0059}}$$

12) $n_1 = 7;\ n_2 = 6$

$\bar{x} = \frac{1}{n_1} \Sigma x_i = \frac{1}{7}(219) = 31.286$

$\bar{y} = \frac{1}{n_2} \Sigma y_i = \frac{169}{6} = 28.16$

$\tilde{S}^2 = \frac{1}{n_1+n_2-2}\left[\Sigma(x_i-\bar{x})^2 + \Sigma(y_i-\bar{y})^2\right]$ (3M)

x	$x-\bar{x}$	$(x-\bar{x})^2$	y	$(y-\bar{y})$	$(y-\bar{y})^2$
28	-3.286	10.8	29	0.84	0.7056
30	-1.286	1.6538	30	1.84	3.3856
32	0.714	0.51	30	1.84	3.3856
33	1.714	2.94	24	-4.16	17.3056
33	1.714	2.94	27	-1.16	1.3456
29	-2.286	5.226	29	0.84	0.7056
34	2.714	7.366			
219		31.4358	169		26.8336 (4M)

$\tilde{S}^2 = \frac{1}{11}[31.4358 + 26.8336] = \frac{1}{11}(58.2694) = 5.23$

$S = \sqrt{5.23} = 2.3$

Null hypothesis $H_0: \mu_1 = \mu_2$

Alternative $H_1: \mu_1 \neq \mu_2$ (Two tail ~~right tail~~)

Level of significance $\alpha = 0.05$ (2M)

test statistic $t = \frac{\bar{x}-\bar{y}}{S\sqrt{\frac{1}{n_1}+\frac{1}{n_2}}}$ $\cancel{= \frac{103-95.8}{(13.51)\sqrt{\frac{1}{10}+}}}$

$= \frac{31.286 - 28.16}{2.3\sqrt{\frac{1}{7}+\frac{1}{6}}} = 2.443$

Table value at $n_1+n_2-2 = 7+6-2 = 11\,df = 2.2$ (3M)

$|t_{cal}| >$ table value 9

We reject the null hypothesis H_0

we accept H_1: i.e, $\mu_1 \neq \mu_2$

The two horses don't have same running capacit

13) a) critical region: The rejection region, is the set of Values for a test statistic for which null hypothesis is rejected. It is a subspace of the sample space $\longrightarrow$ (2M)

b) level of significance: It is denoted by α
It is the Probability of comitting type-I error
(or) size of the critical region $\longrightarrow$ (2M)

$P\{\underset{\frown}{x} \in W \mid H_0\} = \alpha$

13) b) H_0: The accidents are uniformly distributed over the week

H_1: Accidents are not uniformly distributed over the week

N = Total frequency = 84

each cell expected frequency $\longrightarrow$ (2M)

$$e_{ij} = \frac{84}{7} = 12$$

$$\chi^2 = \Sigma\Sigma \frac{(O_{ij}-e_{ij})^2}{e_{ij}} = \frac{(14-12)^2}{12} + \frac{(16-12)^2}{12} + \frac{(8-12)^2}{12}$$

$$+ \frac{(12-12)^2}{12} + \frac{(11-12)^2}{12} + \frac{(9-12)^2}{12} + \frac{(14-12)^2}{12}$$

$$= \frac{50}{12} = 4.17 \longrightarrow (4M)$$

Degress of freedom $= n-1 = 7-1 = 6$

χ^2 table Value at 6; 5% = 12.59

$\chi^2_{cal} = 4.17 <$ table Value (12.59)

We accept the Null hypothesis H_0 $\longrightarrow$ (2M)

$\therefore$ The accidents are uniformly distributed over the week.

10

14.

R.L of Y on X

$$\underset{(\bar{X})}{\text{mean}} = \frac{\Sigma x_i}{n} = 52.33 \cong 52$$

$$\bar{Y} = \frac{\Sigma y_i}{n} = \frac{1684}{12} = 140.33 \cong 140$$

$\rightarrow$ (3M)

$$b_{yx} = \frac{\Sigma XY}{\Sigma X^2} = \frac{1766}{1522} =$$

$$b_{yx} = \frac{n\Sigma XY - (\Sigma X)(\Sigma Y)}{n\Sigma X^2 - (\Sigma X)^2}$$

$$= \frac{12(1766) - 4(4)}{12(1522) - 4^2} = 1.138 \rightarrow (3M)$$

Reg line of Y on X

$$Y - \bar{Y} = b_{yx}(X - \bar{X})$$

$$\Rightarrow (Y - 140.33) = 1.138(X - 52.33)$$

$$\boxed{Y = 1.138X + 80.777} \rightarrow (1M)$$

Reg line of X on Y

$$X - \bar{X} = b_{xy}(Y - \bar{Y})$$

$$b_{xy} = \frac{n\Sigma XY - (\Sigma X)(\Sigma Y)}{n\Sigma Y^2 - (\Sigma Y)^2}$$

$$= \frac{12(1766) - 4(4)}{12(2502) - 4^2} = 0.7057 \rightarrow (3M)$$

$$\Rightarrow X - 52.33 = 0.7057(Y - 140.33)$$

$$\Rightarrow \boxed{X = 0.7057Y - 46.6969} \rightarrow (1M)$$

(11)

When $X = 45$ value of $Y = 131.988 \cong 132 \rightarrow (1M)$

15)

a) $y = a + bx$

normal equations are

$\sum y_i = na + b\sum x_i \longrightarrow$ (1.5M)

$\sum x_i y_i = a\sum x_i + b\sum x_i^2 \longrightarrow$ (1.5M)

b) Second degree Parabola

$y = a + bx + cx^2$

normal equations are

$$\left.\begin{aligned} \sum y_i &= na + b\sum x_i + c\sum x_i^2 \\ \sum x_i y_i &= a\sum x_i + b\sum x_i^2 + c\sum x_i^3 \\ \sum x_i^2 y_i &= a\sum x_i^2 + b\sum x_i^3 + c\sum x_i^4 \end{aligned}\right\} \longrightarrow (3M)$$

$$\left.\begin{aligned} &\sum x_i = 17.5 \quad \sum y_i = 16.2 \quad \sum x_i^2 = 50.75 \\ &\sum x_i^3 = 161.875 \quad \sum x_i^4 = 548.1875 \\ &\sum x_i y_i = 47.65 \quad \sum x_i^2 y_i = 154.475 \end{aligned}\right\} \longrightarrow (3M)$$

equations are

$7a + 17.5b + 50.75c = 16.2$

$17.5a + 50.75b + 161.875c = 47.65$

$50.75a + 161.875b + 548.1875c = 154.475$

on solving $\underline{\underline{a = 1.0357}}$ $b = \underline{\underline{-0.1929}}$ $\longrightarrow$ (2M)

$c = \underline{\underline{0.2429}}$

Required Parabola

$y = \hat{a} + \hat{b}x + \hat{c}x^2$ $\longrightarrow$ (~~1~~M)

$\underline{\underline{y = 1.0357 - 0.1929x + 0.2429x^2}}$ 12

16/6/22

NSRIT

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	Program	CE			Academic Year	2021 - 2022
Course Code	20CE403	Test Duration	3 Hrs.	Max. Marks	70	Semester	IV
Course	CONCRETE TECHNOLOGY						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	List any four properties of cement in field.	20CE403.1	L1
2	Define Segregation and Bleeding.	20CE403.2	L1
3	State the principle of Ultrasonic Pulse Velocity Test.	20CE403.3	L1
4	What are light weight aggregates?	20CE403.4	L1
5	List any four factors affecting the choice of mix proportions.	20CE403.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Write short notes on various types of cement.	12 M	20CE403.1	L2
	OR			
7 (a)	Explain any two tests carried on aggregate.	6M	20CE403.1	L2
7 (b)	Illustrate the briefly note on importance of the quality of water used for concreting.	6M	20CE403.1	L2
8	Briefly discuss the concrete manufacturing process.	12M	20CE403.2	L2
	OR			
9 (a)	Illustrate various factors influencing the Workability of Concrete?	6M	20CE403.2	L2
9 (b)	What are the properties of fresh concrete?	6M	20CE403.2	L2
10 (a)	Discuss the importance of Non-Destructive tests	6M	20CE403.3	L2
10 (b)	List out the factors influencing the strength results in case of hardened concrete.	6M	20CE403.3	L2
	OR			
11	Explain in detail about the determination of Compressive and Flexural strength of concrete.	12M	20CE403.3	L2
12	Give a brief note on polymer concrete.	12M	20CE403.4	L2
	OR			
13 (a)	Write a short note on (a) High performance concrete (b) Fiber reinforced concrete (c) SIFCON	6M	20CE403.4	L2
13 (b)	Describe in detail about Shotcrete and its advantages.	6M	20CE403.4	L2
14	Design a concrete mix for construction of an elevated water tank. The specified design strength of concrete (characteristic strength) is 30 MPa at 28 days measured on standard cylinders. Standard deviation can be taken as 4 MPa. The specific gravity of FA and C.A. are 2.65 and 2.7 respectively. The dry rodded bulk density of C.A. is 1600 kg/m3, and fineness modulus of FA is 2.80. Ordinary Portland cement (Type I) will be used. A slump of 50 mm is necessary. C.A. is found to be absorptive to the extent of 1% and free surface moisture in sand is found to be 2 per	12M	20CE403.5	L3

cent. Assume any other essential data.

OR

15	Explain the concept of mix design and mention the method of proportioning.	12M	20CE403.5	L3

NSRIT

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

Sub: Concrete Technology

Part A (Short Answer Questions 5 x 2 = 10 Marks)

1. List any four properties of cement in field. Content 2 M

Ans: Properties of Cement- Physical & Chemical
Fineness of cement.
Soundness.
Consistency.
Strength.
Setting time.
Heat of hydration.
Loss of ignition.
Bulk density

2. Define Segregation and Bleeding. Content 2 M

Ans: Segregation: There are considerable differences in the sizes and specific gravities of the constituent ingredients of concrete. Therefore, it is natural that the materials show a tendency to fall apart.
Segregation may be of three types
1. Coarse aggregate separating out or settling down from the rest of the matrix.
2. Paste separating away from coarse aggregate. 3. Water separating out from the rest of the material being a material of lowest specific gravity
Bleeding: Concrete bleeding is defined as the appearance of water on the surface of concrete after it has consolidated but before it is set. This is a type of segregation where water appears at the concrete surface after placing and compacting, but before it is set. Water may also form a film under aggregate and reinforcing bar. Some bleeding is useful for finishing operations and to reduce plastic shrinkage cracking.

3. State the principle of Ultrasonic Pulse Velocity Test. Content 2 M

Ans: The ultrasonic pulse velocity test is a non-destructive test used to determine the quality of concrete on site. This test basically involves the assessment of the velocity of electronic pulse passing through the concrete from a transmitting transducer to a receiver transducer.
The principle of the ultrasonic pulse velocity test is that the velocity sound in a solid material is a function of the square root of the ratio of its modulus of elasticity E to its density P. The density and the elastic properties of the material are related to its quality and strength, respectively.

4. What are Light weight Aggregates? Content 2 M

Ans: The lightweight aggregate is a kind of coarse aggregate which is used in the production of lightweight concrete products like concrete block, structural concrete, and pavement.

The shape of the lightweight aggregate used in concrete can be cubical, rounded, angular, or of any other shape. The shape and texture can directly affect its workability. The compressive strength level which is required by the construction industry to design strengths of cast-in-place, precast or prestressed concrete is known to be 3000-5000 psi. It is something that can be easily achieved with the use of lightweight aggregate concrete. The density of the lightweight concretes depends on mixture proportion, air content, water demand, density, and moisture content of the lightweight aggregate. The structural concrete density can be easily achieved with the use of lightweight aggregate concrete. The lightweight concrete is also known to absorb very little water and can maintain the low density. It is known to be equal or lower than that of normal concrete. It also has a high degree of saturation. The lightweight aggregate concrete, structural lightweight concrete slabs, walls, and beams are considered to have greater fire-endurance periods than other types of concrete.

5. List any four factors affecting the choice of mix proportions. Content 2 M

Ans: Factors affecting the choice of mix proportions

Compressive strength.

Workability.

Durability.

Maximum nominal size of aggregate.

Grading and type of aggregate.

Quality Control.

Part B (Long Answer Questions 5 x 12 = 60 Marks)

6. Write short notes on various types of cement.

Ans: There are different types of cement for different construction works. Keep reading to learn more about the most common ones.

1. Ordinary Portland Cement (OPC)

Ordinary Portland Cement also known as OPC is a type of cement that is manufactured and used worldwide. It is widely used for all purposes including:

- Concrete: When OPC is mixed with aggregates and water, it makes concrete, which is widely used in the construction of buildings
- Mortar: For joining masonry
- Plaster: To give a perfect finish to the walls

Cement companies in Malaysia offer OPC in three different grades, namely grades 33, 43, and 53.

Besides the aforementioned purposes, Ordinary Portland cement is also used to manufacture grout, wall putty, solid concrete blocks, AAC blocks, and different types of cement.

2. Portland Pozzolana Cement (PPC)

To prepared PPC or Portland Pozzolana cement, you need to grind pozzolanic clinker with Portland cement.

PPC has a high resistance to different chemical assaults on concrete. It is widely used in construction such as:

- Marine structures
- Sewage works
- Bridges
- Piers
- Dams
- Mass concrete works

Content each-1 M Presentation - 4M

3. Rapid Hardening Cement

Cement suppliers in Malaysia also offer rapid Hardening cement. Rapid Hardening Cement is made when finely grounded C3S is displayed in OPC with higher concrete.

It is commonly used in rapid constructions like the construction pavement.

4. Extra Rapid Hardening Cement

As the name suggests, Extra rapid hardening cement gains strength quicker and it is obtained by adding calcium chloride to rapid hardening cement.

Extra rapid hardening cement is widely used in cold weather concreting, to set the cement fast. It is about 25% faster than that of rapid hardening cement by one or two days.

5. Low Heat Cement

Cement manufacturers in Malaysia offers low heat cement that is prepared by keeping the percentage of tricalcium aluminate below 6% and by increasing the proportion of C2S.

This low heat cement is used in mass concrete construction like gravity dams. It is important to know that it is less reactive and the initial setting time is greater than OPC.

6. Sulfates Resisting Cement

This type of cement is manufactured to resist sulfate attack in concrete. It has a lower percentage of Tricalcium aluminate.

Sulfates resisting cement is used for constructions in contact with soil or groundwater having more than 0.2% or 0.3% g/l sulfate salts respectively.

It can also be used in concrete surfaces subjected to alternate wetting and drying like bridge piers.

7. Quick Setting Cement

Cement suppliers in Malaysia also offer quick setting cement which sets faster than OPC but the strength remains the same. In this formula, the proportion of gypsum is reduced.

Quick setting cement is used for constructions that need a quick setting, like underwater structures and in cold and rainy weather conditions.

8. Blast Furnace Slag Cement

This type of cement is manufactured by grinding the clinker with about 60% slag and it is similar to Portland cement. It is used for constructions where economic considerations are important.

OR

7 (a) Explain any two tests carried on aggregate.

Ans: **Test method: Crushing value test**

The "aggregate crushing value" gives a relative measure of the resistance of an aggregate to crushing under a gradually applied compressive load.

The apparatus, with the test sample and plunger in position, is placed on the compression testing machine and is loaded uniformly upto a total load of 400 kN in 10 minutes time.

The load is then released and the whole of the material removed from the cylinder and sieved on a 2.36 mm I.S. Sieve.

Content each-4 M
Presentation - 2M

- $The\ aggregate\ crushing\ value = \frac{B}{A} \times 100$

B = weight of fraction passing 2.36 mm sieve,

A = weight of surface-dry sample taken in mould.

- The aggregate crushing value should not be more than 45 per cent for aggregate used for concrete other than for wearing surfaces, and 30 per cent for concrete used for wearing surfaces such a runways, roads and air field pavements.

Impact value: Impact value of aggregates measures the toughness of particles by impact

The aggregate impact value gives relative measure of the resistance of an aggregate to sudden shock or impact

The whole sample is filled into a cylindrical steel cup firmly fixed on the base of the machine. A hammer weighing about 14 kgs. is raised to a height of 380 mm above the upper surface of the aggregate in the cup and allowed to fall freely on the aggregate.

The test sample shall be subjected to a total 15 such blows each being delivered at an interval of not less than one second.

The crushed aggregate is removed from the cup and the whole of it is sieved on 2.36 mm I.S. Sieve.

$$The\ Aggregate\ Impact\ Value = \frac{B}{A} \times 100$$

B = weight of fraction passing 2.36 mm I.S. Sieve.

A = weight of oven-dried sample.

The aggregate impact value should not be more than 45 per cent by weight for aggregates used for concrete other than wearing surfaces and 30 per cent by weight for concrete to be used as wearing surfaces, such as runways, roads and pavements.

7 (b) Illustrate the briefly note on importance of the quality of water used for concreting.

Ans: Water is an important ingredient of concrete

Concrete is produced by mixing binding materials and inert materials with water. Thus, water and its quality (and also its quantity) play an important role in determining the quality of concrete. Strength and durability of concrete is to a large extent determined by its water to cementitious materials ratio.

Water is required to wet the surface of aggregates to develop adhesive quality as the cement paste binds quickly and satisfactorily to the wet surface of the aggregates than to a dry surface. Also, water is needed to make plastic mixture of the various ingredients so as to impart workability to concrete to facilitate placing it in the desired position. Ultimately, by chemically reacting with cement, water helps to produce the desired properties of the concrete.

Usually, quality of the water is the highly neglected subject despite it having a very important role to play in determining the durability of the final product. It is a commonly accepted view that any potable water is suitable to be used in concrete making. However, when only non-potable water is available, it is always better to test the water to find out its contents and take suitable steps to contain potential adverse effects on the final concrete.

Content each-4 M Presentation - 2M

Though slightly acidic water is harmless, highly acidic or alkaline water should be avoided as it may have adverse effect over the hardening of concrete. Water mixed with algae should be avoided as such water causes entrainments which in turn results in loss of strength. It is found that sea water reduces the long-term strength of the cement, though reduction in strength is limited to 15%. Water containing large amount of chlorides tends to cause persistent dampness and surface efflorescence and also corrosion of steel used in concrete.

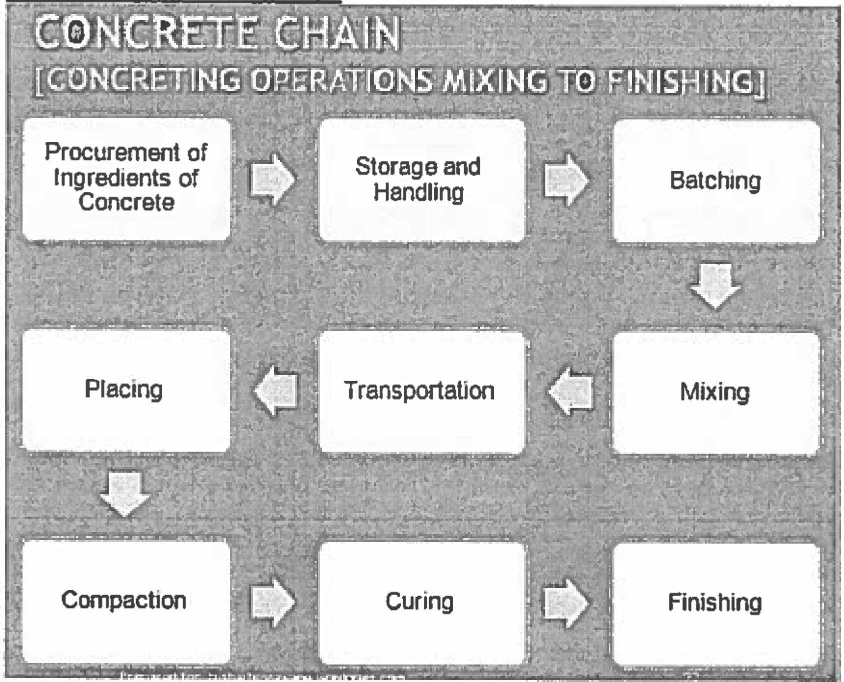
Thus, the chemical constituents present in water may actively participate in the chemical reactions and thus affect the setting, hardening and strength development of concrete. Therefore, It is always better to check water quality for ensuring good quality concrete.

8. Briefly discuss the concrete manufacturing process.

Ans; Concrete Manufacturing Processes Overview

We put together this basic overview of the manufacturing process to provide a high level view of the tasks at hand.

Concrete manufacturing steps:



1. **Procurement of Ingredients of Concrete:**

Procuring all the ingredients like Cement, Sand, Aggregate and Water in required quantity.

Content each-10 M Presentation - 2M

2. **Storage and Handling:**

¡ The procured material is stored in dry and damp free spaces so that they will not get moisture.

3. **Batching:** To measure the materials required concrete is known as batching.

¡ There are two methods

   A. Volumetric Batching
   B. Weigh Batching

4. **Mixing**

The cement is then mixed with the other ingredients: aggregates (sand, gravel, or crushed stone), admixtures, fibers, and water. Aggregates are pre-blended or added at the ready-mix concrete plant under normal operating conditions. The mixing operation uses rotation or stirring to coat the surface of the aggregate with cement paste and to blend the other ingredients uniformly. A variety of batch or continuous mixers are used.

Fibers, if desired, can be added by a variety of methods including direct spraying, premixing, impregnating, or hand laying-up. Silica fume is often used as a dispersing or densifying agent.

5. **Placing and compacting**

Once at the site, the concrete must be placed and compacted. These two operations are performed almost simultaneously. Placing must be done so that segregation of the various ingredients is avoided and full compaction—with all air bubbles eliminated—can be achieved. Whether chutes or buggies are used, position is important in achieving these goals. The rates of placing and of compaction should be equal; the latter is usually accomplished using internal or external vibrators. An internal vibrator uses a poker housing a motor-driven shaft. When the poker is inserted into the concrete, controlled vibration occurs to compact the concrete. External vibrators are used for precast or thin in situ sections having a shape or thickness unsuitable for internal vibrators. These types of vibrators are rigidly clamped to the formwork, which rests on an elastic support. Both the form and the concrete are vibrated. Vibrating tables are also used, where a table produces vertical vibration by using two shafts rotating in opposite directions

6. **Curing**

Once it is placed and compacted, the concrete must cured before it is finished to make sure that it doesn't dry too quickly. Concrete's strength is influenced by its moisture level during the hardening process: as the cement solidifies, the concrete shrinks. If site constraints prevent the concrete from contracting, tensile stresses will develop, weakening the concrete. To minimize this problem, concrete must be kept damp during the several days it requires to set and harden.

OR

9 (a) Illustrate various factors influencing the Workability of Concrete?
Ans; **Factors affecting workability**

- Water content of the mix: Adding water increases workability and decreases strength.
- (or)
- More the water cement ratio more will be workability of concrete. Since by simply adding water the inter particle lubrication is increased. High water content results in a higher fluidity and greater workability but reduces the strength of concrete. Because with increasing w/c ratio the strength decreases as more water will result in higher concrete porosity. So, the lower the w/c, the lower is the void volume/solid volume, and the stronger the hardened cement paste.
- Increased water content also results in bleeding, hence, increased water content can also mean that cement slurry will escape through the joints of the formwork (Shuttering).
- Maximum size of aggregate: Less surface area to be wetted and more water in medium.
- Grading of aggregate: Poor grading reduces the consistency.
- Shape and texture of aggregates: Smooth surfaces give better workability.
- Weather Conditions:

Content each-4 M Presentation - 2M

1. Temperature: If temperature is high, evaporation increases, thus workability decreases.
2. Wind: If wind is moving with greater velocity, the rate of evaporation also increase reduces the amount of water and ultimately reducing workability.

- **Admixtures**
- Chemical admixtures can be used to increase workability.
- Use of air entraining agent produces air bubbles which acts as a sort of ball bearing between particles and increases mobility, workability and decreases bleeding, segregation. The use of fine pozzolanic materials also have better lubricating effect and more workability.

9 (b) What are the properties of fresh concrete?
Ans: PROPERTIES OF FRESH CONCRETE
Concrete remains in its fresh state from the time it is mixed until it sets. During this time the concrete is handled, transported, placed and compacted. Properties of concrete in its fresh state are very important because the influence the quality of the hardened concrete.
The fresh concrete has the following procedure.

- Consistency
- Workability
- Settlement & Bleeding
- Plastic shrinkage
- Loss of consistency

Content each-4 M Presentation - 2M

1. CONSISTENCY

Consistency of a concrete mix is a measure of the stiffness or sloppiness or fluidity of the mix. For effective handling, placing and compacting the concrete, consistency must be the same for each batch. It is therefore necessary to measure consistency of concrete at regular intervals. Slump test is commonly used to measure consistency of concrete.

2. WORKABILITY

The workability of a concrete mix is the relative ease with which concrete can be placed, compacted and finished without separation or segregation of the individual materials.

Workability is not the same thing as consistency. Mixes with the same consistency can have different workabilities, if they are made with different sizes of stone – the smaller the stone the more workable the concrete.

It is not possible to measure workability but the slump test, together with an assessment of properties like stone content, cohesiveness and plasticity, gives a useful indication.

3. SETTLEMENT AND BLEEDING

Cement and aggregate particles have densities about three times that of water. In fresh concrete they consequently tend to settle and displace mixing water which migrates upward and may collect on the top surface of the concrete. This upward movement of mixing water is known as bleeding; water that separates from the rest of the concrete is called bleed water.

4. PLASTIC SHRINKAGE

If water is removed from the compacted concrete before it sets, the volume of the concrete is reduced by the amount of water removed. This volume reduction is called plastic shrinkage.

Water may be removed from the plastic concrete by evaporation or by being absorbed by dry surfaces such as soil or old concrete or by the dry wooden form work.

5. SLUMP LOSS

From the time of mixing, fresh concrete gradually loses consistency. This gives rise to the problems only if the concrete becomes too stiff to handle, place and compact properly.

Slump loss in concrete is caused due to the following reasons.

- Hydration of cement (generating more heat)
- Loss of water by evaporation
- Absorption of water by dry aggregates
- Absorption of water by surfaces in contact with the concrete.

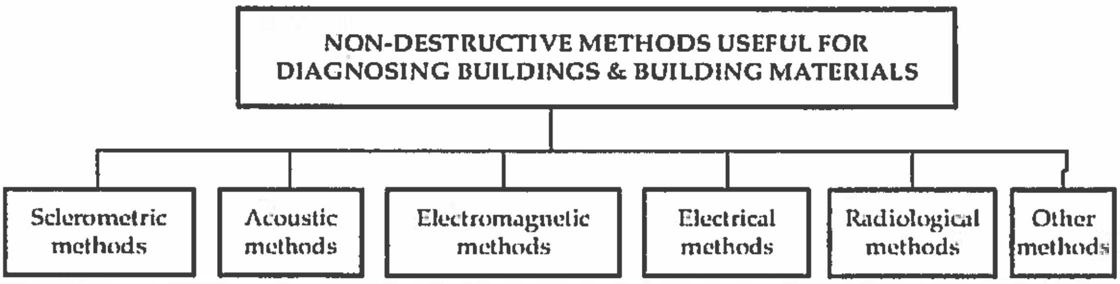
10 (a) Discuss the importance of Non-Destructive tests

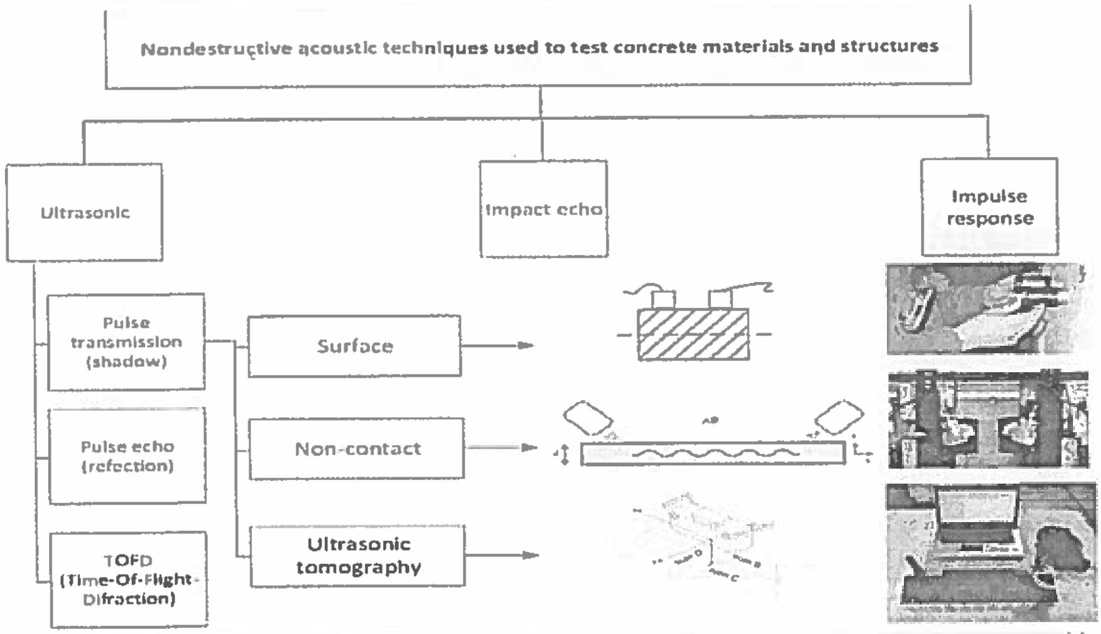
Ans: Non-destructive methods are mainly used to test strength and investigate its changes over time. Usually samples taken from the structure, and sometimes whole members or structures, are tested in this way. Also load tests, which rather rarely are applied to buildings, but more often to bridges and roads, can be put into this category.

Semi-destructive and destructive methods are used to test samples and members. They can also be used to test whole structures. Strength and its changes over time are tested, but mainly other properties are tested in this way.

The difference between semi-destructive and non-destructive methods is that in the case of the former, the material is usually locally and superficially damaged when tested. No such damage occurs in the case of non-destructive methods. This is one of the reasons why they are suitable for testing large surfaces down to a considerable depth, and in general construction. Moreover, in the case of non-destructive methods, measurements can be repeated, whereby the test results can be verified and validated.

Content each-4 M Presentation - 2M





10 (b) List out the factors influencing the strength results in case of hardened concrete.

Ans: Factors affecting concrete strength Concrete strength is effected by many factors, such as quality of raw materials, water/cement ratio, coarse/fine aggregate ratio, age of concrete, compaction of concrete, temperature, relative humidity and curing of concrete.

1. Quality of Raw Materials:

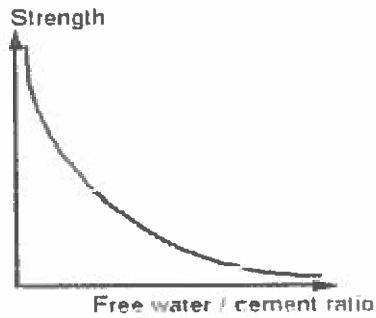
Cement: Provided the cement conforms with the appropriate standard and it has been stored correctly (i.e. in dry conditions), it should be suitable for use in concrete.

Aggregates: Quality of aggregates, its size, shape, texture, strength etc determines the strength of concrete. The presence of salts (chlorides and sulphates), silt and clay also reduces the strength of concrete.

Water: frequently the quality of the water is covered by a clause stating "..the water should be fit for drinking..". This criterion though is not absolute and reference should be made to respective codes for testing of water construction purpose.

2. Water / Cement Ratio

The relation between water cement ratio and strength of concrete is shown in the plot as shown below:



Content each-4 M
Presentation - 2M

The higher the water/cement ratio, the greater the initial spacing between the cement grains and the greater the volume of residual voids not filled by hydration products.

There is one thing missing on the graph. For a given cement content, the workability of the concrete is reduced if the water/cement ratio is reduced. A lower water cement ratio means less water, or more cement and lower workability. However if the workability becomes too low the concrete becomes difficult to compact and the strength reduces. For a given set of materials and environment conditions, the strength at any age depends only on the water-cement ratio, providing full compaction can be achieved.

3. Coarse / fine aggregate ratio: Following points should be noted for coarse/fine aggregate ratio: • If the proportion of fines is increased in relation to the coarse aggregate, the overall aggregate surface area will increase. • If the surface area of the aggregate has increased, the water demand will also increase. • Assuming the water demand has increased, the water cement ratio will increase. • Since the water cement ratio has increased, the compressive

strength will decrease

4. Aggregate / Cement Ratio: Following points must be noted for aggregate cement ratio: • If the volume remains the same and the proportion of cement in relation to that of sand is increased the surface area of the solid will increase. • If the surface area of the solids has increased, the water demand will stay the same for the constant workability.

5. Age of concrete: The degree of hydration is synonymous with the age of concrete provided the concrete has not been allowed to dry out or the temperature is too low. In theory, provided the concrete is not allowed to dry out, then it wil always be increasing albeit at an ever reducing rate. For convenience and for most practical applications, it is generally accepted that the majority of the strength has been achieved by 28 days.

6. Compaction of concrete: Any entrapped air resulting from inadequate compaction of the plastic concrete will lead to a reduction in strength. If there was 10% trapped air in the concrete, the strength will fall down in the range of 30 to 40%.

7. Temperature: The rate of hydration reaction is temperature dependent. If the temperature increases the reaction also increases. This means that the concrete kept at higher temperature will gain strength more quickly than a similar concrete kept at a lower temperature. However, the final strength of the concrete kept at the higher temperature will be lower. This is because the physical form of the hardened cement paste is less well structured and more porous when hydration proceeds at faster rate.

8. Relative humidity: If the concrete is allowed to dry out, the hydration reaction will stop. The hydration reaction cannot proceed without moisture. The three curves shows the strength

9. Curing: It should be clear from what has been said above that the detrimental effects of storage of concrete in a dry environment can be reduced if the concrete is adequately cured to prevent excessive moisture loss.

OR

11. Explain in detail about the determination of Compressive and Flexural strength of concrete.

Ans: Compressive strength of Concrete can be defined as the ability of material or structure to carry the loads on it without any crack or deflection. A material under compressive load tends to reduce the size, while in tension, size elongates.

The compressive strength of concrete can be calculated by dividing the load applied on the concrete cube at the point of failure by the cross-section area of the cube (15x15x15 cm) on which load was applied.

The concrete compressive strength for normal construction work varies from 15 MPa (2200 psi) to 30 MPa (4400 psi) and more in commercial and industrial structures.

The strength of concrete depends on factors such as water-cement ratio, the strength of cement use, quality of concrete materials, quality control during production of concrete, etc.

A compressive strength test of concrete is performed to check the compressive strength of concrete. There are various standard codes that recommend concrete cylinder or concrete cube as the standard specimen for the test.

The American Society for Testing Materials of construction ASTM C39/C39M provides Standard Test Method for Compressive Strength of cube and Cylindrical Concrete Specimens.

Content each-10 M Presentation - 2M

The following is the compressive strength of concrete formula,

Compressive Strength = Load at failure / Cross-sectional Area of element

Flexural strength of Concrete, also known as Modulus of rupture, is an indirect measure of the tensile strength of unreinforced concrete. Modulus of rupture can also be defined as the measure of the extreme fibre stresses when a member is subjected to bending. Apart from external loading, tensile stresses can also be caused by warping, corrosion of steel, drying shrinkage and temperature gradient.

Concrete is strong in compression but weak in tension because of which the flexural strength accounts for only 10% to 20% of the compressive strength.

Experimental Estimation of Flexural Strength using One-point loading test and the Two-point loading test.

Unlike compression, tensile strength of a member can not be found directly as no apparatus or specimen model has been developed to evenly distribute the tensile force to the member. However, the indirect measurement of the flexural strength like the One-point loading test and the Two-point loading test fetch satisfying results.

Principle / Mechanism

Modulus of rupture is the measure of extreme fibre stresses in a member under flexure where the beam can be loaded using One-point loading or the symmetrical Two-point loading. When a simply supported beam is subjected to bending, tensile stresses are developed at the bottom of the beam and once the tensile stresses exceed the flexural strength of the beam, cracks start to occur at the point of maximum bending moment. The load causing the crack and the pattern of the crack can be used to calculate the flexural strength of the given concrete member.

12. Give a brief note on polymer concrete.

Ans: Polymer concrete is an aggregate mixture that uses some type of epoxy binder to cure and harden into place. A polyester, vinyl ester, or normal epoxy mixture is often used, but polymer concrete can be made with many kinds of polymer resins that allow the concrete to be poured or troweled and then hardened. It cures through a chemical reaction with the polymer material. Like traditional concrete, it also has water, sand and gravel or crushed stone as primary ingredients.

Benefits

Polymer concrete offers different benefits depending on the resin used to make it. Acrylic binders set very quickly and offer resistance to weathering, while epoxies create a very strong material that shrinks very little as it cures. Furan resins can withstand high temperatures, and polyurea resins can replace phenolics or formaldehydes in many construction projects.

Considerations

Polymer concrete must be mixed very precisely and very thoroughly. It cannot be mixed beforehand and simply kept turning to avoid curing--the chemical reaction will happen no matter what. Also, the chemicals that this type of concrete uses can be very dangerous and everyone nearby must wear masks and skin protection. 12M

Uses

Polymer concrete is used for many kinds of specialized construction projects. Like other types of concrete, it can be used to join two different components or to provide a structure or base. The material is used in electrical or industrial construction where the concrete needs to last a long time and be resistant to many types of corrosion.

Polymer Mortar

Polymer mortar is a smoother type of polymer concrete made only from a binder and a fine aggregate, like sand. It is used primarily to join objects, like regular mortar, but does not have the same tendency to wear down in harsh climates. It may also be used to coat objects for protection.

OR

13 (a) Write a short note on (a) High performance concrete
(b) Fiber reinforced concrete (c) SIFCON

Ans: **High-performance concrete** (HPC) is concrete that has been designed to be more durable and, if necessary, stronger than conventional concrete. HPC mixtures are composed of essentially the same materials as conventional concrete mixtures, but the proportions are designed, or engineered, to provide the strength and durability needed for the structural and environmental requirements of the project. High-strength concrete is defined as having a specified compressive strength of 8000 psi (55 MPa) or greater. The value of 8000 psi (55 MPa) was selected because it represented a strength level at which special care is required for production and testing of the concrete and at which special structural design requirements may be needed.

Fiber Reinforced Concrete is a composite material consisting of fibrous material which increases its structural integrity. It includes mixtures of cement, mortar or concrete and discontinuous, discrete, uniformly dispersed suitable fibers. Fibers are usually used in concrete to control cracking due to plastic shrinkage and to drying shrinkage. They also reduce the permeability of concrete and thus reduce the bleeding of water.

Content each-4 M Presentation - 2M

Advantages of Fiber-reinforced concrete

- Fibers reinforced concrete may be useful where high tensile strength and reduced cracking are desirable or when conventional reinforcement cannot be placed
- It improves the impact strength of concrete, limits the crack growth and leads to a greater strain capacity of the composite material
- For industrial projects, macro-synthetic fibers are used to improve concrete's durability. Made from synthetic materials, these fibers are long and thick in size and may be used as a replacement for bar or fabric reinforcement
- Adding fibers to the concrete will improve its freeze-thaw resistance and help keep the concrete strong and

attractive for extended periods.

- Improve mix cohesion, improving pumpability over long distances
- Increase resistance to plastic shrinkage during curing
- Minimizes steel reinforcement requirements
- Controls the crack widths tightly, thus improving durability
- Reduces segregation and bleed-water
- FRC, toughness is about 10 to 40 times that of plain concrete
- The addition of fibers increases fatigue strength
- Fibers increase the shear capacity of reinforced concrete beams

SIFCON means slurry infiltrated fiber concrete.

SIFCON is a special type of fiber reinforced composite containing steel fiber from 5% to 20% (by volume).

In this formwork molds are filled to capacity with randomly-oriented steel fiber, usually in loose condition and resulting fiber network is infiltrated by a cement-based slurry.

APPLICATIONS

•Bridge deck rehabilitation.

•Pavement rehabilitation.

•Repairing of structural components such as damaged pre-stressed concrete beams.

13 (b) Describe in detail about Shotcrete and its advantages.

Ans: Shotcrete has proved to be the best method for manufacturing curved surfaces such as dome, tunnel, etc.

Technology improvement gives advanced control the tasks and completed with greater economy in terms of both time and investment.

Why shotcrete is preferred:

- Shotcrete is more economical than conventional concrete because it requires less formwork.
- It only requires a small space for construction and its location.
- In the case of the shotcrete, cement content is high, so it is durable.
- It is resistant to fire, disasters, mold, worms and also has low permeability.
- It also has a good thermal resistance mass.

Content each-4 M Presentation - 2M

Advantages of Shotcrete:

1. In this process, little or no structure is required.
2. It cost-effective process for placing concrete.
3. Ideal for irregular surfaces.
4. It allows for easy material handling in hard-to-reach areas.
5. Also, easy for start-up, shutdown and cleaning.
6. Increases the load-carrying capacity due to redistribution of stress.
7. It provides excellent corrosion resistance.

14. Design a concrete mix for construction of an elevated water tank. The specified design strength of concrete (characteristic strength) is 30 MPa at 28 days measured on standard cylinders. Standard deviation can be taken as 4 MPa. The specific gravity of FA and C.A. are 2.65 and 2.7 respectively. The dry rodded bulk density of C.A. is 1600 kg/m3, and fineness modulus of FA is 2.80. Ordinary Portland cement (Type I) will be used. A slump of 50 mm is necessary. C.A. is found to be absorptive to the extent of 1% and free surface moisture in sand is found to be 2 per cent. Assume any other essential data.

Content each-10 M Presentation - 2M

C-2. TEST DATA FOR MATERIALS

a) Cement used—ordinary Portland cement satisfying the requirements of IS : 269-1976*	
b) Specific gravity of cement	3·15
c) Specific gravity	
1) Coarse aggregate	2·60
2) Fine aggregate	2·60
d) Water absorption	
1) Coarse aggregate	0·5 percent
2) Fine aggregate	1·0 percent
e) Free (surface) moisture	
1) Coarse aggregate	Nil (absorbed moisture also nil)
2) Fine aggregate	2·0 percent
f) Sieve analysis	
1) Coarse aggregate	

*Specification for ordinary and low heat Portland cement (*third revision*).

IS Sieve Sizes mm	*Analysis of Coarse Aggregate Fraction (Percent Passing)* I	II	*Percentage of Different Fractions* I 60 percent	II 40 percent	Combined 100 percent	*Remark*
20	100	100	60	40	100	Conforming to Table 2 of IS : 383-1970*
10	0	71·20	0	28·5	28·5	
4·75		9·40		3·7	3·7	
2·36		0				

2) Fine Aggregate

IS Sieve Sizes	*Fine Aggregate (Percent Passing)*	*Remark*
4·75 mm	100	Conforming to grading Zone III of Table 4 of IS : 383-1970*
2·36 mm	100	
1·18 mm	93	
600 micron	60	
300 micron	12	
150 micron	2	

C-3. TARGET MEAN STRENGTH OF CONCRETE

C-3.1 For a tolerance factor of 1·65 and using Table 1, the target mean strength for the specified characteristic cube strength is $20 + 4{\cdot}6 \times 1{\cdot}65 = 27{\cdot}6$ N/mm².

C-4. SELECTION OF WATER CEMENT RATIO

C-4.1 From Fig. 1, the free water-cement ratio required for the target mean strength of 27·6 N/mm² is 0·50. This is lower than the maximum value of 0·65 prescribed for 'Mild' exposure in Appendix A of IS : 456-1978†.

C-5. SELECTION OF WATER AND SAND CONTENT

C-5.1 From Table 4, for 20 mm nominal maximum size aggregate and sand conforming to grading Zone II, water content per cubic metre of concrete = 186 kg and sand content as percentage of total aggregate by absolute volume = 35 percent.

For change in values in water-cement ratio, compacting factor and sand belonging to Zone III, the following adjustment is required:

Change in Condition (*Ref Table 6*)	*Adjustment Required in*	
	Water Content Percent	Percentage Sand in Total Aggregate
For decrease in water-cement ratio by (0·60 – 0·50) that is 0·1	0	– 2·0
For increase in compacting factor (0·9 – 0·8) that is 0·10	+ 3	0
For sand conforming to Zone III of Table 4 of IS : 383-1970	0	– 1·5
Total	+ 3 percent	– 3·5

Therefore, required sand content as percentage of total aggregate by absolute volume = 35 — 3·5 = 31·5 percent

$$\text{Required water content} = 186 + \frac{186 \times 3}{100} = 186 + 5{\cdot}58 = 191{\cdot}6 \ \text{l/m}^3$$

C-6. DETERMINATION OF CEMENT CONTENT

Water cement ratio = 0·50

Water = 191·6 l

$$\text{Cement} = \frac{191{\cdot}6}{0{\cdot}50} = 383 \ \text{kg/m}^3$$

This cement content is adequate for mild exposure condition, according to Appendix A of IS : 456-1978*.

C-7. DETERMINATION OF COARSE AND FINE AGGREGATE CONTENT

C-7.1 From Table 3, for the specified maximum size of aggregate of 20 mm, the amount of entrapped air in the wet concrete is 2 percent. Taking this into account and applying equations from **3.5.1**,

$$0{\cdot}98 \ \text{m}^3 = \left(191{\cdot}6 + \frac{383}{3{\cdot}15} + \frac{1}{0{\cdot}315} \cdot \frac{f_a}{2{\cdot}60}\right) \times \frac{1}{1\,000}$$

$$\text{and } 0{\cdot}98 \ \text{m}^3 = \left(191{\cdot}6 + \frac{383}{3{\cdot}15} + \frac{1}{0{\cdot}685} \cdot \frac{c_a}{2{\cdot}60}\right) \times \frac{1}{1\,000}$$

$$\text{or } f_a = 546 \ \text{kg/m}^3, \text{ and}$$

$$c_a = 1\,187 \ \text{kg/m}^3$$

*Code of practice for plain and reinforced concrete (*third revision*).

C-8. ACTUAL QUANTITIES REQUIRED FOR THE MIX PER BAG OF CEMENT

C-8.1 The mix is 0·50 : 1 : 1·42 : 3·09 (by mass). For 50 kg of cement, the quantity of materials are worked out as below:

a) Cement = 50 kg

b) Sand = 71·0 kg

c) Coarse aggregate = 154·5 kg (Fraction I = 92·7 kg, fraction II = 61·8 kg)

d) Water

1) For water-cement ratio of 0·50 quantity of water = 25·0 litres

2) Extra quantity of water to be added for absorption in case of coarse aggregate, at 0·5 percent by mass = (+)0·77 l

3) Quantity of water to be deducted for free moisture present in sand, at 2 percent by mass = (—) 1·42 l

4) Actual quantity of water to be added = 25·0 + 0·77 — 1·42 = 24·35 l

e) Actual quantity of sand required after allowing for mass of free moisture = 71·0 + 1·42 = 72·42 kg

f) Actual quantity of coarse aggregate required:

1) Fraction I = 92·7 — 0·46 = 92·24 kg

2) Fraction II = 61·8 — 0·31 = 61·49 kg

Therefore, the actual quantities of different constituents required for the mix are:

Water : 24·35 kg
Cement : 50·00 kg
Sand : 72·42 kg
Coarse aggregate: Fraction I = 92·24 kg
Fraction II = 61·49 kg

OR

15. Explain the concept of mix design and mention the method of proportioning.

Ans: Proportioning of concrete is the process of selecting quantity of cement, sand, coarse aggregate and water in concrete to obtain desired strength and quality. The proportions of coarse aggregate, cement and water should be such that the resulting concrete has the following properties:

Content each-10 M
Presentation- 2M

When concrete is fresh, it should have enough workability so that it can be placed in the formwork economically.

The concrete must possess maximum density or in the other words, it should be strongest and most water-tight.

- The cost of materials and labour required to form concrete should be minimum.

The determination of the proportions of cement, aggregates and water to obtain the required strengths shall be made as follows: a) By designing the concrete mix, such concrete shall be called design mix concrete, or b) By adopting nominal mix, such concrete shall be called nominal mix concrete.

Design mix concrete is preferred to nominal mix.

Concrete of each grade shall be analysed separately to determine its standard deviation.

Standard Deviation $S = \sqrt{\frac{\sum \Delta^2}{n-1}}$ Where, Δ = deviation of the individual test strength from the average strength of n samples. n = Number of sample test results.

Arbitrary Method of Proportioning Concrete

The general expression for the proportions of cement, sand and coarse aggregate is 1 : n : 2n by volume. 1 : 1 : 2 and 1 : 1.2 : 2.4 for very high strength. 1 : 1.5 : 3 and 1 : 2 : 4 for normal works. 1 : 3 : 6 and 1 : 4 : 8 for foundations and mass concrete works.

NSRIT

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	**Program**	Mechanical Engg.			**Academic Year**	2021 - 2022
Course Code	20ME403	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	IV
Course	KINEMATICS OF MACHINERY						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What is meant by degrees of freedom of a mechanism?	20ME403.1	L1
2	State an application of Peaucellier mechanism.	20ME403.2	L1
3	What are the different types of instantaneous centres?	20ME403.3	L1
4	Why a Roller follower is preferred over Knife Edge follower?	20ME403.4	L1
5	What is he Interference in Involute Gears? How to avoid it?	20ME403.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Discuss various types of constrained motion.	6M	20ME403.1	L2
6 (b)	How is the Whitworth quick-return mechanism and crank slotted-lever mechanism different from each other? Explain.	6M	20ME403.1	L2
	OR			
7 (a)	Explain with neat figures the inversions of Double Slider Crank Chain.	6M	20ME403.1	L2
7 (b)	Describe different inversions of quadric cycle chain.	6M	20ME403.1	L2
8(a)	Explain with a neat sketch, Pantograph mechanism. State its applications.	6M	20ME403.2	L2
8 (b)	What is an automobile steering gear? What are its types? Which steering gear is preferred and why?	6M	20ME403.2	L2
	OR			
9 (a)	What is an automobile steering gear? Derive the condition for correct steering of an automobile?	6M	20ME403.2	L3
9 (b)	Draw a neat sketch of the Scott Russell's mechanism, and explain its working. How this mechanism can be modified to produce Grasshopper mechanism.	6M	20ME403.2	L2
10 (a)	Explain how by means of klein's construction the acceleration of a reciprocating engine is determined.	6M	20ME403.3	L2
10 (b)	What is instantaneous centre of rotation? State Kennedy's theorem.	6M	20ME403.3	L2
	OR			
11 (a)	PQRS is a four bar chain with link PS fixed. The lengths of the links are PQ= 62.5 mm; QR = 175 mm; RS = 112.5 mm; and PS = 200 mm. The crank PQ rotates at 10 rad/s clockwise. Draw the when angle QPS = 60° and Q and R lie on the same side of PS. Find the angular velocity and angular acceleration of links QR and RS?	6M	20ME403.3	L3
11 (b)	What is the coriolis acceleration component?	6M	20ME403.3	L2
12 (a)	Explain with sketches the different types of cams and followers.	6M	20ME403.4	L2
12 (b)	Discuss briefly the various types of belts used for the transmission of power.	6M	20ME403.4	L2
	OR			
13	A Cam with 30mm as minimum dia is rotating clockwise at uniform speed of 1300rpm and has to give the following motion	12M	20ME403.4	L3

	to the roller follower12mm in Dia (a)follower to complete the outstroke of 25mm during 120^0 of cam rotation with uniform Acceleration and Retardation (b)follower to dwell for 60^0 of cam rotation (c)follower to return to its initial position during 90^0 of cam rotation with uniform Acceleration and Retardation (d)follower to dwell for the remaining 90^0 of cam rotation. Draw the Cam profile if the axis of the roller follower passes through the axis of the cam, also determine Max. Velocity and Acceleration of the follower during the outstroke and return stroke.			
14	Derive an expression for length of path of contact, length of arc contact and contact ratio for a pair of involute gears in contact.	12M	20ME403.5	L2
	OR			
15	Two 20^0 Involute Spur gears have a Velocity Ratio of 2.5and mesh externally. The module is 4mm and addendum is equal to 1.23 module . The pinion rotates at 150rpm.Find (1)Minimum No. of teeth on each wheel to avoid interference. (2)No. of pairs of teeth in contact. (3)Max. Sliding Velocity	12M	20ME403.5	L3

NSRIT

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

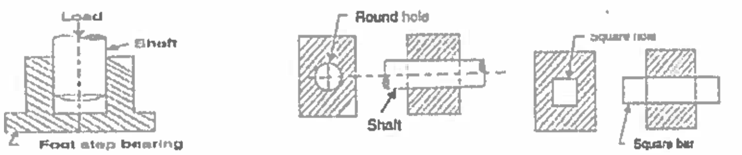
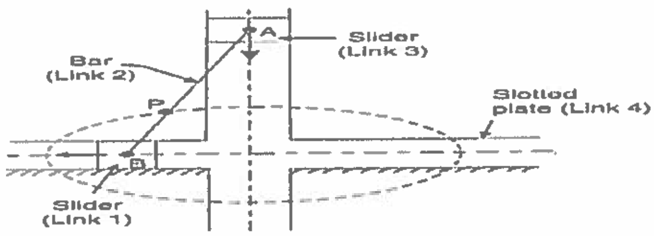
SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

Degree	B.Tech (U.G.)			Year	II	**Academic Year**	2021 - 2022
Course Code	20ME403	**Test Duration**	3 Hrs	Max. Marks	70	**Semester**	IV
Course	KINEMATICS OF MACHINERY						

Part A

No.	Answers	Marks
1.	In the design or analysis of a mechanism, one of the most important concern is the number of degrees of freedom (also called movability) of the mechanism. It is defined as the number of input parameters (usually pair variables) which must be independently controlled in order to bring the mechanism into a useful engineering purpose. It is possible to determine the number of degrees of freedom of a mechanism directly from the number of links and the number and types of joints which it includes. $$n = 3\,(l-1) - 2\,j - h$$	Definition -2M
2.	A straight line motion is a common application in engineering design and manufacture. The Peaucellier mechanism **generates exact straight lines, meeting some restrictions among their links dimensions and the input angle.**	Applications -2M
3.	Number of I-centres:- for two bodies having relative motion between them, there is an I-centre. In a mechanism, the number of I-centres will be equal to possible pairs of bodies or links. $\therefore N = \frac{n(n-1)}{2}$ where N = Number of I-centres. n = number of bodies or links.	Definition -2M
4.	**Roller follower is preferred over the knife edge follower**, the main reason of this is.. Knife edge follower has sliding action on the cam plate due to which there is more friction resulting in rapid wear and tear of cam plate, as well as more power is required for driving the cam due to high friction.	Reason = 2M

5.	**Tooth stubbing** (In this process a portion of the tip of the teeth is removed, thus preventing that portion of the tip of the tooth in contacting the non-involute portion of the other meshing tooth). Increasing the number of teeth on the gear can also eliminate the chances of interference.	Condition = 1M Reason = 1M
6.	**PART – B** **a)** Types of constrained motions :- 1.completely constrained motion :- When the motion between a pair is limited to a definite direction irrespective of the direction of force applied, then the motion is said to be a completely constrained motion. 2.incompletely constrained motion :- When the motion between a pair can take place in more than one direction, the motion is called an incompletely constrained motion. 3 3.successfully constrained motion:- When the motion between the elements, forming a pair, is such that the constrained motion is not completed by it self, but by some other means, then the motion is said to be successfully constained motion.  Shaft in a foot step bearing. Shaft in a circular hole. Square bar in a square hole. **b)** Whitworth is Crack- Crack type mechanism while Slotted bar is Crank-Rocker type mechanism. In Whitworth Crank is the fixed link while in Slotted bar Connecting rod is the fixed link. Whitworth QRM and Slotted QRM bar are different inversions of same mechanism (i.e. Slider Crank Mechanism) We can have better control of Whitworth mechanism comapred to Slotted bar because it's output depends upon 3 links. **OR**	Types with diagrams = 6M Differences = 6m
7	**a) Inversion of double slider crank chain :- (i)Elliptical trammels :-**  It is an instrument used for drawing ellipses. **(ii)scotch yoke mechanism:-**	3 inversions = 6m

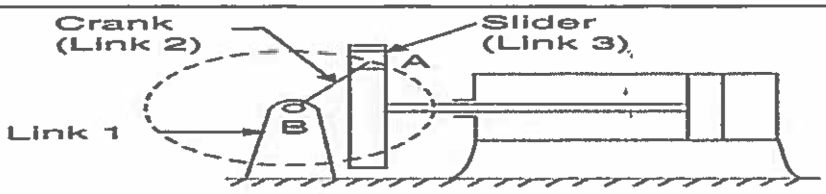
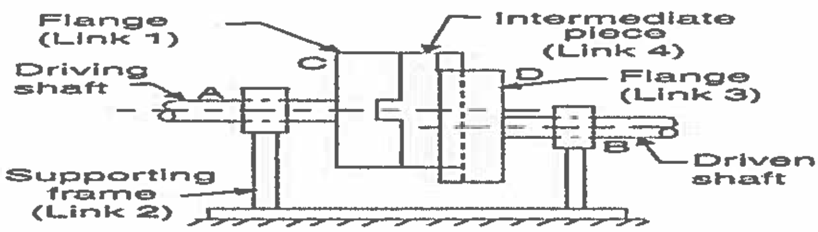


Fig. 5.35. Scotch yoke mechanism.

Converts rotary motion in reciprocating motion.

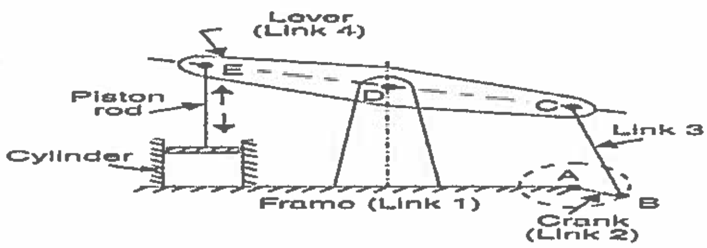
(iii)oldham's coupling:-

An oldham's coupling is used for connecting two parallel shafts whose axes are at a small distance apart. The shafts are coupled in such a way that if one shaft rotates, the other shafts also rotates at the same speed.



b) Inversion of four bar chain:-

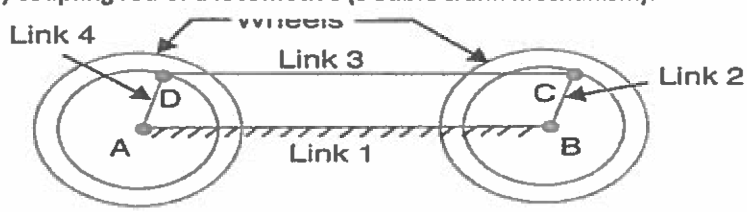
(i) beam engine (crank and lever mechanism) :-



Beam engine.

In other works, the purpose of this mechanism is the convert rotaty motion into reciprocating motion.

(ii) coupling rod of a locomotive (Double crank mechanism):-



Coupling rod of a locomotive.

Here we transfer rotating motion from one wheel to another wheel.

(iii). watt's indicator mechanism (Double lever mechanism):-

3 inversions
= 6m

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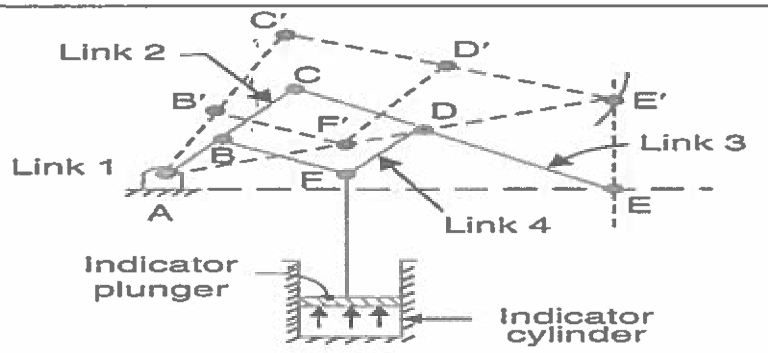

Watt's indicator mechanism.

The displacement of the link BFD is directly proportional to the pressure of gas or steam which acts on the indicator plunger on any small displacement of the mechanism, the tracing point 'E' at the end of the link CE traces out approximately a straight line.
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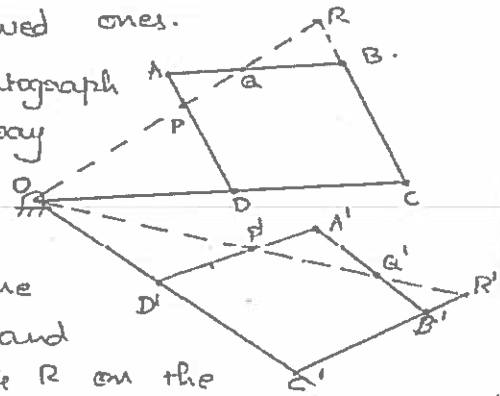
a)

Pantograph:-

A pantograph is a four-bar linkage used to produce paths exactly similar to the ones traced out by a point on the linkage. The paths produced are an enlarged or reduced scale and may be straight or curved ones.

Four links of a pantograph are arranged in such a way that a parallelogram ABCD is formed. Thus, AB = DC and BC = AD

If some point O in one of the links is fixed and three other points P, Q & R on the

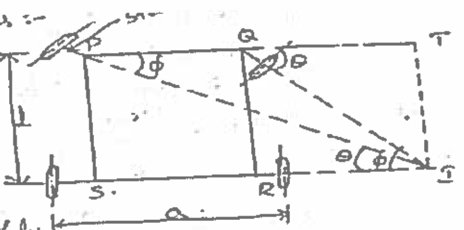


widely used in electrical locomotives to transfer electricity etc..,

b)

Automobile Steering Gears:-

When an automobile takes turn on a road, all the wheels should make concentric circles to ensure that they roll on the road smoothly and there is a line contact b/w the tyres and the surface of the path, preventing the excess wear of tyres.



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Derivation = 4m

Applications = 2m

Derivation types and conclusion = 6m
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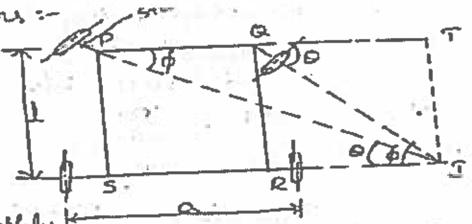
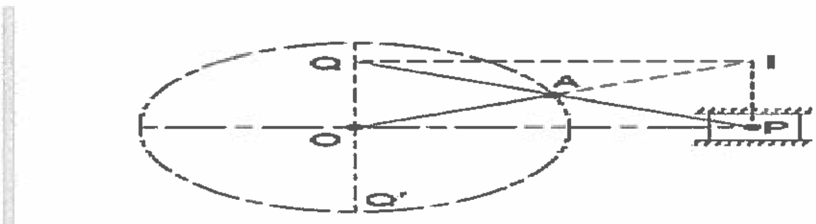
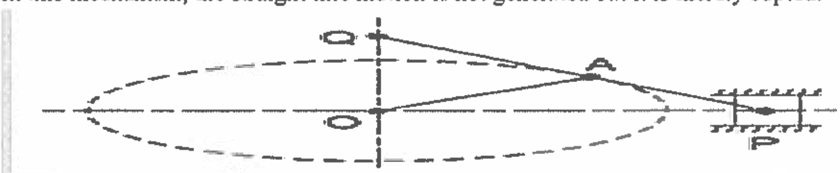
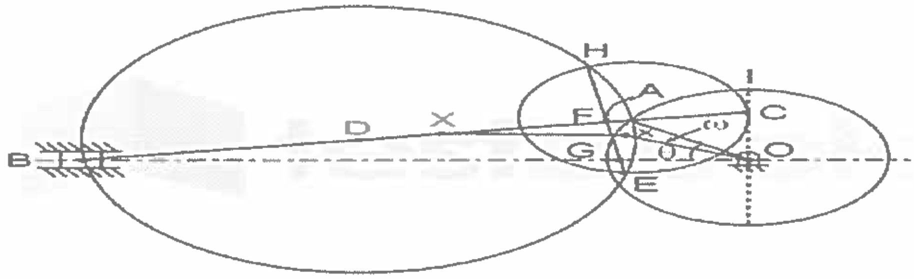
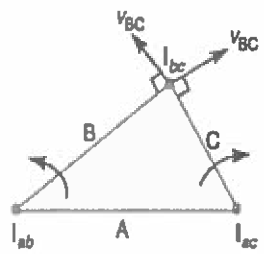
	Types 1. Davis steering gear 2.Ackermann steering gear Ackermann steering gear is mostly preferred.	
	OR	
9	a) Automobile Steering Gears :- When an automobile takes turn on a road, all the wheels should make concentric circles to ensure that they roll on the road smoothly and there is a line contact b/w the tyres and the surface of the path preventing the excess wear of tyres.  Let θ & ϕ = angles turned by the stub axles l = wheel base w = distance b/w the pivots of front axles. Then, $\cot\phi = \frac{PT}{TI}$ $\cot\theta = \frac{QT}{TI}$ $\cot\phi - \cot\theta = \frac{PT-QT}{TI} = \frac{PQ}{TI} = \frac{w}{l}$.	Definition and derivation = 6M
	b)  **Fig. 9.5. Scott Russell's mechanism.** In this mechanism, the straight line motion is not generated but it is merely copied.  **Fig. 9.7. Modified Scott-Russel mechanism.** *Modified Scott-Russel mechanism*. This mechanism, as shown in Fig. 9.7, is similar to Scott-Russel mechanism (discussed in Art. 9.5), but in this case AP is not equal to AQ and the points P and Q are constrained to move in the horizontal and vertical directions. A little consideration will show that it forms an elliptical trammel,	Definitions = 3M Dia grams = 1M Difference = 3M

10	a) Klein's Construction: It is used to draw the velocity and acceleration diagrams for a single slider crank mechanism. The velocity and acceleration of piston of a reciprocating engine mechanism can be determined by the figure given below.  Then, Acceleration of point x, fx = $\omega^2 \times OX$. Though, we can calculate both velocity and acceleration through Klein's construction but mainly it is used in calculating the linear acceleration of the piston.	Derivation = 6M
	The Aronhold Kennedy's theorem states that *if three bodies move relatively to each other, they have three instantaneous centres and lie on a straight line.* Consider three kinematic links A, B and C having relative plane motion. The number of instantaneous centres (N) is given by $N = \frac{n(n-1)}{2} = \frac{3(3-1)}{2} = 3$ where n = Number of links = 3 The two instantaneous centres at the pin joints of B with A, and C with A (*i.e.* I_{ab} and I_{ac}) are the permanent instantaneous centres. According to Aronhold Kennedy's theorem, the third instantaneous centre I_{bc} must lie on the line joining I_{ab} and I_{ac}. In order to prove this,  Fig. 6.7. Aronhold Kennedy's theorem. b)	Derivation = 6M
	OR	
11	a) Problem First step- given data Second step-calculating and speed determination	Given data-1M Formula = 1M Calculation and steps-4M Final answer-2M
	b) Coriolis component of acceleration • The acceleration of point P with respect to O is= sum of acceleration of point Q with respect to O and the acceleration of point P with respect to Q.	Definition = 2m Dia Grams = 2 M Procedure = 2m

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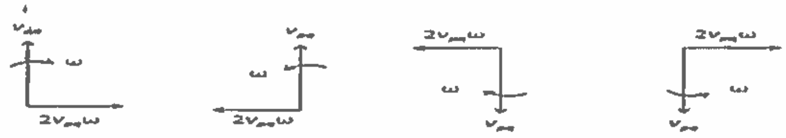
- The direction of coriolis component is such as to rotate the slider velocity vector in the same sense as the angular velocity of the link OP.



Directions of coriolis components

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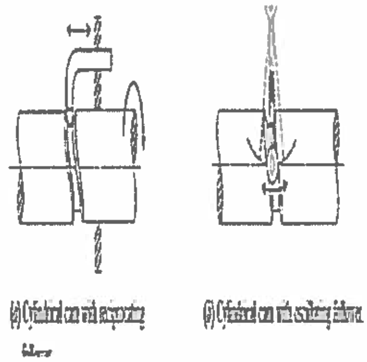
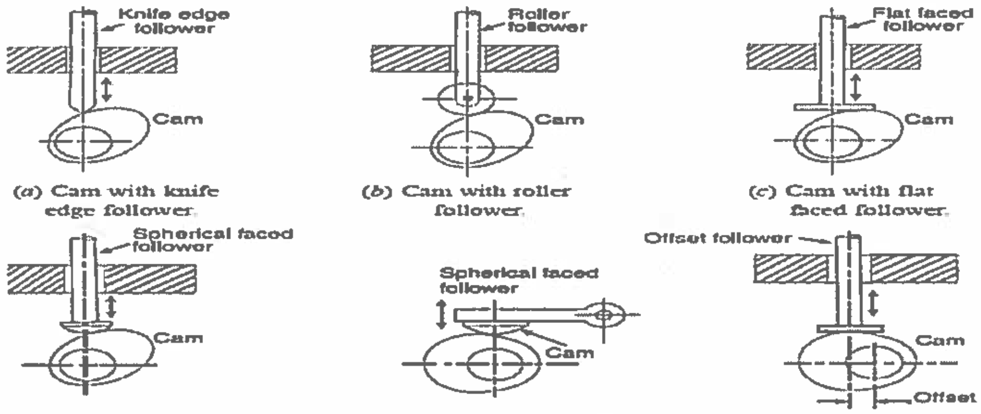
a)

1. *Radial or disc cam.* In radial cams, the follower reciprocates or oscillates in a direction perpendicular to the cam axis. The cams as shown in Fig. 20.1 are all radial cams.

2. *Cylindrical cam.* In cylindrical cams, the follower reciprocates or oscillates in a direction parallel to the cam axis. The follower rides in a groove at its cylindrical surface. A cylindrical grooved cam with a reciprocating and an oscillating follower is shown in Fig. 20.2 (*a*) and (*b*) respectively.

Note : In actual practice, radial cams are widely used. Therefore our discussion will be only confined to radial cams.





(*a*) Cam with knife edge follower. (*b*) Cam with roller follower. (*c*) Cam with flat faced follower. (*d*) Cam with spherical faced follower. (*e*) Cam with spherical faced follower. (*f*) Cam with offset follower.

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<td>Types With Dia Grams = 6m</td>
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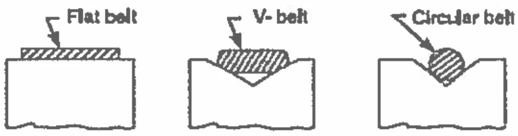
b)



(*a*) Flat belt. (*b*) V-belt. (*c*) Circular belt.

Fig. 11.1. Types of belts.

Though there are many types of belts used these days, yet the following are important from the subject point of view :

1. *Flat belt.* The flat belt, as shown in Fig. 11.1 (*a*), is mostly used in the factories and workshops, where a moderate amount of power is to be transmitted, from one pulley to another when the two pulleys are not more than 8 metres apart.

2. *V-belt.* The V-belt, as shown in Fig. 11.1 (*b*), is mostly used in the factories and workshops, where a moderate amount of power is to be transmitted, from one pulley to another, when the two pulleys are very near to each other.

3. *Circular belt or rope.* The circular belt or rope, as shown in Fig. 11.1 (*c*), is mostly used in the factories and workshops, where a great amount of power is to be transmitted, from one pulley to another, when the two pulleys are more than 8 meters apart.

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<td>Types With Dia Grams = 6m</td>
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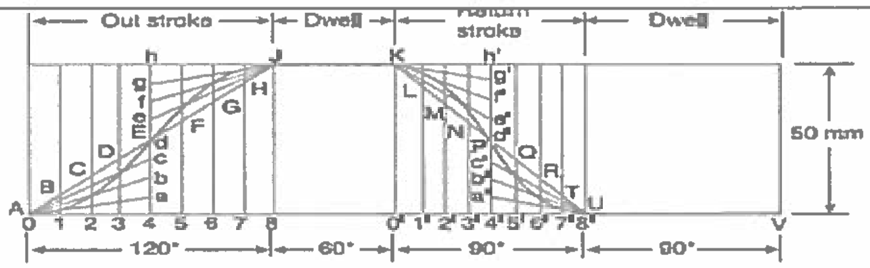

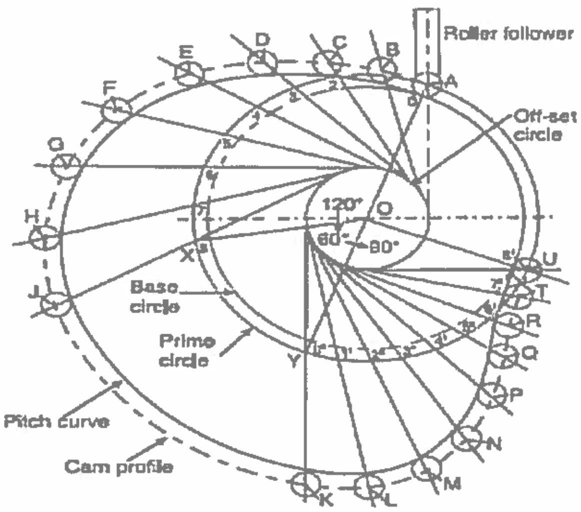

$\omega = \frac{2\pi N}{60}$ $v_O = \frac{2\omega S}{\theta_O}$; $v_R = \frac{2\omega S}{\theta_R}$ $a_O = \frac{4\omega^2 S}{(\theta_O)^2}$ $a_R = \frac{4\omega^2 S}{(\theta_R)^2}$ Calculation
</td>
<td>Cam Profile Drawing = 8 M

Sum = 4m</td>
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We know that the length of the arc of approach (arc GP)

$$= \frac{\text{Length of path of approach}}{\cos\phi} = \frac{KP}{\cos\phi}$$

and the length of the arc of recess (arc PH)

$$= \frac{\text{Length of path of recess}}{\cos\phi} = \frac{PL}{\cos\phi}$$

Since the length of the arc of contact GPH is equal to the sum of the length of arc of approach and arc of recess, therefore,

Length of the arc of contact

$$= \text{arc } GP + \text{arc } PH = \frac{KP}{\cos\phi} + \frac{PL}{\cos\phi} = \frac{KL}{\cos\phi}$$

$$= \frac{\text{Length of path of contact}}{\cos\phi}$$

The contact ratio or the number of pairs of teeth in contact is defined as the ratio of the length of the arc of contact to the circular pitch. Mathematically,

Contact ratio or number of pairs of teeth in contact

$$= \frac{\text{Length of the arc of contact}}{p_c}$$

where p_c = Circular pitch = πm, and

m = Module.
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<td>3 Derivations = 3*4 = 12M</td>
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<td>15</td>
<td>Problem

First step- given data

Second step-calculating and speed determination</td>
<td>Given data-1M
Formula = 1M
Calculation and steps-8M
Final answer-2M</td>
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PDC ✓

KOM ✓

EMWTL sultan

CPM chatnika (CO)

TOC shanbaz

DAA mounli.

	7a	7b	8a
28 –	4	4	4

Mid-Term Question Paper

Degree	B. Tech. (U. G.)	**Program**	EEE	**Test**	I	**Academic Year**	2021- 2022
Course Code	20ESX01	**Test Duration**	90 Min.	Max. Marks	40	**Semester**	II
Course							

Assessment Pattern

R (L1):	13%	U (L2):	37	Apply (L3):	50	Analyze (L4):		E (L5):		C (L6)

Part A (Short Answer Questions 2 x 5 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1.	Construct a regular hexagon of side 40 mm one side is (a) horizontal (b) Vertical.	20ESX01.1	L1
2	A point A is 15mm above HP and 20mm in front of VP another Point B is 25mm behind VP and 40mm below HP. Draw the Projection of A and B keeping the distance between the projectors equal to 90mm draw the straight line joining (I) the top views (ii) the front views.	20ESX01.2	L2

Part B (Long Answer Questions 3 x 10 = 30 Marks)

No.	Questions (6 through 11)	Learning Outcome (s)	DoK
3	Construct a hyperbola when distance between focus and directrix is 50 mm the eccentricity is 3/2. Draw the tangent and normal at any point on the curve	20ESX01.1	L3
	OR		
4	Construct a parabola in Parallelogram of side 100mm and 60mm with an included angle of 75^0	20ESX01.1	L3
5	Draw the Projection of the following points keeping the distance between the projectors as 20mm on the same reference line. (i) Point A 20 mm above the HP and 40 mm in front of VP (ii) Point B is 40mm above HP and 55mm behind the VP (iii) Point C is 40 mm below HP and 30mm behind the VP (iv) Point D is 35 mm below HP and 45 mm in front of VP. (v) Point E is on HP and 30mm in front of VP.	20ESX01.2	L2
	OR		

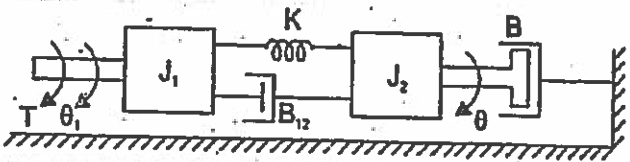
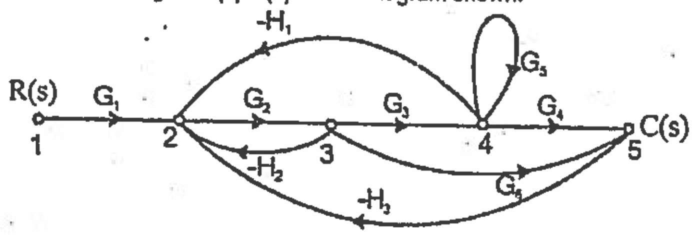
Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	Program	EEE & ECE			Academic Year	2021 - 2022
Course Code	20EE403	Test Duration	3Hrs	Max. Marks	70	Semester	IV
Course	Control Systems						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What are the properties of signal flow graph?	20EE403.1	L1
2	What is steady-state error?	20EE403.2	L1
3	What are the necessary conditions for stability of root locus?	20EE403.3	L1
4	Define corner frequency in bode diagram.	20EE403.4	L1
5	List any three properties of state transition matrix.	20EE403.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 11)	Marks	Learning Outcome (s)	DoK
6 (a)	For the mechanical rotational system derive the transfer function Θ(S)/T(S). 	6M	20EE403.1	L3
6 (b)	Explain about closed loop control system with an example.	6M	20EE403.1	L2
	OR			
7 (a)	Find the overall gain C(s)/R(s) for the diagram shown. 	6M	20EE403.1	L3
7 (b)	Explain the operation of synchro transmitter and receiver.	6M	20EE403.1	L2
8 (a)	A unity feedback control system is characterized by the following open loop transfer function $G(s) = (s+10)/(s+2)(s+6)$. Determine its transient response for unit step input.	6M	20EE403.2	L3
8 (b)	What are the generalized error coefficients?	6M	20EE403.2	L2
	OR			
9 (a)	A unity feedback system has the forward transfer function $G(s) = \frac{K(2s+1)}{s(5s+1)(s+1)^2}$. When the input $r(t) = 1 + 6t$, determine the minimum value of K so that steady state error is less than 0.	10M	20EE403.2	L3
9 (b)	What are the effects of PI controller on system performance?	2M	20EE403.2	L2
10	Construct the Routh array and determine the stability of the system represented by the characteristic equation, $s^7++9s^6+24s^5+24s^4+24s^3+24s^2+23s+15=0$. Comment of the location of roots of the characteristic equation and stability.	12M	20EE403.3	L3

	OR			
11	Sketch the root locus of the system whose open loop transfer function is $G(s) = \frac{K}{s(s^2+4s+13)}$. Commënt on stability.	12M	20EE403.3	L3
12	For the following transfer function, plot the bode plot and comment on stability. $$G(s) = \frac{5(1+2s)}{(4s+1)(0.25s+1)}$$	12M	20EE403.4	L3
	OR			
13	Consider unity feedback system whose open loop transfer function is, $$G(s) = \frac{K}{s(0.2s+1)(0.05s+1)}$$ Sketch the polar plot and determine the range of K so that, (a) gain margin is 18 dB, (b) phase margin is 60°.	12M	20EE403.4	L3
14	Obtain the state transition matrix for the state equation of the continuous control system. $$\begin{bmatrix} \dot{x_1} \\ \dot{x_2} \\ \dot{x_3} \end{bmatrix} = \begin{bmatrix} 1 & 2 & 3 \\ 6 & 2 & 4 \\ 7 & 8 & 1 \end{bmatrix} \begin{bmatrix} x_1 \\ x_2 \\ x_3 \end{bmatrix}$$	12M	20EE403.5	L3
	OR			
15	Check the given system is completely controllbale and completely observable. Comment on it. $$\begin{bmatrix} \dot{x_1} \\ \dot{x_2} \\ \dot{x_3} \end{bmatrix} = \begin{bmatrix} 0 & 2 & 4 \\ 1 & 5 & 2 \\ 1 & -2 & 5 \end{bmatrix} \begin{bmatrix} x_1 \\ x_2 \\ x_3 \end{bmatrix} + \begin{bmatrix} 1 \\ 2 \\ 0 \end{bmatrix} u$$ $$y = \begin{bmatrix} 1 & 1 & 0 \end{bmatrix} \begin{bmatrix} x_1 \\ x_2 \\ x_3 \end{bmatrix}$$	12M	20EE403.5	L3

ANSWER KEY AND SCHEME OF EVALUATION

1A) Properties of signal flow graph.

The basic properties of signal flow graph are the following:-

(i) The algebraic equations which are used to construct signal flow graph must be in the form of cause and effect relationship.

(ii) Signal flow graph is applicable to linear systems only

(iii) A node in the signal flow graph represents the variable or signal.

(iv) A node adds the signals of all incoming branches and transmits the sum to all outgoing branches.

2A) STEADY STATE ERROR :-

The steady state error is the value of error signal e(t), when t tends to infinity. The steady state error is a measure of system accuracy. These errors arise from the nature of inputs, type of system and from non linearity of system components. The steady state performance of a stable control system is generally judged by its steady state error to step, ramp and parabolic inputs.

3A) Stability Criterion in Root Locus :

The poles lying on the ~~left half of~~ the s-plane will give us stability indication. If all the poles are lying on the left half of the s-plane then the system is said to be stable.

(4A). Corner frequency:-

The magnitude plot can be approximated by asymptotic straight lines. The frequencies corresponding to the meeting point of asymptotes are called corner frequency. The slope of the magnitude plot changes at every corner frequency.

(5A) Properties:-

$[\phi(t) \overset{\Delta}{=} e^{At} = L^{-1}[SI - A)^{-1}]$.

1. $\phi(0) = I$
2. $\phi^{-1}(t) = \phi(-t)$
3. $x(0) = \phi(-t)\, x(t)$
4. $\phi(t_2 - t_1)\phi(t_1 - t_0) = \phi(t_2 - t_0)$
5. $\phi(t)^k = \phi(kt)$

LONG ANSWERS

(6a). In the given system, the torque T is the input and the angular displacement θ is the output.

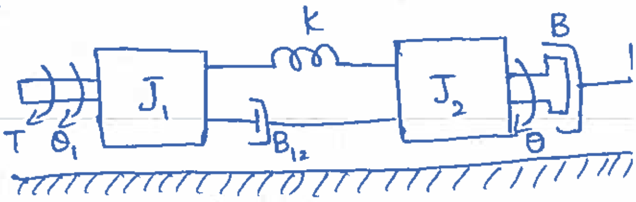


Fig 1.

let, laplace transform of $T = \mathcal{L}\{T\} = T(s)$

laplace transform of $\theta = \mathcal{L}\{\theta\} = \theta(s)$

laplace transform of $\theta_1 = \mathcal{L}\{\theta_1\} = \theta_1(s)$

Hence the required transfer function is $\frac{\theta(s)}{T(s)}$

The system has two nodes and they are masses with moment of inertia J_1 and J_2. The differential equations governing are system are given by torque balance equations at these modes.

Let the angular displacement of mass with moment of inertia J_1 be θ_1. The free body diagram of J_1 is shown in fig.2. The corresponding torques acting on J_1 are marked as T_{J_1}, $T_{b_{12}}$ and T_k

$$T_{jt} = J_1 \frac{d^2\theta_1}{dt^2} \; ; \; T_{b12} = B_{12} \frac{d}{dt}(\theta_1 - \theta_2)$$

$$T_k = k(\theta_1 - \theta)$$

Fig 2: Free body diagram of mass with moment of inertia J_1.

By Newton's second law,

$$T_{j1} + T_{b12} + T_k = T$$

$$J_1 \frac{d^2\theta_1}{dt^2} + B_{12} \frac{d}{dt}(\theta_1 - \theta) + k(\theta_1 - \theta) = T$$

On taking laplace transform of above equation with zero initial conditions we get,

$$J_1 s^2 \theta_1(s) + sB_{12}[\theta_1(s) - \theta(s)] + K\theta_1(s) - K\theta(s) = T(s)$$

$$\theta_1(s)[J_1 s^2 + sB_{12} + K] - \theta(s)[sB_{12} + K] = T(s). \longrightarrow (1)$$

The free body diagram of mass with moment of inertia J_2 is shown in fig 3. The opposing torques are marked as T_{j2}, T_{b12}, T_b and T_k.

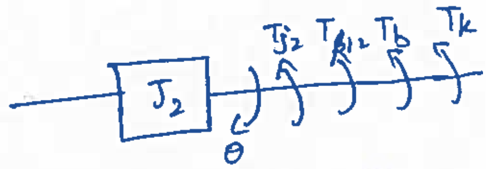


Fig 3: Free body diagram of mass with moment of inertia J_2.

$$T_{j2} = J_2 \frac{d^2\theta}{dt^2} \; ; \; T_{b12} = B_{12} \frac{d}{dt}(\theta - \theta_1)$$

$$T_b = B \frac{d\theta}{dt} \; ; \; T_k = k(\theta - \theta_1)$$

By Newton's second law,

$$T_{j2} + T_{b12} + T_b + T_k = 0$$

$$J_2 \frac{d^2\theta}{dt^2} + B_{12} \frac{d}{dt}(\theta - \theta_1) + B \frac{d\theta}{dt} + k(\theta - \theta_1) = 0$$

$$J_2 \frac{d^2\theta}{dt^2} - B_{12} \frac{d\theta_1}{dt} + \frac{d\theta}{dt}(B_{12} + B) + k\theta - K\theta_1 = 0$$

On taking laplace transform of above equation with zero initial conditions we get,

$$J_2 s^2\theta(s) - B_{12} s\theta_1(s) + s\theta(s)[B_{12} + B] + K\theta(s) - K\theta_1(s) = 0$$

$$\theta(s)[s^2 J_2 + s(B_{12} + B) + K] - \theta_1(s)[sB_{12} + K] = 0$$

$$\theta_1(s) = \frac{[s^2 J_2 + s(B_{12} + B) + K]}{[sB_{12} + K]}\theta(s) \longrightarrow (2)$$

Substituting for $\theta(s)$ from equation (2) in equation (1), we get.

$$[J_1 s^2 + sB_{12} + K] \frac{[J_2 s^2 + s(B_{12} + B) + K]\theta(s)}{(sB_{12} + K)} - (sB_{12} + K)\theta(s) = T(s)$$

$$\left[\frac{(J_1 s^2 + s B_{12} + k)[J_2 s^2 + s(B_{12} + B) + k] - (s B_{12} + k)^2}{(s B_{12} + k)}\right] \theta(s) = T(s)$$

$$\therefore \frac{\theta(s)}{T(s)} = \frac{(s B_{12} + k)}{(J_1 s^2 + s B_{12} + k)[J_2 s^2 + s(B_{12} + B) + k] - (s B_{12} + k)^2}.$$

6 (b)

Closed loop control system:

Control system in which output has an effect upon the input quantity in order to maintain the desired output value are called closed loop systems.

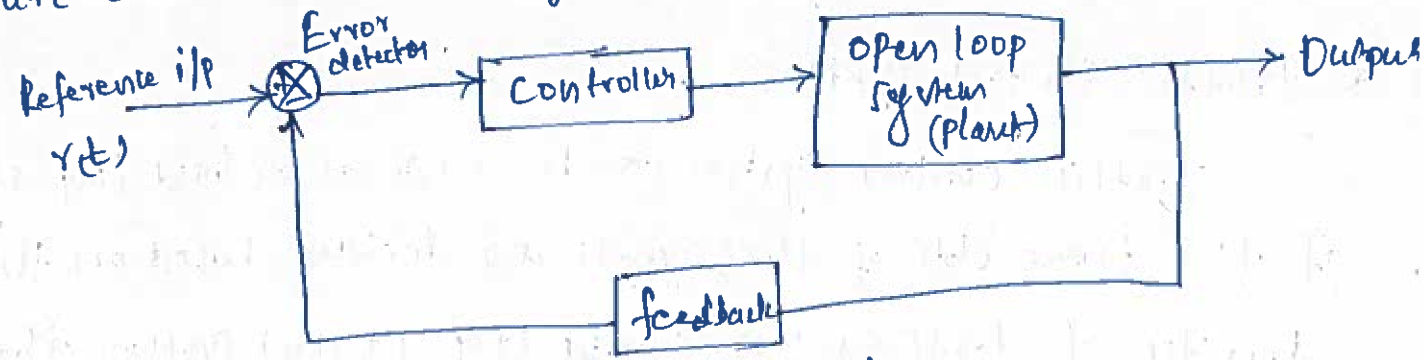


Fig. Closed loop system

The open loop systems are modified by to closed loop system by providing a feedback. The provision of feedback automatically corrects the changes in output due to disturbances. Hence the closed loop system also called ~~as~~ automatic control system. It consists of Error detector, Amplifier, Plant & feed back.

The feedback path element takes the sample output & that is fed to the error detector which will give error signal [~~Com~~ difference b/w Reference signal & feed back signal]. This error signal is fed to plant which is the main processing unit of the system.

Advantages:

1. Accurate
2. More stable
3. Less affected by noise

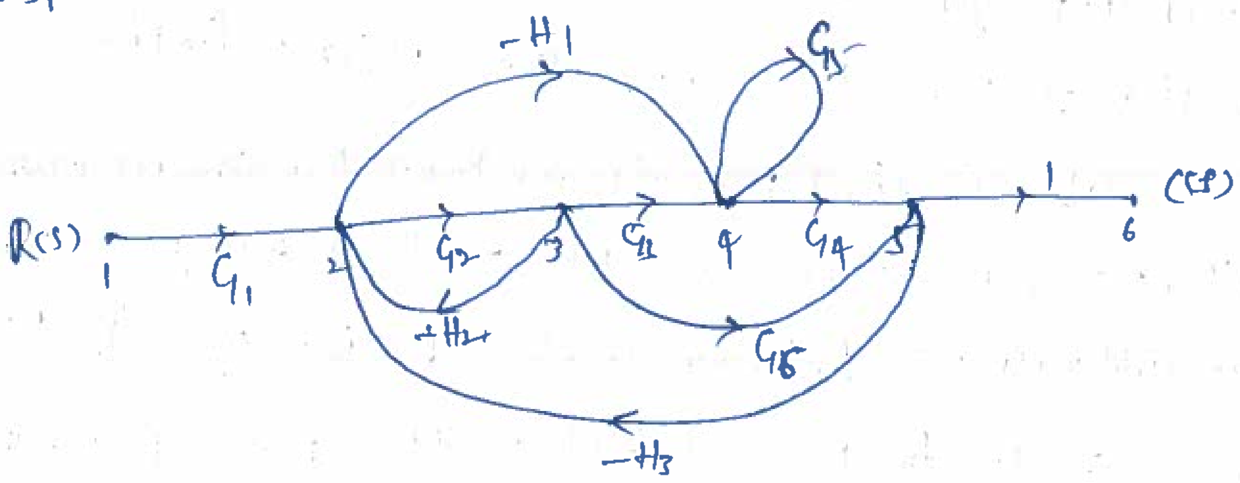
Example: (They can write any example)

1. Temprature control system
2. Numerical control system.
3. Position control system using servo motor
4. Traffic control system.

* Traffic control system:

Traffic control system can be made as a closed loop system if the time slots of the signals are decided based on the density of traffic. In closed loop control system the density of traffic is measured in terms of for each signal by sensors. Since the sensors are used as feed back these signals are fed to error detector according to the actual time will be fixed for respective line.

(7)
(a)

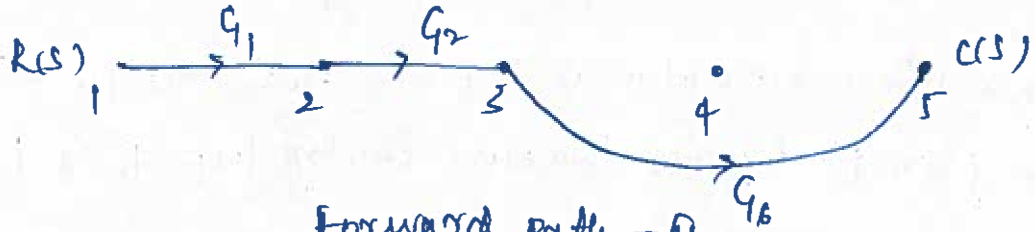


Forward paths

There are two forward paths

∴ $K=2$

Let the forward path gains are P_1 & P_2



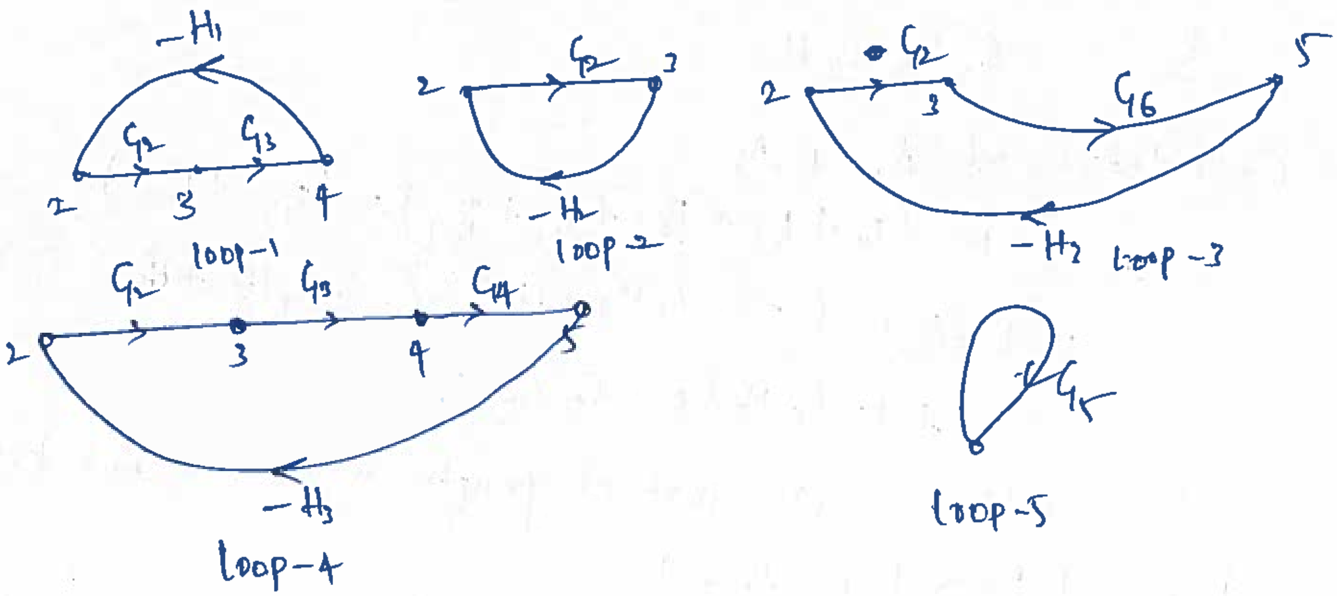
Forward path - II

Gain of forward path -1, $P_1 = G_1 G_2 G_3 G_4$

Gain of forward path -2, $P_2 = G_1 G_2 G_6$

Individual loop gains:

There are five individual loop gains, let The individual loop gains be $P_{11}, P_{21}, P_{31}, P_{41}$ & P_{51}.



Loop gains

$P_{11} = -G_2 G_3 H_1$

$P_{21} = -G_2 H_2$

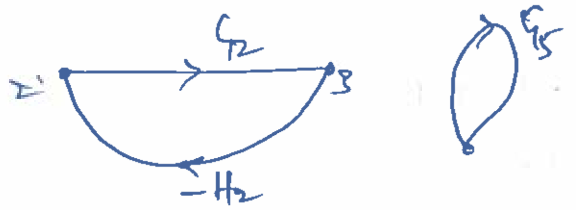
$P_{31} = -G_2 G_6 H_3$

$P_{41} = -G_2 G_3 G_4 H_3$

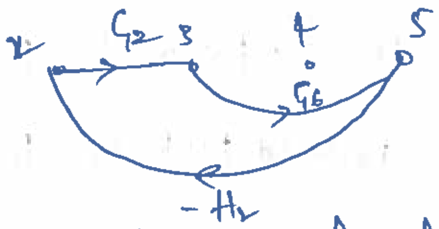
$P_{51} = G_5$

Gain products of two non-touching loops:

There are two combinations of two non-touching loops.
Let the gain products of two non-touching loops P_{12} & P_{22}



1st combination of Two Non-touching loops



2nd combination of Two-non touching loops.

$$P_{12} = P_{21}\,P_{51} = G_2 G_5 H_2$$

$$P_{22} = -G_2 G_5 - G_6 H_3$$

Calculation of Δ & Δ_k

$$\Delta = 1 - (P_{11} + P_{21} + P_{31} + P_{41} + P_{51}) + (P_{12} + P_{22})$$

$$= 1 - (-G_2 G_3 H_1 - H_2 G_2 - G_2 G_3 G_4 H_3 + G_5 - G_2 G_6 H_3)$$
$$+ (-G_2 H_2 G_5 - G_2 G_5 G_6 H_3)$$

Since, there is no part of graph which is not touching forward path-1, $\Delta_1 = 1$

The part of graph which is not touching forward path-2
$$\Delta_2 = 1 - G_5$$

Transfer function: (T)

By Mason's gain formula, 'T' is given by

$$T = \frac{1}{\Delta}\sum_k P_k \Delta_k \quad [\because k = 2]$$

$$T = \frac{G_1 G_2 G_3 G_4 + G_1 G_2 G_6 - G_1 G_2 G_5 G_6}{1 + G_2 G_3 H_1 + H_2 G_2 + G_2 G_3 G_4 H_3 - G_5 + G_2 G_6 H_3 - G_2 H_2 G_5 - G_2 G_5 G_6 H_3}$$

(7)(b) Operation of Synchro transmitter & Receiver:

The term synchro is a generic name for a family of inductive devices which works on the principle of a rotating transformer (I.M). The envlope of the carrier is modulated by the movement of wiper arm. Hence, the information is avaliable in the envlope of the carrier.

The synchro system is formed by interconnection of the devices called synchro transmitter & the synchro control transformer. They are also called synchro pair. The synchro pair measures & compares two angular displacement & its o/p voltage is approximately linear with angular difference of the axis of both the shafts.

1. To control the angular position of load from a remote place / long distance.
2. For Automatic correction of changes due to disturbances in the angular position of the load.

Synchro transmitter:

Construction: The constructional features, electrical circuit & a schematic symbol of synchro transmitter are shown in Fig. Two major parts are synchro transmitter & receiver. Each has stator & Rotor. The stator wdg is concentric type with the axis of three coils of 120° apart. A 1-ϕ AC excitation voltage is applied to rotor through slip rings.

Working principle:

When rotor is excited by AC voltage, the rotor current flows, And a magnetic field induces an emf in the stator coils by Transformer action

Let e_r = Instantaneous value of A.C. voltage applied to rotor.

e_{s1}, e_{s2}, e_{s3} = Instantaneous emf's in induced in stators of S_1, S_2, S_3 w.r.t

E_r = Max. value of rotor excitation voltage.

ω = Angular frequency of rotor excitation voltage.

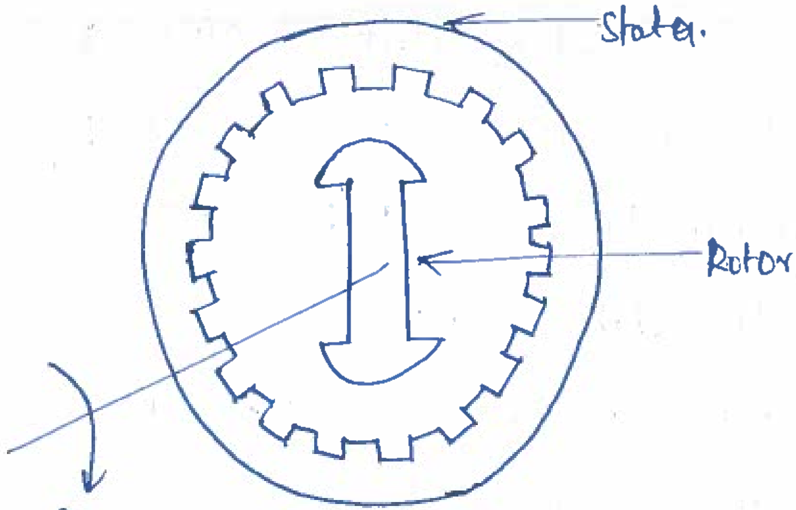
K_t = Turns ratio

K_e = Coupling Coefficient

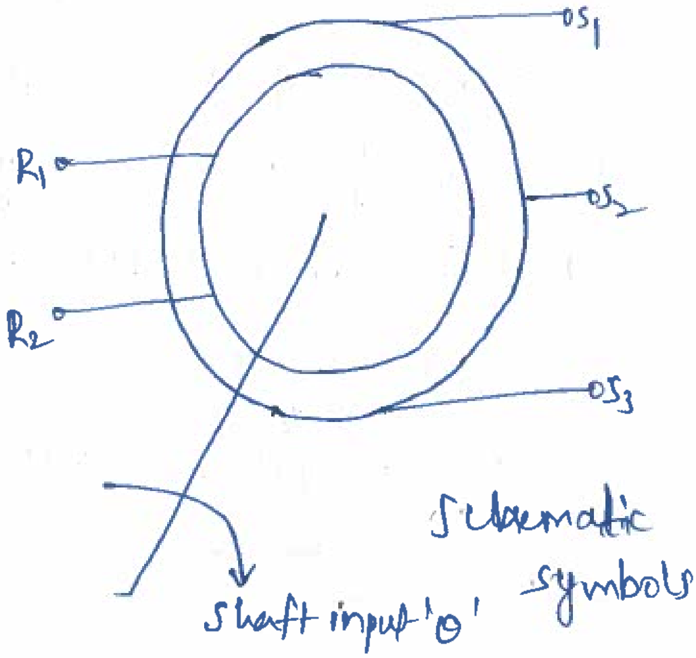
θ = Angular displacement of rotor with r.t to reference.

Let the instantaneous value of ~~stator~~ rotor excitation voltage

$e_r = E_r \sin \omega t$

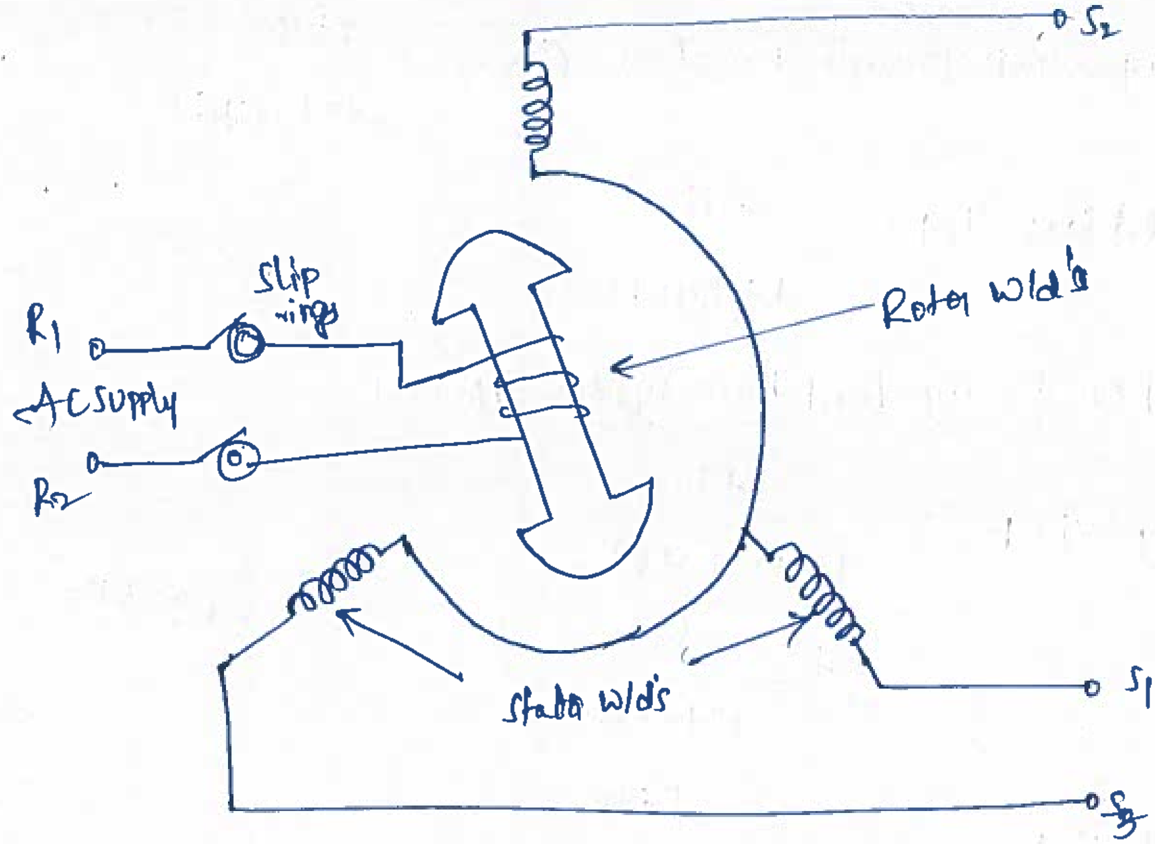


Construction features



Schematic symbols

(5)



Electrical circuit.

Emf induced in stator coil $= k_t k_c k_e \sin\omega t$

$\therefore$ Coupling coefficient, k_c for coil $S_2 = k_1 \cos\theta$

" " , k_c for " $S_3 = k_1 \cos(\theta - 120^\circ)$

" " k_c for " $S_1 = k_1 \cos(\theta - 240^\circ)$

Hence emf equations will be

$e_{S_2} = k_1 k_t \cos\theta \, E_r \sin\omega t = K E_r \cos\theta \sin\omega t$

$e_{S_3} = k_1 k_t \cos(\theta - 120^\circ) E_r \sin\omega t = K E_r \cos(\theta - 120^\circ) \sin\omega t$

$e_{S_1} = k_1 k_t \cos(\theta - 120^\circ) E_r \sin\omega t = K E_r \cos(\theta - 240^\circ) \sin\omega t$

8 (a) Given open loop Transfer function $G(s) = \frac{s+10}{(s+2)(s+6)}$

$$\text{Closed loop T.F} = \frac{G(s)}{1+G(s)H(s)}$$

For unity feed back system $H(s)=1$

$$\text{Closed loop T.F} = \frac{\frac{s+10}{(s+2)(s+6)}}{1+\frac{s+10}{(s+2)(s+6)}} \qquad s^2+8s+12$$

$$\text{C.L.T.F} = \frac{s+10}{s^2+9s+22}$$

The s-domain response, $C(s) = R(s) \times \text{T.F}$

For step i/p, $R(s) = 1/s$

$$C(s) = \frac{1}{s} \times \frac{s+10}{s^2+9s+22}$$

$$\therefore C(s) = \frac{s+10}{s(s^2+9s+22)}$$

By partial fraction expansion $C(s)$ can be expressed as,

$$C(s) = \frac{s+10}{s(s^2+9s+22)} = \frac{A}{s} + \frac{Bs+C}{s^2+9s+22}$$

$$A = C(s) \times s \Big|_{s=0} = \frac{s+10}{s^2+9s+22}\Big|_{s=0} = \frac{10}{22} = 0.454$$

The residue B & C are solved by cross multiplying the following Equation & ~~Pending~~ Equating the coefficients of the like powers of 's'

$$\frac{S+10}{S(S^2+9S+22)} = \frac{A}{S} + \frac{BS+C}{S^2+9S+22}$$

on cross multiplication

$$S+10 = A(S^2+9S+22) + BS^2 + CS$$

$$S+10 = AS^2 + A9S + 22A + BS^2 + CS$$

on Equating 's^2'

$$\cancel{AS^2+BS^2} = A + B = 0$$

$$B = -A = -0.454.$$

$$\boxed{\begin{aligned} A &= 0.454 \\ B &= -0.454 \\ C &= -3.086 \end{aligned}}$$

on Equating 's' power.

$$9A + C = 1$$

$$C = 1 - 9A = 1 - 9(0.454)$$

$$C = -3.086$$

$$\therefore C(S) = \frac{1}{S} + \frac{-0.454S - 3.086}{S^2+9S+22} = \frac{1}{S} - \frac{0.454S+3.086}{S^2+9S+22}$$

$$= \frac{1}{S} - \frac{0.454\left(S + \frac{3.086}{0.454}\right)}{S^2+9S+20.25+1.75}$$

$$\cancel{= \frac{1}{S} - 0.454(S+6.797)}$$

$$C(S) = \frac{1}{S} - \frac{(S+4.5)}{(S+4.5)^2 + \sqrt{(1.75)^2}}$$

Apply inverse laplace transformation

$$c(t) = L^{-1} C(S)$$

$$\therefore \boxed{c(t) = 1 - e^{-4.5t} \cos 4.5t}$$

(8)(b) Generalized error Coefficients:

The generalaized error coefficients gives the steady state error As a function of time. Also using the generalised error coeffi[cient] the steady state error can be found for any type of input.

The error signal in s-domain, E(s) can be expressed as product of two s-domain functions.

$$E(s) = \frac{R(s)}{1+G(s)H(s)} = \frac{1}{1+G(s)H(s)} R(s) = F(s)R(s)$$

$$\text{where } F(s) = \frac{1}{1+G(s)H(s)}$$

the drawback of in 'static' error coefficients is that it does not show the variation of error with time & input should be standard input.

The generalised error coefficients is given by,

$$C_n = (-1)^n \int_0^t T^n f(T)\,dT; \quad \text{where } F(s) = \frac{1}{1+G(s)H(s)}$$

$$F(s) = \int_0^t f(T) e^{-sT} \cdot dT$$

on taking $\lim_{s\to 0}$ on both sides of the Equation

$$\lim_{s\to 0} F(s) = \lim_{s\to 0} \int_0^t f(T) e^{-sT} \cdot dT$$

$$= \int_0^t f(T) \lim_{s\to 0} e^{-sT} dT = \int_0^t f(T) \cdot dT = C_0$$

$$\boxed{\therefore C_0 = \lim_{s\to 0} F(s)}$$

$$C_1 = \underset{s\to 0}{Lt} \frac{d}{ds} F(s)$$

$$C_2 = \underset{s\to 0}{Lt} \frac{d^2}{ds^2} F(s)$$

$$C_n = \underset{s\to 0}{Lt} \frac{d^n}{ds^n} F(s)$$

(9)
(a) given $G(s) = \dfrac{K(2s+1)}{s(5s+1)(s+1)^2}$; input $r(t) = 1+6t$

$$R(s) = L^{-1}(r(t)) = L^{-1}(1+6t) = \frac{1}{s} + \frac{6}{s^2}$$

$$E(s) = \frac{R(s)}{1+G(s)H(s)} = \frac{\frac{1}{s}+\frac{6}{s^2}}{1+\frac{K_1(2s+1)}{s(5s+1)(1+s)^2}} = \frac{\frac{1}{s}+\frac{6}{s^2}}{\frac{s(5s+1)(1+s)^2+K_1(2s+1)}{s(5s+1)(1+s)^2}}$$

$$= \frac{1}{s}\left[\frac{s(5s+1)(1+s)^2}{s(5s+1)(1+s)^2+K_1(2s+1)}\right] + \frac{6}{s^2}\left[\frac{s(5s+1)(1+s)^2}{s(5s+1)(1+s)^2+K_1(2s+1)}\right]$$

The steady state error e_{ss} can be obtained from final value theorem.

$$e_{ss} = \underset{t\to\infty}{Lt}\, e(t) = \underset{s\to 0}{Lt}\, s\,E(s)$$

$$e_{ss} = \underset{s\to 0}{Lt}\, s\left\{\frac{1}{s}\left[\frac{s(5s+1)(1+s)^2}{s(5s+1)(1+s)^2+K_1(2s+1)}\right] + \frac{6}{s^2}\left[\frac{s(5s+1)(1+s)^2}{s(5s+1)(1+s)^2+K_1(2s+1)}\right]\right\}$$

$$e_{ss} = 0 + \frac{6}{K_1} = \frac{6}{K_1}$$

Given that $e_{ss} < 0.1$ $\quad \therefore 0.1 = \dfrac{6}{K_1}$ (or) $K_1 = \dfrac{6}{0.1} = 60$

(9)(b) PI controller effect on system performance:

The proportional plus integral controller (PI), produces an o/p signal, consisting of two terms. one proportional to error signal & the other proportional to the integral of error signal

In PI controller, $u(t) \propto [e(t) + \int e(t)\,dt]$

$$\therefore u(t) = K_p e(t) + \frac{K_p}{T_i} \int e(t)\,dt$$

where K_p = proportional gain

T_i = integral time.

"The proportional Action, increases the loop gain & makes the system less sensitive to variations of system parameters. The integral action eliminates (or) reduces the the steady state Error"

(10)

OK) The characterstic eqn is $s^7+9s^6+24s^5+24s^4+24s^3+24s^2+23s+15=0$

The given characterstic polynomial is 7th order equation and so it has 7 roots. Since the highest power of s is odd number, from the first row of array using the coefficients of odd powers of s and form the second row using the coefficients of even powers of s as shown below.

$$
\begin{array}{lllllll}
s^7 & : & 1 & 24 & 24 & 23 \ldots & \text{Row-1} \\
s^6 & : & 9 & 24 & 24 & 15 \ldots & \text{Row-2}
\end{array}
$$

Divide s^6 row by 3 to simplify the computations.

$$
\begin{array}{llllllll}
s^7 & : & 1 & 24 & 24 & 23 & \ldots & \text{Row-1} \\
s^6 & : & 3 & 8 & 8 & 5 & \ldots & \text{Row-2} \\
s^5 & : & 1 & 1 & 1 & & \ldots & \text{Row-3} \\
s^4 & : & 1 & 1 & 1 & & \ldots & \text{Row-4} \\
s^3 & : & 0 & 0 & & & \ldots & \text{Row-5} \\
s^3 & : & 2 & 1 & & & \ldots & \text{Row-5} \\
s^2 & : & 0.5 & 1 & & & \ldots & \text{Row-6} \\
s^1 & : & 3 & & & & \ldots & \text{Row-7} \\
s^0 & : & 1 & & & & \ldots & \text{Row-8}
\end{array}
$$

+ to -

- to +

column-1

On examining the first column elements of routh array it is found that there are two sign changes. Hence two roots are lying on the right half of s-plane and so the system is unstable.

The row of all zeros indicates the possibility of roots on imaginary axis. this can be tested by evaluating the roots of auxiliary polynomial

The auxiliary equation is $s^4+s^2+1=0$

put $s^2=x$ in the auxiliary equation

$$s^4+s^2+1=x^2+x+1=0$$

The roots of quadratic are, $x = \frac{-1 \pm \sqrt{1-4}}{2} = -\frac{1}{2} \pm j\frac{\sqrt{3}}{2}$

$$= 1\angle 120^\circ \text{ or } 1\angle -120^\circ$$

But $s^2 = x$ $\therefore$ $s = \pm\sqrt{x} = \pm\sqrt{1\angle 120^\circ}$ or $\pm\sqrt{1\angle -120^\circ}$

$$= \pm\sqrt{1}\angle 120^\circ/2 \text{ or } \pm\sqrt{1}\angle -120^\circ/2$$

$$= \pm 1\angle 60^\circ \text{ or } \pm 1\angle -60^\circ$$

$$= \pm(0.5+j0.866) \text{ or } \pm(0.5-j0.866)$$

Two roots of auxiliary polynomial are lying on the right half of s-plane and the remaining two on the left half of s-plane. The roots of auxiliary equation are also the roots of characteristic polynomial. The two roots lying on the right half of s-plane are indicated by two sign changes in the first column of routh array. The remaining five roots are lying on the left half of s-plane. No roots are lying on imaginary axis.

11A

step 1 : To locate poles and Zeros

The poles of open loop transfer function are the roots of the equation, $s(s^2+4s+13)=0$

The roots of the quadratic are $s = \frac{-4 \pm \sqrt{4^2-4\times13}}{2} = -2 \pm j3$

$\therefore$ The poles are lying at $s=0$, $-2+j3$ and $-2-j3$

Let us denote the poles as P_1, P_2, and P_3

Here $P_1 = 0$, $P_2 = -2+j3$ and $P_3 = -2-j3$.

The poles are marked by x as shown in fig. 4.

Step 2 : To find the root locus on real axis

There is only one pole on real axis at the origin. Hence if we choose any test point on the negative real axis then to the right of that point the Total number of real poles and zeroes is one. which is an odd number. Hence the entire negative real axis will be part of root locus. The root locus on real axis is shown as a bold line in fig 4.

Step 3 : To find angles of asymptotes and Centroid

Since there are 3 poles, the number of root locus branches are three. There is no finite zero, Hence all the three root locus branches ends at zeros at infinity. The number of asymptotes required are three.

$$\text{Angles of asymptotes} = \frac{\pm 180^\circ (2q+1)}{n-m}$$

Here $n=3$, and $m=0$ $\therefore q = 0, 1, 2, 3,$

when $q=0$, Angles $= \pm \frac{180^\circ}{3} = \pm 60^\circ$

when $q=1$ Angles $= \pm \frac{180^\circ \times 3}{3} = \pm 180^\circ$

When $q=2$, Angles $= \frac{\pm 180^\circ \times 5}{3} = \pm 300^\circ = \mp 60^\circ$

When $q=3$, Angles $= \frac{\pm 180^\circ \times 7}{3} = \pm 420^\circ = \pm 60^\circ$

$$\text{Centroid} = \frac{\text{Sum of poles} - \text{Sum of zeros}}{n-m} = \frac{0-2+j3-2-j3-0}{3}$$

$$= -4/3 = -1.33$$

The centroid is marked on real axis and from the centroid the angles of asymptotes are marked using a protractor. The asymptotes are drawn as dotted lines are shown in fig 4.

step

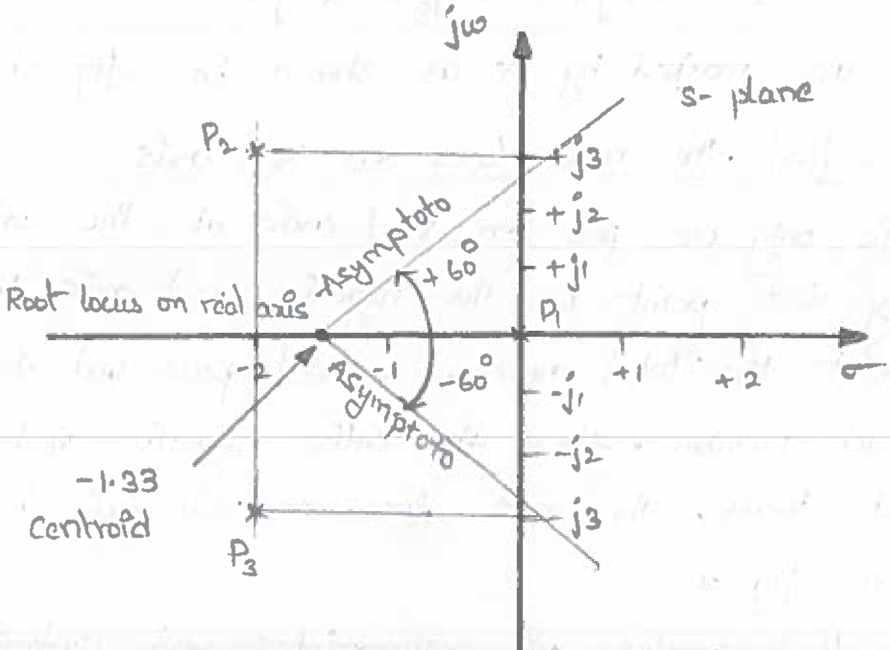


fig 4. :- Figure showing the asymptote, root locus on real axis and location of poles and centroid

step 4 : To find the breakaway and breakin points

$$\left.\begin{array}{l}\text{The closed loop}\\ \text{transfer function}\end{array}\right\} \frac{C(s)}{R(s)} = \frac{G(s)}{1+G(s)} = \frac{\frac{K}{s(s^2+4s+13)}}{1+\frac{K}{s(s^2+4s+13)}} = \frac{K}{s(s^2+4s+13)+K}$$

The characteristic equation is, $s(s^2+4s+13)+K=0$

$$\therefore\ s^3+4s^2+13s+K=0 \implies K = -s^3-4s^2-13s$$

On differentiating the eqn of K with respect s we get

$$\frac{dK}{ds} = -(3s^2+8s+13)$$

put $\frac{dK}{ds} = 0$

$\therefore -(3s^2 + 8s + 13) = 0 \implies (3s^2 + 8s + 13) = 0$

$\therefore s = \frac{-8 \pm \sqrt{8^2 - 4 \times 13 \times 3}}{2 \times 3} = -1.33 \pm j1.6$

check for K : when, $s = -1.33 + j1.6$, the value of K is given by,

$K = -(s^3 + 4s^2 + 13s) = -[(-1.33 + j1.6)^3 + 4(-1.33 + j1.6)^2 + 13(-1.33 + j1.6)]$

$\neq$ positive and real

Also it can be shown that when $s = -1.33 - j1.6$ the value of K is not equal to real and positive.

Since the values of K for, $s = -1.33 \pm j1.6$, are not real and positive, these points are not an actual breakaway or break in points. the root locus has neither breakaway nor breakin point

Step 5 : To find the angle of departure

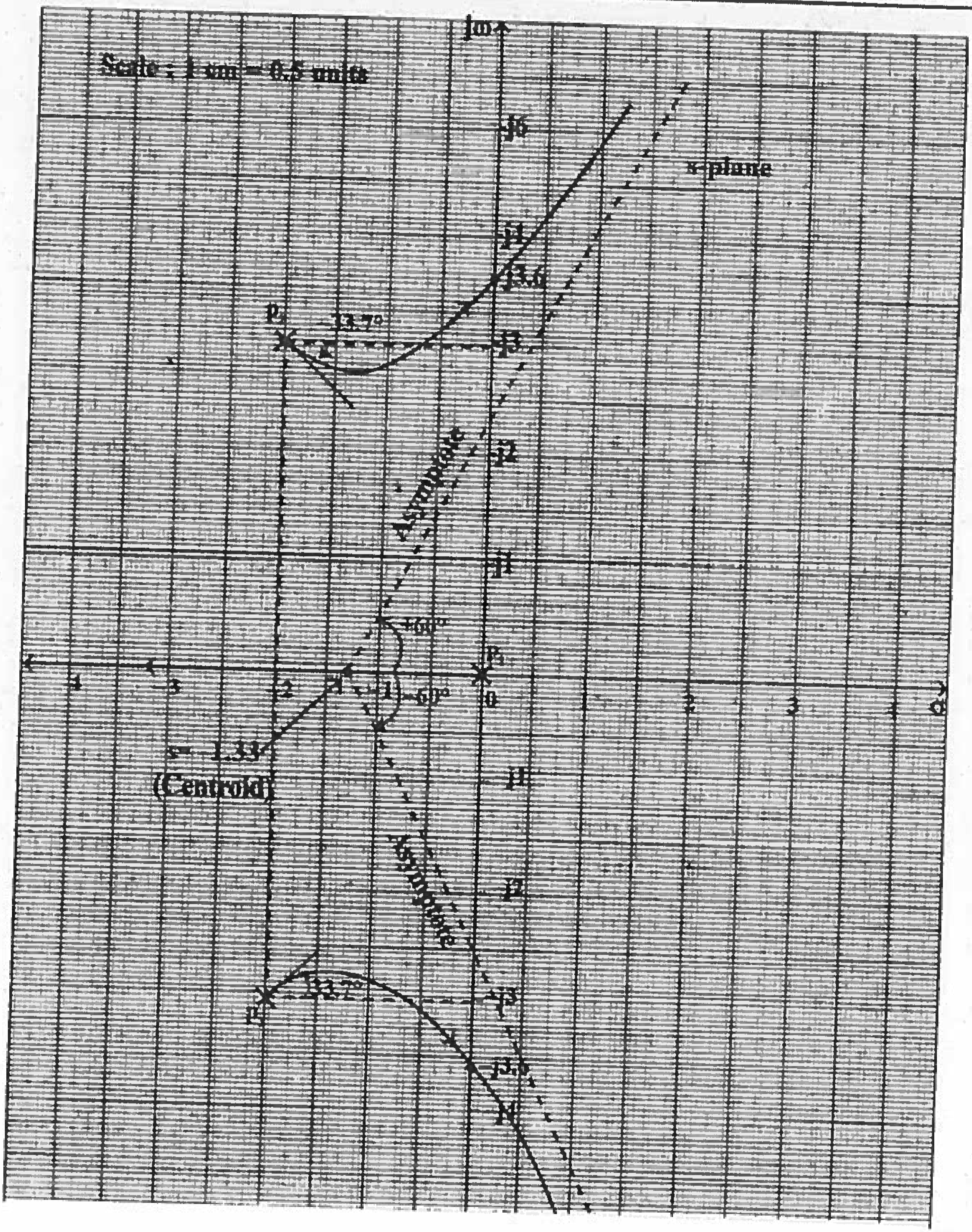
let us consider the complex pole P_2 shown in fig. Draw Vectors from all other poles to the pole P_2 as shown in fig let the angles of these Vectors be θ_1 and θ_2.

Here $\theta_1 = 180^\circ - \tan^{-1}(3/2) = 123.7^\circ$; $\theta_2 = 90^\circ$

Angle of departure from the complex pole $P_2 = 180^\circ - (\theta_1 + \theta_2)$

$= 180^\circ - (123.7^\circ + 90^\circ)$

$= -33.7^\circ$



jω
Scale : 1 cm = 0.5 units
j6
s-plane
j4
j3.6
P₂
13.7°
j3
j2
Asymptote
j1
+60°
P₁
-4
-3
-2
-1
-60°
0
1
2
3
4
σ
s= -1.33
(Centroid)
-j1
Asymptote
-j2
-13.7°
P₃
-j3
-j3.6
-j4

(12) Sol: The sinusoidal transfor function $G(j)$ is obtained by replacing by in $G(s)$

$$\therefore G(j\omega) = \frac{5(1+j2)}{(1+j4)(1+j0.25\omega)}$$

MAGNITUDE PLOT

The corner frequencies are, $c_1 = \frac{1}{4} = 0.25$ rad/sec. $c_2 = \frac{1}{2} = 0.5$ rad/sec, $c_3 = \frac{1}{0.25} = 4$ rad/sec

The verices terms of $G(j)$ are listed in the increasing order of their corner frequencies. Also the table shows the slope contributed by the each term and the change in slope at the corner frequency.

Choose a low frequency ω_l such that $\omega_l < \omega_c$ and choose a high frequency ω_h such that $\omega_h > \omega_c$

Let $\omega_l = 0.1$ rad/sec and $\omega_h = 10$ rad/sec

Let $A = |G(j\omega)|$ in db and let us calculate A at ω_l, ω_{c1}, ω_{c2}, ω_{c3} and ω_h

TABLE-1

Term	Corner frequency rad/sec	Slope db/dec	Change in slope db/deg
5	—	0	—
$\frac{1}{1+j4\omega}$	$\omega_{c1} = \frac{1}{4} = 0.25$	-20	$0-20=-20$
$1+j2\omega$	$\omega_{c2} = \frac{1}{2} = 0.5$	20	$-20+20=0$
$\frac{1}{1+j0.25\omega}$	$\omega_{c3} = \frac{1}{0.25} = 4$	-20	$0-20=-20$

At $\omega = \omega_1$, $A = |G(j\omega)| = 20\log 5 = +14\,db$

At $\omega = \omega_{c1}$, $A = |G(j\omega)| = 20\log 5 = +14\,db$

At $\omega = \omega_{c2}$, $A = \left[\text{Slope from } \omega_{c1} \text{ to } \omega_{c2} \times \log \frac{\omega_{c2}}{\omega_{c1}}\right] + A_{(at\,\omega=\omega_{c1})}$

$$= -20 \times \log \frac{0.5}{0.25} + 14 = +8\,db$$

At $\omega = \omega_{c3}$, $A = \left[\text{Slope from } \omega_{c2} \text{ to } \omega_{c3} \times \log \frac{\omega_{c3}}{\omega_{c2}}\right]$

$$+A_{(at\,\omega=\omega_{c2})} = 0 \times \log \frac{4}{0.5} + 8 = +8\,db$$

At $\omega = \omega_h$, $A = \left[\text{slope from } \omega_{c3} \text{ to } \omega_h \times \log \frac{\omega_h}{\omega_{c3}}\right]$

$$+A_{(at\,\omega_h)} = -20\log \frac{10}{4} + 8 = 0\,db$$

Let the points a, b, c, d and e be the points corresponding to frequencies $\omega_1, \omega_{c1}, \omega_{c2}, \omega_{c3}$ and ω_h respectively on the megritude plot. In a sermiog graph street choose a scale of 1unit = 5db on Y axis.

PHASE PLOT

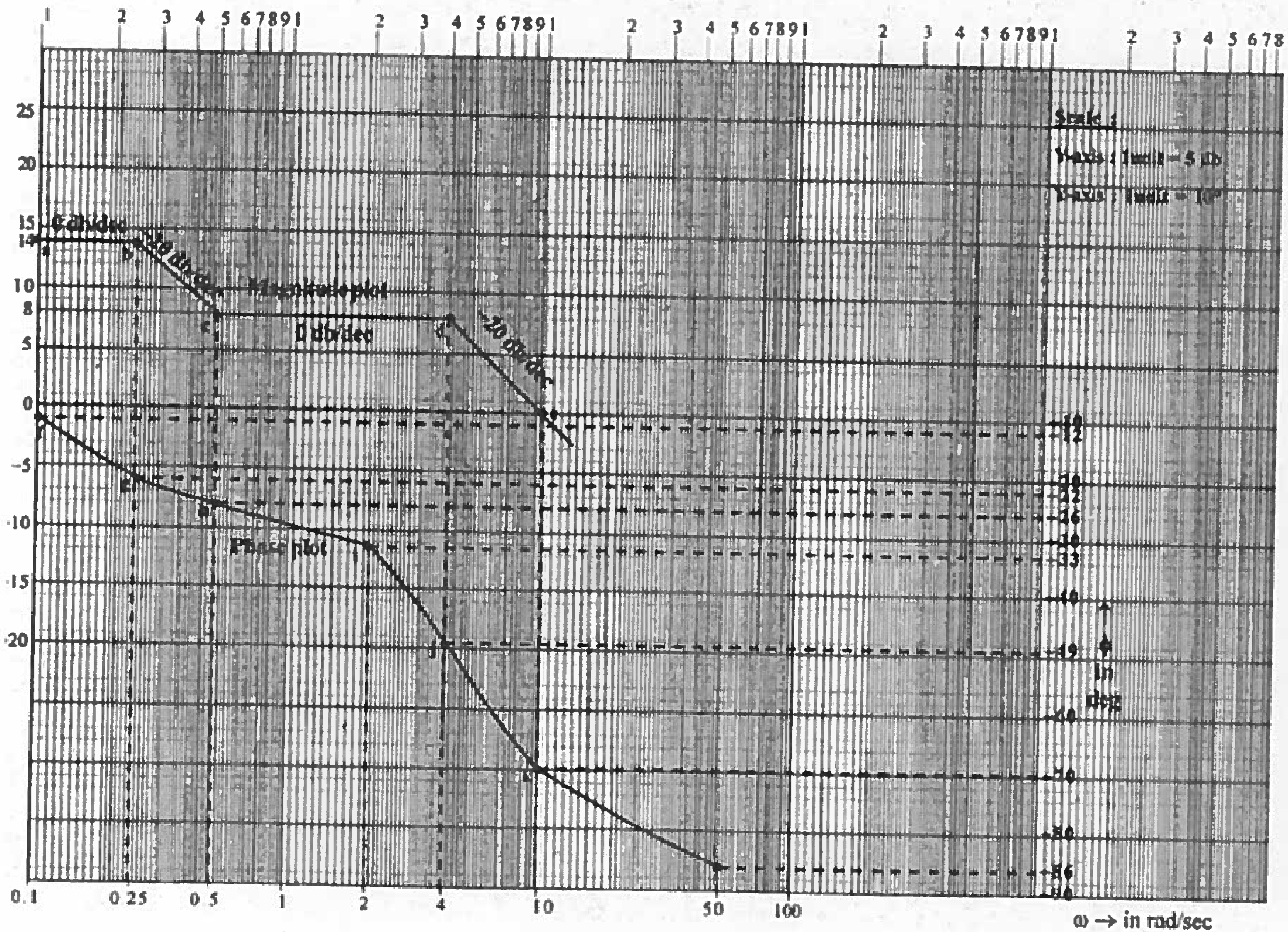
The phase angle of $G(j\omega)$, $\phi = \tan^{-1}(2\omega) - \tan^{-1}(4\omega) - \tan^{-1}(0.25\omega)$

The phase angle of $G(j\omega)$ are calculated for various values of ω and listed in the table-2.

TABLE-2

ω	$\tan^{-1} 2\omega$ deg	$\tan^{-1} 4\omega$ deg	$\tan^{-1} 0.25\omega$ deg	$\phi = \angle G(j\omega)$	Points in phase plot
0.1	11.3	21.8	1.43	$-11.93 \approx -12$	f
0.25	26.56	45.0	3.5	$-21.94 \approx -22$	g
0.5	45.0	63.43	7.1	$-25.53 \approx -26$	h
2	75.96	82.87	26.56	$-33.47 \approx -33$	i
4	82.87	86.42	45.0	$-48.55 \approx -49$	j
10	87.13	88.56	68.19	$-69.62 \approx -70$	k
50	89.42	89.71	85.42	$-85.71 \approx -96$	l

On the same semilog graph sheet choose a scale of 1unit = 10° on y-axis on the right side of the semilog graph sheet mark the calculated phase angle on the graph sheet, join the points by a smooth curve. The magnitude and phase plots are shown in fig 3.6.1.



0 db/dec
-20 db/dec
Phase plot
in deg
ω → in rad/sec

13. Consider a unity feedback having an open loop transfer function $G(s) = \frac{k}{s(1+0.2s)(1+0.05s)}$. Sketch the polar and determine the value of k so that (i) Gain margin is 18 db (ii) phase margin is 60°.

SOLUTION

Given that, $G(s) = \frac{k}{s(1+0.2s)(1+0.05s)}$. The polar plot is sketched by taking $k=1$.

$\therefore$ put $k=1$ and $s=j\omega$ in $G(s)$. ~~$G(j\omega)$ =~~

$$\therefore G(j\omega) = \frac{1}{j\omega(1+j0.2\omega)(1+j0.05\omega)}$$

The corner frequencies are $\omega_{c1} = 1/0.2 = 5$ rad/sec and $\omega_{c2} = 1/0.05 = 20$ rad/sec. The magnitude and phase angle of $G(j\omega)$ are calculated for various frequencies and tabulated in table-1.

$$G(j\omega) = \frac{1}{j\omega(1+j0.2\omega)(1+j0.05\omega)}$$

$$= \frac{1}{\omega\angle 90^\circ \sqrt{1+(0.2\omega)^2}\angle \tan^{-1}0.2\omega \sqrt{1+(0.05\omega)^2}\angle \tan^{-1}0.05\omega}$$

$$= \frac{1}{\omega\sqrt{1+(0.2\omega)^2}\sqrt{1+(0.05\omega)^2}} \angle\left(-90^\circ - \tan^{-1}0.2\omega - \tan^{-1}0.05\omega\right)$$

$$\therefore |G(j\omega)| = \frac{1}{\omega\sqrt{1+(0.2\omega)^2}\sqrt{1+(0.05\omega)^2}} \text{ and } \angle G(j\omega) = -90^\circ - \tan^{-1}0.2\omega - \tan^{-1}0.05\omega$$

TABLE-1: Magnitude and phase of $G(j\omega)$ at various frequencies

ω rad/sec	0.6	0.8	1	2	3	4
$\lvert G(j\omega)\rvert$	1.65	1.65	1.0	0.5	0.3	0.2
$\angle G(j\omega)$ deg	-98	-101	-104	-117.5	-129.4	-140

ω rad/sec	5	6	7	9	10	11	14
$\lvert G(j\omega)\rvert$	0.14	0.1	0.07	0.05	0.04	0.03	0.02
$\angle G(\omega)$ deg	-149	-157	-164	-176	-180	-184	-195

TABLE-2: Real and imaginary parts of $G(j\omega)$ at various frequencies

ω rad/sec	0.6	0.8	1	2	3	4
$G_R(j\omega)$	-0.23	-0.23	-0.24	-0.23	-0.19	-0.15
$G_I(j\omega)$	-1.63	-1.21	-0.97	-0.44	-0.23	-0.13

ω rad/sec	5	6	7	9	10	11	14
$G_R(j\omega)$	-0.120	-0.092	-0.067	-0.050	-0.04	-0.030	-0.019
$G_I(j\omega)$	-0.072	-0.039	-0.019	-0.0034	0	0.002	0.005

In the polar plot shown in fig 3.11.1 and 3.11.2 there are two plots, marked as curve-I and curve-II. These two plots are sketched with different scales to clearly determine the gain margin and phase margin.

From the polar plot, with $K=1$

Gain margin, $K_n = 1/0.04 = 25$

Gain margin in db $= 20 \log 25 = 28$ db

phase margin $\gamma = 76^\circ$

<u>case (i)</u>

with $K=1$, let $G(j\omega)$ cut the -180° axis at point B and gain corresponding to the point be G_n. From the polar plot $G_n = 0.04$. The gain margin of 28 db with $K=1$ has to be reduced to 18 db and so K has to be increased to a value greater than one.

Let G_A be the gain at -180° for a gain margin of 18 db.

Now, $20 \log \frac{1}{G_A} = 18 \Rightarrow \log \frac{1}{G_A} = \frac{18}{20} \Rightarrow \frac{1}{G_A} = 10^{18/20}$

$\therefore G_A = \frac{1}{10^{18/20}} = 0.125$

The value of K is given by, $K = \frac{G_A}{G_B} = \frac{0.125}{0.04} = 3.125$

<u>case (ii)</u>

with $K=1$, the phase margin is 76°. This has to be reduced to 60°. Hence gain has to be increased.

Let ϕ_{c2} be the phase of $G(j\omega)$ for a phase margin of 60°.

$\therefore\ 60^\circ = 180^\circ + \phi_{gc2}$

$\phi_{c2} = 60^\circ - 180^\circ = -120^\circ$

In the polar plot the -120° line cut the locus of $G(j\omega)$ at point C and cut the unity circle at point D.

Let, G_C = Magnitude of $G(j\omega)$ at point C.

G_D = Magnitude of $G(j\omega)$ at point D.

From the polar plot, $G_C = 0.425$ and $G_D = 1$.

Now, $K = \frac{G_D}{G_C} = \frac{1}{0.425} = 2.353$

RESULT

(a) when $K=1$, Gain margin, $K_G = 25$

Gain margin in db $= 28$ db

(b) when $K=1$, phase margin, $\gamma = 76^\circ$

(c) For a gain margin of 18db, $K = 3.125$

(d) For a phase margin of 60°, $K = 2.353$

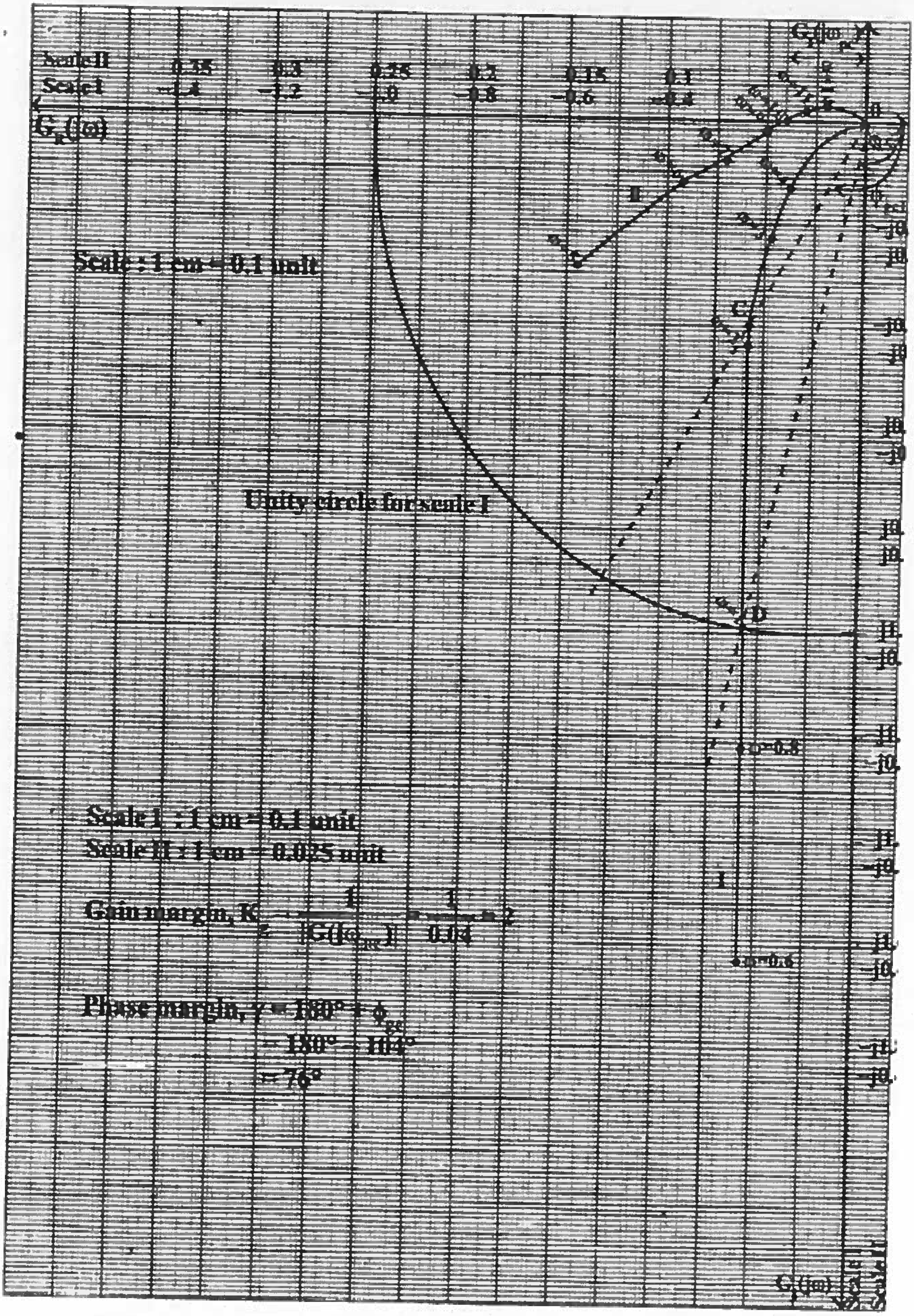


Fig 3.11.2: Polar plot of G(jω) = 1/jω (1+j0.2ω) (1+j0.05ω), *(using rectangular coordinates)*

(14)A

From Given state Equation : Matrix 'A' is

$$A = \begin{bmatrix} 1 & 2 & 3 \\ 6 & 2 & 4 \\ 7 & 8 & 1 \end{bmatrix}$$

To get state transition matrix $e^{At} = \phi(t) = L^{-1}(SI-A)^{-1}$

$$[SI-A] = \begin{bmatrix} s & 0 & 0 \\ 0 & s & 0 \\ 0 & 0 & s \end{bmatrix} - \begin{bmatrix} 1 & 2 & 3 \\ 6 & 2 & 4 \\ 7 & 8 & 1 \end{bmatrix}$$

$$= \begin{bmatrix} s-1 & -2 & -3 \\ -6 & s-2 & -4 \\ -7 & -8 & s-1 \end{bmatrix}$$

$$[SI-A]^{-1} = \frac{Adj(SI-A)}{|SI-A|}$$

$$|SI-A| = (s-1)\left((s-2)(s-1) - 32\right) + 2\left(-6s+6-28\right) - 3\left(48+7(s-2)\right)$$

$$= (s-1)(s^2-3s+2-32) + 2(-6s-22) - 3(48+7s-42)$$

$$= (s-1)(s^2-3s-30) + 2(-6s-22) - 3(7s+6)$$

$$= s^3 - 3s^2 - 30s - s^2 + 3s + 30 - 12s - 44 - 21s - 18$$

$$|SI-A| = s^3 - 4s^2 - 50s - 32$$

Adjoint $(SI-A)$ = $[\text{Cofactor Matrix}]^T$

$$\text{Adjoint } [SI-A] = \begin{bmatrix} s^2-3s-30 & -12s-12 & -21s-14 \\ 12s+44 & s^3-4s^2-16s+40 & -32s-112 \\ -21s-102 & -32s-24 & s^3-4s^2-7s+10 \end{bmatrix}$$

$$e^{At} = L^{-1}\left((SI-A)^{-1}\right) =$$

(5) $A = \begin{bmatrix} 0 & 2 & 4 \\ 1 & 5 & 2 \\ 1 & -2 & 5 \end{bmatrix}$; $B = \begin{bmatrix} 1 \\ 2 \\ 0 \end{bmatrix}$

$C = \begin{bmatrix} 1 & 1 & 0 \end{bmatrix}$

For Observability $|Q_0| \neq 0$

$$Q_0 = \begin{bmatrix} C^T & AC^T & A^2 C^T \end{bmatrix}$$

For controllability $|Q_c| \neq 0$

$$Q_c = \begin{bmatrix} B & AB & A^2B \end{bmatrix}$$

$$|Q_c| \neq 0$$

Observability: (Q_0)

$C^T = \begin{bmatrix} 1 \\ 1 \\ 0 \end{bmatrix}$ so $Q_0 = \begin{bmatrix} C^T & AC^T & A^2C^T \end{bmatrix}$

$AC^T = \begin{bmatrix} \ \end{bmatrix}$ For observable $|Q_0| \neq 0$.

$A^2C^T = \begin{bmatrix} \ \end{bmatrix}$

controllability (Q_c):

$B = \begin{bmatrix} 1 \\ 2 \\ 0 \end{bmatrix}$

$AB = \begin{bmatrix} \ \end{bmatrix}$ so controllability $Q_c = \begin{bmatrix} B & AB & A^2B \end{bmatrix}$

$|Q_c| \neq 0$

$A^2B = \begin{bmatrix} \ \end{bmatrix}$

Course Co-ordinator

HOD 29/06/22

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	Program	CSE			Academic Year	2021 - 2022
Course Code	20CS402	Test Duration	3 Hrs.	Max. Marks	70	Semester	IV
Course	Data Warehousing and Data Mining						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	List any four OLAP operations.	20CS402.1	L1
2	List any four data mining tools.	20CS402.2	L1
3	What is a decision tree?	20CS402.3	L1
4	What is meant by association rule?	20CS402.4	L1
5	Define Agglomerative Clustering.	20CS402.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Illustrate the schemas of the data warehouse.	12M	20CS405.1	L2
	OR			
7 (a)	Explain in detail about the multidimensional data model.	6M	20CS405.1	L2
7 (b)	Differentiate OLTP and OLAP with features.	6M	20CS405.1	L2
8 (a)	Show with diagrammatic illustration of the steps involved in the process of the Knowledge Discovery from Data.	6M	20CS405.2	L2
8 (b)	Discuss the major issues of Data Mining.	6M	20CS405.2	L2
	OR			
9 (a)	Explain in detail about the Data Transformation method with suitable example	6M	20CS405.2	L2
9 (b)	Elaborate the different Data Reduction techniques.	6M	20CS405.2	L2
10	Discuss in detail about Decision tree induction algorithm with an example.	12M	20CS405.3	L2
	OR			
11	Explain the Naive Bayesian Classification algorithm.	12M	20CS405.3	L2
12	Analyze the steps involved in Apriori Algorithm.	12M	20CS405.4	L2
	OR			
13 (a)	Discuss about FP-growth algorithm for the following given example{M,O,N,K,E,Y} {D,O,N,K,E,Y} {M,A,K,E} {M,U,C,K,Y} {C,O,O,K,I,E},Support= 60 %, Confidence = 80 %.	6M	20CS405.4	L2
13 (b)	Discuss about Quantitative association mining.	6M	20CS405.4	L2
14	Elaborate the various Clustering methods with an example.	12M	20CS405.5	L2
	OR			
15	Discuss in detail about K – MEANS algorithm with an example.	12M	20CS405.5	L2

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

Course Code: 20CS402

Subject:Data Mining and Data Warehouse

PART A (Short Answers)

Question 1: List four OLAP operations

Answer: **Any four from the following:**

a) Drill Down
b) Roll Up
c) Slice
d) Dice
e) Pivot

Question 2: List any four data mining tools

Answer **Any Four of the following:**

a) Weka
b) Orange
c) Oracle Data Miner
d) Knime
e) Rapid Miner
f) SAS Data Miner
i) IBM SPSS
j) Sisence
h) Apache Spark

Question 3: What is a Decision Tree?

Answer:

1) **Decision tree is a Classification Technique.**
2) A decision Tree is a Tree structure that includes Root node, Branches and Leaf Nodes.
3) The series of questions & possible answers can be organized in the form of Decision Tree, which is hierarchical Structure containing Nodes, Branches & Leaves.

4) Root Node: No Incoming edges i.e. Zero and More outgoing edges.
5) Internal Node: One Income edge and Two or More Outgoing edges.
6) Leaf (or) Terminal: One Income edge No outgoing.
7) In decision Tree each Leaf Node is assigned a Class Label.
8) The Non-Terminal Node i.e., Root & Internal node, contain attribute test conditions to separate records that have different characteristics.

Question 4: What is meant by Association Rule

Answer:

1) An association rule is an implication expression of the form X ----> Y, where X and Y are disjoint item sets i.e. $X \cap Y = 0$ (zero).
2) The relationships between co-occurring items are expressed as Association Rules.

3) { Diapers } -------------> { Coffee }

Note: Student can use an two goods or even "A" and "B".

4) The rule suggests a strong relationship exists between the sale of Diapers and Coffee.
5) Association Rules are used for business decisions.

Question 5: Define Agglomerative Clustering?

Answer:

This is a "bottom-up" approach: each observation starts in its owncluster, and pairs of Clusters are merged as one moves up thehierarchy.

a) One of the types of hierarchical Clustering
b) Start with each data point as individual clusters
c) At each step merge the closest pair of data points into clusters.
d) This requires defining a notion of cluster proximity.

PART A (Long Answers)

Question 6:

Illustrate the Schemas of the data warehouse.

Answer:

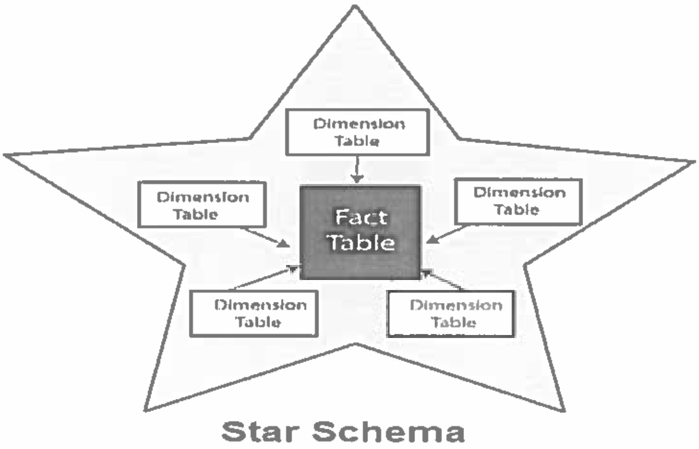
6) There are Three (3) types of Multi-dimensional Schema's they are:

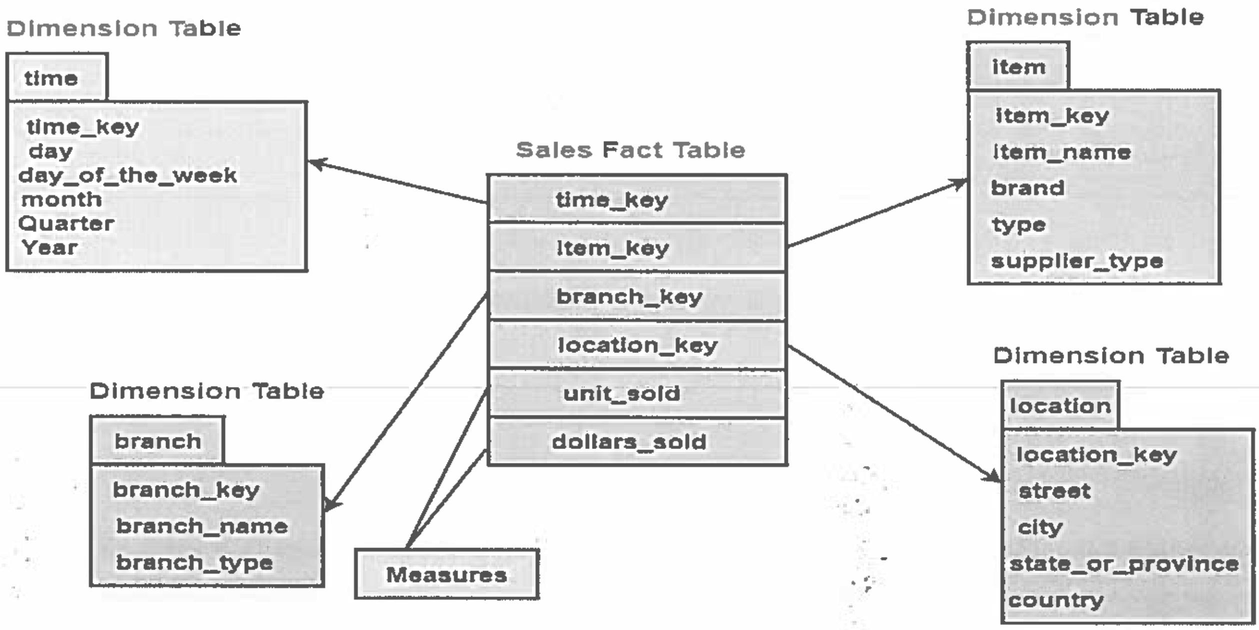
A) Star Schema B) Snowflake Schema
C) Galaxy Schema or Fact Constellation schema.

A) Star Schema :

1) Star Schema in data warehouse, is the one in which the center of the star can have one fact table and a number of associated dimension tables.
2) It is known as star schema as its ER - structure resembles a star.
3) The Star Schema data model is the simplest type of Data Warehouse schema.
4) Every dimension in a star schema is represented with the only one-dimension table.
5) The dimension table should contain the set of attributes.
6) The dimension table is joined to the fact table using a foreign key
7) The dimension table are not joined to each other
8) Fact table would contain key and measure
9) The Star schema is easy to understand and provides optimal disk usage.
10) The dimension tables are not normalized i.e. Denormalisation.

The Following Diagram explains the STAR Schema :





B) Snowflake Schema

1) "A schema is known as a snowflake if one or more dimension tables do not connect directly to the fact table but must join through other dimension tables."

2) Snowflake Schema in data warehouse is a logical arrangement of tables in a multidimensional database such that the ER diagram resembles a snowflake shape.

3) A Snowflake Schema is an extension of a Star Schema,

4) When we normalize all the dimension tables entirely, the resultant structure resembles a snowflake with the fact table in the middle.

5) Snowflake Schema adds additional dimensions.

6) The dimension tables are normalized which splits data into additional tables.

7) A snowflake schema is designed for flexible querying across more complex dimensions and relationship. It is suitable for many to many and one to many relationships between dimension levels.

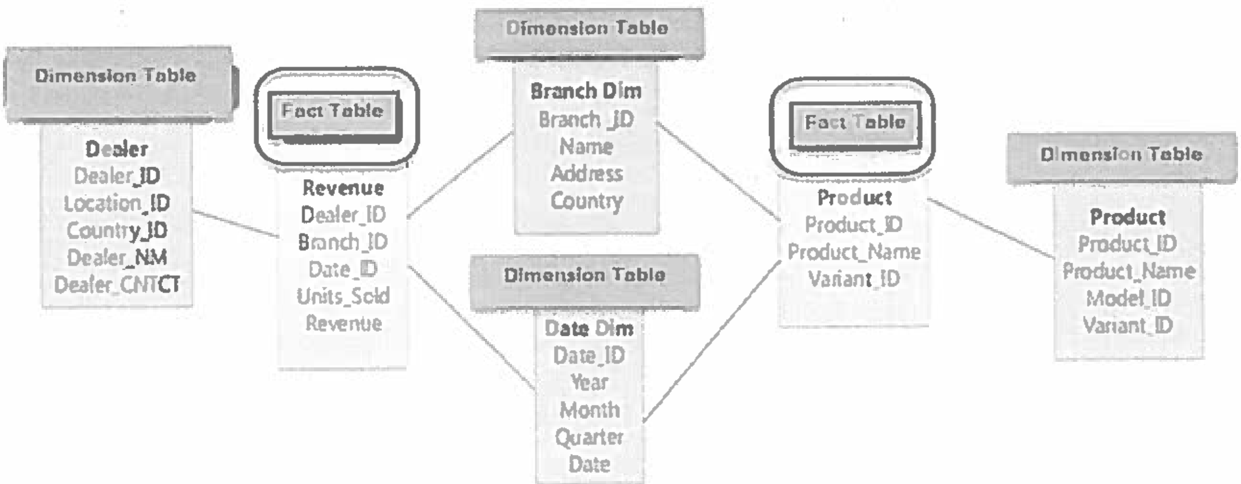
8) The following diagram shows the Snowflake Schema.

Snowflake Schema

Galaxy Schema or Fact Constellation Schema:

1) A Galaxy Schema contains two fact table that share dimension tables between them.
2) Galaxy Schema is also called Fact Constellation Schema.
3) The schema is viewed as a collection of stars hence the name Galaxy Schema.
4) A Fact constellation means two or more fact tables sharing one or more dimensions. It is also called **Galaxy schema.**
5) Fact Constellation Schema describes a logical structure of data warehouse or data mart.

6) Following diagram shows the Galaxy Schema.



5) In the above Schema we have two fact Tables "Revenue" and "Product"

Question 7:

A) Explain in detail about the multidimensional data model

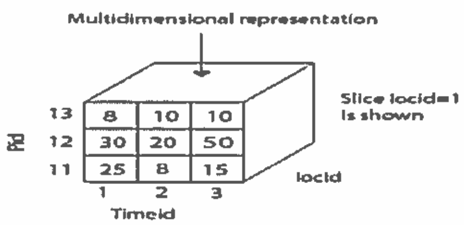
Answer

1) A multidimensional model views data in the form of a data-cube.
2) Data cube enables data to be modeled and viewed in multiple dimensions and perspectives.
3) It is defined by dimensions and facts.
4) The dimensions are the perspectives or entities concerning which an organization keeps records. Example: Time , Region , Supplier , Item etc.,
5) Facts are measures that are values. Example: Sale value, Number of respondents in a survey, Number of units sold, number of defective units etc.,
6) The Multi Dimensional Data Model allows customers to interrogate analytical Questions associated with market or business trends
7) OLAP (online analytical processing) and data warehousing uses multi dimensional Models
8) It is used to show multiple dimensions of the data to users.
9) The data is stored and represented in Dimensional and fact table (database Table).
10) Facts are numerical measures and fact tables contain measures of the related Dimensional tables.
11) A multi dimensional model can have any number of Data Cubes.
12) Multi dimensional Models are different from Relational Models.
13) Following diagram shows both two dimensional and multidimensional models and how data is arranged or organized in the computer:

The sales for Visakhapatnam are shown for the time dimension (organized in quarters) and the item dimension (classified according to the types of an item sold). The fact or measure displayed in rupee_sold (in thousands).

Tabular representation

Pid	Timeid	locid	Sales
11	1	1	25
11	2	1	8
11	3	1	15
12	1	1	30
12	2	1	20
12	3	1	50
13	1	1	8
13	2	1	10
13	3	1	10
11	1	2	35



Relational Data Model　　　Multi Dimensional Model

14) In Relational model the data is organized in Rows and colums.
15) The data is shown in the table and in 2D representation The sales of in a shop are shown in tabular representation.
16) In Multi dimensional Model the data is represented cube form.
17) The time is organized into quarters (3 months one quarter) , the dimensions are pid (product ID) ,locid (location ID) and Timeid (Time ID) .
18) The Fact Measure the sale value is also shown in the diagram.

B) Differentiate OLTP and OLAP with features

Answer:

OLTP:

1) The full form of OLTP is Online Transaction Processing.
2) OLTP or Online Transaction Processing is a type of data processing that consists of executing a number of transactions occurring concurrently
3) Examples: Online banking, shopping, order entry, or sending text messages etc.,
4) Onlinetransaction processing typically involves inserting, updating, and/or deleting small amounts of data in a data store or database.

OLAP:

1) The full form of OLAP is Online Analytical Processing
2) OLAP can be used to perform complex analytical queries.
3) OLAP is software for performing multidimensional analysis
4) OLAP is used for Business Intelligence Applications.

Comparisons of OLTP vs. OLAP –

OLTP (Online transaction processing)	OLAP (Online analytical processing)
OLTP Consists only operational current data	OLAP Consists of historical data from various Databases
OLPT is application oriented. Used for business tasks.	OLAP subject oriented. Used for Data Mining, Analytics, Decision making, etc
OLTP data is used to perform day to day fundamental operations.	OLAP data is used in planning, problem solving and decision making.
Design is Relational Data Model. Data is stored in Two dimensional Tables. Rows and Columns.	Design is Multi-Dimensional Model. Data is stored as Facts and Dimensions,
OLTP Applications are developed using programming Languages like Java, VB, C++ etc.,	OLAP based Applications used ETL tools to load data, use Analytical processing tools like Business Objects, Micro Strategy, Tableau a Data visualization tool.
OLTP Applications the data size is relatively small as the historical data is archived. For ex MB, GB	OLAP that is in Data Warehouse ,the size of data Large typically in TB, PB
OLTP is Very Fast as the queries operate on 5% of the data.	OLAP is Relatively slow as the amount of data involved is large. Queries may take hours.
In OLTP Applications Backup and recovery process is maintained every day and regularly.	In OLAP , only need backup from time to time as compared to OLTP.
OLTP data is managed by clerks, managers.	OLAP data is generally managed by CEO, Managing Director and General Manager.
In OLTP Applications both read and write operations are performed i.e. select , insert,	OLAP applications uses only read and rarely

update and Delete.	write operation. No Updates only select and insert.

Question 8:

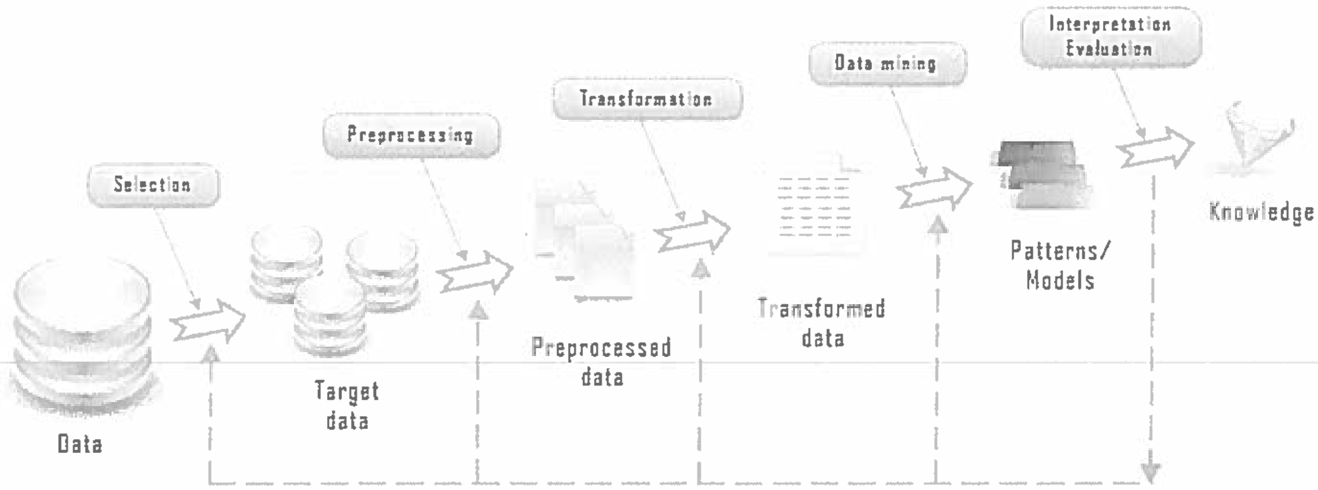
A)Show with diagrammatic illustration of steps involved in the process of the Knowledge Discovery from Data.

Answer:

Knowledge Discovery Process

1) Knowledge discovery is a process that extracts implicit, potentially useful or previously unknown information from the data.

2) The knowledge discovery process is described as follows:



Knowledge Discovery Process

3) Knowledge discovery process in the aboe diagram provides us the following details:

a) Data comes from variety of sources is integrated into a single data store called target data

b) Data then is pre-processed and transformed into the standard format.

c) The *data mining* algorithms process the data to the output in the form of patterns or rules.

d) Then those patterns and rules are interpreted to new or usefulknowledge orinformation.

4) The ultimate goal of knowledge discovery and data mining process is to find the patterns that are hidden among the huge sets of data and interpret them to useful knowledge and information.

5) Data mining is a central part of knowledge discovery process

B) Discuss the major issues of Data Mining.

Answer:

1. Data mining has several issues and need out attention.

2) Data mining issues can be solved using several types of techniques and methods.

3) As DM deals with large volumes and different types of data the DM issues are complex.

4) DM issues can be broadly classified into:

A) Mining methodology and user Interface
B) Performance issues
C) Diverse data type issues

A. Mining of methodology and user interface:

a. Mining different kinds of knowledge in Databases like Bioinformatics, streams, web.

b. Interactive mining of knowledge at multiple level of abstraction.

c. Incorporation of background knowledge.

d. DM query language and adhoc on DM.

e. Presentation and visualization of DM results.

f. Handling noisy or incomplete data.

g. Pattern evaluation and interpreting the problem and Results.

B. Performance Issues:

a. Scalability and efficiency DM algorithms.

b. Parallel, distributed and incremental mining algorithms.

c. The various DM algorithms include:

a) K-means DM algorithm b) Apriori DM algorithm c) Support vector Machine d) K-NN (nearest neighbor) e) Naive Bayes

C.Diverse data type issues:

a. Handling relational and complex types of data.
b. Mining information from heterogeneous DB's and global information systems.

c. DM issues may also include.

1) Protection of data security. 2) Integrity. 3) Privacy.

Question 9

A) Explain in detail about the data Transformation method with suitable example,

Answer:

DATA TRANSFORMATION:

1) In DT, the data are transformed of consolidated into forms appropriate for mining.

2) DT involves the following techniques:

a) Smoothening b) Aggregation c) Generalization d) Normalization

e) Attribute construction

Smoothening:

1) This works to remove noise involves techniques like

a) Binning b) Clustering c) Regression,

2) Smoothening is form of data cleansing using Extraction Transformation Loading (ETL)tools by specifying transformations to correct data inconsistencies.

Aggregation :

1) Summary or aggregation operations are applied on data

2) This is also a data reduction technique

3) In aggregation we summarize the attribute value

Example: Daily sales are aggregated into Monthly or yearly sales.

4) This technique is used for construction of data cube for analysis of data at multiple Granularity

Generalization:

3) In Generalization we replace lower level or "primitive" raw data by higher level concepts by using concept hierarchy.

Example: white papers, pen, book, pencil we replace these with term "stationary".

Or

Street to city or age to "youth", middle aged or Sr.Citizen.

Normalization:

1) An attribute is normalized by scaling its values, so that the values fall within small specified range such as 0.0 to 1.0.

2) There are many methods of normalization

a) Min-max normalization b) Z-score normalization c) Normalization by decimal scaling

3) Min-max normalization performs a linear transformation of original data.

4) Min-max normalization preserves the relationships among the original Data values.

5) It will encounter an "out-of-bounds" error if a future input case for normalization falls outside of the original data range.

Min-Max normalization:

1) Suppose min & max marks of students in a class is 12 and 98
2) We would like to map marks to the range of [0.0,1.0], min-max normalization of a student marks of 73 is transformed as follows:

3) Min_Max Normalization = ((Attribute Value - Min) / Max - Min) (1.0 - 0) + 0

((73 -12) / 98 -12) (1.0 -0) + 0 = 61/86 = 0.709

Z-Score Normalization:

1) Z-score normalization is also called as zero-mean normalization.

2) The value for an attribute "A" is normalized based on the mean and standard deviation of "A" Z-Score Normalization = (Value of an Attribute - Mean of Attribute) / Standard deviation

3) This method of normalization is useful when the actual min & max of attribute are unknown.

4) Suppose Mean & Standard Deviation of the values of the attribute income are 540 and 160 respectively and if, we have to find the Z-score normalization for a tuple with income value of 736 then we transform this value as follows:

5) Z-Score Normalization For Income 736 = (736 - 540) / 160 = 1.225

Normalization by Decimal Scaling:

1) Normalizes by moving the decimal point values of an attribute.
2) The Number of decimal points moved depends on the maximum absolute values of Attribute "A".

3) Normalized Value V^1 = $V / 10^{j1}$
Where j is the smallest integer such as max (v1) <1
4) Recorded value of Attribute "A" ranges from -986 to 917.
The maximum absolute value = 986 (therefore J=3, divided by 1000)
Normalizes to - 0.986 to 0.917
5) Normalization can change the original data a bit especially last two methods.

Attribute Reconstruction:

1) In Attribute Reconstruction new attributes are constructed from the given attributes.

2) These new attributes helps to improve accuracy & understanding of the structure of high- dimensional data.

3) From Date of birth we may have new attribute of "Year" from Full name we may have a new attribute "First Name"

4) New attributes may be added & used on height & width.

B) Elaborate the different Data Reduction Techniques

Answer:

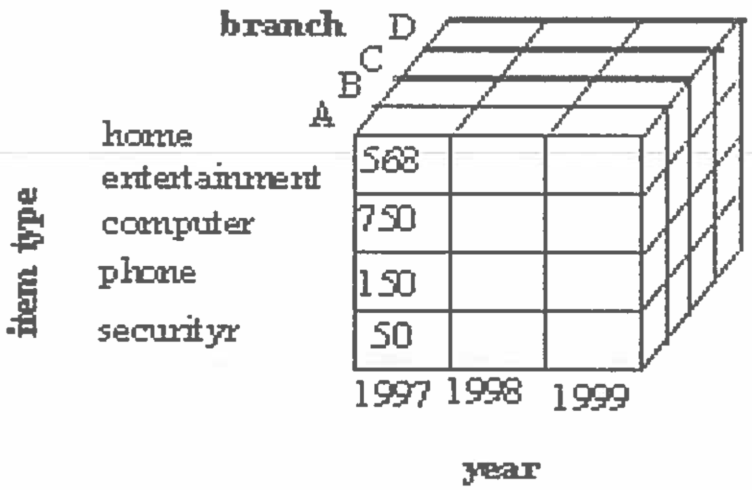
1) Data Reduction technique can be applied to obtain reduced representation of the data set that is much smaller in volume, yet closely maintains the integrity of the original data.
2) The mining of the reduced dataset should be more efficient yet produce the same (are almost same) analytical results.

3) The various strategies of data reduction include:

a) Data cube aggregation b) Attribute sub-selection c) Dimensionality reduction
d) Numerosity reduction e) Discretization and concept hierarchy generation.

Data cube aggregation:

1) Data cube store multi dimensional aggregated information.
2) Example: Annual Sales per item type for each branch.



3) Hierarchy may exist for each attribute.
4) This allows the analysis of the data at multiple level of abstraction.
5) Example: Branches can be grouped into regions etc.
6) The cube created at the lowest level of abstraction referred to as the base cuboids.
7) The cube of the highest level of abstraction is the apex cuboids.

Attribute Sub-Selection:

1) Data sets for analysis may contain hundreds of attribute, many of which maybe irrelevant to the mining task.

2) The domain experts may pick some useful attributes, leaving other irrelevant

attributes.

3) Sometimes keeping irrelevant attributes may cause confusion for DM algorithms.
4) Attribute subset selection reduces the dataset by removing irrelevant or redundant attributes.
5) The goal of attribute sub selection is to bind a minimum set of attributes.
6) Mining reduced set of attributes has additional benefits.
7) Reduces number of patterns and helps to make patterns easier to understand.
8) Question arises how we can find a "good" subset of original attributes.
9) Attributes are selected based on statistical significance.
10) Various techniques used for Attribute Sub-Selection are:

a) Stepwise forward selection b) Stepwise backward elimination c) Combination of forward and backward elimination d) Decision tree induction

Dimensionality Reduction (DMR)

1) In DMR, the encoding or transformation is applied so as to obtain a reduced or "compressed" representation of original data.

2) If, the original data can be reconstructed from the compressed data without any loss of information, the data reduction is called Lossless.
3) If, instead we can reconstruct an approximation of original data, the reduction is called as Lossy.
4) The two popular and effective methods of Lossy Dimensionality reduction are

a) Wavelet Transforms b) Principle component Analysis

5) The discrete Wavelet Transform is a linear signal processing Technique that, when applied to data vector X, transforms it to numerically different vector $X^{1,}$, of Wavelet coefficient.

6) The two vectors are of same length.
7) Wavelet transforms have many real world applications:
   a) Compression of Finger Prints b) Computer Vision
   c) Analysis of Time Series d) Data Cleansing.
8) The principle component Analysis (PCA) search for k, n-dimensional orthogonal vectors that can be best used to represent data.

Nuemorosity Reduction:

1) Numerosity reduction can reduce the data volume by choosing alternative, "smaller" forms of data representation.

2) This technique can be:
   a) Parametric b) Non-parametric
3) For parametric methods , a model is used to estimate data so that typically

only the data parameter need to be stored, instead of original data.

4) Ex: Outliers may be stored example: Log linear Models

5) Non parametric methods used for storing reduced representation of data include histogram, clustering and sampling.

Question 10:

Discuss in detail about Decision tree induction algorithm with an example.

Answer:

Decision Tree Induction:

1) Decision tree classifier is simple and widely used classification Technique.
2) A decision Tree is a Tree structure that includes Root node, Branches and Leaf Nodes.
3) The series of questions & possible answers can be organized in the form of Decision Tree, which is hierarchical Structure containing Nodes, Branches & Leaves.
4) Root Node: No Incoming edges i.e. Zero and More outgoing edges.
5) Internal Node: One Income edge and Two or More Outgoing edges.
6) Leaf (or) Terminal: One Income edge No outgoing.
7) In decision Tree each Leaf Node is assigned a Class Label.
8) The Non-Terminal Node i.e., Root & Internal node, contain attribute test conditions to separate records that have different characteristics.

Building a Decision Tree:

1) In principle, there are exponentially many decision trees that can beconstructed from a given set of attributes..

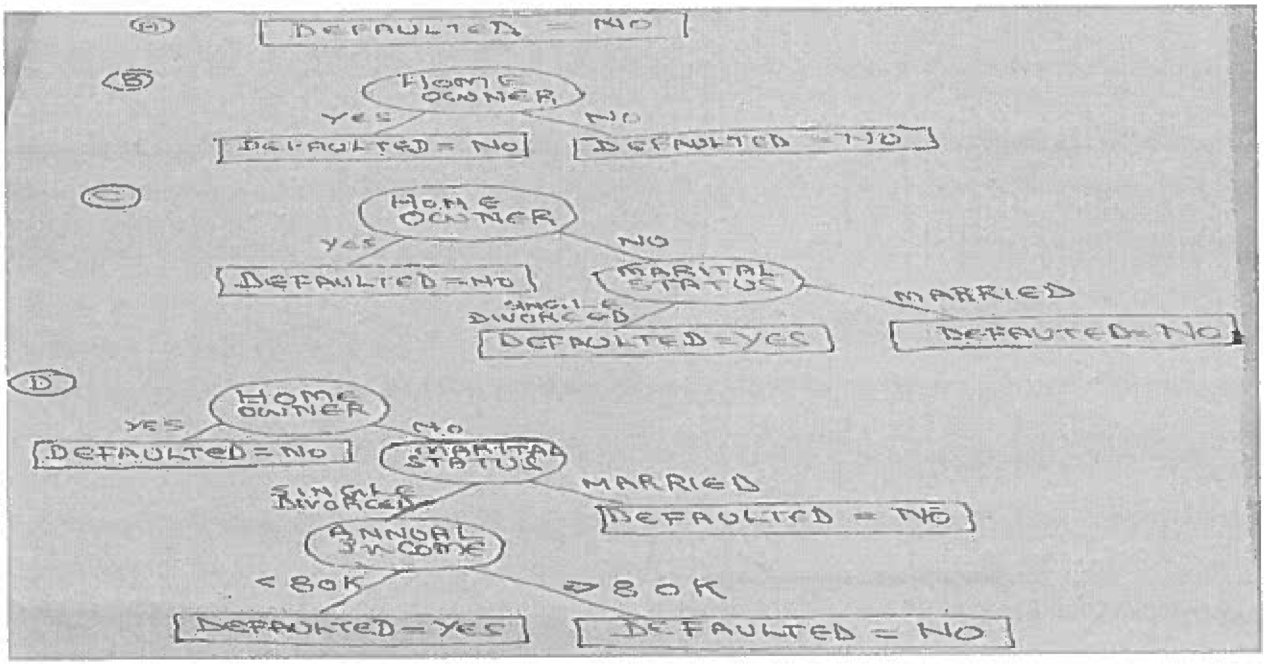
2) Hunt's Algorithm is basis of many existing Decision tree induction algorithms such as ID3, C4.5 and CART.

3) For example: Consider the problem of predicting whether a loan applicant will repay the loan of obligations or to become delinquent, subsequently defaulting the loan.

4) The training set for this problem is:

Transaction ID	Home Owner	Marital Status	Annual Income	Defaulted Borrow
1	Yes	Single	125 K	No
2	No	Married	100 K	No
3	No	Divorced	75	Yes
4	No	Married	60 K	No
5	No	Single	65 K	Yes
6	No	Divorced	90 K	No

9) In the above Training Set"Home owner" is Binary Attribute ,"Marital Status" is Categorical Attribute, "Annual income" is Continuous Attribute, "Defaulted Borrower "is CLASS LABEL.



EXPLANATION:

1) Initially the tree for classification problems contain a single node with class label i.e., Defaulted=NO.

2) Therefore tree needs to be refined.

3) The records are subsequently divided into smallest subsets based on the outcome of Home owner test condition.

4) Tree resulting by applying recursively steps of HUNT algorithm shown indiagrams.

Algorithm for Decision Tree Induction

1) Decision tree algorithm falls under the category of supervised learning.
2) Decision tree algorithm can be used to solve both classification and regression problems.
3) A skeleton of decision tree algorithm called Tree Growth.
4) The algorithm has 4 functions:

- **The createNode():**
  a) This function expands the decision tree by creating a new node.
  b) A node in the Decision tree has either a test condition denoted as **"Node.test_cond"** or class label denoted as **"Node.child"**.
- **The find_leaf_split():**
  a) This function determines which attribute should be selected as the test condition for splitting of training records.
  b) The choice depends upon measurement of impurity either by using entropy or gini index.
- **The classify():**
  a) This function determines how class labels to be assigned to a leaf

node.

b) For each leaf node t, let p (i / t) denote the fraction of training record from class i associated with node t.

c) In most cases, the leaf node is assigned to the class that has the majority of number of training records.

Leaf.node = argmax (p (i/ t))

d) Theargmaxfurnction returns the argument i that maizimizestheexpression p(i/t)

e) The faction p(i/t) can be used to estimate the probability that a record assigned to the leaf node t belongs to class i.

- **The stopping_cond():**

a) This Function is used to terminate the tree growing process by testing whether all the records have either the same class label or the same attribute values.

b) Another way to terminate the recurssive function is to test whether the number of records have fallen below some minumum threshold.

Question 11:

Explain the Naïve Bayesian Classification Algorithm

Answer:

Note: As student's have Bayesian Classification in syllabus not Naïve Bayesian Classification Algorithm. Please give marks if, they write the following Bayesian Classification Algorithm.

BAYE'S THEOREM:

1) In many applications the relationship between attribute set and the class variable is non-deterministic.

2) The class label of a test record cannot be predicted with certainty even though it's attribute set is identical to some of the training example.

3) This situation may arise because of noisy data or the presence of certain confounding factors that affect classification, but are not included in the analysis.

4) Example: - Heart disease based on diet and workout frequency------ but still some other factors like heredity , excess smoking and alcohol may also effect the prediction or results.

5) So, this may introduce uncertainties to the learning algorithm.

6) BAYE'S theorem, is a statistical principle for combining prior knowledge of classes with new evidence gathered from data.

7) Let X and Y be a pair of random variables.

8) Their joint probability of p(X=x, Y=y), refers to the probability that variable 'X' will take on the values of 'x' and variable 'Y' will take on the value 'y'.

9) A conditional probability is the probability that a random variable will take on a

particular value, given that the outcome for another random variable is known.

10) For example, the conditional probability p(Y=y / X=x) refers to the probability that the variable 'Y' will taken on the value 'y' , given that the variable 'X' is observed to have the value 'x'.

11) The joint and conditional probabilities for 'X' and 'Y' are

$$P(x, y) = p(y/x) * p(x) = p(x/y) * p(y)$$

12) Rearranging these expressions, it leads to following known as baye's theorem.

$$p(y/x) = (p(x/y) * p(y)) / p(x)$$

13) Consider a football game between two rival teams: team0 and team1

14) Suppose team0 wins 65% times and team1 wins the remaining matches.

15) Among the games won by team0 only 30% of them came from playing on team1 's football field.

16) On the other hand, 75% of the victories of the team1 are obtained while playing at home.

17) If team1 is to host the next match between the two teams, which team will most likely emerge as the winner.

18) The baye's theorem can be used to solve the prediction problem.

19) For notational convenience, let 'X' be the random variable that represents the team hosting the match and 'Y' be random variable that represents the winner of the match.

20) Both 'X' and 'Y' can take on values from the set {0,1}

21) We can summarize the information given in the problem as follows:

22) a) Probability team0 wins is p(Y=0) =0.65
b) Probability team1 wins is p(y=1) = 1- p(y=0) =0.35
c) Probability team1 hosted the match and it won is p (x=1 /y=1) = 0.75
d) Probability team1 hosted the match won by team0 is p(x=1 / y=0) = 0.3

23) Our objective is to compute p(y=1/x=1) which is conditional probability that team1 wins the next match it will be hosting and compare it against p(y=0 /x=1) using BAYES THEOREM.

24)

$$P(Y=1/X=1) = (P(X=1 / Y=1) * P(Y=1)) / P(X=1)$$

$$=> (P(X=1/Y=1) * P(Y=1)) / P(X=1,Y=1) + P(X=1,Y=0)$$

$$=> (P(x=1/y=1) * p(y=1)) / p(x=1/y=1) * p(y=1) + p(x=1/y=0) * p(y=0)$$

$$=> 0.75 * 0.35 / 0.75 * 0.35 + 0.3 * 0.65$$

$$P(Y=1/X=1) => 0.5738$$

25) Law of total probability was applied in the 2nd line.

26) Further more

$$= 1 - 0.5738$$

$$= 0.4262$$

$0.5738 > 0.4262$

27 Since $p(y=1/x=1) > p(y=0/x=1)$ team1 has a better chance than team0 of winning the next match.

Question 12:

Analyze the steps involved in Apriori Algorithm

Answer:

THE APRIORI PRINCIPLE:

1) Apriori is an algorithm for frequent item set mining and association rule learning over relational databases.
2) Name of the algorithm is Apriori because it uses prior knowledge of frequent itemset properties.
3) The property of Apriori Algorithm is :

All subsets of a frequent itemset must be frequent(Aprioripropertry).
If an itemset is infrequent, all its supersets will be infrequent.

4) Apriori proceeds by identifying the frequent individual items in the database and extending them to larger and larger item sets as long as those item sets appear sufficiently often in the database.

5) The Apriori principle is if, an itemset is frequent, then all of its subsets must also be frequent.
6) Suppose {c,d,e} is frequent itemset.
7) Clearly, any transaction that contains {c,d,e} must also contain its subsets, {cd}, {ce} , {de}, {c} , {d} and {e}.
8) As a result if, {c,d,e} is frequent then all subsets of {c,d,e} must also be frequent.

The following is the Flow of Apriori Algorithm

Data Set → Candidate Item Sets → Frequent Item Sets→ Association Rules

Frequent Itemset Generation in the Apriori Algorithm:

1) Apriori is the first Association Rule mining Algorithm that pioneered the use of support based pruning to control the exponential growth of candidate item sets

2) A high level illustration of the Frequent itemset generation part of theApriori Algorithm is for the Market Basket Transactions.

3) The following Data set is considered to illustrate the Aprirori Algorithm:

Example is

TID	ITEMS
1	{bread, milk}
2	{bread, diapers, coffee, eggs}
3	{milk, diapers, coffee, cola}
4	{bread, milk, diapers, coffee}
5	{bread, milk, diapers, cola}

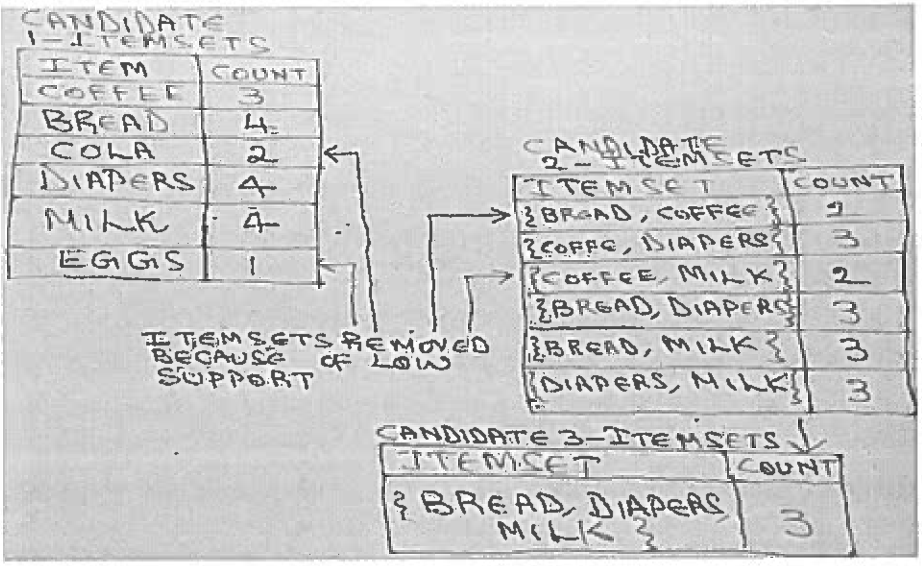
4) The Assumption of support threshold is 60 % which is equivalent to a minimum support of count equal to 3 out of 5 transactions.

5) Initially, every itemset is considered as a Candidate 1-itemset.
6) After counting their support, the candidate itemsets{ cola} and {eggs} are discarded because they appear in fewer than Three.

7) As with the assumption of the Apriori Principle, only need to keep candidate 3-itemsets whose subsets are frequent.

8) The only candidate that has this property is {Bread,Diapers,Milk}

9) Illustration of Frequent itemset generation using the Apriori Algorithm:



10) With the Apriori Principle, we onlly need to keep candidate 3-itemsets whose subsets are frequent.
11) The only candidate that has the property is {Bread,Diapers,Milk}
12) The effectiveness of the apriori pruning strategy can be shown by counting the number of candidateitemsets generated.
13) A brute force strategy enumerating all the itemsets (upto size 3) as candidates will produce:

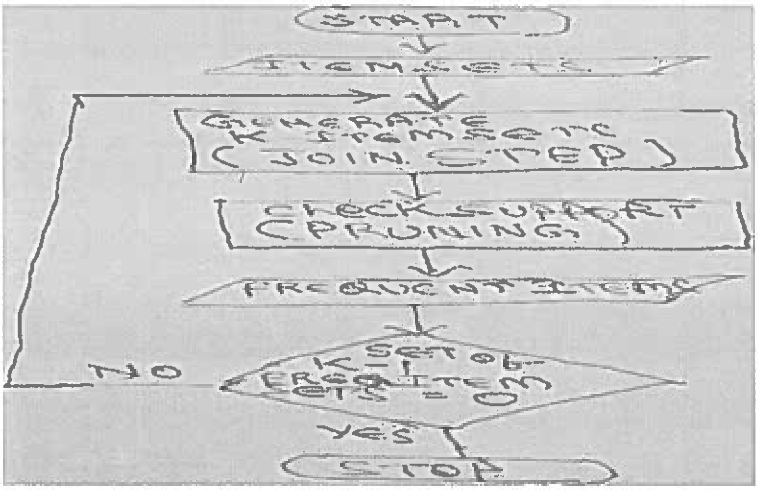
$$\left(\frac{6}{1}\right) + \left(\frac{6}{2}\right) + \left(\frac{6}{3}\right) = 6 + 15 + 20 = 41$$

14) **With the Apriori principle, the number decreases to:**

$$\binom{6}{1} + \binom{4}{2} + 1 = 6+6+1 = 13$$

15) **This represents 68 % reduction in the number of candidate itemsets even in this simple example.**

Flow Chart for Frequent Itemset Generation in Apriori Algorithm



Apriori Algorithm Pseudo CodeFor Frequent Itemset Generation:

C: **candidate itemset of size 'K'**

L: **Frequent itemset of size 'K'**

Join Step: Ck is generated by joining Lk-1with itself

Prune Step: Any (k-1)-itemset that is not frequent cannot be a subset of a frequent k-itemset

Pseudo-code:
Ck: Candidate itemset of size k

Lk: frequent itemset of size k

L1 = {frequent items};

for (k = 1; Lk !=0; k++) **do begin**
  Ck+1 = candidates generated from Lk;
  for each transaction 't' in database
do
  increment the count of all candidates in Ck+1 that are contained in transaction 't'
  Lk+1 = candidates in Ck+1 with min_support
  end
return = Lk;

Question 13:

A) Discuss about FP growth Algorithm for the following given Example

{M,O,N,K,E,Y} {D,O,N,K,E,Y} {M,A,K,E} {M,U,C,K,Y} {C,O,O,K,I,E},Support= 60 %, Confidence = 80 %.

Answer:

FP-GROWTH ALGORITHM:

1) FP growth algorithm is an improvement of apriori algorithm.
2) FP growth algorithm is used for finding frequent itemset in a transaction database without candidate generation.
3) FP growth represents frequent items in frequent pattern trees or FP-tree.
4) FP-Growth Algorithm is an alternative that takes a radically different approach to discover frequent itemsets.
5) This encodes the data set using compact data structure called FP-Tree and extracts frequent itemsets directly from the Tree-Structure.
6) The FP-Tree is a compressed representation of the input data.
7) It is constructed by reading the data set one transaction at a time and mapping each transaction into Path in the FP-Tree.

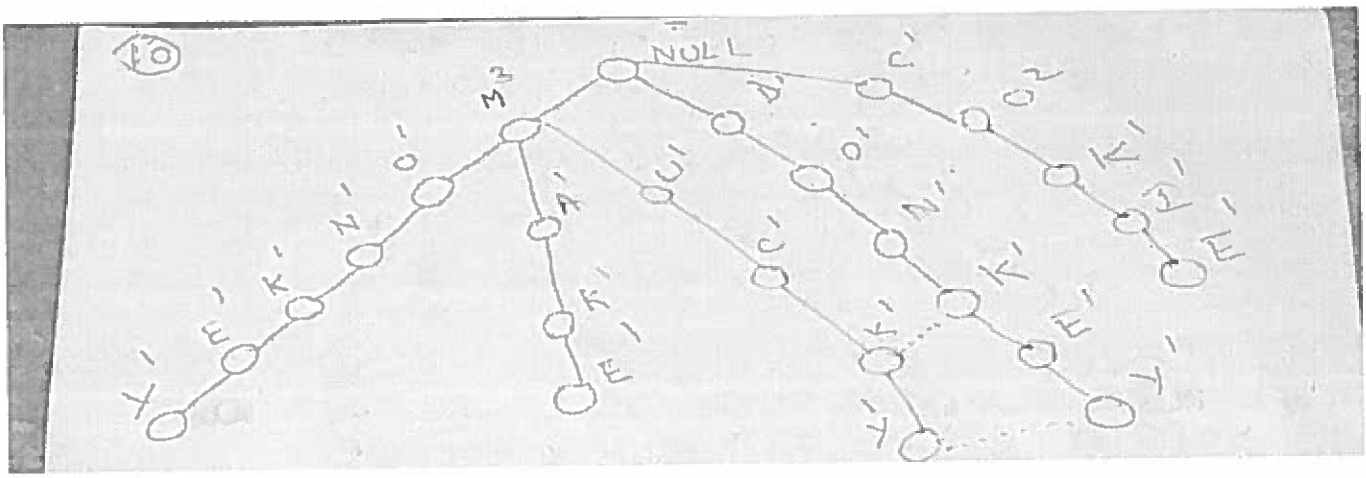
8) The following is the Flow of Apriori Algorithm

Data Set → FP-Tree→ Frequent Item Sets→ Association Rules

9) The following is the data set that contains 5 transactions used for the FP-Tree Construction.

{M,O,N,K,E,Y} {D,O,N,K,E,Y} {M,A,K,E} {M,U,C,K,Y} {C,O,O,K,I,E}

The FP Tree is :



(11) The FP-Tree is generated by drawing path for each Transaction. Some paths have common prefix of first Item.

(12) To Find Frequent Items we have to consider Support.

Support = 60%

ie. 5 (Transactions) $\times \frac{60}{100} = 3$

(13) To generate Association Rules we consider Confidence.

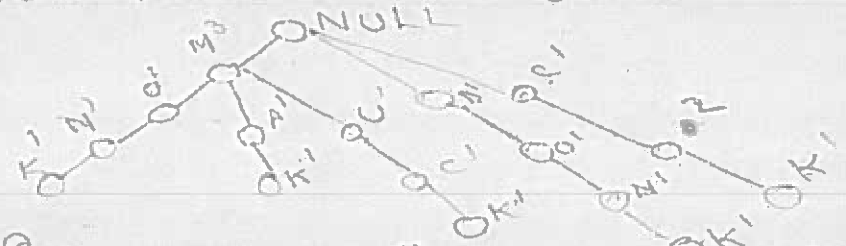
Confidence = 80%

ie = 5 (Transactions) $\times \frac{80}{100}$

Confidence = 5

14) FP-Growth finds all the frequent item sets ending with a particular suffix by employing a divide and conquer strategy to split the problem into smaller sub-problems.

(15) Let us take an item "K". The suffix path is as follows.



(16) As Support is "3" and "K" repeated in 4 transactions "K" is a Frequent item.

(17) Now we can in similar way have to find "K" item with other items or element ie NK, OK, MK, CK, AK. These become "2" Item set

(18) For Three item set of "K" the combinations can be.

MOK, NOK, MAK, DOK, DNK etc,

(19) we have to check their support to determine whether such combination of Items are Frequent

(20) In same way Confidence can be used to generate Association Rules.

B) Discuss about Quantitative association mining.

Answer:

1) Quantitative association rules refer to a special type of association rules in the form of $X \rightarrow Y$, with X and Y consisting of a set of numerical and/or categorical attributes.

2) Different from general association rules where both the left-hand and the right-hand sides of the rule should be categorical (nominal or discrete) attributes, at least one attribute of the quantitative association rule (left or right) must involve a numerical attribute.

3) Examples of this type of association rule can be categorized into the following two classes, depending on whether the rules are measured by the frequency of the supporting data records or by some distributional features of some numerical attributes.

4) It can be used to improve decision making in a wide variety of applications such as:

   a) market basket analysis,

   b) medical diagnosis,

   c) bio-medical literature,

   d) protein sequences,

   e) census data,

   f) logistic regression,

   g) fraud detection in web,

   h) CRM of credit card business

5) Quantitative association rules are multidimensional association rules in which the numeric attributes are dynamically discretized

6) Student can give any example of Association Rule but it should contain at least one numeric attribute.

Question 14: Elaborate the various Clustering methods with an example.

Answer:

1) A cluster is a group of objects that lie under the same class, or in other words, objects with similar properties are grouped in one cluster, and dissimilar objects are collected in another cluster.
2) Clustering is the process of classifying objects into a number of groups wherein each group, objects are very similar to each other than those objects in other groups.
3) Simply, segmenting groups with similar properties/behaviour and assign them into clusters.
4) Clustering is unsupervised learning.
5) Clustering is one of the important technique of Data Mining
6) There are various clustering methods they are:

A) Partitioning-based Clustering

B) Hierarchical Clustering

C) Density based Clustering

D) Grid-based Clustering

A) **Partitioning-based Clustering**

1) Partitioning objects into k number of clusters where each partition makes/represents one cluster,
2) these clusters hold certain properties such as each cluster should consist of at least one data object and each data object should be classified to exactly one cluster.

3) Example : K-means clustering,

B) **Hierarchical-based Clustering**

1) Depending upon the hierarchy, these clustering methods create a cluster having a tree-type structure where each newly formed clusters are made using priorly formed clusters.
2) Hierarchical clustering is categorized into two categories:

a) **Agglomerative (bottom-up approach)**

b) **Divisive (top-down approach).**

C) Density-based Clustering

1) Density based clustering recognize clusters of dense regions that possess some similarity and are distinct from low dense regions of the space.
2) These methods have sufficient accuracy and the high ability to combine two clusters.
3) Example is : DBSCAN (Density-based Spatial Clustering of Applications with Noise)

D) **Grid-based Clustering**

1) Grid-based Clustering follows a grid-like structure, i.e, data space is organized into a finite number of cells to design a grid-structure.
2) Various clustering operations are conducted on such grids (i.e quantized space) and are quickly responsive and do not rely upon the quantity of data objects. Its examples are;
3) Examples of Grid-based Clustering are:

a) STING (Statistical Information Grid),
b) Wave cluster,
c) CLIQUE (Clustering In Quest)

Question 15: Discuss in detail about K – MEANS algorithm with an example.

Answer:

1) This is a prototype based, partitioned clustering technique that attempts to find a user specified number of clusters (K) , which are represented by their centroids.
2) K-Means defines a prototype in terms of a centroid, which is usually the mean of a group of points.
3) K-Means is oldest and widely used algorithm.
4) K-Means is a popular, unsupervised Machine Learning Algorithm.
5) Unsupervised algorithms make inferences from using only input vectors without

known or labeled outcomes.

6) The approach K-Means follows to solve the problem is called expectation Maximization.

THE BASIC K-MEANS ALGORITHM:

1) K-Means clustering is an unsupervised Iterative clustering technique.
2) It partitions the given data into "K" predefined distinct clusters.
3) Each data point belong to one cluster with the nearest mean.
4 First we choose 'K' intial centroids where 'K' is user specified parameter.
5 'K' describes the number of clusters required.
6) Each point is then assigned to the closet centroid.
7) Each collection of points assigned to a centroid is a cluster.
8) The centroid of each cluster is then updated based on the points assigned to the cluster.
9) We repeat assignment and update steps until no point changes clusters or until centroid remanins same.
10) K-Means Algorithm:

Step 1: Choose Number of clusters "K"

Step 2:

a) Randomly select any "K" data points as cluster centers.
b) Select cluster centers in such a way that they are as farther as possible from each other.

Step 3:

a) Calculate the distance between each data point and each cluster center.
b) Distance may be calculated either by using given distance function or by using Euclidean distance formula.

Step 4:

a) Assign each data point to some cluster.
b) A data point is assigned to that cluster whose center is nearest to that data point.

Step 5:

a) Re-compute the center of newly formed clusters.
b) The center of a cluster is computed by taking mean of all the data points contained in that cluster.

Step 6:

Keep repeating procedure from step 3 to step 5 until any of the following stopping criteria is met.

a) Center of newly formed cluster do not change.
b) Data point remain present in the same cluster.
c) Maximum number of iterations reached.

11) Example: Suppose we want to group the visitors to a website using just their Age.
(One Dimensional Space) as follows:

12) The number of visitors (n) = 5 and their respective ages are
15 , 16 , 22 , 28 and 30

13) Initial clusters random centroid of average
K = 2 and C1 = 16 and C2 = 22

14) Distance D1 = $|X_i - C_1|$
Distance D2 = $|X_i - C_2|$

ITERATION 1 :

X	C1	C2	D1	D2	NEAREST CLUSTER	NEW CENTROID
15	16	22	1	7	1	
16	16	22	0	6	1	15.5
22	16	22	6	0	2	
28	16	22	12	6	2	26.6
30	16	22	14	8	2	

15) The New Cluster Centroids are:

C1 = 15.5 C2 = 26.6

ITERATION 2 :

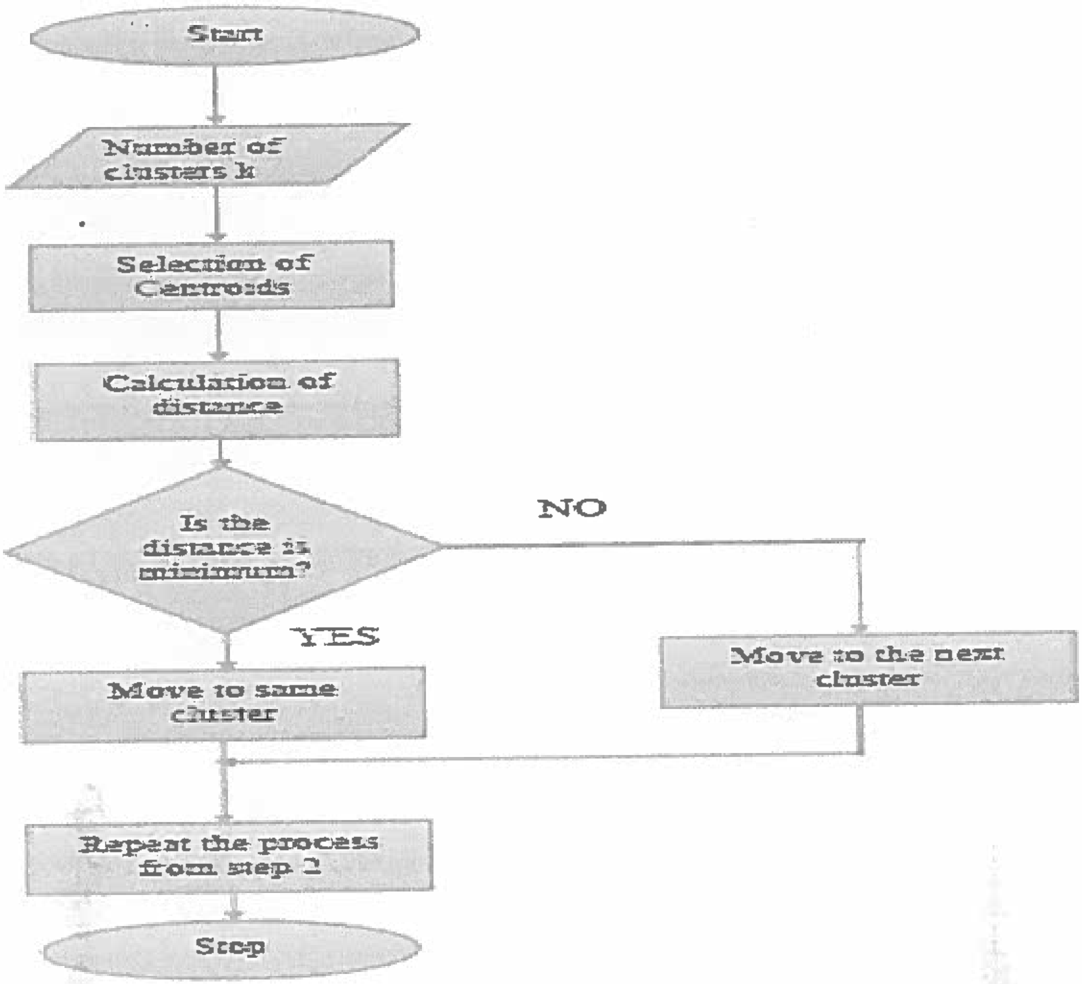
X	C1	C2	D1	D2	NEAREST CLUSTER	NEW CENTROID
15	15.5	26.6	0.5	11.6	1	
16	15.5	26.6	0.5	10.6	1	15.5
22	15.5	26.6	6.5	4.6	2	
28	15.5	26.6	12.5	1.4	2	26.6
30	155	26.6	14.5	3.4	2	

16) No change between iteration 1 and iteration 2 has been noticed.

17) By clustering 2 groups have been identified

a) 15-16 b) 22-30

<u>Flow Chart For K-Means:</u>

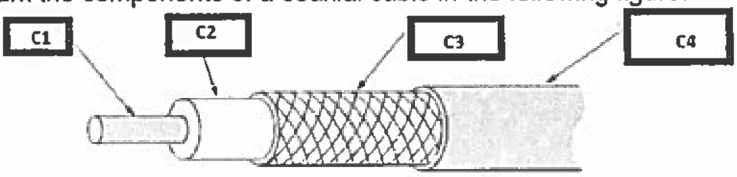

Start
Number of clusters k
Selection of Centroids
Calculation of distance
Is the distance is minimum?
NO
YES
Move to same cluster
Move to the next cluster
Repeat the process from step 2
Stop

NSRIT

Semester End Regular Examination, June, 2022

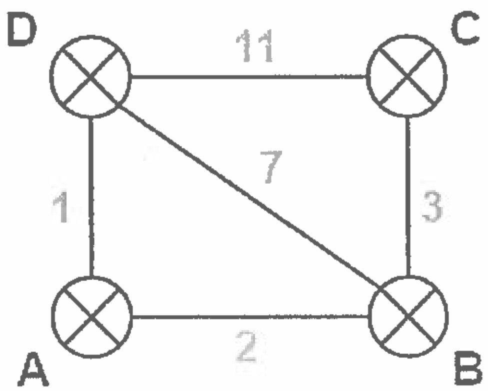
Degree	B. Tech. (U. G.)	Program	CSM/CSD			Academic Year	2021 - 2022
Course Code	20CS502	Test Duration	3 Hrs.	Max. Marks	70	Semester	IV
Course	COMPUTER NETWORKS						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Mark the components of a coaxial cable in the following figure.  	20CS502.1	L1
2	Identify any two error correcting code mechanism in data link layer.	20CS502.2	L1
3	How IPv6 differs from IPv4?	20CS502.3	L4
4	What is multiplexing?	20CS502.4	L2
5	List the typical element of email user agents.	20CS502.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

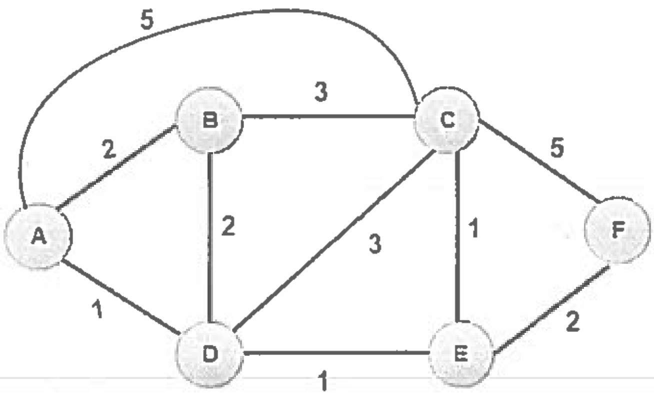
No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Draw the ISO-OSI Architecture and outline the functions performed by each layer.	8M	20CS502.1	L2
6 (b)	Discuss some of the design issues of the computer network layers in general.	4M	20CS502.1	L1
	OR			
7 (a)	Explain any four network topologies with a neat sketch	8M	20CS502.1	L2
7 (b)	Interpret the terms in brief about unicasting, broadcasting and multicasting with respect to network hardware.	4M	20CS502.1	L1
8 (a)	Explain with a neat sketch about the sliding window protocol.	8M	20CS502.2	L2
8 (b)	For CRC polynomial, each of the following, explain whether the errors during message transmission will be detected by the receiver: (a) There was a single-bit error. (b) There were two isolated bit errors.	4M	20CS502.2	L3
	OR			
9 (a)	Explain about any one of the Multiple access protocols.	8M	20CS502.2	L2
9 (b)	Sixteen-bit messages are transmitted using a Hamming code. How many check bits are needed to ensure that the receiver can detect and correct single-bit errors? Show the bit pattern transmitted for the message 1101001100110101. Assume that even parity is used in the Hamming code.	4M	20CS502.2	L3
10 (a)	Tabulate the shortest path for all nodes for the following network using distance vector routing.	7M	20CS502.3	L3



10 (b)	Discuss the Four issues must be addressed to ensure quality of service in network layer	5M	20CS502.3	L2

OR

11 (a)	Tabulate the shortest path for all nodes for the following network using Links state routing.	7M	20CS502.3	L3



11 (b)	Briefly discuss about the approaches to Congestion Control with its timeline.	5M	20CS502.3	L2
12 (a)	Explain about TCP Addressing with respect to transport layer	6M	20CS502.4	L2
12 (b)	Explain about TCP Congestion control with respect to transport layer.	6M	20CS502.4	L2
	OR			
13 (a)	Explain the UDP header and its components with a neat sketch	6M	20CS502.4	L2
13 (b)	Explain the TCP Segment header and its components with a neat sketch.	6M	20CS502.4	L2
14 (a)	Write a short note on DNS Namespace.	6M	20CS502.5	L1
14 (b)	Explain about Domain Resource Record with its format	6M	20CS502.5	L2
	OR			
15 (a)	Explain about the architecture of email with a neat sketch.	6M	20CS502.5	L1
15 (b)	Explain about any one of mail transport protocols with its purpose.	6M	20CS502.5	L2

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

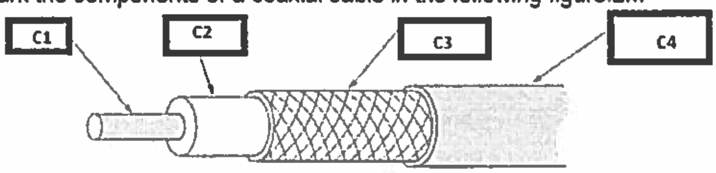
ANSWER KEY AND SCHEME OF EVALUATION

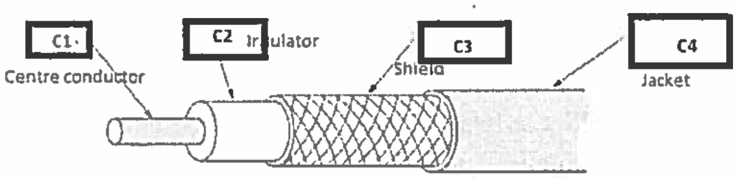
Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	**Program**	CSM/CSD			**Academic Year**	2021 - 2022
Course Code	20CS502	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	IV
Course	COMPUTER NETWORKS						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Mark the components of a coaxial cable in the following figure.2M	20CS502.1	L1





No.	Questions	Learning Outcome (s)	DoK
2	**Identify any two error correcting code mechanism in data link layer.**	20CS502.2	L1

Scheme :
Any of the following two mechanisms with uses -2M
1. Parity Check
2. Checksum
3.CRC

Parity Check

The parity check is done by adding an extra bit, called parity bit to the data to make a number of 1s either even in case of even parity or odd in case of odd parity.

While creating a frame, the sender counts the number of 1s in it and adds the parity bit in the following way

- In case of even parity: If a number of 1s is even then parity bit value is 0. If the number of 1s is odd then parity bit value is 1.

- In case of odd parity: If a number of 1s is odd then parity bit value is 0. If a number of 1s is even then parity bit value is 1.

The parity check is suitable for single bit error detection only.

Checksum

In this error detection scheme, the following procedure is applied

- Data is divided into fixed sized frames or segments.
- The sender adds the segments using 1's complement arithmetic to get the sum. It then complements the sum to get the checksum and sends it along with the data frames.
- The receiver adds the incoming segments along with the checksum using 1's complement arithmetic to get the sum and then complements it.
- If the result is zero, the received frames are accepted; otherwise, they are discarded.

Cyclic Redundancy Check (CRC)

Cyclic Redundancy Check (CRC) involves binary division of the data bits being sent by a predetermined divisor agreed upon by the communicating system. The divisor is generated using polynomials.

- Here, the sender performs binary division of the data segment by the divisor. It then appends the remainder called CRC bits to the end of the data segment. This makes the resulting data unit exactly divisible by the divisor.
- The receiver divides the incoming data unit by the divisor. If there is no remainder, the data unit is assumed to be correct and is accepted. Otherwise, it is understood that the data is corrupted and is therefore rejected.

3 **How IPv6 differs from IPv4?** 20CS502.3 L4

Scheme : any two Main differences out of the following differences- 2M

IPv4 has a 32-bit address length	IPv6 has a 128-bit address length
It Supports Manual and DHCP address configuration	It supports Auto and renumbering address configuration
In IPv4 end to end, connection integrity is Unachievable	In IPv6 end to end, connection integrity is Achievable
It can generate 4.29×109 address space	Address space of IPv6 is quite large it can produce 3.4×1038 address space
The Security feature is dependent on application	IPSEC is an inbuilt security feature in the IPv6 protocol
Address representation of IPv4 is in decimal	Address Representation of IPv6 is in hexadecimal

4 **What is multiplexing?** 20CS502.4 L2

Scheme : Definition of multiplexing – 2M

Multiplexing is the process of combining multiple signals into one signal, over a shared medium. If analog signals are multiplexed, it is Analog Multiplexing and if digital signals are multiplexed, that process is Digital Multiplexing.

Multiplexing and Demultiplexing

Types of Multiplexers

There are mainly two types of multiplexers, namely analog and digital. They are further divided into FDM, WDM, and TDM.

5	**List the typical element of email user agents.** **Scheme : two elements – 2M** **Electronic Mail** (e-mail) is one of most widely used services of Internet. This service allows an Internet user to send a message in formatted manner (mail) to the other Internet user in any part of world. The user agent is normally a program which is used to send and receive mail. Sometimes, it is called as mail reader. It accepts variety of commands for composing, receiving and replying to messages as well as for manipulation of the mailboxes.	20CS502.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	**Draw the ISO-OSI Architecture and outline the functions performed by each layer.** Scheme : Daigram-2 M ,Functions-6M	8M	20CS502.1	L2

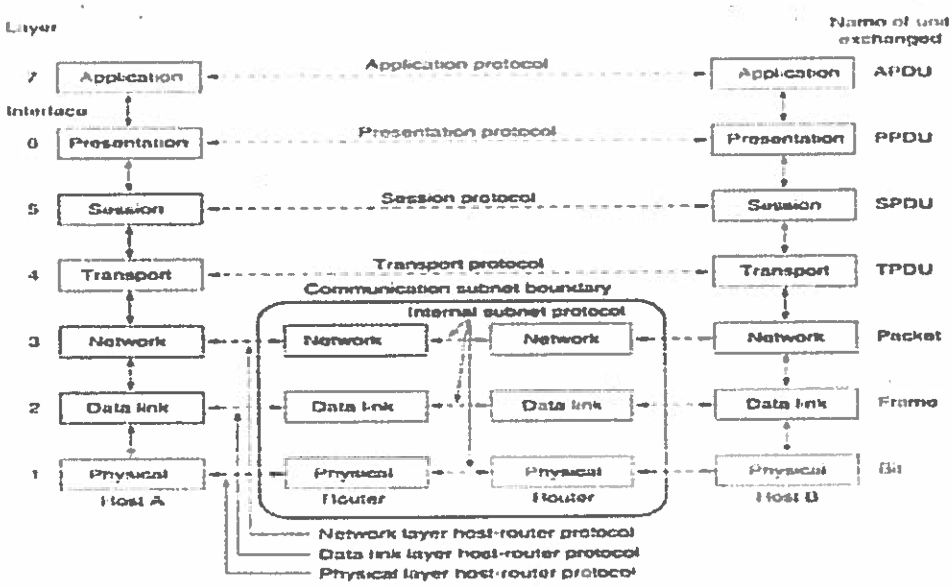


Fig.4: The OSI reference model

Following are the functions performed by each layer of the OSI model.

OSI Model Layer 1: The Physical Layer

1. Physical Layer is the lowest layer of the OSI Model.
2. It activates, maintains and deactivates the physical connection.
3. It is responsible for transmission and reception of the unstructured raw data over network.
4. Voltages and data rates needed for transmission is defined in the physical layer.
5. It converts the digital/analog bits into electrical signal or optical signals.
6. Data encoding is also done in this layer.

OSI Model Layer 2: Data Link Layer

1. Data link layer synchronizes the information which is to be transmitted over the physical layer.
2. The main function of this layer is to make sure data transfer is error free from one node to another, over the physical layer.
3. Transmitting and receiving data frames sequentially is managed by this layer.
4. This layer sends and expects acknowledgements for frames received and sent respectively. Resending of non-acknowledgement received frames is also handled by this layer.
5. This layer establishes a logical layer between two nodes and also manages the Frame traffic control over the network. It signals the transmitting node to stop, when the frame buffers are full.

OSI Model Layer 3: The Network Layer

1. Network Layer routes the signal through different channels from one node to other.
2. It acts as a network controller. It manages the Subnet traffic.
3. It decides by which route data should take.
4. It divides the outgoing messages into packets and assembles the incoming packets into messages for higher levels.

OSI Model Layer 4: Transport Layer

1. Transport Layer decides if data transmission should be on parallel path or single path.

2. Functions such as Multiplexing, Segmenting or Splitting on the data are done by this layer

3. It receives messages from the Session layer above it, convert the message into smaller units and passes it on to the Network layer.

4. Transport layer can be very complex, depending upon the network requirements.

Transport layer breaks the message (data) into small units so that they are handled more efficiently by the network layer.

OSI Model Layer 5: The Session Layer

1. Session Layer manages and synchronize the conversation between two different applications.

2. Transfer of data from source to destination session layer streams of data are marked and are resynchronized properly, so that the ends of the messages are not cut prematurely and data loss is avoided.

OSI Model Layer 6: The Presentation Layer

1. Presentation Layer takes care that the data is sent in such a way that the receiver will understand the information (data) and will be able to use the data.

2. While receiving the data, presentation layer transforms the data to be ready for the application layer.

3. Languages(syntax) can be different of the two communicating systems. Under this condition presentation layer plays a role of translator.

4. It perfroms Data compression, Data encryption, Data conversion etc.

OSI Model Layer 7: Application Layer

1. Application Layer is the topmost layer.

2. Transferring of files disturbing the results to the user is also done in this layer. Mail services, directory services, network resource etc are services provided by application layer.

3. This layer mainly holds application programs to act upon the received and to be sent data.

6 (b)	**Discuss some of the design issues of the computer network layers in general.** **Scheme : Any 2 design issues – 4M**	4M	20CS502.1	L1

A number of design issues exist for the layer to layer approach of computer networks. Some of the main design issues are as follows –

Reliability

Network channels and components may be unreliable, resulting in loss of bits while data transfer. So, an important design issue is to make sure that the information transferred is not distorted.

Scalability

Networks are continuously evolving. The sizes are continually increasing leading to congestion. Also, when new technologies are applied to the added components, it may lead to incompatibility issues. Hence, the design should be done so that the networks are scalable and can accommodate such additions and alterations.

Addressing

At a particular time, innumerable messages are being transferred between large numbers of computers. So, a naming or addressing system should exist so that each layer can identify the sender and receivers of each message.

Error Control

Unreliable channels introduce a number of errors in the data streams that are communicated. So, the layers need to agree upon common error detection and error correction methods so as to protect data packets while they are transferred.

Flow Control

If the rate at which data is produced by the sender is higher than the rate at which data is received by the receiver, there are chances of overflowing the receiver. So, a proper flow control mechanism needs to be implemented.

Resource Allocation

Computer networks provide services in the form of network resources to the end users. The main design issue is to allocate and deallocate resources to processes. The allocation/deallocation should occur so that minimal interference among the hosts occurs and there is optimal usage of the resources.

Statistical Multiplexing

It is not feasible to allocate a dedicated path for each message while it is being transferred from the source to the destination. So, the data channel needs to be multiplexed, so as to allocate a fraction of the bandwidth or time to each host.

Routing

There may be multiple paths from the source to the destination. Routing involves choosing an optimal path among all possible paths, in terms of cost and time. There are several routing algorithms that are used in network systems.

Security

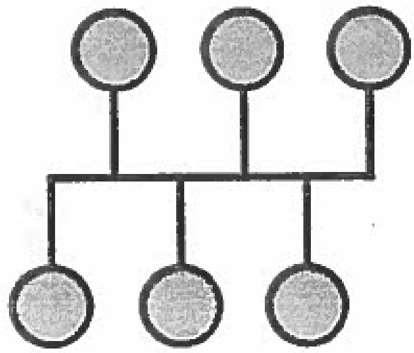
A major factor of data communication is to defend it against threats like eavesdropping and surreptitious alteration of messages. So, there should be adequate mechanisms to prevent unauthorized access to data through authentication and cryptography.

OR

Explain any four network topologies with a neat sketch 8M 20CS502.1 L2

7 (a) Scheme :
4 diagrams -4 M
Explanation -4M

Bus

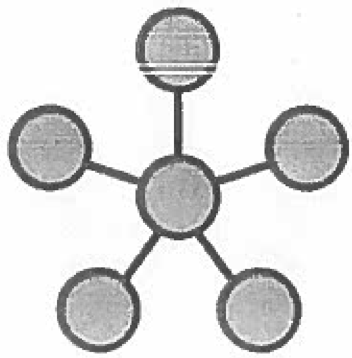


Bus network topology

In local area networks using bus topology, each node is connected by interface connectors to a single central cable. This is the 'bus', also referred to as the backbone, or trunk – all data transmission between nodes in the network is transmitted over this common transmission medium and is able to be received by all nodes in the network simultaneously.[1]

A signal containing the address of the intended receiving machine travels from a source machine in both directions to all machines connected to the bus until it finds the intended recipient, which then accepts the data. If the machine address does not match the intended address for the data, the data portion of the signal is ignored. Since the bus topology consists of only one wire it is less expensive to implement than other topologies, but the savings are offset by the higher cost of managing the network. Additionally, since the network is dependent on the single cable, it can be the single point of failure of the network. In this topology data being transferred may be accessed by any node.

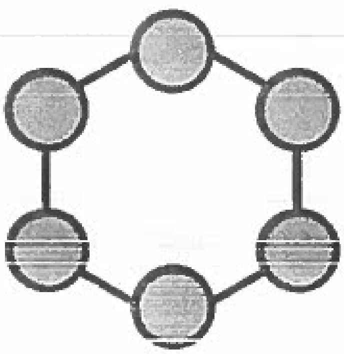
Star



In star topology, every peripheral node (computer workstation or any other peripheral) is connected to a central node called a hub or switch. The hub is the server and the peripherals are the clients. The network does not necessarily have to resemble a star to be classified as a star network, but all of the peripheral nodes on the network must be connected to one central hub. All traffic that traverses the network passes through the central hub, which acts as a signal repeater.

The star topology is considered the easiest topology to design and implement. One advantage of the star topology is the simplicity of adding additional nodes. The primary disadvantage of the star topology is that the hub represents a single point of failure. Also, since all peripheral communication must flow through the central hub, the aggregate central bandwidth forms a network bottleneck for large clusters.

Ring



Ring network topology

A ring topology is a daisy chain in a closed loop. Data travels around the ring in one direction. When one node sends data to another, the data passes through each intermediate node on the ring until it reaches its destination. The intermediate nodes repeat (retransmit) the data to keep the signal strong.[5] Every node is a peer; there is no hierarchical relationship of clients and servers. If one node is unable to retransmit data, it severs communication between the nodes before and after it in the bus.

Advantages:

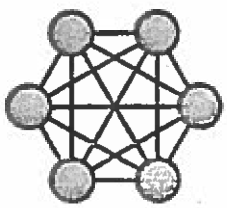
- When the load on the network increases, its performance is better than bus topology.
- There is no need of network server to control the connectivity between workstations.

Disadvantages:

- Aggregate network bandwidth is bottlenecked by the weakest link between two nodes.

Mesh

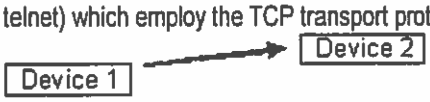
The value of fully meshed networks is proportional to the exponent of the number of subscribers, assuming that communicating groups of any two endpoints, up to and including all the endpoints, is approximated by Reed's Law.



7 (b)	Interpret the terms in brief about unicasting, broadcasting and multicasting with respect to network hardware. Explanation-3M ,Diagrams-3M	6M	20CS502.1	L1

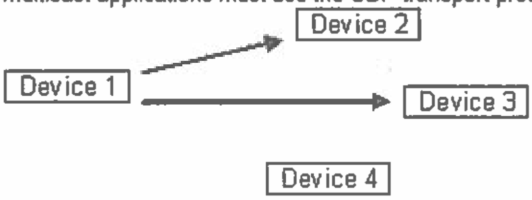
A Unicast communication is from one device on the network to another device on the network.

Unicast transmission, in which a packet is sent from a single source to a specified destination, is still the predominant form of transmission on LANs and within the Internet. All LANs (e.g. Ethernet) and IP networks support the unicast transfer mode, and most users are familiar with the standard unicast applications (e.g. http, smtp, ftp and telnet) which employ the TCP transport protocol.



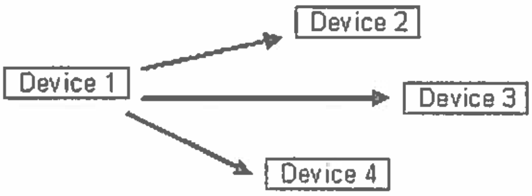
A MultiCasting communication is from one device on the network to many, but not all, devices on the network.

Multicasting is the networking technique of delivering the same packet simultaneously to a group of clients. IP multicast provides dynamic many-to-many connectivity between a set of senders (at least 1) and a group of receivers. The format of IP multicast packets is identical to that of unicast packets and is distinguished only by the use of a special class of destination address (class D IPv4 address) which denotes a specific multicast group. Since TCP supports only the unicast mode, multicast applications must use the UDP transport protocol.



A Broadcast communication is from one device on the network to all devices on the network.

Broadcast transmission is supported on most LANs (e.g. Ethernet), and may be used to send the same message to all computers on the LAN (e.g. the address resolution protocol (arp) uses this to send an address resolution query to all computers on a LAN, and this is used to communicate with an IPv4 DHC server). Network layer protocols (such as IPv4) also support a form of broadcast that allows the same packet to be sent to every system in a logical network (in IPv4 this consists of the IP network ID and an all 1's host number).



8 (a)	Explain with a neat sketch about the sliding window protocol **Scheme :.** **Daigram-2M** **,Explanation-4M,** **Example-2M**	8M	20CS502.2	L2

Sliding window protocols are data link layer protocols for reliable and sequential delivery of data frames. The sliding window is also used in Transmission Control Protocol.

In this protocol, multiple frames can be sent by a sender at a time before receiving an acknowledgment from the receiver. The term sliding window refers to the imaginary boxes to hold frames. Sliding window method is also known as windowing.

Working Principle

In these protocols, the sender has a buffer called the sending window and the receiver has buffer called the receiving window.

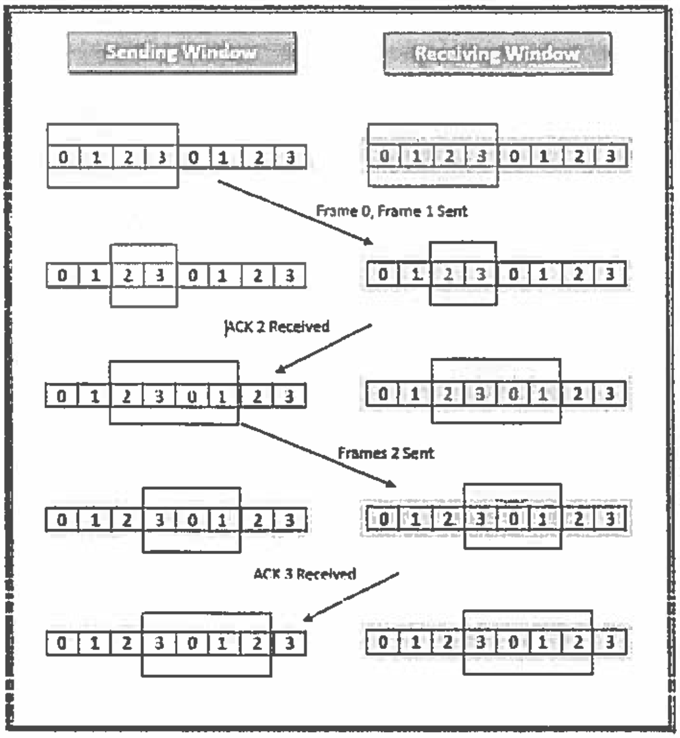
The size of the sending window determines the sequence number of the outbound frames. If the sequence number of the frames is an n-bit field, then the range of sequence numbers that can be assigned is 0 to 2^n-1. Consequently, the size of the sending window is 2^n-1. Thus in order to accommodate a sending window size of 2^n-1, a n-bit sequence number is chosen.

The sequence numbers are numbered as modulo-n. For example, if the sending window size is 4, then the sequence numbers will be 0, 1, 2, 3, 0, 1, 2, 3, 0, 1, and so on. The number of bits in the sequence number is 2 to generate the binary sequence 00, 01, 10, 11.

The size of the receiving window is the maximum number of frames that the receiver can accept at a time. It determines the maximum number of frames that the sender can send before receiving acknowledgment.

Example

Suppose that we have sender window and receiver window each of size 4. So the sequence numbering of both the windows will be 0,1,2,3,0,1,2 and so on. The following diagram shows the positions of the windows after sending the frames and receiving acknowledgments.



8 (b)	For CRC polynomial, each of the following, explain whether the errors during message transmission will be detected by the receiver: **Scheme :** **(a) There was a single-bit error.2M** **(b) There were two isolated bit errors.2M**	4M	20CS502.2	L3

a) Yes, Cyclic Redundancy Check (CRC) checksum detects all single bit errors during the data transmission of 1024 bit.• Single bit errors is detected by CRC method.It produces 100 percentage of error detection.

b) Cyclic Redundancy Check (CRC) checksum detects all double bit errors during the data transmission of 1024 bit. That is, two isolated error bits for long messages.

• Double bit errors are detected by CRC method.It produces 100 percentage of error detection and creates the generating polynomial for at least three 1s.

OR

9 (a)	Explain about any one of the Multiple access protocols. **Scheme :** **Explanation-6M,** **Daigrams-2M.**	8M	20CS502.2	L2

CSMA was developed to improve performance and minimize the chance of collision. Each station is required to test the state of the medium before transmission. In other words, CSMA is based on the principle "sense before transmit." CSMA can reduce the risk of collision, but it cannot remove it.

CSMA primary access mode:

1-Persistent: The 1-persistent method is very easy and quick. In this method, after the station finds the line empty, it immediately transmits its frame. The chances of the collision are very high in this method because two or more stations immediately transmit their frames as soon as the line is found empty.

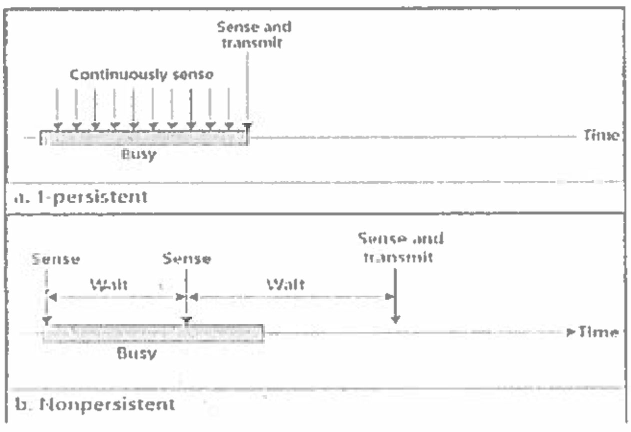
Non-Persistent: In the non-persistent method, if found the line is empty, it transmits the frames immediately. If the line isn't clear, it waits for a random period and detects the line again. This approach decreases the risk of a collision.

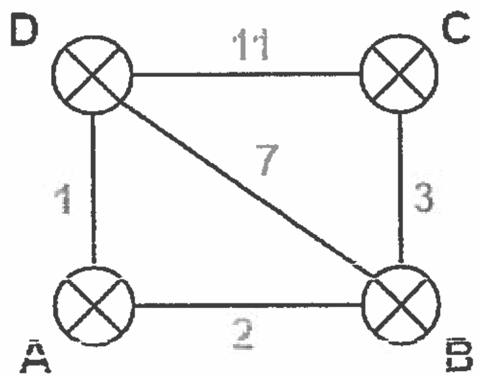
P-Persistent: This method is a combination of 1-Persistent and Non-Persistent advantages. The p-persistent approach decreases the risk of collision and increases performance.

In the P-Persistent approach, the following steps follow after the station finds the line-empty:

1. With probability (p), the station transmits its frame.
2. With probability (q = 1 ? p), the station waits for the starting of the next time slot and re-test the line.
3. If the line is empty, it goes to step 1.
4. If the line is not empty, it behaves as though a collision has happened, and it uses the back-off process.

In later, CSMA is divided into two parallel methods: CSMA/CD and CSMA/CA. When a collision is detected, CSMA/CD tells the station what to do, and CSMA/CA attempts to stop a collision.



9 (b)	Sixteen-bit messages are transmitted using a Hamming code. How many check bits are needed to ensure that the receiver can detect and correct single-bit errors? Show the bit pattern transmitted for the message 1101001100110101. **Scheme :** **Assume that even parity is used in the Hamming code.Explanation:4M** 011110110011001110101	4M	20CS502.2	L3
10 (a)	Tabulate the shortest path for all nodes for the following network using distance vector routing.  Distance Vector is a simple routing protocol which takes routing decisions on the number of hops between source and destination. A route with less number of hops is considered as the best route. Every router advertises its set best routes to other routers. Ultimately, all routers build up their network topology based on the advertisements of their peer routers, For example Routing Information Protocol (RIP).	7M	20CS502.3	L3

Final Routing table:

Destination	Distance	Next Hop
A	0	A
B	2	B
C	5	B
D	1	D

10 (b) Discuss the Four issues must be addressed to ensure quality of service in network layer. 5M 20CS502.3 L2

Scheme :
Explanation-1 M,
Issues-4M

Quality of service (QoS) is the use of mechanisms or technologies that work on a network to control traffic and ensure the performance of critical applications with limited network capacity. It enables organizations to adjust their overall network traffic by prioritizing specific high-performance applications.

QoS is typically applied to networks that carry traffic for resource-intensive systems. Common services for which it is required include internet protocol television (IPTV), online gaming, streaming media, videoconferencing, video on demand (VOD), and Voice over IP (VoIP).

Using QoS in networking, organizations have the ability to optimize the performance of multiple applications on their network and gain visibility into the bit rate, delay, jitter, and packet rate of their network. This ensures they can engineer the traffic on their network and change the way that packets are routed to the internet or other networks to avoid transmission delay. This also ensures that the organization achieves the expected service quality for applications and delivers expected user experiences.

These are the Four issues must be addressed to ensure quality of service in network layer

1. Bandwidth: The speed of a link. QoS can tell a router how to use bandwidth. For example, assigning a certain amount of bandwidth to different queues for different traffic types.
2. Delay: The time it takes for a packet to go from its source to its end destination. This can often be affected by queuing delay, which occurs during times of congestion and a packet

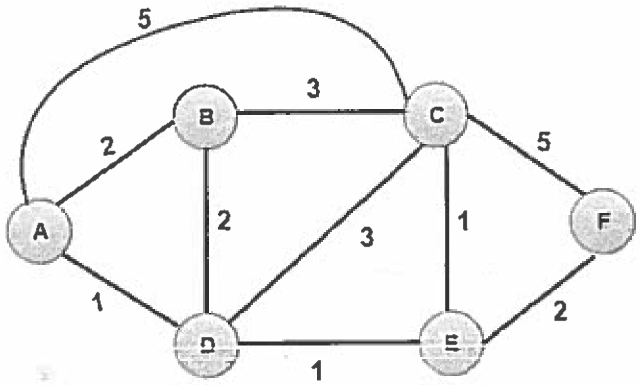
waits in a queue before being transmitted. QoS enables organizations to avoid this by creating a priority queue for certain types of traffic.

3. Loss: The amount of data lost as a result of packet loss, which typically occurs due to network congestion. QoS enables organizations to decide which packets to drop in this event.

4. Jitter: The irregular speed of packets on a network as a result of congestion, which can result in packets arriving late and out of sequence. This can cause distortion or gaps in audio and video being delivered.

OR

11 (a) Tabulate the shortest path for all nodes for the following network using Links state routing.Explanation-4M,Tables-3M 7M 20CS502.3 L3



Link state routing is a technique in which each router shares the knowledge of its neighborhood with every other router in the internetwork. The three keys to understand the Link State Routing algorithm: o Knowledge about the neighborhood: Instead of sending its routing table, a router sends the information about its neighborhood only. A router broadcast its identities and cost of the directly attached links to other routers. o Flooding: Each router sends the information to every other router on the internetwork except its neighbors. This process is known as Flooding. Every router that receives the packet sends the copies to all its neighbors. Finally, each and every router receives a copy of the same information. o Information sharing: A router sends the information to every other router only when the change occurs in the information.

Link State Routing has two phases:
Reliable Flooding o Initial state: Each node knows the cost of its neighbors.
Final state: Each node knows the entire graph

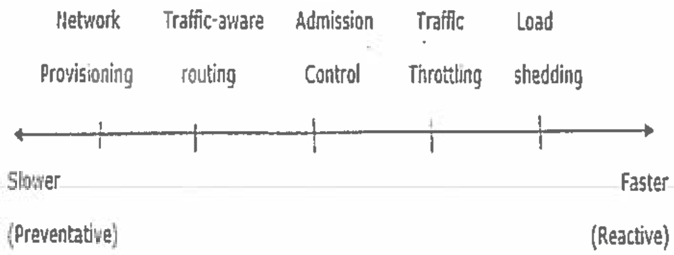
Final table:

Step	N	D(B),P(B)	D(C),P(C)	D(D),P(D)	D(E),P(E)	D(F),P(F)
1	A	2,A	5,A	1,A	∞	∞
2	AD	2,A	4,D		2,D	∞
3	ADE	2,A	3,E			4,E
4	ADEB		3,E			4,E
5	ADEBC					4,E
6	ADEBCF					

11 (b) Briefly discuss about the approaches to Congestion Control with its timeline. 5M 20CS502.3 L2

Approaches to Congestion Control

There are some approaches for congestion control over a network which are usually applied on different time scales to either prevent congestion or react to it once it has occurred.



Time scale of approaches to congestion control

Let us understand these approaches step wise as mentioned below −

Step 1 − The basic way to avoid congestion is to build a network that is well matched to the traffic that it carries. If more traffic is directed but a low-bandwidth link is available, definitely congestion occurs.

Step 2 − Sometimes resources can be added dynamically like routers and links when there is serious congestion. This is called provisioning, and which happens on a timescale of months, driven by long-term trends.

Step 3 − To utilise most existing network capacity, routers can be tailored to traffic patterns making them active during daytime when network users are using more and sleep in different time zones.

Step 4 – Some of local radio stations have helicopters flying around their cities to report on road congestion to make it possible for their mobile listeners to route their packets (cars) around hotspots. This is called traffic aware routing.

Step 5 – Sometimes it is not possible to increase capacity. The only way to reduce the congestion is to decrease the load. In a virtual circuit network, new connections can be refused if they would cause the network to become congested. This is called admission control.

Step 6 – Routers can monitor the average load, queueing delay, or packet loss. In all these cases, the rising number indicates growing congestion. The network is forced to discard packets that it cannot deliver. The general name for this is Load shedding. The better technique for choosing which packets to discard can help to prevent congestion collapse.

12 (a)	Explain about TCP Addressing with respect to transport layer. **Scheme :** **Diagram-2M,** **Explanation-5M**	6M	20CS502.4	L2

TCP

- TCP stands for Transmission Control Protocol.
- It provides full transport layer services to applications.
- It is a connection-oriented protocol means the connection established between both the ends of the transmission. For creating the connection, TCP generates a virtual circuit between sender and receiver for the duration of a transmission.

TCP Segment Format

<table>
<tr><td colspan="8">Source port address 16 bits</td><td>Destination port address 16 bits</td></tr>
<tr><td colspan="9">Sequence number 32 bits</td></tr>
<tr><td colspan="9">Acknowledgement number 32 bits</td></tr>
<tr><td>HLEN 4 bits</td><td>Reserved 6 bits</td><td>URG</td><td>ACK</td><td>PSH</td><td>RST</td><td>SYN</td><td>FIN</td><td>Window size 16 bits</td></tr>
<tr><td colspan="8">Checksum 16 bits</td><td>Urgent pointer 16 bits</td></tr>
<tr><td colspan="9">Options & padding</td></tr>
</table>

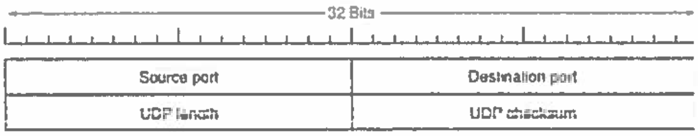
Where,

- **Source port address:** It is used to define the address of the application program in a source computer. It is a 16-bit field.
- **Destination port address:** It is used to define the address of the application program in a destination computer. It is a 16-bit field.
- **Sequence number:** A stream of data is divided into two or more TCP segments. The 32-bit sequence number field represents the position of the data in an original data stream.
- **Acknowledgement number:** A 32-field acknowledgement number acknowledge the data from other communicating devices. If ACK field is set to 1, then it specifies the sequence number that the receiver is expecting to receive.
- **Header Length (HLEN):** It specifies the size of the TCP header in 32-bit words. The minimum size of the header is 5 words, and the maximum size of the header is 15 words. Therefore, the maximum size of the TCP header is 60 bytes, and the minimum size of the TCP header is 20 bytes.
- **Reserved:** It is a six-bit field which is reserved for future use.
- **Control bits:** Each bit of a control field functions individually and independently. A control bit defines the use of a segment or serves as a validity check for other fields.

There are total six types of flags in control field:

- **URG:** The URG field indicates that the data in a segment is urgent.
- **ACK:** When ACK field is set, then it validates the acknowledgement number.
- **PSH:** The PSH field is used to inform the sender that higher throughput is needed so if possible, data must be pushed with higher throughput.
- **RST:** The reset bit is used to reset the TCP connection when there is any confusion occurs in the sequence numbers.
- **SYN:** The SYN field is used to synchronize the sequence numbers in three types of segments: connection request, connection confirmation (with the ACK bit set), and confirmation acknowledgement.
- **FIN:** The FIN field is used to inform the receiving TCP module

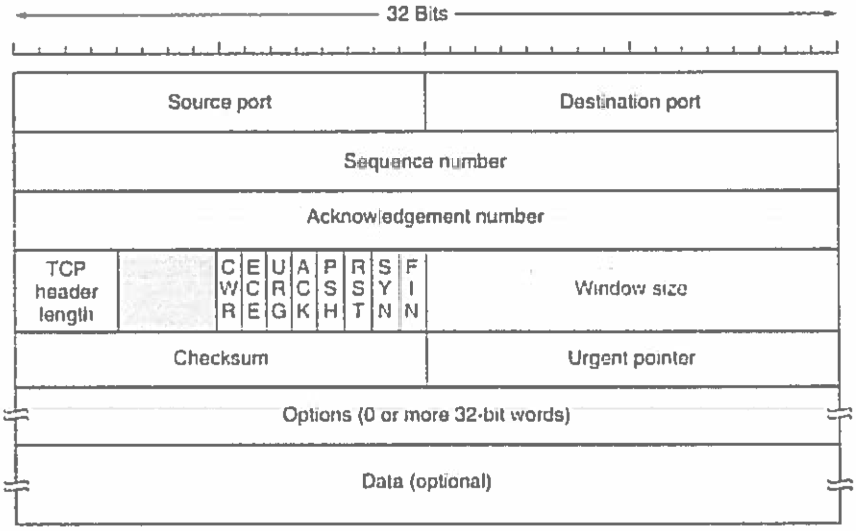
	that the sender has finished sending data. It is used in connection termination in three types of segments: termination request, termination confirmation, and acknowledgement of termination confirmation. o **Window Size:** The window is a 16-bit field that defines the size of the window. o **Checksum:** The checksum is a 16-bit field used in error detection. o **Urgent pointer:** If URG flag is set to 1, then this 16-bit field is an offset from the sequence number indicating that it is a last urgent data byte. o **Options and padding:** It defines the optional fields that convey the additional information to the receiver.			
12 (b)	Explain about TCP Congestion control with respect to transport layer. **Scheme :** **Expalanation-4M,** **Daigrams-2M** If the transport entities on many machines send too many packets into the net-work too quickly, the network will become congested, with performance degraded as packets are delayed and lost. Controlling congestion to avoid this problem is the combined responsibility of the network and transport layers. Congestion oc-curs at routers, so it is detected at the network layer. However, congestion is ultimately caused by traffic sent into the network by the transport layer. The only effective way to control congestion is for the transport protocols to send packets into the network more slowly. Desirable Band width Allocation Efficiency and Power Regulating the sending Rate Wireless Issues	6M	20CS502.4	L2
	OR			
13 (a)	Explain the UDP header and its components with a neat sketch **Scheme :** **Explanation-4M** **Daigram-2M** • The Internet protocol suite supports a connectionless transport protocol called • UDP (User Datagram Protocol). UDP provides a way for applications to send • Encapsulated IP datagrams without having to establish a connection. UDP is de- scribed in RFC 768.	6M	20CS502.4	L2



13 (b)	Explain the TCP Segment header and its components with a neat sketch **Scheme :** **.Explanation-4M** **Daigram-2M**	6M	20CS502.4	L2

TCP Segment header

- Every segment begins with a fixed-format, 20-byte header. The fixed header may be followed by header options. After the options, if any, up to 65,535 − 20 − 20 = 65,495 data bytes may follow, where the first 20 refer to the IP header and the second to the TCP header.
- Segments without any data are legal and are commonly used for acknowledgements and control messages.



14 (a)	Write a short note on DNS Namespace. **Scheme :** **Explanation-4M** **Daigrams-2M**	6M	20CS502.5	L1

The entire collection of DNS administrative domains throughout the world are organized in a hierarchy called the DNS namespace. This section shows how the namespace organization affects both local domains and the Internet.

The domain hierarchy is, conceptually, a "leaf" of the huge DNS namespace supported on the global Internet.

	The DNS namespace for the Internet is organized hierarchically. It consists of the root directory, represented as a dot (.) and two top level domain hierarchies, one organizational and one geographical. Note that the com domain introduced in one of a number of top-level organizational domains in existence on the Internet.			
14 (b)	Explain about Domain Resource Record with its format . **Scheme :** **Explanation-4M** **Format-2M**	6M	20CS502.5	L2

- Every domain, whether it is a single host or a top-level domain, can have a set of resource records associated with it. These records are the DNS database. For a single host, the most common resource record is just its IP address, but many other kinds of resource records also exist. When a resolver gives a domain name to DNS, what it gets back are the resource records associated with that name. Thus, the primary function of DNS is to map domain names onto resource records.A resource record is a five-tuple. Although they are encoded in binary for efficiency, in most expositions resource records are presented as ASCII text, one line per resource record. The format we will use is as follows:

  Domain name Time to live Class Type Value

- The Domain name tells the domain to which this record applies. Normally, many records exist for each domain and each copy of the database holds information about multiple domains. This field is thus the primary search key used to satisfy queries. The order of the records in the database is not significant.

- The Time to live field gives an indication of how stable the record is. Information that is highly stable is assigned a large value, such as 86400 (the number of seconds in 1 day). Information that is highly volatile is assigned a small value,such as 60 (1 minute). We will come back to this point later when we have discussed caching.

- The third field of every resource record is the Class. For Internet information,it is always IN. For non-Internet information, other codes can be used, but in practice these are rarely seen.

- The Type field tells what kind of record this is. There are many kinds of DNS records.

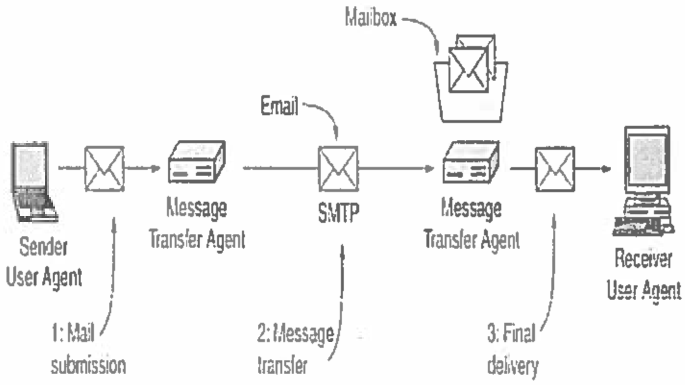
- The most important record type is the A (Address) record. It holds a 32-bit IPv4 address of an interface for some host. The corresponding AAAA, or "quad A," record holds a 128-bit IPv6 address.

OR

15 (a)	Explain about the architecture of email with a neat sketch. **Sol :**	6M	20CS502.5	L1

Explanation: 4M
Daigram-2M

- Electronic mail, or more commonly email, has been around for over three decades. Faster and cheaper than paper mail, email has been a popular application since the early days of the Internet. Before 1990, it was mostly used in academia. During the 1990s, it became known to the public at large and grew exponentially, to the point where the number of emails sent per day now is vastly more than the number of snail mail (i.e., paper) letters. Other forms of network communication, such as instant messaging and voice-over-IP calls have expanded greatly in use over the past decade, but email remains the workhorse of Internet communication. It is widely used within industry for intracompany communication, for example, to allow far-flung employees all over the world to cooperate on complex projects. Unfortunately, like paper mail, the majority of email—some 9 out of 10 messages—is junk mail or spam (McAfee, 2010).



Explain about any one of mail transport protocols with its purpose.

15 (b) **Sol :** 6M 20CS502.5 L2

Any one Mail transport protocol – 6M

Email is emerging as one of the most valuable services on the internet today. Most internet systems use SMTP as a method to transfer mail from one user to another. SMTP is a push protocol and is used to send the mail whereas POP (post office protocol) or IMAP (internet message access protocol) are used to retrieve those emails at the receiver's side.

SMTP Fundamentals

SMTP is an application layer protocol. The client who wants to send the mail opens a TCP connection to the SMTP server and then sends the mail across the connection. The SMTP server is an always-on listening mode. As soon as it listens for a TCP connection from any client, the SMTP process initiates a connection through port 25. After successfully establishing a TCP connection the client process sends the mail instantly.

SMTP Protocol

The SMTP model is of two types:

1. End-to-end method
2. Store-and- forward method

NSRIT

Semester End Regular Examination, June, 2022

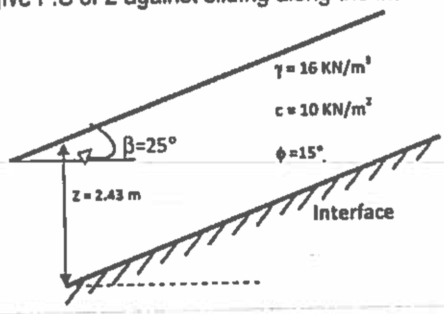
Degree	B. Tech. (U. G.)	**Program**		CE		**Academic Year**	2021 - 2022
Course Code	20CE404	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	IV
Course	**SOIL MECHANICS**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Relate between void ratio, degree of saturation, specific gravity and water content from fundamental.	20CE404.1	L2
2	List any four factors affecting permeability of soil.	20CE404.2	L1
3	List any four assumptions made in Boussinesq's theory for point load.	20CE404.3	L1
4	Categorize any two merits and demerits of direct shear test.	20CE404.4	L1
5	State the significance of stability number	20CE404.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain the salient features of Indian standard soil classification system.	6M	20CE404.1	L2
6 (b)	A natural soil deposit has a bulk unit weight of18 kN/m^3 and a water content of 5%. Calculate the amount of water in liters required to be added to 1 Cu.m of soil to raise the water content to 15%. Assume the void ratio to remain constant. Take Gs = 2.60	6M	20CE404.1	L3
	OR			
7 (a)	Explain the Effect of compaction on soil properties.	6M	20CE404.1	L2
7 (b)	The mass of wet soil when compacted in a mould was 19.55 kN. The water content of the soil was 16%. If the volume of the mould is 0.95 m^3 , If G= 2.68 then determine: (i) dry unit weight (ii) void ratio (iii) degree of saturation (iv) percent air voids	6M	20CE404.1	L3
8 (a)	What is flow net? State its properties and application. What is the quantity of seepage between two successive flow line and equipotential line? Explain with neat sketch.	6M	20CE404.2	L2
8 (b)	In a falling head permeability test, head causing flow was initially 500 mm and it drops to 20 mm in 5 minutes. Calculate the time required for the head to fall 250 mm.	6M	20CE404.2	L3
	OR			
9 (a)	Derive an expression for coefficient of permeability using variable head permeameter test with a neat sketch.	6M	20CE404.2	L2
9 (b)	In order to compute the seepage loss through the foundation of a cofferdam, flownets were constructed. The result of the flownet study gave N_f= 6, N_d = 16. The head of water lost during seepage was 19.68 m. If the hydraulic conductivity of the soil is k = 13.12x10^{-5} m/s, compute the seepage loss per metre length of dam per day	6M	20CE404.2	L3
10 (a)	Write the assumptions of Terzaghi's one - dimensional consolidation theory.	5M	20CE404.3	L2
10 (b)	Find the intensity of vertical pressure at a point 3 m directly below 25 kN point load acting on a horizontal ground surface. What will be the vertical pressure at a distance 2 m horizontally away from the axis of loading and at a same depth of 3 m? Use Boussinesq's equation.	7M	20CE404.3	L3
	OR			
11 (a)	Explain any one method of Computation of Rate of Settlement.	5M	20CE404.3	L2
11 (b)	A Subsurface exploration at the site of a proposed building reveals the existence of 2.4 m thick layer of soft clay below a stratum of coarse	7M	20CE404.3	L3

	sand which is 4 m thick and extends from the ground surface up to the top of the clay layer. The ground water table is at 2.5 m below the ground surface. The laboratory tests indicate the natural water content of the clay 40%, average liquid limit as 45% and specific gravity of solids as 2.75. The unit weight of the sand above and below water table is 17.8 kN/m^3 and 21 kN/m^3 respectively. Estimate the probable settlement of the building, if its construction will increase average vertical pressure on the clay layer by 71 KPa.			
12 (a)	What are the various types of shear tests based on drainage conditions? Explain them.	6M	20CE404.4	L2
12 (b)	A sample of dry sand was subjected to triaxial test, with a confining pressure of 150 kN/m^2. The angle of shearing resistance was found to be 33°. At what value of major principal stress, the sample is likely to fail.	6M	20CE404.4	L3
	OR			
13 (a)	A sample of dry cohesion less soil was tested in a triaxial machine. If the angle of shearing resistance is 36° and the confining pressure 100 kN/m^2, determine the deviator stress at which the sample failed.	6M	20CE404.4	L3
13 (b)	Explain with neat sketches the procedure of conducting Direct Shear test.	6M	20CE404.4	L2
14 (a)	What are the various methods of analysis of infinite slopes? Explain briefly any one of method.	6M	20CE404.5	L1
14 (b)	Find the Factor of safety against sliding along the interface for the infinite slope shown in Figure below. Also find the height Z that will give F.S of 2 against sliding along the interface. 	6M	20CE404.5	L2
	OR			
15	Explain the friction Circle method of analysis of stability of slopes.	12M	20CE404.5	L1

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM, ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

SOIL MECHANICS 20CE404

PART-A

1. Relate between void ratio, degree of saturation, specific gravity and water content from fundamental. 2M

Ans) Relation between *e, w, G & S*

$$w = \frac{W_w}{W_s} = \frac{\gamma_w \times V_w}{\gamma_s \times V_s} = \frac{V_w\, V_v}{V_s\, V_v\, G} = \frac{S\, e}{G}$$

$$Se = wG$$

2. List any four factors affecting permeability of soil. 2M

Ans) Factors affecting permeability of soil

1. Particle size
2. Structure of soil mass
3. Shape of Particles
4. Void ratio
5. Properties of water
6. Degree of Saturation
7. Adsorbed water
8. Impurities in water

3. List any four assumptions made in Boussinesq's theory for point load. (2M)

Ans) Boussinesq's theory assumptions:

1. The soil medium is an elastic, homogeneous, isotropic and semi-infinite medium, which extends infinitely in all directions from a level surface.
2. The medium obeys Hooke's law.
3. The self-weight of the soil is ignored.
4. The soil is initially unstressed
5. The change in volume of the soil upon application of the loads on to it is neglected.
6. The top surface of the medium is free of shear stress and is subjected to only the point load at a specified location. Continuity of stress is considered to exist in the medium.
7. The stresses are distributed symmetrically with respect to z axis.

4. Categorize any two merits and demerits of direct shear test. (2M)

Ans) Merits of direct shear test:

1. The sample preparation is easy. The test is simple and convenient.
2. As the thickness of the sample is relatively small, the drainage is quick and the pore pressure dissipates very rapidly.
3. Direct shear test is ideally suited for conducting drained tests on cohesionless soils.
4. The apparatus is relatively cheap

Demerits of direct shear test:

1. The stress condition is known only at failure. The conditions prior to failure are indeterminate and, therefore, the Mohr circle cannot be drawn.
2. In direct shear test, the stress distribution on the failure plane (horizontal plane) is not uniform.
3. The area under shear gradually decreases as the test progresses. But the corrected area cannot be determined and therefore, the original area is taken for the computation of stresses.
4. The orientation of failure plane is fixed. This plane may not be the weakest plane.
5. Control on the drainage conditions is very difficult. So, only drained tests can be conducted on highly permeable soils.
6. The measurement of pore water pressure is not possible in direct shear test.
7. The side walls of the shear box cause lateral restraint on the specimen and do not allow it to deform laterally.

5. State the significance of stability number (2M)

Ans) Significance of stability number:

- It is the method used to evaluate slope stability for homogeneous soils having cohesion. This method is proposed by the Taylor.
- It is based on the principle resistance of soil mass against sliding, because of cohesion and internal friction acting over the failure plane.
- This failure surface is assumed to be circular arc. The factors affecting the stability of soil slope is expressed with the parameter stability number.
- Stability number (S_n) is given by

$$S_n = \frac{c}{F_c \gamma H}$$

PART- B

6. a) Explain the salient features of Indian standard soil classification system. (6M)

Ans) INDIAN STANDARD CLASSIFICATION (ISC) SYSTEM

- Indian Standard Classification (ISC) system adopted by Bureau of Indian Standards is in many respects similar to the Unified Soil Classification (USC) system.
- In classification of fine-grained soil, there is one difference between USC and ISC system
- The fine-grained soils in ISC system are subdivided into three categories of low, medium and high compressibility instead of two categories of low and high compressibility in USC system.
- Soils are divided into three broad divisions:

1. Coarse-grained soils, when 50% or more of the total material by weight is retained on 75-micron sieve.
2. Fine-grained soils, when more than 50% of the total material passes 75 micron IS sieve.
3. If the soil is highly organic and contains a large percentage of organic matter and particles d decomposed vegetation, it is kept in a separate category marked as peat (P_t).

In this system, the soils are classified in to 18 groups

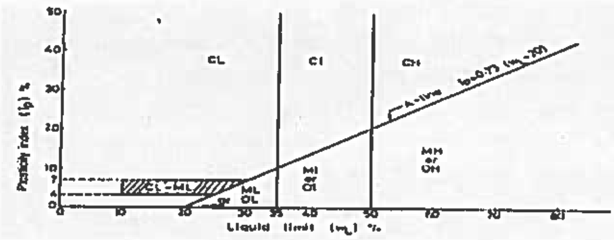
1. 8 groups of coarse-grained soils
2. 9 groups of fine-grained and
3. One of peat

Coarse-grained Soils

- Coarse-grained soils are subdivided into gravel and sand.
- The soil is termed gravel (G) when more than 50% of coarse fraction (plus 75μ) is retained on 4.75 mm IS sieve, and termed sand (S) if more than 50% of the coarse fraction is smaller than 4.75 mm IS sieve.
- Coarse-grained soils are further subdivided as given in to 8 groups.

Fine-grained Soils

- The fine-grained soils are further divided into three subdivisions, depending upon the values of the liquid limit:
  1. **Silts and clays of low compressibility:** These soils have a liquid limit less than 35 (represented by symbol L).
  2. **Silts and clays of medium compressibility:** These soils have a liquid limit greater than 35 but less than 50 (represented by symbol I).
  3. **Silts and clays of high compressibility:** These soils have a liquid limit greater than 50 (represented by symbol H).
- Fine-grained soils are further subdivided, in 9 groups based on plasticity as shown in figure



6. b) A natural soil deposit has a bulk unit weight of18 kN/m³ and a water content of 5%. Calculate the amount of water in litres required to be added to 1 Cu.m of soil to raise the water content to 15%. Assume the void ratio to remain constant. Take G_s = 2.60 (6M)

Sol)

6b) Given data,

bulk unit weight $\gamma_{b_1} = 18\ kN/m^3$

Water content $w_1 = 5\% = 0.05$

$G_s = 2.60$

$$\gamma_d = \frac{\gamma_b}{1+w} = \frac{18}{1+0.05} = 17.14\ kN/m^3$$

Bulk unit weight (γ_{b_2}) corresponding to $w_2 = 15\%$ is

$$\gamma_d = \frac{\gamma_{b_2}}{1+w}$$

$$17.14 = \frac{\gamma_{b_2}}{1+0.15}$$

$$\gamma_{b_2} = 19.711\ kN/m^3$$

Amount of water required to raise $w = 15\%$ is

$$\gamma_{b_2} - \gamma_{b_1} = 19.711 - 18 = 1.711\ kN/m^3$$

For one cum of soil water added should be 1.711 kN

Specific wt. of water $\gamma_w = \frac{W_w}{V_w}$

$$V_w = \frac{1.711}{9.81} = 0.1744\ m^3$$

$= 174.4$ litres

7. a) Explain the Effect of compaction on soil properties. (6M)

Ans) EFFECT OF COMPACTION ON SOIL PROPERTIES

1. Soil Structure
- Soils compacted to dry of optimum have flocculated structure due to the attraction between soil particles because of low water content.
- Soils compacted to wet of optimum have dispersed structure due to repulsive force between soil particles because of high water content.
2. Permeability
- Compaction reduces the voids present in the soil hence permeability also reduces.
- At a particular density, for the same soil sample, permeability is more for soils which are compacted to dry of optimum than those compacted to wet of optimum
- If the compactive effort is increased, the permeability of the soil decreases due to increased dry density and better orientation of particles.
3. Swelling of soils
- When the soil is compacted to dry of optimum, the soil is in need of water and it swells easily when contacted with water.
- When soil is compacted to wet of optimum, the soil particles are oriented in a dispersed manner and swelling does not occur.
- So, to avoid swelling, soils should be compacted to wet of optimum
4. Pore Water Pressure
- Pore water pressure is high for those soil whose water content is high.
- Hence, soils compacted to wet of optimum compaction will exhibit more pore water pressure than soil compacted dry of optimum
5. Shrinkage of Soil
- Shrinkage is more for the soil compacted to wet of optimum than dry of optimum.
- In case of dry of optimum compaction, soil particles are in random orientation and they are in stable condition.
- But in case of wet of optimum, soil particles are in parallel orientation and they are unstable which makes it easy for packing of particles causing shrinkage
6. Compressibility
- The Compressibility of compacted soil varies according to the amount of pressure applied.
- For low-pressure range, compressibility is more for soils which are compacted to wet of optimum than soil compacted to dry of optimum.
- Similarly, for high-pressure ranges, compressibility is more for soils which are compacted to dry of optimum than soil compacted to wet of optimum

7. b) The mass of wet soil when compacted in a mould was 19.55 kN. The water content of the soil was 16%. If the volume of the mould is 0.95 m³ , If G= 2.68 then determine: (i) dry unit weight (ii) void ratio (iii) degree of saturation (iv) percent air voids (6M)

Sol) Bulk unit weight, $\gamma = \frac{W}{V} = \frac{19.55}{0.95} = 20.58$ kN/m³

(i) dry unit weight, $\gamma_d = \frac{\gamma}{1+w} = \frac{20.58}{1+0.16} = 17.74$ kN/m³

(ii) void ratio,

$$\gamma_d = \frac{G\gamma_w}{1+e}$$

$$e = \frac{G\gamma_w}{\gamma_d} - 1 = \frac{2.68 \cdot 9.81}{17.74} - 1 = 0.482$$ (Taking $\gamma_w = 9.81$ kN/m³)

(iii) degree of saturation

$$Se = wG$$

$$S = \frac{wG}{e} = \frac{0.16 * 2.68}{0.48} = 0.8933 = 89.33\%$$

(iv) percent air voids

$$\gamma_d = (1 - n_a)\frac{G\gamma_w}{1 + wG}$$

$$17.74 = (1 - n_a)\frac{2.68 * 9.81}{1 + (0.16 * 2.68)}$$

$$n_a = 0.036 = 3.6\,\%$$

8. a) What is flow net? State its properties and application. What is the quantity of seepage between two successive flow line and equipotential line? Explain with neat sketch (6M)

Ans) Flow net

- The entire pattern of flow lines and equipotential lines is referred as a flow net.
- The flow lines and equipotential lines together form a flow net.
- The flow net gives a pictorial representation of the path taken by water particles and the head variation along that path.

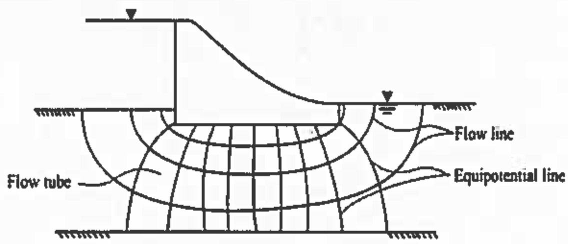
Properties of flow net are as follows:

- The angle of intersection between each flow line and an equipotential line must be 90^0 which means they should be orthogonal to each other.
- Two flow lines or two equipotential lines can never cross each other.
- Equal quantity of seepage occurs in each flow channel. A flow channel is a space between two flow lines.
- Head loss is the same between two adjacent potential lines.
- Flow nets are drawn based on the boundary conditions only. They are independent of the permeability of soil and the head causing flow.
- The space formed between two flow lines and two equipotential lines is called a flow field. It should be in a square form.
- Either flow lines or equipotential lines are smoothly drawn curves.

Applications of Flow Net

Flow net is useful determine the following parameters in seepage analysis of soil :

1. Rate of Seepage loss
2. Seepage Pressure
3. Uplift Pressure
4. Exit Gradient



Flow net

8. b) In a falling head permeability test, head causing flow was initially 500 mm and it drops to 20 mm in 5 minutes. Calculate the time required for the head to fall 250 mm. (6M)

Sol) Coefficient of permeability (k) using variable head permeability test is :

$$k = \frac{2.303\, a\, L\, \log_{10}\left(\frac{h_1}{h_2}\right)}{At}$$

In first case, $h_1 = 500\ mm; h_2 = 20\ mm;$ t = 5 minutes

In second case, $h_1 = 500\ mm; h_2 = 250\ mm;$ t = ?

a, L & A *are* constant for an equipment, equating coefficient of permeability for both the cases,

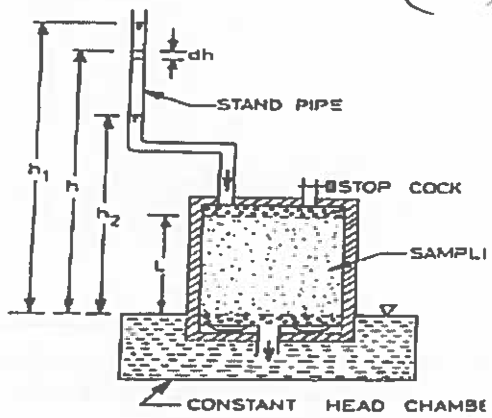
$$\frac{2.303\, a\, L\, \log_{10}\left(\frac{500}{20}\right)}{A * 5} = \frac{2.303\, a\, L\, \log_{10}\left(\frac{500}{250}\right)}{A * t}$$

Solving we get, the time required for the head to fall 250 mm from 500 mm is

$t = 1.08\, minutes$

9. a) Derive an expression for coefficient of permeability using variable head permeameter test with a neat sketch (6M)

Ans)



Expression for coefficient of permeability using variable head permeameter test:

- For relatively less permeable soils, the quantity of water collected in the graduated jar of the constant-head permeability test is very small and cannot be measured accurately.
- For such soils, the variable-head permeability test is used.
- Water flows through the sample from a standpipe attached to the top of the cylinder.
- The head of water (h) changes with time as flow occurs through the soil. At different times the head of water is recorded.

Let us consider the instant when the head is h. For the infinitesimal small time dt, the head falls by dh. Let the discharge through the sample be q. From continuity of flow,

$$a\, dh = -q\, dt$$

where a is cross-sectional area of the standpipe.

or $$a\, dh = -(A \times k \times i) \times dt$$

or $$a\, dh = -A\, k \times \frac{h}{L} \times dt$$

or $$\frac{A\, k\, dt}{a\, L} = \frac{-dh}{h}$$

Integrating, $$\frac{A\, k}{a\, L} \int_{t_1}^{t_2} dt = -\int_{h_1}^{h_2} \frac{dh}{h}$$

or $$\frac{A\, k}{a\, L}(t_2 - t_1) = \log_e (h_1/h_2)$$

or $$k = \frac{a\, L}{A\, t} \log_e (h_1/h_2)$$

9. b) In order to compute the seepage loss through the foundation of a cofferdam, flow nets were constructed. The result of the flow net study gave N_f= 6, N_d = 16. The head of water lost during seepage was 19.68 m. If the hydraulic conductivity of the soil is k = 13.12x10^{-5} m/s, compute the seepage loss per metre length of dam per day (6M)

Sol) Seepage loss per unit length of dam is given by following equation

$$q = kH\frac{N_f}{N_d}$$

Given, N_f= 6; N_d = 16; h = 19.68 m; k = 13.12x10^{-5} m/s

$$q = 13.12 \times 10^{-5} \times 19.68 \times \frac{6}{16} = 9.68256 \times 10^{-4}\ \text{m}^3/\text{s/m length}$$

$= 83.66\ m^3/day/m$ length

10. a) Write the assumptions of Terzaghi's one - dimensional consolidation theory. (5M)

Ans) Assumptions of Terzaghi's one - dimensional consolidation theory

1. The soil is homogeneous and isotropic.
2. The soil is fully saturated (S = 100%)
3. The solid particles and water in the voids are incompressible. The consolidation occurs due to expulsion of water from the voids.
4. The coefficient of permeability of the soil has the same value at all points, and it remains constant during the entire period of consolidation.
5. Darcy's law is valid throughout the consolidation process.
6. Soil is laterally confined, and the consolidation takes place only in axial direction. Drainage of water also occurs only in the vertical direction.
7. The time lag in consolidation is due entirely to the low permeability of the soil.
8. There is a unique relationship between the void ratio and the effective stress, and this relationship remains constant during the load increment

10. b) Find the intensity of vertical pressure at a point 3 m directly below 25 kN point load acting on a horizontal ground surface. What will be the vertical pressure at a distance 2 m horizontally away from the axis of loading and at a same depth of 3 m? Use Boussinesq's equation. (7M)

Sol) Intensity of vertical pressure using Boussinesq's equation is:

$$\sigma_z = \frac{3Q}{2\pi\, z^2}\left[\frac{1}{1+\left(\frac{r}{z}\right)^2}\right]^{\frac{5}{2}}$$

i) Intensity of vertical pressure at a point z = 3 m directly below Q = 25 kN point load

r = 0

$$\sigma_z = \frac{3\times 25}{2\pi\, 3^2}\left[\frac{1}{1+\left(\frac{0}{3}\right)^2}\right]^{\frac{5}{2}} = 1.326\ kPa$$

ii) Intensity of vertical pressure at a point z = 3 m and x = 2 m horizontally due point load Q = 25 kN, r = 2 m

$$\sigma_z = \frac{3\times 25}{2\pi\, 3^2}\left[\frac{1}{1+\left(\frac{2}{3}\right)^2}\right]^{\frac{5}{2}} = 0.529\ kPa$$

11. a) Explain any one method of Computation of Rate of Settlement. (5M)

Ans)

(1) Final Settlement Using Coefficient of Volume Change

Let us consider a small element of thickness Δz at a depth z in the clay deposit of total thickness H_0 12.18). Let the effective pressure increment causing the settlement be $\Delta\bar{\sigma}$. From Eq. 12.15.

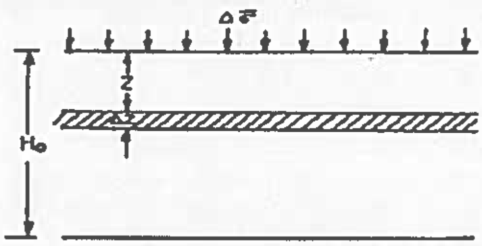


Fig. 12.18. Layer Subjected to $\Delta\bar{\sigma}$.

$$\Delta H = m_v H_0 (\Delta\bar{\sigma})$$

Representing the final settlement as Δs_f and taking $H_0 = \Delta z$,

$$\Delta s_f = m_v \, \Delta z \, (\Delta\bar{\sigma})$$

Total settlement of the complete layer,

$$s_f = \int_0^{H_0} \Delta s_f = \int_0^{H_0} m_v \, \Delta\bar{\sigma} \, dz$$

If both m_v and $\Delta\bar{\sigma}$ are constant,

$$s_f = m_v \, \Delta\bar{\sigma} \quad H_0$$

(2) Final settlement using Void Ratio

If $e - \bar{\sigma}$ plot for the soil is available, it can be used to determine the final settlement. The value of Δe corresponding to the given load increment is read off from the plot and substituted in Eq. 12.11.

$$\Delta H = H_0 \left(\frac{\Delta e}{1 + e_0}\right)$$

or

$$s_f = H_0 \cdot \left(\frac{\Delta e}{1 + e_0}\right)$$

where e_0 is the initial void ratio.

$$s_f = \frac{C_c}{1 + e_0} \cdot H_0 \cdot \log_{10}\left(\frac{\bar{\sigma}_0 + \Delta\bar{\sigma}}{\bar{\sigma}_0}\right)$$

11. b) A Subsurface exploration at the site of a proposed building reveals the existence of 2.4 m thick layer of soft clay below a stratum of coarse sand which is 4 m thick and extends from the ground surface up to the top of the clay layer. The ground water table is at 2.5 m below the ground surface. The laboratory tests indicate the natural water content of the clay 40%, average liquid limit as 45% and specific gravity of solids as 2.75. The unit weight of the sand above and below water table is 17.8 kN/m³ and 21 kN/m³ respectively. Estimate the probable settlement of the building, if its construction will increase average vertical pressure on the clay layer by 71 KPa. (7M)

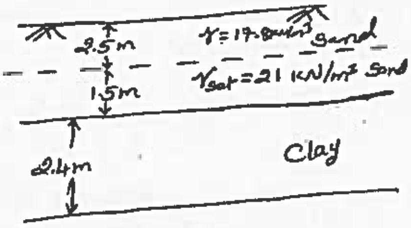
Sol)

11 b) Given data,

Clay, water content $w = 40\%$

Avg liquid limit $W_L = 45\%$

$G_s = 2.75$



Avg increase in vertical pressure due to construction in clay layer is $\Delta\sigma = 71$ kPa

Void ratio $e = \frac{wG}{S} = \frac{0.4 \times 2.75}{1} = 1.1$

Unit weight of clay layer $\gamma_{sat} = \left[\frac{G+Se}{1+e}\right]\gamma_w = \left[\frac{2.75 + (1 \times 1.1)}{1+1.1}\right] \times 9.81$

$\gamma_{sat} = 17.985$ kN/m³

Initial effective overburden pressure at the top of clay layer is

$\sigma_0 = (17.8 \times 2.5) + (11.19 \times 1.5) = 61.285$ kN/m²

Probable settlement of building is

$S = \frac{C_c}{1+e} H \log_{10}\left[\frac{\sigma_0 + \Delta\sigma}{\sigma_0}\right]$ $\quad \left(\because C_c = 0.07(W_L - 10)\right]$

$= 0.07[45-10]$

$= 2.45$

$= \frac{2.45}{1+1.1} \times 2.4 \log_{10}\left[\frac{61.285 + 71}{61.285}\right]$

$= 0.936$ m

936 mm

12. a) What are the various types of shear tests based on drainage conditions? Explain them. (6M)

Ans) Different types of shear test based on drainage conditions:

a. **Unconsolidated–Undrained.**
   - The specimen is subjected to a specified all-round pressure and then the principal stress difference is applied immediately, with no drainage being permitted at any stage of the test.

b. **Consolidated–Undrained:**
   - Drainage of the specimen is permitted under a specified all-round pressure until consolidation is complete
   - The principal stress difference is then applied with no drainage being permitted. Pore water pressure measurements may be made during the undrained part of the test.

c. **Consolidated- Drained:**
   - Drainage of the specimen is permitted under a specified all-round pressure until consolidation is complete; with drainage still being permitted
   - The principal stress difference is then applied at a rate slow enough to ensure that the excess pore water pressure is maintained at zero.

12. b) A sample of dry sand was subjected to triaxial test, with a confining pressure of 150 kN/m^2. The angle of shearing resistance was found to be 33^0. At what value of major principal stress, the sample is likely to fail. (6M)

Ans) Given data, $\sigma_3 = 150$ kN/m^2 ; $\emptyset = 33°$

$$\sigma_1 = \sigma_3 \tan^2\left(45 + \frac{\emptyset}{2}\right) + 2\,c\,\tan\left(45 + \frac{\emptyset}{2}\right)$$

For sand c =0 then above equation becomes,

$$\sigma_1 = \sigma_3 \tan^2\left(45 + \frac{\emptyset}{2}\right) = 150 \tan^2\left(45 + \frac{33}{2}\right) = 508.82 \text{ kN/m}^2$$

The sample is likely to fail at major principal stress, σ_1 = 508.82 kN/m^2

13. a) A sample of dry cohesion less soil was tested in a triaxial machine. If the angle of shearing resistance is 36^0 and the confining pressure 100 kN/m^2, determine the deviator stress at which the sample failed. (6M)

Ans) Given data, $\sigma_3 = 100$ kN/m^2 ; $\emptyset = 36°$

$$\sigma_1 = \sigma_3 \tan^2\left(45 + \frac{\emptyset}{2}\right) + 2\,c\,\tan\left(45 + \frac{\emptyset}{2}\right)$$

For cohesion less soil, c =0 then above equation becomes,

$$\sigma_1 = \sigma_3 \tan^2\left(45 + \frac{\emptyset}{2}\right) = 100 \tan^2\left(45 + \frac{36}{2}\right) = 385.18 \text{ kN/m}^2$$

$$\sigma_1 = \sigma_3 + \sigma_d$$ where σ_d = deviator stress

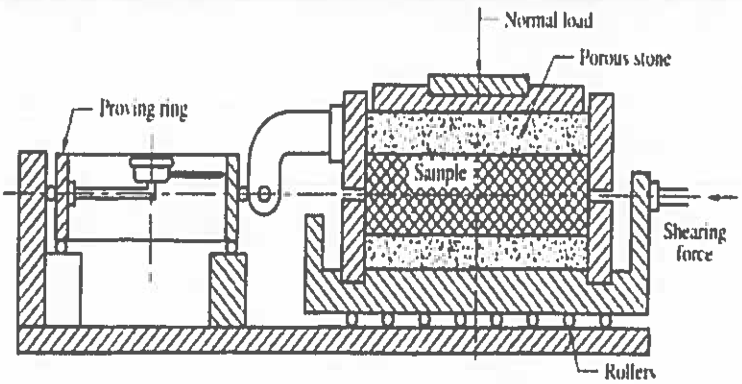
Deviator stress at which the sample failed is:

$$\sigma_d = \sigma_1 - \sigma_3 = 385.18 - 100 = 285.18 \text{ kN/m}^2$$

13. b) Explain with neat sketches the procedure of conducting Direct Shear test. (6M)

Ans) **Direct Shear Test**

- The original form of apparatus for the direct application of shear force is the shear box.
- The box shear test, though simple in principle, has certain shortcomings
- The apparatus consists of a square metal box split horizontally at the level of the center of the soil sample, which is held between metal grilles and porous stones as shown in Figure



Direct Shear Test

- Vertical load is applied to the sample as shown in the figure and is held constant during a test.

- A gradually increasing horizontal load is applied to the lower part of the box until the sample fails in shear.
- The shear load at failure is divided by the cross-sectional area of the sample to give the ultimate shearing strength.
- The vertical load divided by the area of the sample gives the applied vertical stress σ.
- The test may be repeated with a few more samples having the same initial conditions as the first sample. Each sample is tested with a different vertical load.
- The horizontal load is applied at a constant rate of strain.
- The lower half of the box is mounted on rollers and is pushed forward at a uniform rate by a motorized gearing arrangement.
- The upper half of the box bears against a steel proving ring, the deformation of which is shown on the dial gauge indicating the shearing force.
- To measure the volume change during consolidation and during the shearing process another dial gauge is mounted to show the vertical movement of the top plate.
- The horizontal displacement of the bottom of the box may also be measured by another dial gauge which is not shown in the figure.

14. a) What are the various methods of analysis of infinite slopes? Explain briefly any one of method. (6M)

Ans) A slope that extends for a relatively long distance and has a consistent subsurface profile may be analyzed as an infinite slope. The failure plane for this case is parallel to the surface of the slope and the limit equilibrium method can be applied readily

Case 1 : Infinite slope in Cohesionless soil $(c = 0)$

F S against sliding on a given plane at depth 'z' below slope surface

$$F_S = \frac{\gamma z \cos^2\beta \tan\phi}{\gamma z \cos\beta \sin\beta} = \frac{\tan\phi}{\tan\beta}$$

Note : For $\beta = \phi$, Factor of safety $= 1$

For $\beta > \phi$, $F.S < 1.0$

* The maximum inclination of infinite slope in cohensionless soil is equal to angle of internal friction of soil.

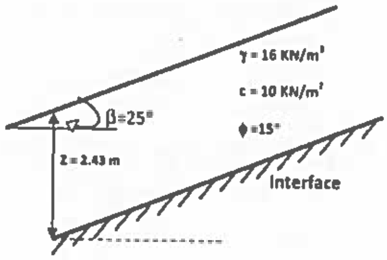
* Infinite slope is stable as long as $\beta < \phi$

Case 2 : Infinite slope in Pure cohesive soil (saturated soil) $\phi_u = 0$

Fs against sliding on a plane at depth 'z' below slope surface

$$\boxed{F_S = \frac{C_u}{\gamma_z \cos\beta \sin\beta}} \longrightarrow (1)$$

14. b) Find the Factor of safety sliding along the interface for the infinite slope shown in Figure below. Also find the height Z that will give F.S of 2 against sliding along the interface. 6M



Sol: i) For infinite slope, Factor of safety against sliding is

$$F_s = \frac{C + \gamma z \cos^2\beta \tan\phi}{\gamma z \cos\beta \sin\beta}$$

$$= \frac{10 + (16 \times 2.43)\cos^2 25 \tan(15^\circ)}{16 \times 2.43 \cos 25 \cdot \sin 25}$$

$$F_s = 1.25$$

ii) Heigh Z, corresponding to F.S = 2

$$2 = \frac{10 + (16 \times z)\cos^2 25 \tan 15^\circ}{16 \times z \times \cos 25 \sin 25}$$

$$12.26z = 10 + 3.52z$$

$$z = 1.14\,m$$

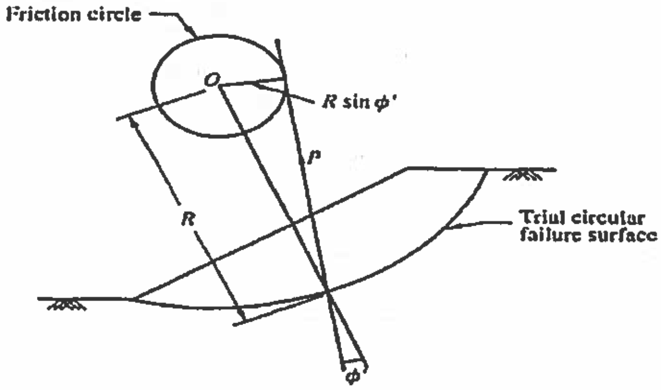
15. Explain the friction Circle method of analysis of stability of slopes. (~~6M~~) (12 M)

Ans) FRICTION-CIRCLE METHOD

Physical Concept of the Method

- The principle of the method is explained with reference to the section through a dam shown in Figure.
- A trial circle with center of rotation O is shown in the figure. With center O and radius sin Φ', where R is the radius of the trial circle, a circle is drawn.
- Any line tangent to the inner circle must intersect the trial circle at an angle Φ' with R.
- Therefore, any vector representing an intergranular pressure at obliquity Φ' to an element of the rupture arc must be tangent to the inner circle.
- This inner circle is called the friction circle or Φ-circle.
- The friction circle method of slope analysis is a convenient approach for both graphical and mathematical solutions.
- It is given this name because the characteristic assumption of the method refers to the Φ-circle.
- The forces considered in the analysis are:

a. The total weight W of the mass above the trial circle acting through the center of mass. The center of mass may be determined by any one of the known methods.
b. The resultant boundary neutral force U. The vector U may be determined by a graphical method from flow net construction.
c. The resultant intergranular force, P, acting on the boundary.
d. The resultant cohesive force C



Principle of friction circle method.

Semester End Regular Examination, June, 2022

Degree	B.Tech.(U.G.)	**Program**	Mechanical Engineering	**Academic Year**	2021–2022
Course Code	20ME404	**Test Duration**	3 Hrs. **Max. Marks** 70	**Semester**	IV
Course	**Fluid Mechanics and Hydraulic Machines**				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define viscosity	20ME404.1	L1
2	Compare between Stream line and Streak line.	20ME404.2	L2
3	List the three types of Similarities.	20ME404.3	L1
4	Derive an expression for the force exerted by the jet on a stationary vertical plate.	20ME404.4	L3
5	Write the difference between impulse turbines and reaction turbines	20ME404.5	L2

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Derive the expression for the hydrostatic force exerted on the vertically submerged plane and obtain the total pressure and position of center of pressure.	12M	20ME404.1	L3
	OR			
7 (a)	Enumerate list of manometers and explain any one with neat sketch	6M	20ME404.1	L2
7 (b)	Briefly explain the conditions for stability of a floating body and submerged body.	6M	20ME404.1	L2
8(a)	Explain the Reynold's experiment with the help of a neat sketch.	6M	20ME404.2	L2
8 (b)	Derive Bernoulli's equation with assumptions.	6M	20ME404.2	L3
	OR			
9 (a)	Define and explain the terms Hydraulic gradient line and Total energy line.	6M	20ME404.2	L2
9 (b)	Explain the characteristics of Laminar and Turbulent boundary layer over a thin flat plate with a neat figure.	6M	20ME404.2	L2
10 (a)	What are the different laws on which models are designed for dynamic similarity? Where are they used?	6M	20ME404.3	L2
10 (b)	State and Explain Buckingham's π - theorem	6M	20ME404.3	L2
	OR			
11 (a)	Define Buckingham's π-theorem. What are the advantages of Pi theorem over the Rayleigh's method for dimension analysis?	6M	20ME404.3	L2
11 (b)	What do you mean by dimensionless numbers? Explain various types of dimensionless numbers	6M	20ME404.3	L2
12 (a)	A jet of water of diameter 75 mm moving with a velocity of 25 m/s strikes a fixed plate in such a way that the angle between the jet and plate is 60 °. Find the force exerted by the jet on the plate (i) in the direction normal to the plate (ii) in the direction of the jet.	6M	20ME404.4	L3
12 (b)	The internal and external diameters of the impeller of a centrifugal pump are 200mm and 400mm respectively. The pump is running at 1200 rpm. The vane angles of the impeller at inlet and outlet are 20^0 and 30^0 respectively. The water enters the impeller radially and velocity of flow is constant. Determine the work done by the impeller per unit weight of water. .	6M	20ME404.4	L3

	OR			
13 (a)	Explain the working of Centrifugal pump with a neat sketch.	6M	20ME404.4	L2
13 (b)	A centrifugal pump with impeller of outer dia. 45 cm and inner dia. of 25 cm, is required to develop a net head of 20 m. Find the lowest speed to start the pumping.	6M	20ME404.4	L3
14 (a)	Explain in detail about performance curves of turbines	6M	20ME404.5	L2
14 (b)	A Pelton wheel has a mean bucket speed of 35 m/s with a jet of water flowing at the rate of 1 m^3/s under a head of 270 m. The buckets deflect the jet through an angle of 170°. Calculate the power delivered to the runner and the hydraulic efficiency of the turbine. Assume co-efficient of velocity as 0.98.	6M	20ME404.5	L3
	OR			
15 (a)	A single jet Pelton wheel develops 2 MW power under a gross head of 360m, while running at 560 rpm. The water is supplied through a penstock which is 1200 m long. Take Cv = 0.98 and friction factor f= 0.03, hydraulic efficiency as 85% . The head lost in the penstock is 12 m of water. Find out the quantity of water supplied to the turbine, diameter of the nozzle and the diameter of the penstock .	6M	20ME404.5	L3
15 (b)	Explain in detail the working principle of Kaplan Turbine with neat sketch.	6M	20ME404.5	L2

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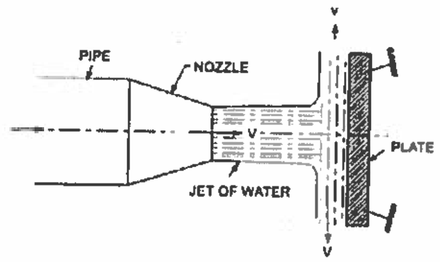
ANSWER KEY AND SCHEME OF EVALUATION

Part A (Short Answer Questions 5 x 2 = 10 Marks)

1. **Viscosity:** It is a property of liquid that is closely related to the resistance of flow. It is defined as in terms of the force required to move on plane surface continuously past another specified steady state conditions when the space between them is filled with a liquid.
   It is also defined as the shear stress required to produce unit rate of shear strain

$$\tau = \mu \frac{du}{dy}$$

2. **Streamline:** It is an imaginary line showing the positions of various fluid particles. Streamlines cannot intersect with each other, they are always parallel
   Streak line: It is a real line showing instantaneous positions of various particles. Streak line changes with time. Two streak lines may intersect each other
3. The 3 types of similarities are A) Geometric Similarity B) Kinematic Similarity C) Dynamic Similarity
4. **Expression for the force exerted by the jet on a stationary vertical plate:** Let us consider a jet of water, which is coming from the outlet of nozzle fitted at the pipe, strikes a flat vertical flat plate as displayed here in following figure. V = Velocity of the jet, d = Diameter of the jet, a = Area of cross-section of the jet = $(\pi/4) \times d^2$



Force exerted by liquid jet on the plate in the direction of jet will be determined by using the concept of impulse momentum equation.

F_x = Rate of change of momentum in the direction of force

$= \dfrac{\text{Initial momentum} - \text{Final momentum}}{\text{Time}}$

$= \dfrac{(\text{Mass} \times \text{Initial velocity} - \text{Mass} \times \text{Final velocity})}{\text{Time}}$

$= \dfrac{\text{Mass}}{\text{Time}} [\text{Initial velocity} - \text{Final velocity}]$

= (Mass/sec) × (velocity of jet before striking − velocity of jet after striking)

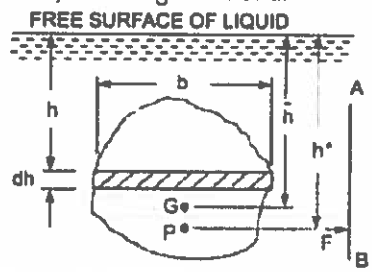
$= \rho a V[V - 0]$ $\quad (\because \text{ mass/sec} = \rho \times a\, V)$

$= \rho a V^2$

5. The basic and main difference between impulse and reaction turbine is that there is pressure change in the fluid as it passes through runner of reaction turbine while in impulse turbine there is no pressure change in the runner. In the impulse turbine first all pressure energy of water convert into the kinetic energy through a nozzle and generate a high speed jet of water. This water jet strikes the blade of turbine and rotates it. In the reaction turbine there is pressure change of water when it passes through the rotor of turbine. So it uses kinetic energy as well as pressure energy to rotate the turbine. Due to this it is known as reaction turbine

Part B (Long Answer Questions 5 x 12 = 60 Marks)

6. **Expression for the hydrostatic force exerted on the vertically submerged plane:**
Let us consider the small strip of thickness dh, width b and at a depth of h from free surface of liquid as displayed here in above figure.
Intensity of pressure on small strip,
dp = ρgh Area of strip,
dA = b x dh Total pressure force on small strip,
dF = dP x dA Total pressure force on small strip,
dF = ρgh x b x dh
Total pressure force on whole surface, F = Integration of dF



$$F = \int dF = \int \rho g h \times b \times dh = \rho g \int b \times h \times dh$$

But $\int b \times h \times dh = \int h \times dA$

= Moment of surface area about the free surface of liquid

= Area of surface × Distance of C.G. from free surface

$= A \times \bar{h}$

$\therefore \quad F = \rho g A \bar{h}$

Where,
ρ = Density of liquid (Kg/m^3)

g = Acceleration due to gravity (m/s^2), A = Area of surface (m^2), $\bar{h}$ = Height of C.G from free surface of liquid (m)

Unit of total pressure: As total pressure is basically a hydrostatic force and therefore total pressure will be measured in terms of N or KN.

Centre of pressure: Centre of pressure is basically defined as a single point through which or at which total pressure or total hydrostatic force will act. Let us consider that we have one tank filled with liquid e.g. water.

Let us consider that there is one object of arbitrary shape immersed inside the water as displayed here in above figure.

Let us consider G is the centre of gravity and P is the centre of pressure. $\bar{h}$ is the height of C.G from free surface of liquid and h* is the height of centre of pressure from free surface of liquid.

Derivation of Centre of Pressure: In order to determine the centre of pressure, we will consider the object in terms of small strips as displayed here in above figure. We will use the concept of "principle of moments" to determine the centre of pressure. According to the principle of moments, moment of the resultant force about an axis will be equal to the sum of the moments of components about the same axis. As we have shown above in figure, total hydrostatic force F is applied at centre of pressure P which is at height of h* from the free surface of liquid. Therefore, let us determine the moment of resultant force F about the free surface of liquid and it will be determined as F x h*. As we have considered here the object in terms of small strips as displayed here in above figure and hence we will determine the moment of force dF acting on small strip about the free surface of liquid.

Moment of force dF = dF x h

Moment of force dF = ρ g h x b dh x h

Let us sum of all moments of such small forces about the free surface of liquid and it will be written as mentioned here.

$$= \int \rho g h \times b \times dh \times h = \rho g \int b \times h \times h dh$$

$$= \rho g \int b h^2 \, dh = \rho g \int h^2 \, dA \qquad (\because \quad b dh = dA)$$

But $$\int h^2 dA = \int b h^2 dh$$

= Moment of Inertia of the surface about free surface of liquid

$= I_0$

∴ Sum of moments about free surface

$$= \rho g I_0$$

$$F \times h^* = \rho g I_0$$

But $$F = \rho g A \bar{h}$$

∴ $$\rho g A \bar{h} \times h^* = \rho g I_0$$

or $$h^* = \frac{\rho g I_0}{\rho g A \bar{h}} = \frac{I_0}{A \bar{h}}$$

By the theorem of parallel axis, we have

$$I_0 = I_G + A \times \bar{h}^2$$

where I_G = Moment of Inertia of area about an axis passing through the C.G. of the area and parallel to the free surface of liquid.

Substituting I_0 in equation (3.4), we get

$$h^* = \frac{I_G + A\bar{h}^2}{A\bar{h}} = \frac{I_G}{A\bar{h}} + \bar{h}$$

$$h^* = \frac{I_G}{A\bar{h}} + \bar{h}$$

7. A) **Types of manometers are:**
   1. Simple manometer
   2. Differential Manometer

Simple Manometer: A simple manometer has a glass tube that's one end is connected to a point where pressure is to be measured and the other end remains open to the atmosphere.

The simple manometer is further classified into four types:

1. Piezometer
2. U-tube manometer
   1. For gauge pressure
   2. For vacuum pressure
3. Single Column Manometer
4. Inclined tube manometer or Sensitive Manometer

2.6.1 piezometer: It is the simplest form of manometer used for measuring guage pressures. one End of this manometer is connected to the point where pressure is to be measured and other End is open to the atmosphere as shown in fig 2.8. The raise of liquid gives the pressure head at the point. Is at a point A. the height of liquid say water is h in piezometer tube, then pressure at A

$$= \rho \times g \times h \ \frac{N}{m^2}.$$

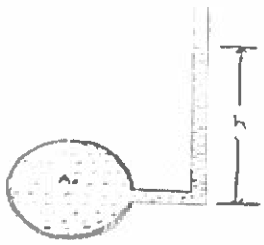


Fig 2.8. piezometer.

2.6.2 U-tube manometer. It consists of glass tube bent in U-shape, one end of which is connected to a point at which pressure is to be measured and other End remains open to the atmosphere as shown in fig 2.9. The tube generally contains mercury or any other liquid whose specific gravity is ~~grater~~ greater than the specific gravity of the liquid whose pressure is to be measured.

a) For guage pressure.

b) For vaccum pressure

Fig 2.9 U-tube manometer.

a) For guage pressure

let B is the point at which pressure is to be measured, whose value is P. The datum line is A-A

let

h_1 = Height of light liquid above the datum line

h_2 = Height of heavy liquid above the datum line

S_1 = Sp. gr. of light liquid.

ρ_1 = Density of light liquid = $1000 \times S_1$

S_2 = Sp. gr. of heavy liquid

ρ_2 = Density of heavy liquid = $1000 \times S_2$

As the pressure in the same for the horizontal surface. Hence pressure above the horizontal datum line A-A in the left column and in the right column of U-tube manometer should be same.

pressure above A-A in the left column = $p + \rho_1 \times g \times h_1$

pressure above A-A in the right column = $\rho_2 \times g \times h_2$

Hence equating the two pressures $p + \rho_1 g h_1 = \rho_2 g h_2$

$\therefore \quad p = (\rho_2 g h_2 - \rho_1 \times g \times h_1) \quad \ldots (1)$

2) For vaccum pressure. For measuring vaccum pressure, the level of the heavy liquid in the manometer will be shown in Fig 2.9(b) Then

Pressure above A-A in the left column = $\rho_2 g h_2 + \rho_1 g h_1 + p$

pressure head in the right column above A-A = 0

$\therefore \quad \rho_2 g h_2 + \rho_1 g h_1 + p = 0$

$\therefore \quad p = -(\rho_2 g h_2 + \rho_1 g h_1) \quad \ldots (2)$

2.6.3 Single column manometer. Single column manometer is a modified form of a U-tube manometer in which a reservoir, having a large cross-sectional area (about 100 times) as compared to the area of the tube is connected to one of the limbs (say left limb) of the manometer as shown in Fig 2.15. Due to large cross sectional area of the reservoir, for any variation in pressure, the change in the liquid level in the reservoir will be very small which may be neglected and hence the pressure is given by the height of liquid in the other limb. The other limb may be vertical or inclined. Thus there are two types of single column manometer as:

1) Vertical single column manometer.
2) Inclined single column manometer.

1) Vertical single column manometer

Fig 2.15 shows the vertical single column manometer. Let X-X be the datum line in the reservoir and in the right limb of the manometer, when it is not connected to the pipe. When the manometer is connected to the pipe due to high pressure at A, the heavy liquid in the reservoir will be pushed downward and will rise in the right limb.

Let Δh = Fall of heavy liquid in reservoir.
h_2 = Raise of heavy liquid in right limb.
h_1 = Height of centre of pipe above X-X
P_A = pressure at A, which is to be measured
A = cross-sectional area of the reservoir
a = cross-sectional area of the right limb
S_1 = Sp. gr. of liquid in pipe
S_2 = Sp. gr. of heavy liquid in reservoir and right limb
ρ_1 = Density of liquid in pipe
ρ_2 = Density of liquid in reservoir.

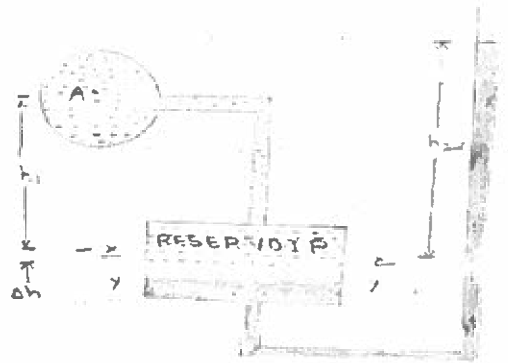


Fig 2.15 Vertical single column manometer.

Fall of heavy liquid in reservoir will cause a rise of heavy liquid level in the right limb

$$\therefore \quad A \times \Delta h = a \times h_2$$

$$\therefore \quad \Delta h = \frac{a \times h_2}{A}$$

Now consider the datum line Y-Y as shown in Fig 2.15. Then pressure in the right limb above Y-Y

$$\rho_2 \times g \times (\Delta h + h_2)$$

Pressure in the left limb above Y-Y $= \rho_1 \times g \times (\Delta h + h_1) + p_A$

Equating these pressures, we have

$$\rho_2 \times g \times (\Delta h + h_2) = \rho_1 \times g \times (\Delta h + h_1) + p_A$$

or

$$p_A = \rho_2 g (\Delta h + h_2) - \rho_1 g (\Delta h + h_1)$$

$$= \Delta h [\rho_2 g - \rho_1 g] + h_2 \rho_2 g - h_1 \rho_1 g$$

But from equation (i), $\Delta h = \frac{a \times h_2}{A}$

$$p_A = \frac{a \times h_2}{A} [\rho_2 g - \rho_1 g] + h_2 \rho_2 g - h_1 \rho_1 g \quad \ldots (2.9)$$

As the area A is very large as compared to a, hence ratio $\frac{a}{A}$ becomes very small and can be neglected

Then $p_A = h_2 \rho_2 g - h_1 \rho_1 g \quad \ldots (2.10)$

From equation (2.10), it is clear that as h_1 is known and hence by knowing h_2 or rise of heavy liquid in the right limb, the pressure at A can be calculated

2. Inclined single column manometer

Fig 2.16 shows the inclined single column manometer. This manometer is more sensitive. Due to inclination the distance moved by the heavy liquid in the right limb will be more.

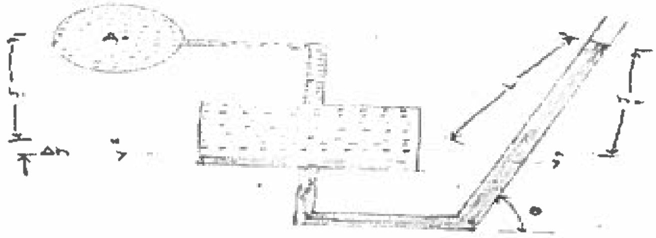


Fig 2.16 Inclined single column manometer

let

L = length of heavy liquid moved in right limb from X-X

θ = Inclination of right limb with horizontal

h_2 = vertical rise of heavy liquid in right limb from X-X $= L \times \sin\theta$

From equation (2.10), the pressure at A is

$$p_A = h_2 \rho_2 g - h_1 \rho_1 g$$

Substituting the value of h_2, we get

$$p_A = \sin\theta \times \rho_2 g - h_1 \rho_1 g \quad (2.11)$$

2.7 Differential manometers

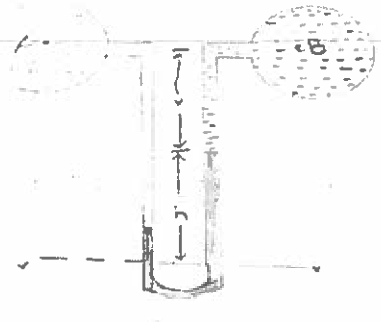
Differential manometers are the devices used for measuring the difference of pressures between two points in a pipe or in two different pipes. A differential manometers consists of a U-tube, containing a heavy liquid, whose two ends are connected to the points, whose difference of pressure is to be measured. most commonly types of differential manometers are:

1. U-tube differential manometer and
2. Inverted U-tube differential manometer.

2.7.1 U-tube differential manometer.

Fig 2.18 shows the differential manometers of ~~[illegible]~~ U-tube type.

(a) Two pipes at different levels



(b) A and B are at the same level.

Fig 2.18 U-tube differential manometers.

In Fig 2.18 (a), the two points A and B are at different level and also contains liquids of different sp.gr. These points are connected to the U-tube differential manometer let the pressure at A and B are P_A and P_B.

let

h = difference of mercury level in the U-tube.

y = distance of the centre of B, from the mercury level in the right limb.

x = distance of the centre of A, from the mercury level in the right limb.

ρ_1 = Density of liquid at A.

ρ_2 = Density of liquid at B.

ρ_g = Density of heavy liquid & mercury.

Taking datum line at X - X

pressure above X - X in the left limb = $\rho_1 g (h+x) + p_A$

where p_A = pressure at A.

pressure above X - X in the right limb = $\rho_g \times g \times h + \rho_2 \times g \times y + p_B$

where p_B = pressure at B.

Equating the two pressure, we have

$$\rho_1 g (h+x) + p_A = \rho_g \times g \times h + \rho_2 g y + p_B$$

$$\therefore \quad p_A - p_B = \rho_g \times g \times h + \rho_2 g y - \rho_1 g (h+x)$$

$$= h \times g (\rho_g - \rho_1) + \rho_2 g y - \rho_1 g x \quad \cdots\cdots (2.12)$$

$\therefore$ Difference of pressure at A and B = $h \times g (\rho_g - \rho_1) + \rho_2 g y - \rho_1$

In Fig 2.18 (b), the two points A and B are at the same level and contains the same liquid of density ρ_1. Then

pressure above X - X in right limb = $\rho_g \times g \times h + \rho_1 \times g \times x + p_B$

Pressure above X - X in left limb = $\rho_1 \times g \times (h+x) + p_A$

Equating the two pressure

$$\rho_g \times g \times h + \rho_1 g x + p_B = \rho_1 \times g \times (h+x) + p_A$$

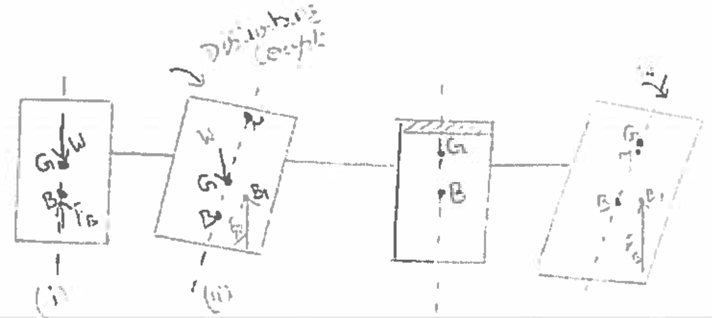
$$\therefore \quad p_A - p_B = \rho_g \times g \times h + \rho_1 g x - \rho_1 g (h+x)$$

$$= g \times h (\rho_g - \rho_1).$$

7.B) Conditions for stability of a floating body:

The stability of a floating body is determined from the position of meta-centre (M). In case of floating body, the weight of the body is equal to the weight of liquid displaced.

a) Stable Equilibrium. If the point M is above G. The floating body will be in stable equilibrium as shown in Fig 4.13 (a). If a slight angular displacement is given to the floating body in the clockwise direction, the centre of buoyancy shifts from B to B_1 such that the vertical line through B_1 cuts at M. Then the buoyant force F_B through B_1 and weight W through G constitute a couple acting in the anti-clockwise direction and thus bringing the floating body in the original position.



a) Stable Equilibrium M is above G

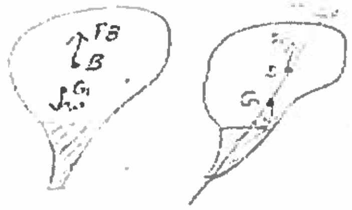
b) unstable Equilibrium M is below G

b) Unstable Equilibrium. If the point M is below G, the floating body will be in unstable equilibrium as shown in Fig 4.13 (b). The disturbing couple is acting in the clockwise direction. The couple due to buoyant force F_B and W is also acting in the clockwise direction and thus overturning the floating body.

c) Neutral Equilibrium. If the point M is at the centre of gravity of the body. The floating body will be in neutral equilibrium.

Conditions for stability of a Submerged body:

The position of centre of gravity and centre of buoyancy in case of a completely sub-merged body are fixed. Consider a balloon, which is completely sub-merged in air. Let the lower portion of the balloon contains heavier material, so that its centre of gravity is lower than its centre of buoyancy as shown in Fig 4.12(a). Let the weight of the balloon is W. The weight W is acting through G, vertically in the downward direction, while the buoyant force F_B is acting vertically up, through B. For the equilibrium of the balloon $W = F_B$. If the balloon is given an angular displacement in the clockwise direction as shown in Fig 4.12(a), then W and F_B constitute a couple acting in the anti-clockwise direction and brings the balloon in the original position. Thus the balloon in the position, shown by fig 4.12(a) is in stable equilibrium.



(a) Stable Equilibrium.

b) Unstable Equilibrium

c) Neutral Equilibrium

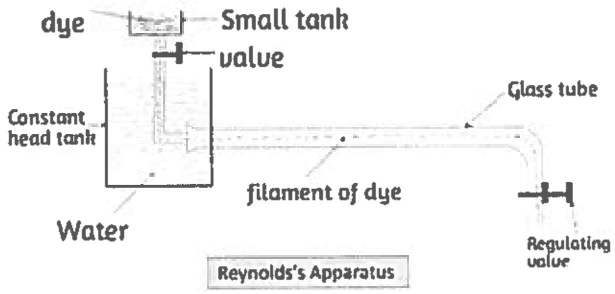
Stabilities of Sub-merged bodies

a) Stable equilibrium. When $W = F_B$ and point B is above G, the body is said to be in stable equilibrium.

b) Unstable equilibrium. If $W = F_B$ but the centre of buoyancy (B) is below centre of gravity (G), the body is in unstable equilibrium as shown in Fig 4.12(b). A slight displacement of the body, in the clockwise direction, gives the couple due to W and F_B also in the clockwise direction. Thus the body does not return to its original position and hence the body is in unstable equilibrium.

c) Neutral equilibrium. If $F_B = W$ and B and G are at the same point, as shown in Fig 4.12(c) the body is said to be in neutral equilibrium.

8.A) Reynolds Experiment: Osborne Reynolds an English Scientist in 1833 who first confirmed the existence of the laminar and turbulent flow experimentally. He performed the experiment where the apparatus consist of (a) a long glass tube which is 1.5m long and 50mm is diameter with bell mouth entrance, (b) a tank(big one) filled with water at constant head(means head will remain constant during the experiment) and (c) a small tank containing dye. The glass tube fitted with the constant heat tank filled with water. the regulating valve is used to control the flow of water in the glass tube. A jet of dye supplied by a small tank. water in the tank is allowed to stand for a long time so that it become completely at rest before conducting the experiment.



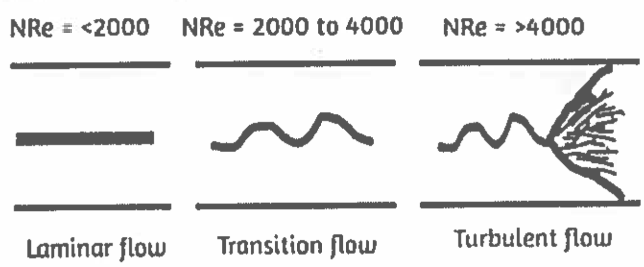
Reynolds's Apparatus

Observations by Reynolds

Liquid dye was introduced into the glass tube when the water was allowed to flow through the glass tube and then following observation observed by the Osborne Reynold's:

1. When the velocity was low, a fine filament of dye was carried by the flowing water in a straight line. The dye filament was moved so steadily that it hardly appeared in the motion. This was the case of **laminar flow.**
2. As the velocity increased, the dye filament becomes wavy and irregular and now it does not maintain the straight line path. This state is known as the **transitional state** and the velocity at this state is called the **lower critical velocity.**
3. If we further increase the velocity then the wavyness and irregularity of the dye filament increases and finally diffused into the water. this type of flow is known as **the turbulent flow** and velocity at this state is known as **the upper critical velocity.**

Types of flow based on Reynold's Experiment



8.B) Assumptions in Bernoulli's Equation: The following are the assumptions made while deriving equation:

1. Flow is ideal
2. Flow is steady
3. Flow is incompressible
4. Flow is irrotational

Euler's Equation of Motion:

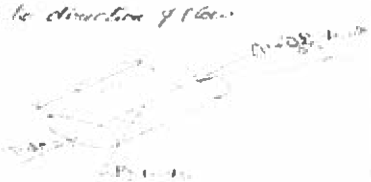
This is the Equation of motion in which the forces due to gravity and pressure are taken into Consideration.

Consider a streamline in which the flow is taking place in S-direction as shown in diagram. Consider a Cylindrical element Section of dA and length ds. Forces acting on the Cylindrical Element are:

a) Pressure pdA in the direction of Flow

b) Pressure force $(p + \frac{\partial p}{\partial s} ds)dA$ opposite to direction of flow

c) Weight of Element $\rho g \, dA \, ds$



Let 'θ' be the angle b/w direction of flow line & action of weight of Element

$$F = ma$$

$$pdA - (p + \frac{\partial p}{\partial s} ds)dA - \rho g \, dA \, ds \cos\theta = \rho \, dA \, ds \times a_s \text{ --- } 5$$

a_s - acceleration in direction of s

$$a_s = \frac{dv}{dt}$$

$$= \frac{dv}{ds}\frac{ds}{dt} + \frac{\partial v}{\partial t}$$

$$= v\frac{\partial v}{\partial s} + \frac{\partial v}{\partial t}$$

For steady flow

$$\frac{\partial v}{\partial t} = 0$$

$$\therefore a_s = v\frac{\partial v}{\partial s}$$

Sub a_s in Eq. 5

$$-\frac{\partial p}{\partial s} ds dA - \rho g \, dA ds \cos\theta = \rho dA \, ds \frac{\partial v}{\partial s}$$

~~Dividing~~ on simplifying

$$-\frac{\partial p}{\partial s} ds \, dA - \rho g \, dA \, ds \cos\theta = \rho \, dA \, ds \, v \frac{\partial v}{\partial s} \text{ --- } 6$$

Eq. 6 Divide $\rho ds dA$ $-\frac{\partial p}{\partial s} ds - g\cos\theta = v\frac{\partial v}{\partial s}$

$$\frac{\partial p}{\partial s} \cancel{\ldots} \rho g$$

$$\frac{\partial p}{\rho \partial s} + g\cos\theta - v\frac{\partial v}{\partial s} = 0$$

From Diagram $\theta = \frac{dz}{ds}$

$$\therefore \frac{1}{\rho}\frac{\partial p}{\partial s} + g\frac{dz}{ds} + \frac{v dv}{ds} = 0$$

$$\boxed{\frac{dp}{\rho} + g dz + VdV = 0}$$

Is known as Euler's Equation of Motion

Bernoulli's Equation from Euler's Equation

Bernoulli's Equation is obtained by integrating the Euler's Eq. of motion

$$\int \frac{dp}{\rho} + \int g dz + \int v dv = \text{Constant}$$

If flow is incompressible, ρ is Constant

$$\frac{p}{\rho} + gz + \frac{v^2}{2} = \text{Constant}$$

$$\frac{p}{\rho g} + \frac{v^2}{2g} + z = \text{Constant}$$

$\frac{p}{\rho g}$ - Pressure Energy per unit weight of fluid or Pressure head

$\frac{v^2}{2g}$ - Kinetic Energy per unit weight or Kinetic head

z - potential Energy per unit weight or Potential head

9.A) Hydraulic gradient line

Hydraulic gradient line is basically defined as the line which will give the sum of pressure head and datum head or potential head of a fluid flowing through a pipe with respect to some reference line.

Hydraulic gradient line = Pressure head + Potential head or datum head

$$H.G.L = P/\rho g + Z$$

Where,

H.G.L = Hydraulic gradient line

$P/\rho g$ = Pressure head

Z = Potential head or datum head

Total Energy Line

Total energy line is basically defined as the line which will give the sum of pressure head, potential head and kinetic head of a fluid flowing through a pipe with respect to some reference line.

Total energy line = Pressure head + Potential head + Kinetic head

$$H.G.L = P/\rho g + Z + V^2/2g$$

Where,

T.E.L = Total energy line

$P/\rho g$ = Pressure head

Z = Potential head or datum head

$V^2/2g$ = Kinetic head or velocity head

9.B)

13.2.1 Laminar Boundary Layer. For defining the boundary layer (*i.e.*, laminar boundary layer or turbulent boundary layer) consider the flow of a fluid, having free-stream velocity (U), over a smooth thin plate which is flat and placed parallel to the direction for free stream of fluid as shown in Fig. 13.2. Let us consider the flow with zero pressure gradient on one side of the plate, which is stationary.

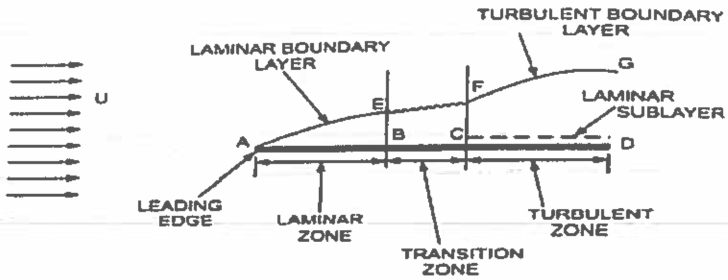


Fig. 13.2 *Flow over a plate.*

The velocity of fluid on the surface of the plate should be equal to the velocity of the plate. But plate is stationary and hence velocity of fluid on the surface of the plate is zero. But at a distance away from the plate, the fluid is having certain velocity. Thus a velocity gradient is set up in the fluid near the surface of the plate. This velocity gradient develops shear resistance, which retards the fluid. Thus the fluid with a uniform free stream velocity (U) is retarded in the vicinity of the solid surface of the plate and the boundary layer region begins at the sharp leading edge. At subsequent points downstream the leading edge, the boundary layer region increases because the retarded fluid is further retarded. This is also referred as the growth of boundary layer. Near the leading edge of the surface of the plate, where the thickness is small, the flow in the boundary layer is laminar though the main flow is turbulent. This layer of the fluid is said to be laminar boundary layer. This is shown by AE in Fig. 13.2. The length of the plate from the leading edge, upto which laminar boundary layer exists, is called laminar zone. This is shown by distance AB. The distance of B from leading edge is obtained from Reynold number equal to 5×10^5 for a plate. Because upto this Reynold number the boundary layer is laminar. The Reynold number is given by $(R_e)_x = \dfrac{U \times x}{\nu}$

where x = Distance from leading edge,

U = Free-stream velocity of fluid,

ν = Kinematic viscosity of fluid.

Hence for laminar boundary layer, we have $5 \times 10^5 = \frac{U \times x}{\nu}$ ...(13.1)

If the values of U and ν are known, x or the distance from the leading edge upto which laminar boundary layer exists can be calculated.

13.2.2 Turbulent Boundary Layer. If the length of the plate is more than the distance x, calculated from equation (13.1), the thickness of boundary layer will go on increasing in the down-stream direction. Then the laminar boundary layer becomes unstable and motion of fluid within it, is disturbed and irregular which leads to a transition from laminar to turbulent boundary layer. This short length over which the boundary layer flow changes from laminar to turbulent is called transition zone. This is shown by distance BC in Fig. 13.2. Further downstream the transition zone, the boundary layer is turbulent and continues to grow in thickness. This layer of boundary is called turbulent boundary layer, which is shown by the portion FG in Fig. 13.2.

13.2.3 Laminar Sub-layer. This is the region in the turbulent boundary layer zone, adjacent to the solid surface of the plate as shown in Fig. 13.2. In this zone, the velocity variation is influenced only by viscous effects. Though the velocity distribution would be a parabolic curve in the laminar sub-layer zone, but in view of the very small thickness we can reasonably assume that velocity variation is linear and so the velocity gradient can be considered constant. Therefore, the shear stress in the laminar sub-layer would be constant and equal to the boundary shear stress τ_0. Thus the shear stress in the sub-layer is

$$\tau_0 = \mu \left(\frac{\partial u}{\partial y}\right)_{y=0} = \mu \frac{u}{y} \qquad \left\{\because \text{ For linear variation, } \frac{\partial u}{\partial y} = \frac{u}{y}\right\}$$

10.A) The dynamic similarity is said to exist between model and prototype, if the ratios of corresponding forces acting at the corresponding points are the same. We must note it here that the direction of forces at the corresponding points in the model and prototype must be same.

$$\boxed{\frac{F_p}{F_m} = F_r}$$

where F_r is Force Ratio

F_m = Force at a point in model, F_P = Force at respective point in prototype

1. Reynolds Law:

Reynolds number is the ratio of inertia force to the viscous force. It describes the predominance of inertia

$$R_e = \frac{\rho . v . d}{\mu}$$

forces to the viscous forces occurring in the flow systems.

Where,

ρ = Density of fluid (kg/m³)

μ = viscosity of fluid (kg/m.s)

d = diameter of pipe (m)

v = velocity of flow (m/s)

Importance

Reynolds number is applicable for closed surface flows as well as for free surface flows. Some applications where Reynolds number is significant for finding the flow behavior are incompressible flow through small pipes, the motion of a submarine completely under water, flow through low-speed turbomachines, etc.

2. Froude Law

Froude number is the ratio of inertia force to the gravitational force. Froude number is significant in case of free surface flows where the gravitational force is predominant compared to other forces.

$$F_r = \frac{v}{\sqrt{g . L}}$$

Where,

L = length of flow (m)
v = velocity of flow (m/s)
g = acceleration due to gravity (m/s^2)
Importance
Froude number is useful to describe the flow in open channels, flow over notches and weirs, the motion of a ship in turbulent sea conditions (ship resistance), flow over spillways, etc.

10.B. **Buckingham's π – theorem:** a. List all the '*n*' physical quantities or variables involved in the phenomenon. Note their dimensions and the number '*m*' of the fundamental dimensions comprised in them. So that there will be (*n-m*). π terms.
b. Select '*m*' variables out of these to serve as repeating variables with the following guidelines:
i. These variables should be such that none of them is dimensionless.
ii. No two variables have the same dimensions.
iii. They themselves do not form a dimensionless parameter.
iv. The entire '*m*' fundamental are included collectively in them.
v. The dependent variable should not be taken as repeating variable.
c. Write the general equations for different π terms. These may be expressed as the product of the repeating variables each raised to an unknown exponent and one of the remaining variables taken in turn, with a known power (usually taken as one).
d. Write the dimensional equations of the equations of the π terms obtained in the step (c) above.
Compute the value of the unknown exponents by equating the exponents of the respective fundamental dimensions on both the sides of each of the dimensional equations. Thereby different dimensional groups or π terms are formed.
e. Write the final general equation for the phenomenon in terms of the π terms.
In order to obtain the final expression the following additional may be considered
i. If the quantity is dimensionless, it is a π term with out going through the above procedure.
ii. If any two physical quantities have the same dimensions, their ratio will be one of the π term. For example (*H/d*) is dimensionless and hence it is a π term.
iii. Any π term may be replaced by any power of that term, including negative as well as fractional powers. For example, π_1 may be replaced by π^{-1}, or π_2 may be replaced by π_2^2, or π_3 may be replaced by $\frac{1}{\pi_3^{\frac{1}{2}}}$ etc.
iv. Any π term may be replaced by multiplying it by numerical constant. For example, π_1 may be replaced by $3\pi_1$ or so.
v. Any π term may be replaced by another π term obtained by adding or subtracting an absolute numerical from it.
vi. Any π term may be replaced by multiplying it by another π term. For example, π_1 may be replaced by (π_1 x π_2).
Mathematically, if any variable Q_1 depends on the independent variables Q_2, Q_3, Q_4......Q_n, ; the fundamental equation may be written as,

$$Q_1 = f\,(Q_2,\ Q_3,\ Q_4 \ldots\ldots Q_n)$$

which can be transformed to another functional relationship as,

$$f_1\,(Q_1,\ Q_2,\ Q_3,\ Q_4 \ldots\ldots Q_n) = C$$

where 'C" is the dimensionless constant.
In accordance with the π theorem, a non dimensional equation can thus be obtained in the form,

$$f_2\,(\pi_1,\ \pi_2,\ \pi_3, \ldots\ldots\ldots \pi_{n-m}) \quad = \quad C_1$$

wherein, each dimensionless π-term is formed by combining *m* variables out of the total *n* variables with one of the remaining (*n-m*) variables. These '*m*' variables which appear repeatedly in each of the π terms, are called repeating variables. These are '*m*' fundamentals quantities. They themselves do not form a dimensionless parameter. Thus the different π terms may be established as,

$$\pi_1 \quad = \quad Q_1^{a1},\ Q_2^{b1},\ Q_3^{c1} \ldots\ldots Q_m^{m1}\ Q_{m+1}$$

$$\pi_2 = Q_1^{a2}, Q_2^{b2}, Q_3^{c2} ..... Q_m^{m3} Q_{m+2}$$

....

$$\pi_{n-m} = Q_1^{an-m}, Q_2^{b\, n-m}, Q_3^{c\, n-m} ..... Q_m^{m\, n-m}, Q_n$$

In the above equation, each individual equation is dimensionless and the exponents *a, b, c, d* .... *m* etc are determined by considering dimensional homogeneity for each equation such away that each π term is dimensionless.

The final general equation for the phenomenon may then be obtained by expressing any one of the π terms as a function of the others.

$$\pi_1 = f_1 (\pi_1, \pi_2, \pi_3, ...\pi_{n-m})$$

$$\pi_2 = f_2 (\pi_1, \pi_2, \pi_3, ...\pi_{n-m})$$

or any other desired relationship may be obtained.

11.A) **Buckingham Pi Theorem:** If there are n variables in a problem and these variables contain m primary dimensions (for example M, L, T) the equation relating all the variables will have (n-m) dimensionless groups. Buckingham referred to these groups as π groups. The final equation obtained is in the form of : πl = f(π2, π3 ,..... πn-m) The π groups must be independent of each other and no one group should be formed by multiplying together powers of other groups. This method offers the advantage of being more simple than the method of solving simultaneous equations for obtaining the values of the indices (the exponent values of the variables).

Advantages of Pi theorem over the Rayleigh's method for dimension analysis: In the Rayleigh's method of dimensional analysis, solution becomes more and more cumbersome and laborious if number of influencing variables become more than the fundamental units (M, L, T and θ) involved in the physical phenomenon.. The use of Buckingham's π-theorem method enables to overcome this limitation and states that if there are 'n' variables (independent and dependent) in a physical phenomenon and if these variables contain 'm' number of fundamental dimensions (M, L, T and θ), then the variables are arranged in to (n-m) dimensionless terms called π-terms.

11.B) Dimensionless Numbers: A number representing a property of a physical system, but not measured on a scale of physical units (as of time, mass, or distance). Drag coefficients and stress, for example, are measured as dimensionless numbers.

Some important dimensionless numbers used in fluid mechanics and their importance is explained below.

1. Reynolds Number
2. Froude Number
3. Weber Number
4. Mach Number
5. Euler's Number

1. Reynolds number

Reynolds number is the ratio of inertia force to the viscous force. It describes the predominance of inertia forces to the viscous forces occurring in the flow systems.

$$R_e = \frac{\rho . v . d}{\mu}$$

Where,

ρ = Density of fluid (kg/m³)

μ = viscosity of fluid (kg/m.s)

d = diameter of pipe (m)

v = velocity of flow (m/s)

Importance

Reynolds number is applicable for closed surface flows as well as for free surface flows. Some applications where Reynolds number is significant for finding the flow behavior are incompressible flow through small pipes, the motion of a submarine completely under water, flow through low-speed turbomachines, etc.

2. Froude number

Froude number is the ratio of inertia force to the gravitational force. Froude number is significant in case of free surface flows where the gravitational force is predominant compared to other forces.

$$F_r = \frac{v}{\sqrt{g.L}}$$

Where,

L = length of flow (m)

v = velocity of flow (m/s)

g = acceleration due to gravity (m/s^2)

Importance

Froude number is useful to describe the flow in open channels, flow over notches and weirs, the motion of a ship in turbulent sea conditions (ship resistance), flow over spillways, etc.

3. Weber number

Weber number is the ratio of inertia force to the surface tension. The formation of droplets or water bubbles in a fluid is normally due to surface tension. If Weber number is small, surface tension is larger and vice versa.

$$W_e = \frac{\rho.d.v^2}{\sigma}$$

Where,

ρ = Density of fluid (kg/m^3)

σ = Surface tension of fluid (N/m)

d = diameter of water droplet (m)

v = velocity of flow (m/s)

Applications

Weber number is less than 1 when surface tension is predominant. It happens when the curvature of the liquid surface is small compared to its depth. This can be seen in different situations such as the flow of blood in veins and arteries, atomization of liquids, capillary flow of water in soils, thin layers of fluid passing over surface, etc.

4. Mach number

Mach number is the ratio of inertia force to the elastic force. If the Mach number is one, then the flow velocity is equal to the velocity of sound in the fluid. If it is less than one, then the flow is called subsonic flow, and if it is greater than one the flow is called supersonic flow.

$$M_a = \frac{v}{c}$$

Where,

v = Velocity of flow (m/s)

c = Velocity of sound in fluid (m/s)

Applications

Mach number is useful to describe problems in high flow velocities. It is also used in aerodynamics to describe the speed of jet plane or missile in terms of speed of sound.

5. Euler's number

Euler number is the ratio of pressure force to the inertia force.

$$E_u = \frac{F}{\rho . v^2 . L^2}$$

Where,

F = pressure force

ρ = Density of fluid (kg/m³)

L = Characteristic length of flow (m)

v = velocity of flow (m/s)

Applications

Euler's number is significant in cases where pressure gradient exists such as flow through pipes, water hammer pressure in penstocks, discharge through orifices and mouthpieces, etc.

12.A)

Given.

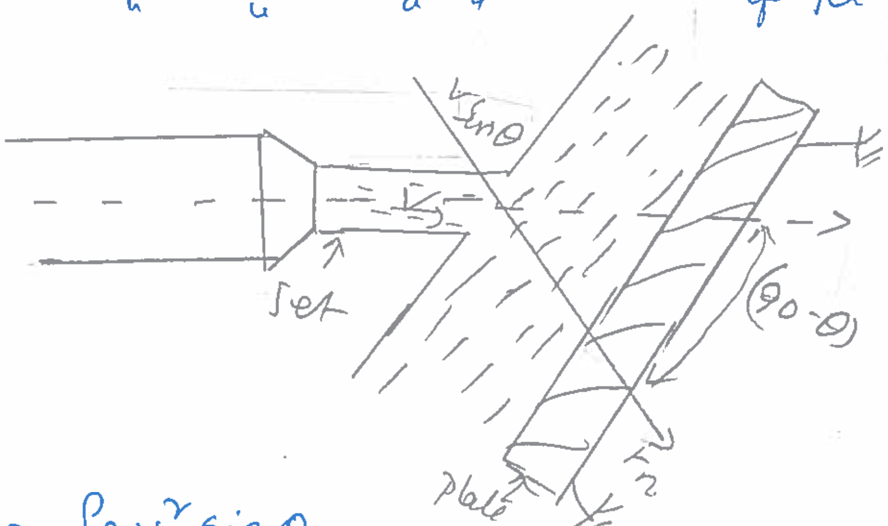
Water diameter = 75 mm.

Velocity of jet (V) = 25 m/s

$\theta = 60^\circ$.

Force exerted by jet in direction normal to plate = ? (F_n)

" " " " " of jet = ? = F_x)



$F_n = \rho a v^2 \sin\theta$

$= 1000 \times \frac{\pi}{4}(75\times10^{-3})^2 \times (25)^2 \sin 60$

$= 2,391.2\,N$

$F_x = \rho a v^2 \sin^2\theta = F_n \sin\theta$

$= 2391.2 \times \sin 60$

$= 2070.87\,N.$

12) b) Given.

Internal diameter of Impeller $D_1 = 0.2m$

External diameter of Impeller $D_2 = 0.4m$

Speed $N = 1200$ rpm

Vane angle of impeller at inlet $\theta = 20^0$

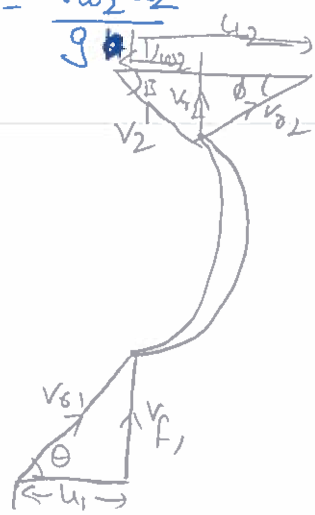
" " " " outlet $\phi = 30^0$

W.D / wt of water = ?

$u_1 = \frac{\pi D_1 N}{60} = \frac{\pi \times 0.2 \times 1200}{60} = 12.5 \ m/s$

$u_2 = \frac{\pi \times 0.4 \times 1200}{60} = 25.13 \ m/s$

W.D by Impeller / wt of water $= \frac{V_{w_2} u_2}{g}$



$\tan\theta = \frac{V_{f_1}}{u_1}$

$\tan 20 = \frac{V_{f_1}}{12.5}$

$V_{f_1} = 4.5 \ m/s = V_{f_2}$

$\tan\phi = \frac{V_{f_2}}{u_2 - V_{w_2}}$

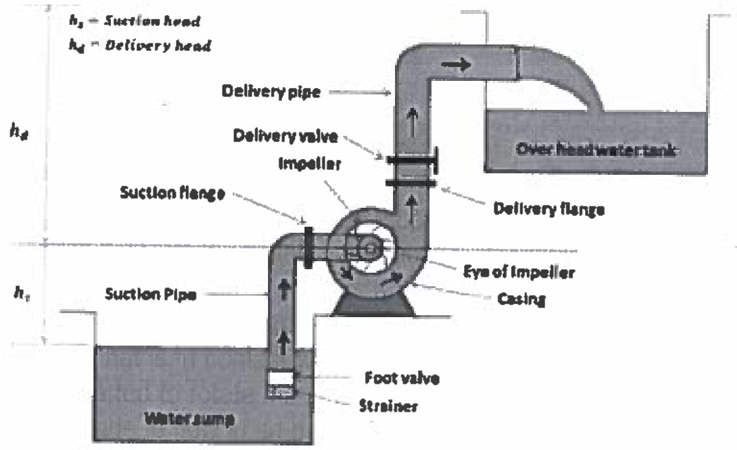
$\tan 30 = \frac{4.5}{25.13 - V_{w_2}}$

$V_{w_2} = 17.3 \ m/s$

W.D / wt $= \frac{17.3 \times 25.13}{9.81}$

$= 44.3 \ N$

13.A) Working of Centrifugal Pump: The first step in the operation of a centrifugal pump is priming. Priming is the operation in which suction pipe casing of the pump and the position of fluid with the liquid which is to be pumped so that all the air from the position of pump is driven out and no air is left. The necessity of priming of a centrifugal pump is due to the fact that the pressure generated at the centrifugal pump impeller is directly proportional to density of fluid that is in contact with it. After the pump is primed the delivery valve is still kept closed and electric motor is started to rotate the impeller. The delivery valve is kept closed in order to reduce valve is opened the liquid is made to flow in an outward radial direction there by vanes of impeller at the outer circumference with high velocity at outer circumference due to centrifugal action vacuum is created. This cause liquid from sump to rush through suction pipe to eye of impeller thereby replacing long discharge from center circumference of the impeller is utilized in lifting liquid to required height through delivery pipe.



Centrifugal Pump Working

13.B)

Given.

Outer dia of centrifugal pump = 45 Cm

inner dia of " " = 25 Cm

Head = 20 M.

Min. starting speed of pump = ?

For Minimum starting speed $\frac{u_2^2 - u_1^2}{2g} \geq H_m$

i e $\frac{1}{2g}\left[\frac{\pi^2 D_2^2 N^2}{(60)^2} - \frac{\pi^2 D_1^2 N^2}{(60)^2}\right] H_m$

$\Rightarrow \frac{1}{2 \times 9.81} \times \frac{\pi^2 \times N^2}{(60)^2}\left[(45 \times 10^{-2})^2 - (25 \times 10^{-2})^2\right] = 20$

$\Rightarrow N^2 = 1022359.6$

$N = 1011$ rpm.

14.A) Performance Curves of Turbines: Characteristic curves of hydraulic turbines are are the curves with the help of which the exact behavior and performance of the turbine under different working conditions can be known. These are plotted from the results of the tests performed. The important parameters which are varied during a test are -1.speed 2.head 3.discharge 4.power 5.overall efficiency 6.Gate opening

There are 3 main and very important curves
1. Constant Head Curve
2. Constant Speed Curve
3. Constant Efficiency Curve.

1. Constant Head Curve: In this head and gate opening is kept constant. Thus for every value of speed we get corresponding values of power and discharge. Thus overall efficiency can be calculated.

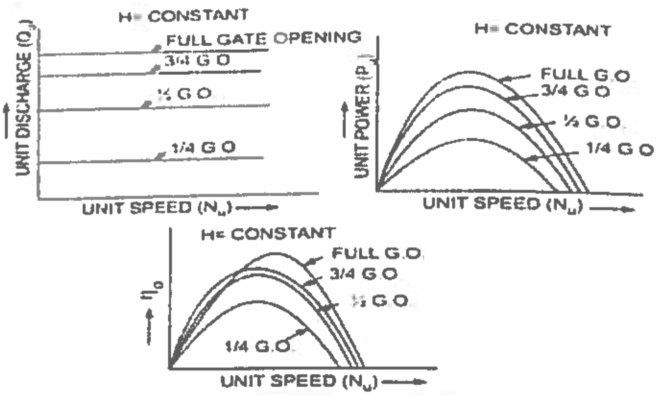
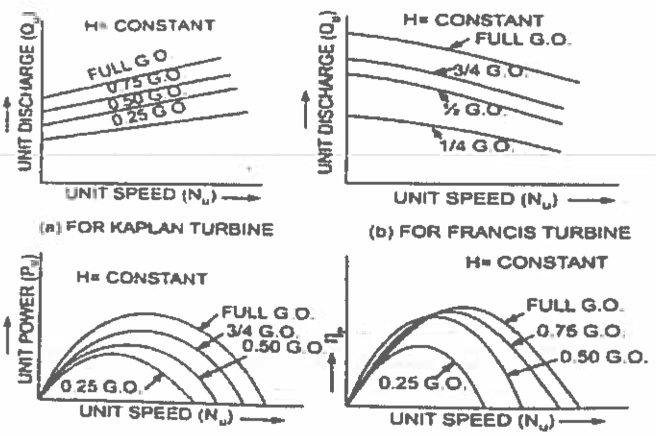


Fig. 18.35 *Main characteristic curves for a Pelton wheel.*



2. Constant Speed Curve: In this speed and head is kept constant. Thus we can find out variation in power and efficiency with respect to discharge. It also helps tells us about the minimum discharge needed to overcome the friction

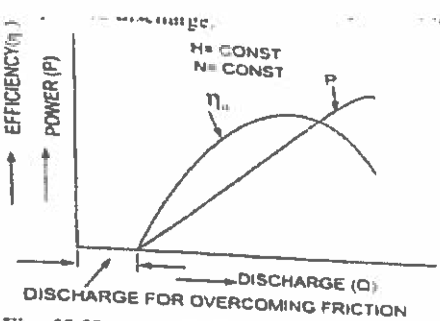


Fig. 18.37 *Operating characteristic curves.*

3. Constant Efficiency Curve: Obtained at different gate openings. Thus for a given efficiency there are two values of discharge and speeds. If efficiency is maximum then we get only one value.
They are helpful in determining the zone of constant efficiency and for prediction of performance of turbines at different efficiencies

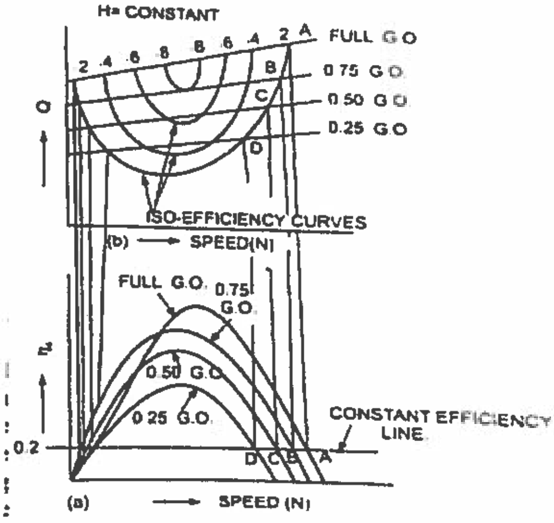


Fig. 18.38 *Constant efficiency curve.*

14) B) Given

Pelton wheel

Bucket speed $(u) = 35 m/s = u_1 = u_2$

$Q = 1 m^3/s$

$H = 270 mt$

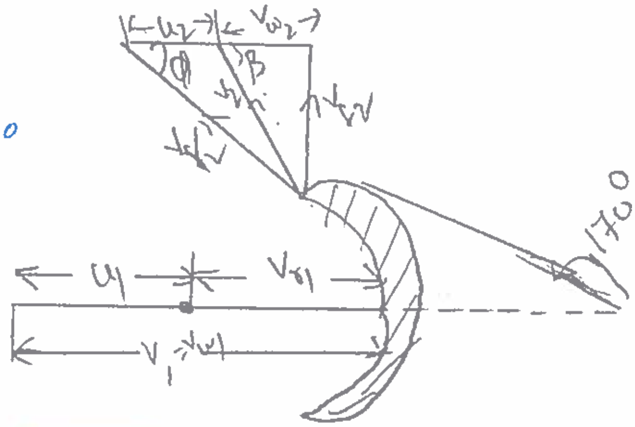
$\phi = 170^0$

i.e $\phi = 180 - 170 = 10^0$

Power = ?

$\eta_H = ?$

$C_v = 0.98$



$V_1 = C_v \sqrt{2gH} = 0.98 \sqrt{2 \times 9.81 \times 270}$

$= 71.3 m/s = V_{w1}$

$V_{r1} = V_1 - u_1 = 71.3 - 35$

$= 36.3 m/s$

$V_{r2} = V_{r1}$

$V_{w2} = V_{r2} \cos\phi - u_2$

$= 36.3 \cos 10 - 35$

$= 0.77 m/s$

$$\text{Power W.D/sec} = \rho a v_1 [V_{w_1} + V_{w_2}] u.$$

$$= 1000 \times 1 [71.3 + 0.77] \times 35$$

$$= 2,52,3500 \text{ N-m/s}$$

$$\text{Hydraulic Efficiency } (\eta_H) = \frac{2(V_{w_1} + V_{w_2}) u}{V_1^2}$$

$$= \frac{2(71.3 + 0.77) \times 35}{(71.3)^2}$$

$$\simeq 0.75$$

$$= 75\%$$

15) A) Given

Pelton wheel

Power developed $= 2$ MW

Gross head $= 360$ mt

Speed $= 560$ rpm

$L = 1200$ mt

$C_v = 0.98$ $f = 0.03$

$\eta_H = 85\%$

head lost in penstock $= 12$ m.

Quantity of water supplied $= ?$

Dia of nozzle $= ?$

" " Penstock $= ?$

Let V^* = velocity of water in penstock

V_1 = " " Jet of water

~~Area of Penstock $\times V^*$ = Area of Jet $\times V_1$~~

$$\cancel{\frac{\pi}{4} D^2 \times V^* = \frac{\pi}{4} d^2 \times V_1}$$

$$\cancel{V^* = \frac{d^2}{D^2} \times V_1} \Rightarrow \cancel{V^* = 150 \times \frac{d^2}{D^2}}$$

$$\cancel{V_1 = C_v \sqrt{2gH} = 0.98 \times \sqrt{2 \times 9.81 \times 1200}}$$

$$\cancel{= 150 \text{ m/s}}$$

Net Head $= H_g - h_f \left(V^2/2g\right)$

$= 360 - \left(4 \times \cancel{0.03}^{12} \times \cancel{1200} \times V^2\right)$

$= 348\,mt$

$\frac{V^2}{2g} = 12$

$V^2 = 12 \times 2 \times 9.81$

$V_1 = 15.3\,m/s.$

$u_1 = K_u\sqrt{2gH} = 0.45\sqrt{2 \times 9.81 \times 348}$

$= 37.18\,m/s$

$V_1 = C_v\sqrt{2gH} = 0.98\sqrt{2 \times 9.81 \times 348}$

$= 80.9\,m/s$

$V_{r1} = V_1 - u_1 = 43.7\,m/s = V_{r2}$

$\eta_H = \frac{2\,(V_{w1} + V_{w2})\,u}{V_1^2}$

$\Rightarrow 0.85 = \frac{2\,(80.9 + V_{w2})\,37}{(80.9)^2}$

$V_{w2} = 5.7\,m/s$

Runner Power $= (V_{w1} + V_{w2})\,u$

$= (80.9 + 5.7)\,37$

$= 3204.2\,W$

$\boxed{\cancel{\eta_o = \frac{S.P}{W.P} = \frac{2 \times 10^6}{3204.2}}}$

Assume $\eta_o = 85\%$

$\eta_o = \frac{S.P}{W.P} \Rightarrow 0.85 = \frac{2 \times 10^6}{W.P}$

$W.P = 2352941.1\,W$

$W.P = \rho g Q H$

$2352941.1 = 1000 \times 9.81 \times Q \times 348$

$Q = 6.89\,m^3/s$

$Q = A \times V$

$6.8 = A \times 80.9$

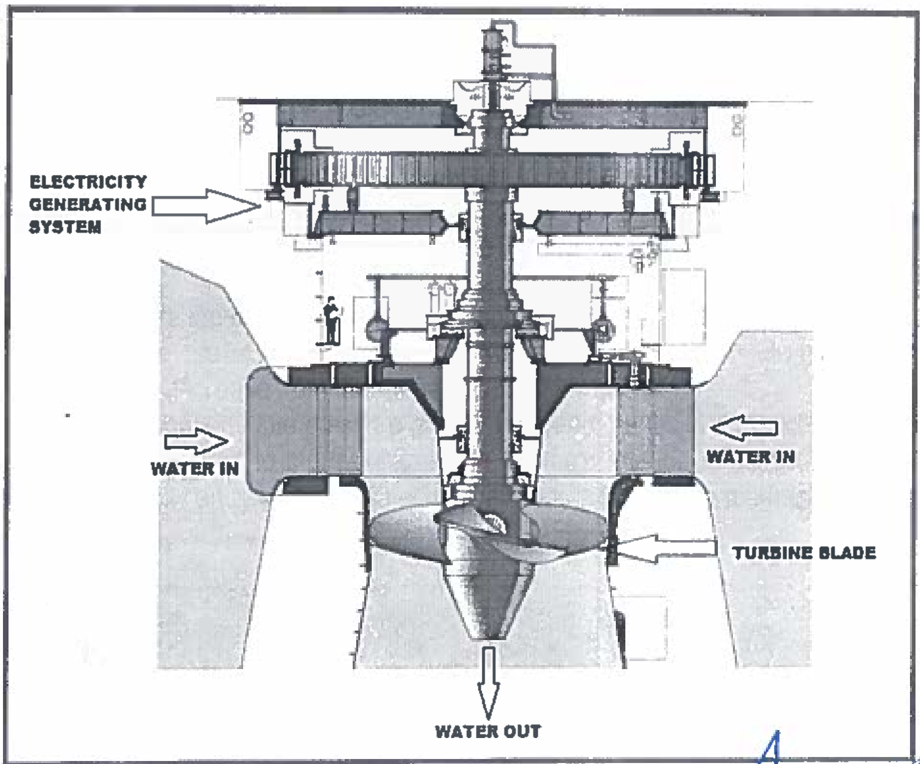
$6.8 = \frac{\pi}{4} d^2 \times 80.9$

$d = 0.329$ mt

15.B) Working principle of Kaplan Turbine: This turbine is one sort of axial flow reaction turbines. In this way, the working fluid, which is often water, changes the pressure while moving within the turbine and produces energy. The power is the combination of both the kinetic energy of the flowing water and hydrostatic head In order to understand the way it produces power, the water from the pen-stock enters the scroll casing. After that, the water passes through the scroll casing, and the guide vanes guide the water from the casing to the runner's blades. The critical point is that the vanes are flexible and can adjust themselves based on the required flow rate.

While the water moves across the blades, it begins rotating because of the water's reaction force. Besides, the blades in the Kaplan turbine are also adjustable. The water goes through the draft tube, where its kinetic energy and pressure energy decrease from the runner blades.

Actually, the kinetic energy converts into pressure energy here and results in enhanced pressure of the water. Eventually, the water is discharged to the trail race. The turbine's rotation is utilized to rotate the shaft of a generator to produce electricity and some extra mechanical work. Below is the diagram of Kaplan Turbine



21/6/22

21/06/2022

NSRIT

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	**Program**	EEE			**Academic Year**	2021 - 2022
Course Code	20EE404	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	IV
Course	INDUCTIONS MOTORS AND SYNCHRONOUS MACHINES						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Why induction motor called as asynchronous motor?	20EE404.1	L1
2	Identify the need of starter for induction motor	20EE404.2	L1
3	Identify the role of centrifugal switches provided in many single-phase induction motors	20EE404.3	L1
4	Define the term voltage regulation in alternator	20EE404.4	L1
5	Enlist the applications of synchronous condenser	20EE404.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain the construction and working principle of 3 phase induction motor	6M	20EE404.1	L2
6 (b)	Explain in detail the equivalent circuit of 3 phase induction motor.	6M	20EE404.1	L2
	OR			
7	Derive an expression for the torque of an induction motor and torque-slip characteristics and obtain the condition for maximum torque.	12M	20EE404.1	L2
8	Describe the following: (i) Rotor resistance starter for starting slip ring induction motor. (ii) Speed control of an induction motor by changing the frequency and poles	12M	20EE404.2	L2
	OR			
9 (a)	With a neat diagram, Discuss about the slip power recovery scheme of induction motor	6M	20EE404.2	L3
9 (b)	Compare the relative merits and demerits of stator resistance starter	6M	20EE404.2	L2
10 (a)	Explain the principle of operation of single-phase induction motor based on “double revolving field theory”.	6M	20EE404.3	L2
10 (b)	Explain in detail the operation of capacitor start and run induction motor	6M	20EE404.3	L3
	OR			
11 (a)	The equivalent impedance of the main and auxiliary winding in a capacitor motor are $(15 + j\,25)\Omega$ and $(50 + j120)\Omega$ respectively, while the capacitance of the capacitor is $12\ \mu F$. Estimate the line current at starting a 230 V, 50Hz supply.	8M	20EE404.3	L2
11 (b)	Identify the features of no load and blocked rotor test	4M	20EE404.3	L2
12 (a)	Demonstrate the POTIER method of determining the regulation of an alternator.	8M	20EE404.4	L1
12 (b)	Compare the Constructional details of rotor of both non-salient and salient pole synchronous machine.	4M	20EE404.4	L2
	OR			
13 (a)	Describe the role of voltage regulation in alternator. Also explain synchronous impedance method for determining regulation of an alternator.	8M	20EE404.4	L2
13 (b)	Identify the features of synchronizing of alternator. Describe any one method of synchronizing	4M	20EE404.4	L1

14 (a)	Explain the principle of operation of a 3-phase synchronous motor with neat sketch	6M	20EE404.5	L2
14 (b)	Derive an expression for the power developed in an synchronous motor.	6M	20EE404.5	L3
	OR			
15 (a)	Explain any two starting methods of synchronous motor in detail.	8M	20EE404.5	L2
15 (b)	Illustrate the performance of a synchronous motor using V and inverted V curves.	4M	20EE404.5	L2

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

SUBJECT: INDUCTION MOTORS & SYNCHRONOUS MACHINES

IV SEMESTER END EXAMINATION DATE:20-06-2022

1.Why induction motor called as asynchronous motor? (2M)

An induction motor is called asynchronous motor because the actual speed of the motor is not equal to the synchronous speed of the motor. The synchronous speed of the motor is always more than the actual speed of the motor. If the actual speed of the motor(N) is equal to the synchronous speed (Ns), then no torque will be produced and motoring function not possible.

2.Identify the need of starter for induction motor (2M)

An induction motor is similar to a poly-phase transformer whose secondary is short circuited. Thus, at normal supply voltage, like in transformers, the initial current taken by the primary is very large for a short while. Unlike in DC motors, large current at starting is due to the absence of back emf. If an induction motor is directly switched on from the supply, it takes 5 to 7 times its full load current and develops a torque which is only 1.5 to 2.5 times the full load torque. This large starting current produces a large voltage drop in the line, which may affect the operation of other devices connected to the same line. Hence, it is not advisable to start induction motors of higher ratings (generally above 25kW) directly from the mains supply.

3.Identify the role of centrifugal switches provided in many single-phase induction motors (2M)

The centrifugal switch turns on a circuit, providing the needed boost to start the motor. Once the motor comes up to its operating speed, the switch turns off (disconnects auxiliary winding) the boost circuit, and the motor runs normally

4. Define the term voltage regulation in alternator **(2M)**

The voltage regulation of an alternator or synchronous generator is defined as the rise in the terminal voltage when the load is decreased from full-load rated value to zero. The speed and field current of the alternator remain constant.

5. Enlist the applications of synchronous condenser **(2M)**

1. Power factor improvement
2. Use for leading power
3. Use for lagging power factor
4. Use for unity power factor
5. For synchronous speed
6. Less than synchronous speed
7. More than synchronous speed

6(a). Explain the construction and working principle of 3 phase induction motor (6M)

Working of Three Phase Induction Motor

Production of Rotating Magnetic Field

The stator of the motor consists of overlapping winding offset by an electrical angle of 120o. When we connect the primary winding, or the stator to a 3 phase AC source, it establishes rotating magnetic field which rotates at the synchronous speed.

Secrets Behind the Rotation:

According to Faraday's law an emf induced in any circuit is due to the rate of change of magnetic flux linkage through the circuit. As the rotor winding in an induction motor are either closed through an external resistance or directly shorted by end ring, and cut the stator rotating magnetic field, an emf is induced in the rotor copper bar and due to this emf a current flows through the rotor conductor.
Here the relative speed between the rotating flux and static rotor conductor is the cause of current generation; hence as per Lenz's law, the rotor will rotate in the same direction to reduce the cause, i.e., the relative velocity.

3 Phase Induction Motor Construction

Like any other type of electrical motor induction motor, a 3 phase induction motor is constructed from two main parts, namely the rotor and stator:

1. Stator: As its name indicates stator is a stationary part of induction motor. A stator winding is placed in the stator of induction motor and the three phase supply is given to it.
2. Rotor: The rotor is a rotating part of induction motor. The rotor is connected to the mechanical load through the shaft

The rotor of the three phase induction motor are further classified as

1. Squirrel cage rotor
2. Slip ring rotor or wound rotor or phase wound rotor.

Depending upon the type of rotor construction used the three phase induction motor are classified as:

1. Squirrel cage induction motor
2. Slip ring induction motor or wound induction motor or phase wound induction motor.

The other parts of a 3 phase induction motor are:

1. Shaft for transmitting the torque to the load. This shaft is made up of steel.
2. Bearings for supporting the rotating shaft.
3. One of the problems with electrical motor is the production of heat during its rotation. To overcome this problem, we need a fan for cooling.
4. For receiving external electrical connection Terminal box is needed.
5. There is a small distance between rotor and stator which usually varies from 0.4 mm to 4 mm. Such a distance is called air gap.

Stator of Three Phase Induction Motor

The stator of the three-phase induction motor consists of three main parts :

1. Stator frame,
2. Stator core,
3. Stator winding or field winding.

6(b). Explain in detail the equivalent circuit of 3 phase induction motor
(6M)

Equivalent Circuit of Three Phase Induction Motor

Fig. 3.10 (i) shows the equivalent circuit per phase of the rotor at slip s. The rotor phase current is given by;

$$I'_2 = \frac{s E_2}{\sqrt{R_2^2 + (s X_2)^2}}$$

Mathematically, this value is unaltered by writing it as:

$$I'_2 = \frac{E_2}{\sqrt{(R_2 / s)^2 + (X_2)^2}}$$

As shown in Fig. 3.10 (ii), we now have a rotor circuit that has a fixed reactance X2 connected in series with a variable resistance R2/s and supplied with constant voltage E2. Note that Fig. 3.10 (ii) transfers the variable to the resistance without altering power or power factor conditions.

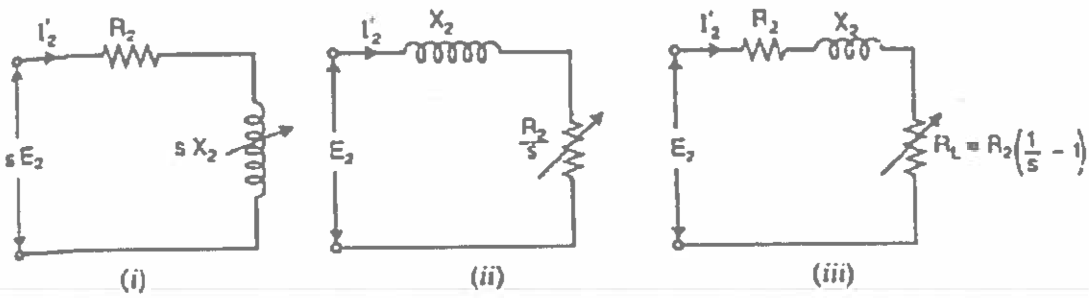


Fig: 3.10

The quantity R_2/s is greater than R_2 since s is a fraction. Therefore, R_2/s can be divided into a fixed part R_2 and a variable part $(R_2/s - R_2)$ i.e.,

$$\frac{R_2}{s} = R_2 + R_2\left(\frac{1}{s} - 1\right)$$

(i) The first part R_2 is the rotor resistance/phase, and represents the rotor Cu loss.

(ii) The second part $R_2\left(\frac{1}{s} - 1\right)$ is a variable-resistance load. The power delivered to this load represents the total mechanical power developed in the rotor. Thus mechanical load on the induction motor can be replaced by a variable-resistance load of value $R_2\left(\frac{1}{s} - 1\right)$. This is

$$\therefore \quad R_L = R_2\left(\frac{1}{s} - 1\right)$$

Fig. 3.10 (iii) shows the equivalent rotor circuit along with load resistance RL.

Now Fig: 3.11 shows the equivalent circuit per phase of a 3-phase induction motor. Note that mechanical load on the motor has been replaced by an equivalent electrical resistance RL given by;

$$R_L = R_2\left(\frac{1}{s} - 1\right) \qquad \text{—(i)}$$

The circuit shown in Fig. 3.11 is similar to the equivalent circuit of a transformer with secondary load equal to R2 given by eq. (i). The rotor e.m.f. in the equivalent circuit now depends only on the transformation ratio K (= E2/E1).

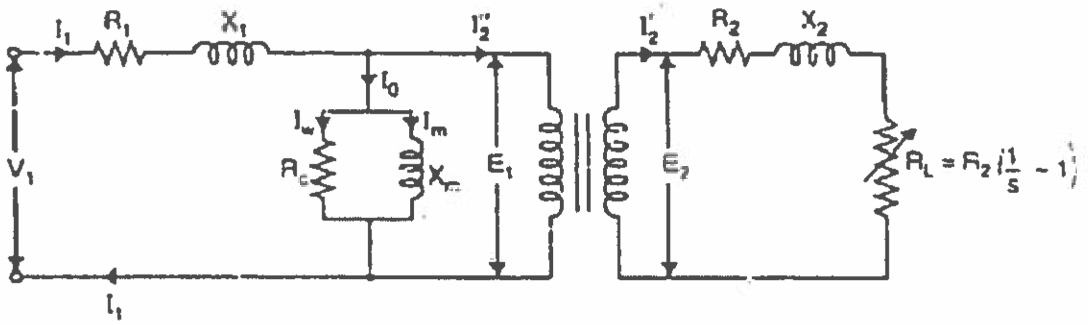


Fig: 3.11

Therefore; induction motor can be represented as an equivalent transformer connected to a variable-resistance load RL given by eq. (i). The power delivered to RL represents the total mechanical power developed in the rotor. Since the equivalent circuit of Fig. 3.11 is that of a transformer, the secondary (i.e., rotor) values can be transferred to primary (i.e., stator) through the appropriate use of transformation ratio K. Recall that when shifting resistance/reactance from secondary to primary, it should be divided by K2 whereas current should be multiplied by K. The equivalent circuit of an induction motor referred to primary is shown in Fig. 3.12.

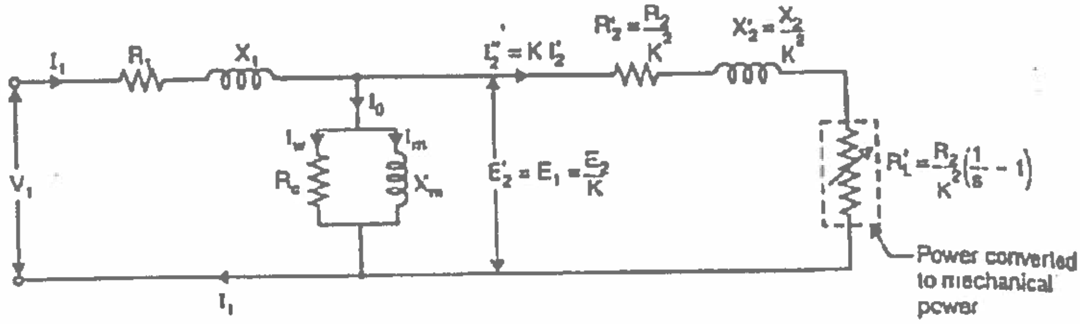


Fig. 3.12

Note that the element (i.e., R'L) enclosed in the dotted box is the equivalent electrical resistance related to the mechanical load on the motor. The following points may be noted from the equivalent circuit of the induction motor:

(i) At no-load, the slip is practically zero and the load R'L is infinite. This condition resembles that in a transformer whose secondary winding is open-circuited.

(ii) At standstill, the slip is unity and the load R'L is zero. This condition resembles that in a transformer whose secondary winding is short-circuited

(iii) When the motor is running under load, the value of R'L will depend upon the value of the slip s. This condition resembles that in a transformer whose secondary is supplying variable and purely resistive load.

(iv) The equivalent electrical resistance R'L related to mechanical load is slip or speed dependent. If the slip s increases, the load R'L decreases and the rotor current increases and motor will develop more mechanical power. This is expected because the slip of the motor increases with the increase of load on the motor shaft.

7.Derive an expression for the torque of an induction motor and torque-slip characteristics and obtain the condition for maximum torque (12M)

Torque Equation of Three Phase Induction Motor

he torque produced by three phase induction motor depends upon the following three factors:

Firstly the magnitude of rotor current, secondly the flux which interact with the rotor of three phase induction motor and is responsible for producing emf in the rotor part of induction motor, lastly the power factor of rotor of the three phase induction motor. Combining all these factors, we get the equation of torque as-

$T \propto \phi I_2 \cos\theta_2$

Where, T is the torque produced by the induction motor,

φ is flux responsible for producing induced emf,
I2 is rotor current,
cosθ2 is the power factor of rotor circuit.

The flux φ produced by the stator is proportional to stator emf E1.
i.e $\varphi \propto E1$
We know that transformation ratio K is defined as the ratio of secondary voltage (rotor voltage) to that of primary voltage (stator voltage).

$$K = \frac{E_2}{E_1}$$

$$or.\ K = \frac{E_2}{\phi}$$

$$or.\ E_2 = \phi$$

Rotor current I2 is defined as the ratio of rotor induced emf under running condition , sE2 to total impedance, Z2 of rotor side, de,

$$i.e\ I_2 = \frac{sE_2}{Z_2}$$

and total impedance Z2 on rotor side is given by ,

$$Z_2 = \sqrt{R_2^2 + (sX_2)^2}$$

Putting this value in above equation we get,

$$I_2 = \frac{sE_2}{\sqrt{R_2^2 + (sX_2)^2}}$$

s = slip of induction motor
We know that power factor is defined as ratio of resistance to that of impedance. The power factor of the rotor circuit is

$$\cos\theta_2 = \frac{R_2}{Z_2} = \frac{R_2}{\sqrt{R_2^2 + (sX_2)^2}}$$

Putting the value of flux φ, rotor current I2, power factor cosθ2 in the equation of torque we get,

$$T \propto E_2 \frac{sE_2}{\sqrt{R_2^2 + (sX_2)^2}} \times \frac{R_2}{\sqrt{R_2^2 + (sX_2)^2}}$$

Combining similar term we get,

$$T \propto sE_2^2 \frac{R_2}{\sqrt{R_2^2 + (sX_2)^2}}$$

Removing proportionality constant we get,

$$T = KsE_2^2 \frac{R_2}{\sqrt{R_2^2 + (sX_2)^2}}$$

$$This\ constant\ K = \frac{3}{2\pi n_s}$$

Where, ns is synchronous speed in r. p. s, ns = Ns / 60. So, finally the equation of torque becomes,

$$T = sE_2^2 \times \frac{R_2}{R_2^2 + (sX_2)^2} \times \frac{3}{2\pi n_s} N - m$$

Derivation of K in torque equation.

In case of three phase induction motor, there occur copper losses in rotor. These rotor copper losses are expressed as

Pc = 3I22R2

We know that rotor current,

$$I_2 = \frac{sE_2}{\sqrt{R_2^2 + (sX_2)^2}}$$

Substitute this value of I2 in the equation of rotor copper losses, Pc. So, we get

$$P_c = 3R_2\left(\frac{sE_2}{\sqrt{R_2^2 + (sX_2)^2}}\right)^2$$

$$On\ simplifying\ P_c = \frac{3R_2s^2E_2^2}{R_2^2 + (sX_2)^2}$$

The ratio of P2 : Pc : Pm = 1 : s : (1 – s)

Where, P2 is the rotor input,

Pc is the rotor copper losses,

Pm is the mechanical power developed.

$$\frac{P_c}{P_m} = \frac{s}{1-s}$$

$$or\ \ P_m = \frac{(1-s)P_c}{s}$$

Substitute the value of Pc in above equation we get,

$$P_m = \frac{1}{s} \times \frac{(1-s)3R_2s^2E_2^2}{R_2^2 + (sX_2)^2}$$

On simplifying we get,

$$P_m = \frac{(1-s)3R_2sE_2^2}{R_2^2 + (sX_2)^2}$$

The mechanical power developed Pm = Tω,

$$\omega = \frac{2\pi N}{60}$$

$$or\ P_m = T\frac{2\pi N}{60}$$

Substituting the value of Pm

$$\frac{1}{s} \times \frac{(1-s)3R_2s^2E_2^2}{R_2^2 + (sX_2)^2} = T\frac{2\pi N}{60}$$

$$or\ T = \frac{1}{s} \times \frac{(1-s)3R_2s^2E_2^2}{R_2^2 + (sX_2)^2} \times \frac{60}{2\pi N}$$

We know that the rotor speed N = Ns(1 – s)
Substituting this value of rotor speed in above equation we get,

$$T = \frac{1}{s} \times \frac{(1-s)3R_2 s^2 E_2^2}{R_2^2 + (sX_2)^2} \times \frac{60}{2\pi N_s(1-s)}$$

Ns is speed in revolution per minute (rpm) and ns is speed in revolution per sec (rps) and the relation between the two is

$$\frac{N_s}{60} = n_s$$

Substitute this value of Ns in above equation and simplifying it we get

$$Toeque.\ T = \frac{s\ E_2^2 R_2}{R_2^2 + (sX_2)^2} \times \frac{3}{2\pi N_s}$$

$$or.\ T = KsE_2^2 \frac{R_2}{R_2^2 + (sX_2)^2}$$

Comparing both the equations, we get, constant K = 3 / 2πns

Equation of Starting Torque of Three Phase Induction Motor

Starting torque is the torque produced by induction motor when it starts. We know that at the start the rotor speed, N is zero.

$$So.\ slip\ s = \frac{N_s - N}{N_s}\ becomes\ 1$$

So, the equation of starting torque is easily obtained by simply putting the value of s = 1 in the equation of torque of the three phase induction motor,

$$T = \frac{E_2^2 R_2}{R_2^2 + X_2^2} \times \frac{3}{2\pi n_s}\ N - m$$

The starting torque is also known as standstill torque.

Maximum Torque Condition for Three-Phase Induction Motor

In the equation of torque,

$$T = \frac{sE_2^2 R_2}{R_2^2 + (sX_2)^2} \times \frac{3}{2\pi n_s}$$

The rotor resistance, rotor inductive reactance and synchronous speed of induction motor remain constant. The supply voltage to the three phase induction motor is usually rated and remains constant, so the stator emf also remains the constant. We define the transformation ratio as the ratio of rotor emf to that of stator emf. So if stator emf remains constant, then rotor emf also remains constant.
If we want to find the maximum value of some quantity, then we have to differentiate that quantity concerning some variable parameter and then put it equal to zero. In this case, we have to find the condition for maximum torque, so we have to differentiate torque concerning some variable quantity which is the slip, s in this case as all other parameters in the equation of torque remains constant.
So, for torque to be maximum

$$\frac{dT}{ds} = 0$$

$$T = KsE_2^2 \frac{R_2}{R_2^2 + (sX_2)^2}$$

Now differentiate the above equation by using division rule of differentiation. On differentiating and after putting the terms equal to zero we get,

$$s^2 = \frac{R_2^2}{X_2^2}$$

Neglecting the negative value of slip we get

$$s^2 = \frac{R_2^2}{X_2^2}$$

So, when slip s = R2 / X2, the torque will be maximum and this slip is called maximum slip Sm and it is defined as the ratio of rotor resistance to that of rotor reactance.
NOTE: At starting S = 1, so the maximum starting torque occur when rotor resistance is equal to rotor reactance.

Equation of Maximum Torque

The equation of torque is

The torque will be maximum when slip s = R2 / X2

Substituting the value of this slip in above equation we get the maximum value of torque as,

In order to increase the starting torque, extra resistance should be added to the rotor circuit at start and cut out gradually as motor speeds up.

Conclusion

From the above equation it is concluded that

1. The maximum torque is directly proportional to square of rotor induced emf at the standstill.
2. The maximum torque is inversely proportional to rotor reactance.
3. The maximum torque is independent of rotor resistance.
4. The slip at which maximum torque occur depends upon rotor resistance, R2. So, by varying the rotor resistance, maximum torque can be obtained at any required slip.

Torque Slip Characteristics of 3-Phase Induction Motor

The graph plotted between the torque and slip for a particular value of rotor resistance and reactance is known as torque-slip characteristics of the induction motor.

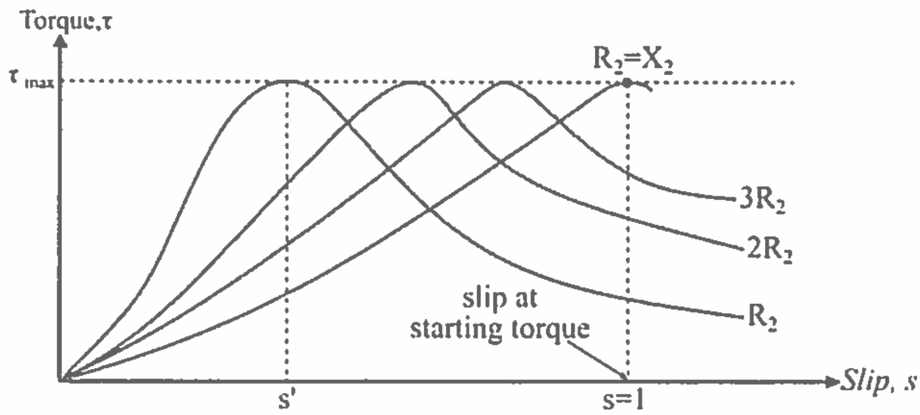
The torque of a 3-phase induction motor under running conditions is given by,

τr=KsE22R2 /(R22+(sX22))...(1)

From the eqn. (1), it can be seen that if R2 and X2 are kept constant, the torque depends upon the slip 's'. The torque-slip characteristics curve can be divided into three regions, viz.

- Low-slip region

- Medium-slip region
- High-slip region



Low-Slip Region

At synchronous speed, the slip s = 0, thus, the torque is 0. When the speed is very near to the synchronous speed, the slip is very low and the term $(sX2)2$ is negligible in comparison with R2. Therefore,

$$\tau r \propto s/R2$$

If R2 is constant, then

$$\tau r \propto s \ldots(2)$$

Eqn. (2) shows that the torque is proportional to the slip. Hence, when the slip is small, *the torque-slip curve is straight line.*

Medium-Slip Region

When the slip increases, the term ($sX2$)2 becomes large so that $R22$ may be neglected in comparison with ($sX2$)2. Therefore,

$$\tau r \propto s/(sX2)2=1/sX22$$

If X2 is constant, then

$$\tau r \propto 1/s \ldots (3)$$

Thus, the torque is inversely proportional to slip towards standstill conditions. Hence, for intermediate values of the slip, the torque-slip characteristics is represented by a *rectangular hyperbola*. The curve passes through the point of *maximum torque* when R2 = $sX2$.

The maximum torque developed by an induction motor is known as *pull-out torque* or *breakdown torque*. This breakdown torque is a measure of the short time overloading capability of the motor.

High-Slip Region

The torque decreases beyond the point of maximum torque. As a result of this, the motor slows down and eventually stops. The induction motor operates for the values of slip between s = 0 and s = sm, where sm is the value of slip corresponding to maximum torque. For a typical 3-phase induction motor, the breakdown torque is 2 to 3 times of the full-load torque. Therefore, the motor can handle overloading for a short period of time without stalling.

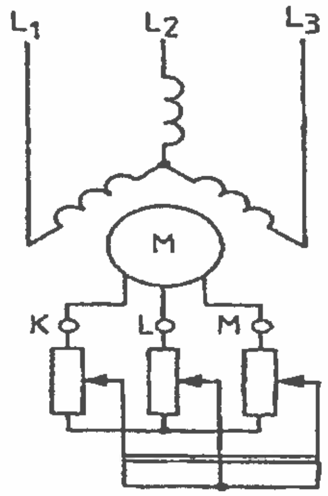
It may be seen from the torque-slip characteristics that addition of resistance to the rotor circuit does not change the value of maximum torque but it only changes the value of slip at which maximum torque occurs.

8.Describe the following:
(i) Rotor resistance starter for starting slip ring induction motor.
(ii) Speed control of an induction motor by changing the frequency and poles
(12M)

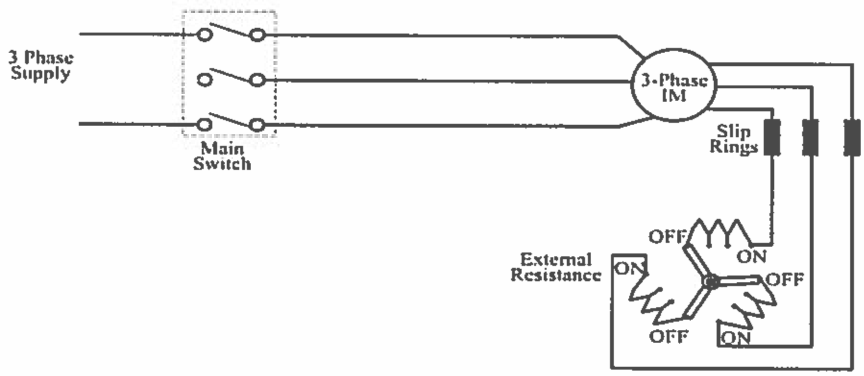
(i)Rotor resistance starter for starting slip ring induction motor

If it is necessary to start a three phase induction motor on load then a wound rotor machine also known as slip ring motor will normally be selected.
Such a machine allows an external resistance to be connected to the rotor of the machine through slip rings and brushes. A 3-phase rheostat is connected in series with the rotor circuit through brushes.

At start-up the rotor resistance is set at maximum but is reduced as speed increases until eventually it is reduced to zero and the machine runs as if it is a cage rotor machine.
By inserting external resistance in the rotor circuit, not only the starting current is reduced but at the same time starting torque is increased due to improvement of power factor.



In a *rotor resistance starter*, a star connected variable resistance is connected in the rotor circuit through slip-rings. The full voltage is applied to the stator windings. The connection arrangement of the rotor resistance starter is shown in the figure.



At the instant of starting, the handle of variable resistance (rheostat) is set to 'OFF' position. This inserts maximum resistance in series with each phase of the rotor circuit. This reduces the starting current and at the same time starting torque is increased due to external rotor resistance.

As the motor accelerates, the external resistance is gradually removed from the rotor circuit. When the motor attains rated speed, the handle is switched in the 'ON' position, this removes the whole external resistance from the rotor circuit.

Slip Ring induction motors are mainly used for driving high inertia loads or the loads which require a starting torque across a full speed range. Hence, by correctly selecting the starting resistors inserted into the rotor circuit, the maximum torque can be obtained from the motor at a relatively low starting current. Thus, the magnitude of starting torque is controlled by the value starting resistance.

The starting resistors need to be designed to allow for the starting duty cycle without overheating or too large change in the value of the resistance when hot.

(ii)Speed control of an induction motor by changing the frequency and poles

Changing The Number Of Stator Poles

From the above equation of synchronous speed, it can be seen that synchronous speed (and hence, running speed) can be changed by changing the number of stator poles.

This method is generally used for squirrel cage induction motors, as squirrel cage rotor adapts itself for any number of stator poles. Change in stator poles is achieved by two or more independent stator windings wound for different number of poles in same slots.

For example, a stator is wound with two 3phase windings, one for 4 poles and other for 6 poles. for supply frequency of 50 Hz

i) synchronous speed when 4 pole winding is connected,

Ns = 120*50/4 = 1500 RPM

ii) synchronous speed when 6 pole winding is connected,

Ns = 120*50/6 = 1000 RPM

By Changing The Applied Frequency

Synchronous speed of the rotating magnetic field of an induction motor is given by,

$$Ns = \frac{120\,f}{P} \quad \text{(RPM)}$$

where, f = frequency of the supply and

P = number of stator poles.

Hence, the synchronous speed changes with change in supply frequency. Actual speed of an induction motor is given as N = Ns (1 - s). However, this method is not widely used. It may be used where, the induction motor is supplied by a dedicated generator (so that frequency can be easily varied by changing the speed of prime mover). Also, at lower frequency, the motor current may become too high due to decreased reactance. And if the frequency is increased beyond the rated value, the maximum torque developed falls while the speed rises.

9(a).With a neat diagram, Discuss about the slip power recovery scheme of induction motor (6M)

This system is mainly used for Induction motor speed control. The speed control in induction motor has poor efficiency due to wasting of slip power in the rotor circuit. By using recovery schemes the induction motor speed is controlled to avoid slip power loss.

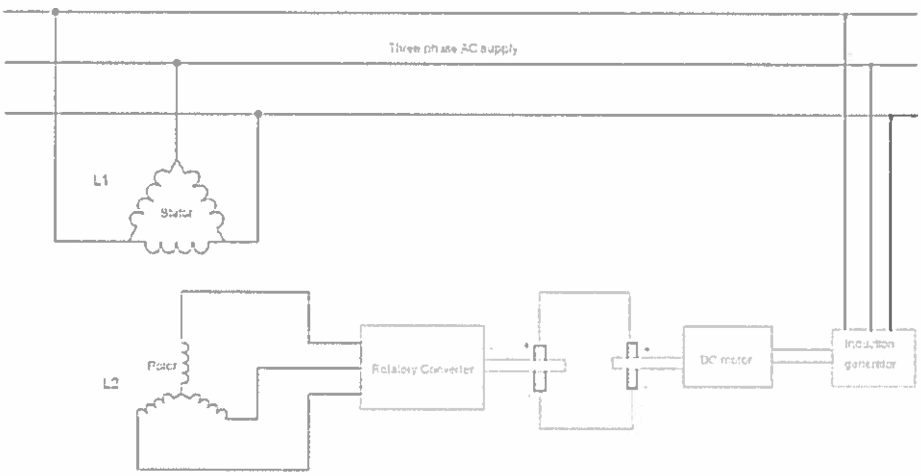
The slip power is classified into two types
1. Scherbius system
2. Kramer system

1. Static scherbius drive system:

This system provides feedback path i.e. the wastage of slip power is again fed to AC mains supply. The static scherbius system is of two types
i) Conventional Scherbius system
ii) Static Scherbius system

i) Conventional scherbius system:

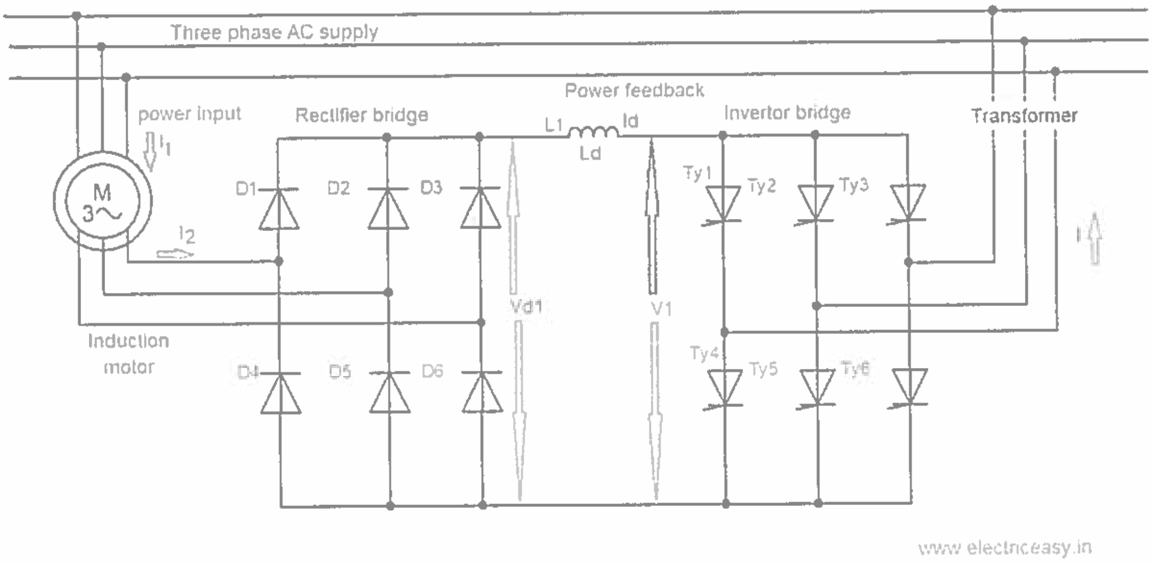
In this system the recovery scheme is done by feedback path. The output of three phase Induction motor is connected to the DC motor by coupling them the mechanical power input of DC motor is converted into electrical power and fed to Induction generator and again back to mains.



ii) Static Scherbius drive system:

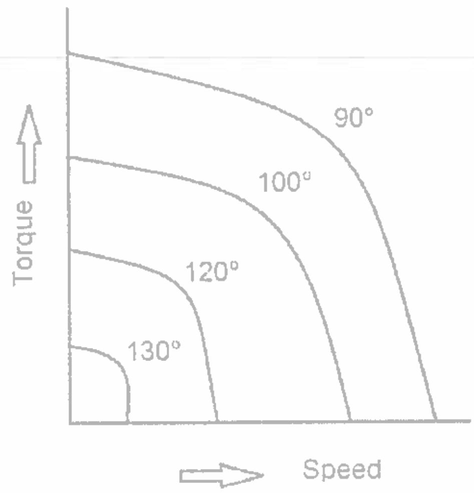
The phenomenon of this system is same as conventional type but the only difference is this system provides with diode bridge rectifier along with thyristor bridge inverter. This is also known as Sub-synchronous cascade drive.

When Induction motor is operating at slip frequency the rotor slip power is rectified by the diode rectifier. The output of rectifier is fed to inverter three phase bridge again the output is fed back to supply lines with the help of transformer.



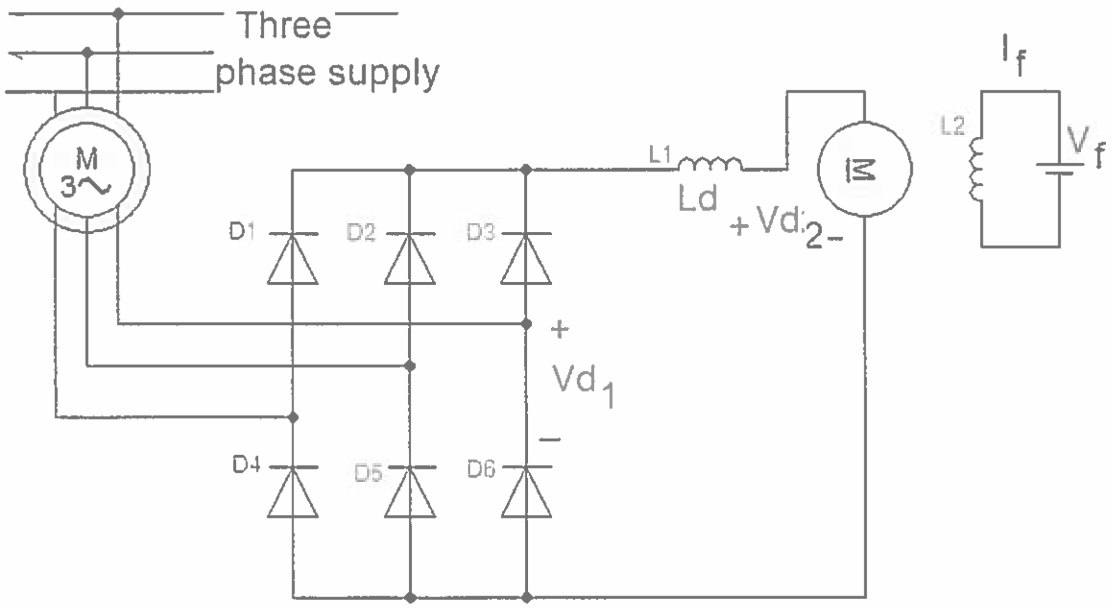
Natural commutation proves involves across slip rings bus-bars. The induced emf frequency is made equal to rotor emf frequency by rectification of slip ring voltage to obtain speed control at injected voltage.

*In this circuit if commutation overlap is negligible the output voltage of uncontrolled three phase bridge rectifier is obtained as



2. Static Kramer drive:

In this method the rotatory slip power is coverted into DC by a diode bridge. The DC power is fed to the DC motor which is mechanically coupled with the Induction motor. The speed control is done by varying the field current If.



From the characteristics you can easily observe the voltage and field current differences. The steady state operation is possible at Vd1 = Vd2

For large speed applications the diode bridge is replaced by using thyristor bridge, the speed can be controlled by varying the firing angle. Upto standstill condition the speed can be controlled

9(b). Compare the relative merits and demerits of stator resistance starter
(6M)

Applications of Stator Resistance Starter:

1. This types of starter are suitable for more than 7.5 KW rating motors.

Advantages of Stator Resistance Starter:

1. It is very simple in construction.
2. It reduces the terminal voltage of the motor during the starting time hence the starting current also decreases.
3. It is also a low-cost starter.

Disadvantages of Stator Resistance Starter:

1. It lowers the starting torque of the motor.
2. It increases the accelerating time of the motor
3. A huge amount of power loss in by the starting resistance.

10(a).Explain the principle of operation of single-phase induction motor based on "double revolving field theory"

(6M)

DOUBLE REVOLVING FIELD THEORY

Let us see why single phase induction motors are not self starting with the help of a theory called double revolving field theory.

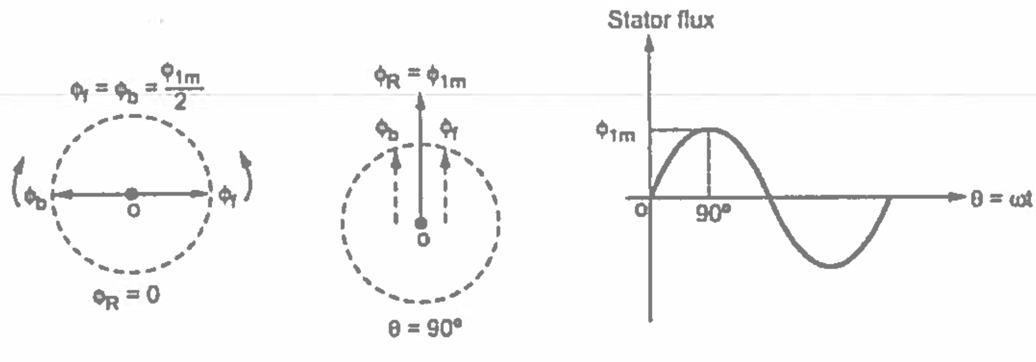
According to this theory, any alternating quantity can be resolved into two rotating components which rotate in opposite directions and each having magnitude as half of the maximum magnitude of the alternating quantity.

In case of single phase induction motors, the stator winding produces an alternating magnetic field having maximum magnitude of (ϕ_{1m}.

According to double revolving field theory, consider the two components of the stator flux, each having magnitude half of maximum magnitude of stator flux i.e. ($\phi_{1m}/2$). Both these components are rotating in opposite directions at the synchronous speed Ns which is dependent on frequency and stator poles.

Let ϕ_f is forward component rotating in anticlockwise direction while ϕ_b is the backward component rotating in clockwise direction. The resultant of these two components at any instant gives the instantaneous value of the stator flux at that instant. So resultant of these two is the original stator flux.

Following figure shows the stator flux and its two components ϕ_f and ϕ_b.



(a) (b) (c)

At start both the components are shown opposite to each other in the

Fig. (a). Thus the resultant $\phi R = 0$. This is nothing but the instantaneous value of stator flux at start. After 90^0, as shown in the

Fig. (b), the two components are rotated in such a way that both are pointing in the same direction. Hence the resultant ϕR is the algebraic sum of the magnitudes of the two components. So $\phi R = (\phi 1m/2) + \phi 1m/2) = \phi 1m$. This is nothing but the instantaneous value of the stator flux at $0 = 90^0$ as shown in the

Fig. (c). Thus continuous rotation of the two components gives the original alternating stator flux.

Both the components are rotating and hence get cut by the rotor conductors. Due to cutting of flux, e.m.f. gets induced in rotor which circulates rotor current. The rotor current produces rotor flux. This flux interacts with forward component 0f to produce a torque in one particular direction say anticlockwise direction. While rotor flux interacts with backward component ob to produce a torque in the clockwise direction. So if anticlockwise torque is positive then clockwise torque is negative.

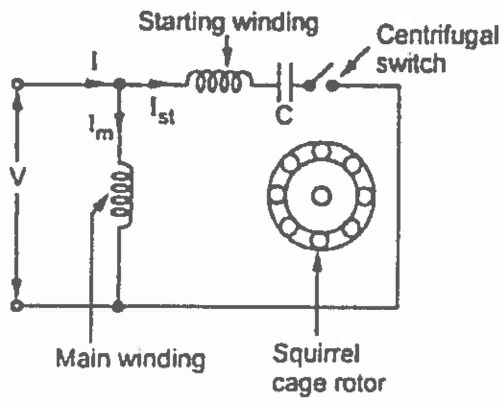
10(b).Explain in detail the operation of capacitor start and run induction motor (6M)

Capacitor Start Induction Motors

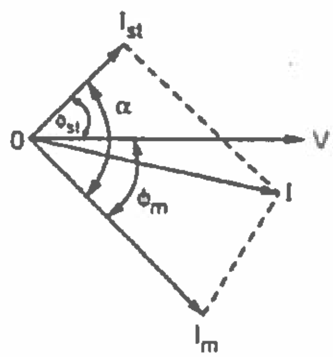
The construction of this type of motor is similar to the resistance split phase type. The difference is that in series with the auxiliary winding the capacitor is connected. The capacitive circuit draws a leading current, this feature used in this type to increase the split phase angle a between the two currents I_m and I_{st}.
Depending upon whether capacitor remains in the circuit permanently or is disconnected from the circuit using centrifugal switch, these motors are classified as,

1. Capacitor start motors and
2. Capacitor start capacitor run motors

The construction of capacitor start motor is shown in the Fig. (a). The current I_m lags the voltage by angle ϕ_m while due to capacitor the current I_{st} leads the voltage by angle ϕ_{st}. Hencethere exists a large phase difference between the two currents which is almost 90^0, which is an ideal case. The phasor diagram is shown in the Fig. (b).



(a) Schematic re presentation

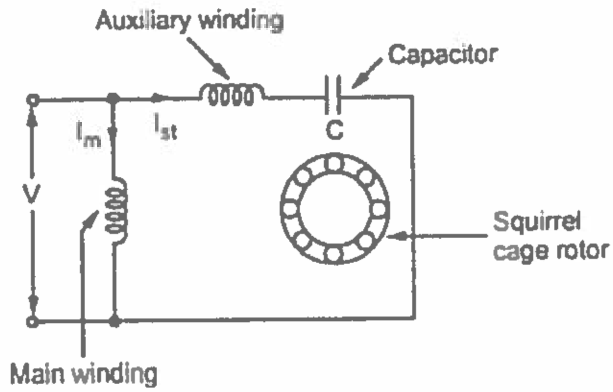


(b) Phasor diagram

(b) The starting torque is proportional to 'α' and hence such motors produce very high starting torque.

(c) When speed approaches to 75 to 80% of the synchronous speed, the starting winding gets disconnected due to operation of the centrifugal switch. The capacitor remains in the circuit only at start hence it is called capacitor start motors.

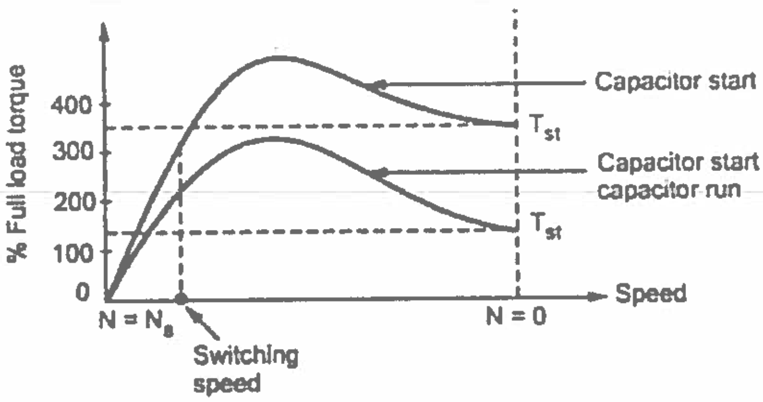
(d) The schematic representation of such motor is shown in following figur



(e)

(f) The phasor diagram remains same as shown in the Fig. (b). The performance not only at start but in running condition also depends on the capacitor C hence its value is to be designed soas to compromise between best starting and best running condition. Hence the starting torque

(g) available in such type of motor is about 50 to 100% of full load torque. The torque-slip characteristics is shown in following figure.



(h) These motors have high starting torque and hence are used for hard starting loads. These areused for compressors, conveyors, grinders, fans, blowers, refrigerators, air conditioners etc. These are most commonly used motors. The capacitor start capacitor run motors are used inceiling fans, blowers and air-circulators. These motors are available upto 6 kW.

(i) At start these two torques are equal in magnitude but opposite in direction. Each torque tries to rotate the rotor in its own direction.

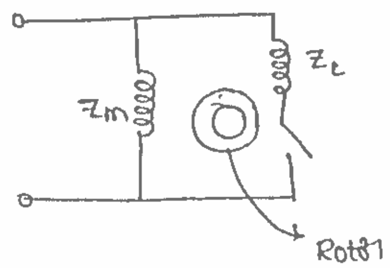
(j) Thus net torque experienced by the rotor is zero at start. And hence the single phase induction motors are not self starting

11(a). The equivalent impedance of the main and auxiliary winding in a capacitor motor are (15 + j 25)Ω and (50+ j120)Ω respectively, while the capacitance of the capacitor is 12 μ F. Estimate the line current at starting a 230 V, 50Hz supply (8M)

① a) Given data

$Z_m = (15+j25)\,\Omega$

$Z_A = (50+j120)\,\Omega$



$Z_m = (15+j25)\Omega$

$Z_M = Z_m \angle \phi_m$

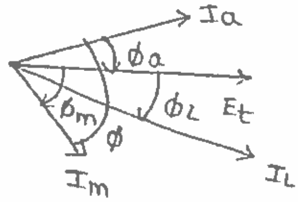
$$I_m = \frac{E_t}{Z_m} = I_m \angle -\phi_m$$

$Z_a = R_a - jX = Z_a \angle -\phi_a$

$X = X_C - X_a$

$I_L = I_m + I_a$

$\phi = \phi_m + \phi_a$



$Z_a = 50 - j(X_C - X_a)$

$Z_a = 50 - j\left(2\pi \times 50 \times 12\times10^{-6} - 120\right)$

$= 50 - j(-119.99)$

$Z_a = 50 + j\,119.99$

$$I_m = \frac{10\angle 0}{\sqrt{15^2+25^2}\ \tan^{-1}\left(\frac{25}{15}\right)} = \frac{10\angle 0}{29.15 \times 59.03}$$

$$= \frac{10\angle 0}{1720.72+j0} = \frac{10\angle 0}{1720.72\angle 0}$$

$I_m = 5.811\times10^{-3}\angle 0^\circ$

$$I_a = \frac{E_t}{Z_a}$$

$$= \frac{10\angle 0}{Z_a}$$

$$I_a = \frac{10\angle 0}{50+j119.99}$$

$$I_a = \frac{10\angle 0}{129.9\angle 67.3} = 0.076\angle{-67.3}$$

$$I_L = I_m + I_a$$

$$I_L = 5.811\times10^{-3}\angle 0^\circ + 0.046\angle{-67.3}$$

$$I_L = 5.81\times10^{-3} + j0 + 0.029 - j0.07$$

$$I_L = 0.0348 - j0.07$$

$$I_L = 0.07\angle 64.09$$

11(b). Identify the features of no load and blocked rotor test (4M)

The *no load test and the blocked rotor test* are two main induction motor tests, which are performed on induction motor to know the different losses, power factor and efficiency of the induction motor

No load test: The **no load test of 3 phase induction motor** is performed on induction motor when it is running without load. **This test tells us the magnitude of constant losses occurring in the motor.**

The machine is started in the usual way and runs unloaded from normal voltage mains. On the mains side suitable instruments are connected between supply mains and motor terminals to measure power, line current and line voltage.

Blocked rotor test: It is performed by locking the rotor (by keeping the rotor not to rotate). This is carried out to know the copper losses, power factor at short circuit current; total equivalent resistance and reactance.

This test is just similar to short circuit test of the transformer.

Starting with zero voltages across the stator, the applied voltage is gradually increased in steps until the full load current is circulated. The readings of voltmeter ammeter and watt meters are noted. While performing this test the following points are taken care of:

- Means of holding tight (not to rotate) the rotor should be of proper strength.
- The direction of rotation of a rotor should be established prior to start the test and direction of force, which is to keep the rotor blocked (unmoved) should be in opposite direction.
- As the windings get heated the test should be carried out quickly.
- The short circuit current should not be more than the full load current.

12(a). Demonstrate the POTIER method of determining the regulation of an alternator (8M)

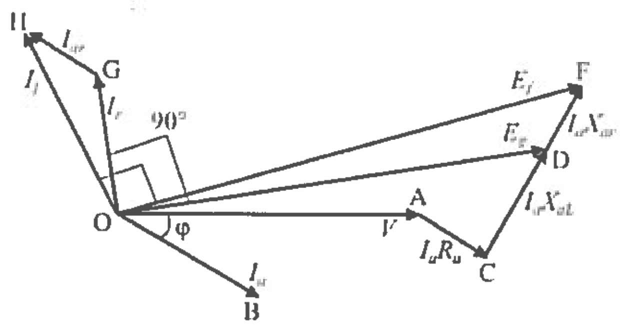
The Potier Triangle Method is used in determining the voltage regulation of alternators. It is also known as the Zero Power Factor (ZPF) method. The following assumptions are made in the *Potier triangle method* −

- The armature reaction MMF is constant.
- The open-circuit characteristic (O.C.C.) taken on no-load accurately represents the relation between MMF and voltage under loaded conditions.
- The voltage drop due to the armature leakage reactance (IaXaLI_aX_{aL}) is independent of the excitation.

Procedure to Obtain Voltage Regulation by ZPF Method

The following procedure is followed to determine the voltage regulation of an alternator or synchronous generator by the zero power factor (ZPF) method −

First of all, we draw the phasor diagram of the alternator at lagging power factor as shown in the figure.



In the phasor diagram,

- The phasor OA represents the terminal phase voltage (V) at full-load. It is taken as the reference phasor, thus drawn horizontally.
- The phasor OB represents the full-load current (IaI_a). Since, the load power factor is φ lagging, hence, it is drawn lagging behind the voltage (V) by the power factor angle (φ).
- The phasor AC is representing the voltage drop (IaRaI_aR_a) in the armature resistance. As it is in phase with the armature current (IaI_a) and hence is drawn parallel to current phasor (IaI_a). If the armature resistance is neglected, then, this phasor will not be drawn in the phasor diagram.
- The voltage drop due to leakage reactance (IaXaLI_aX_{aL}) is represented by the phasor CD. It is perpendicular to the phasor AC.

- Now, join O and D, the phasor OD represents the generated EMF (EgEg).

Now, find the field excitation current (IrIr) of the resultant MMF which is corresponding to the generated EMF (EgEg) from the open-circuit characteristic (O.C.C.). Then, draw the phasor OG equal to the current (IrIr) and it is perpendicular to the phasor OD.

Draw the phasor GH parallel to the load current (IaIa) to represent the field current equivalent to the full-load armature reaction current (IarIar). Therefore, the phasor OH gives the total field current (IfIf).

When the load on the alternator is thrown off, then the terminal voltage will be equal to the generated EMF, corresponding to the field current (If$If = OH$).

Now, determine the EMF (EfEf) represented by the phasor OF corresponding to the field current (If=OH$If=OH$) from the O.C.C. The phasor OF will lag behind the phasor OH by 90°. Also, the voltage drop due to the armature reaction is represented by the phasor DF.

Therefore, the *voltage regulation* of the alternator or synchronous generator can be obtained from the following expression –

%voltage regulation=(Ef−V)/V×100

12(b).Compare the Constructional details of rotor of both non-salient and salient pole synchronous machine (4M)

Salient pole Synchronous generator

- Salient pole Generators will have large diameter and short axial length
- Pole shoes cover 2/3 of the pitch
- Salient Poles are laminated in order to reduce eddy currents
- They are used in hydraulic turbines or diesel engines
- Salient pole generators will have typical speed about 100 to 375 rpm.
- As the speed of the water turbine is slow hence more number of poles are required to attain the frequency. Therefore Salient pole machines will have typically number of poles will be between 4 to 60.
- Cheaper compared to cylindrical rotor machines for speeds below 1000rpm.
- Causes excessive windage losses
- Flux distribution is not uniform due to the presence of salient poles, hence emf waveform generated is not good compared to cylindrical machine
- Salient Pole Synchronous Generators are employed in Hydro-Power plants.

Non-Salient pole Synchronous Alternator:

- Non-Salient pole generators will have smaller diameter and longer axial length
- They are used for High speed operation (typically speed will be 1500 and 3000 rpm)
- Better in dynamic balancing because of absence of salient poles
- Less windage loss
- Robust construction and noiseless operation

- Nearly sinusoidal flux distribution around the periphery, therefore gives a better emf waveform than salient pole machine
- No need to provide damper windings (except in special case to assist the synchronising) because the field poles themselves acts as efficient dampers.
- Non-Salient pole generators are used in Thermal,Gas and in Nuclear Power plants.

13(a). Describe the role of voltage regulation in alternator. Also explain synchronous impedance method for determining regulation of an alternator. (8M)

Synchronous Impedance Method

The **Synchronous Impedance Method or Emf Method** is based on the concept of replacing the effect of armature reaction with an imaginary reactance. For calculating the regulation, the synchronous method requires the following data; they are the armature resistance per phase and the open-circuit characteristic. The open-circuit characteristic is the graph of the circuit voltage and the field current. This method also requires a short circuit characteristic which is the graph of the short circuit and the field current.

For a synchronous generator following are the equation given below:

$$V = E_a - Z_s I_a \qquad \text{Where,} \qquad Z_s = R_a + jX_s$$

For calculating the synchronous impedance, Zs is measured, and then the value of Ea is calculated. From the values of Ea and V, the voltage regulation is calculated.

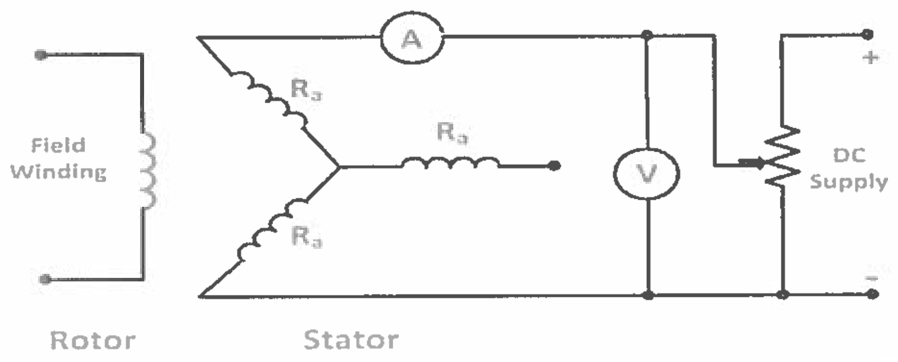
Measurement of Synchronous Impedance

The measurement of synchronous impedance is done by the following methods. They are known as:

- DC resistance test
- Open circuit test
- Short circuit test

DC resistance test

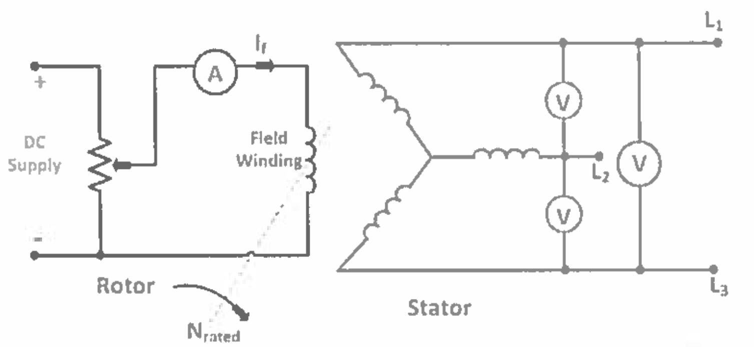
In this test, it is assumed that the alternator is star connected with the DC field winding open as shown in the circuit diagram below:



It measures the DC resistance between each pair of terminals either by using an ammeter – voltmeter method or by using the Wheatstone's bridge. The average of three sets of resistance value Rt is taken. The value of Rt is divided by 2 to obtain a value of DC resistance per phase. Since the effective AC resistance is larger than the DC resistance due to skin effect. Therefore, the effective AC resistance per phase is obtained by multiplying the DC resistance by a factor 1.20 to 1.75 depending on the size of the machine. A typical value to use in the calculation would be 1.25.

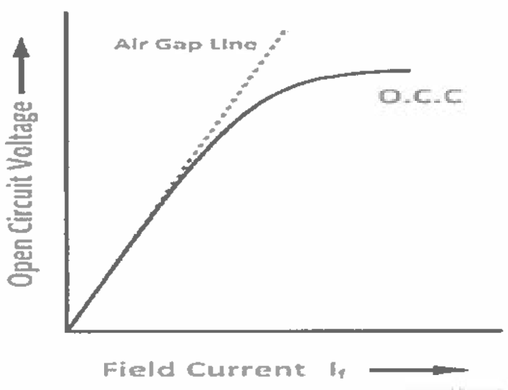
Open Circuit Test

In the **open-circuit test** for determining the synchronous impedance, the alternator is running at the rated synchronous speed, and the load terminals are kept open. This means that the loads are disconnected, and the field current is set to zero. The circuit diagram is shown below:



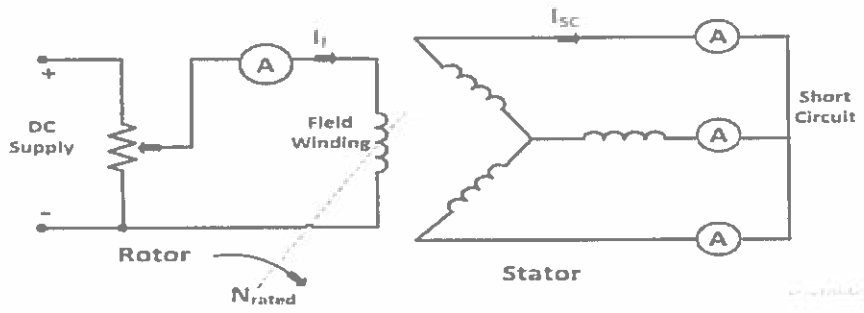
After setting the field current to zero, the field current is gradually increased step by step. The terminal voltage Et is measured at each step. The excitation current may be increased to get 25% more than the rated voltage. A graph is drawn between the open circuit phase voltage $Ep = Et/\sqrt{3}$ and the field current If. The curve so obtained called Open Circuit Characteristic (O.C.C). The shape is the same as the normal magnetization curve. The linear portion of the O.C.C is extended to form an air gap line.

The Open Circuit Characteristic (O.C.C) and the air gap line is shown in the figure below:



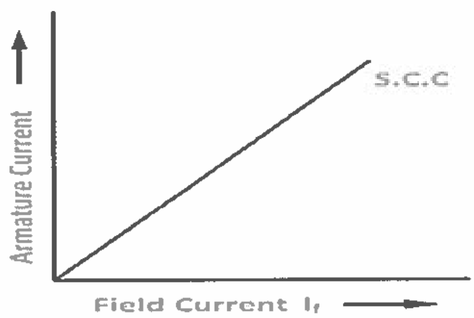
Short Circuit Test

In the **short circuit test**, the armature terminals are shorted through three ammeters as shown in the figure below:



The field current should first be decreased to zero before starting the alternator. Each ammeter should have a range greater than the rated full load value. The alternator is then run at synchronous speed. Same as in an open circuit test that the field current is increased gradually in steps and the armature current is measured at each step. The field current is increased to get armature currents up to 150% of the rated value.

The value of field current If and the average of three ammeter readings at each step is taken. A graph is plotted between the armature current Ia and the field current If. The characteristic so obtained is called Short Circuit Characteristic (S.C.C). This characteristic is a straight line as shown in the figure below.



Calculation of Synchronous Impedance

The following steps are given below for the calculation of the synchronous impedance.

- The open-circuit characteristics and the short circuit characteristic are drawn on the same curve.
- Determine the value of short circuit current Isc and gives the rated alternator voltage per phase.
- The synchronous impedance ZS will then be equal to the open-circuit voltage divided by the short circuit current at that field current which gives the rated EMF per phase.

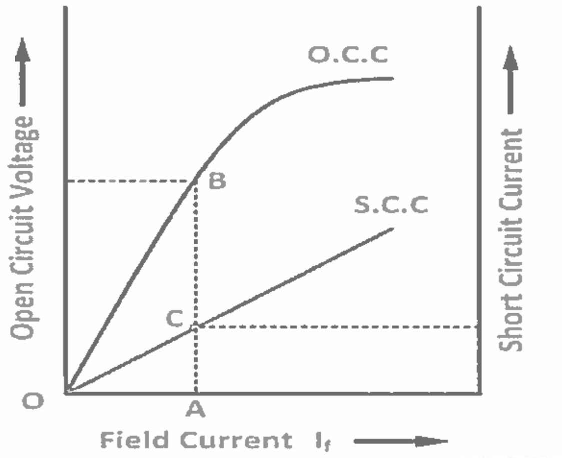
$$Z_S = \frac{\text{Open circuit voltgae per phase}}{\text{Short circuit armature current}} \quad \text{(for the same value of field curren}$$

The synchronous reactance is determined as

$$X_S = \sqrt{Z_S^2 - R_a^2}$$

as

The graph is shown below:



From the above figure consider the field current If = OA that produces rated alternator voltage per phase. Corresponding to this field current, the open-circuit voltage is AB

Therefore,

$$Z_S = \frac{\text{AB (in volts)}}{\text{AC (in amperes)}}$$

13(b).Identify the features of synchronizing of alternator. Describe any one method of synchronizing (4M)

Synchronization of alternator means connecting an alternator into grid in parallel with many other alternators, that is in a live system of constant voltage and constant frequency. Many alternators and loads are connected into a grid, and all the alternators in grid are having same output voltage and frequency (whatever may be the power). It is also said that the alternator is connected to infinite bus-bar.

A stationary alternator is never connected to live bus-bars, because it will result in short circuit in the stator winding (since there is no generated emf yet). Before connecting an alternator into grid, following conditions must be satisfied:

1. Equal voltage: The terminal voltage of incoming alternator must be equal to the bus-bar voltage.
2. Similar frequency: The frequency of generated voltage must be equal to the frequency of the bus-bar voltage.
3. Phase sequence: The phase sequence of the three phases of alternator must be similar to that of the grid or bus-bars.
4. Phase angle: The phase angle between the generated voltage and the voltage of grid must be zero.

The first condition of voltage equality can be satisfied by a voltmeter. To satisfy the conditions of equal frequency and identical phases, one of the following two methods can be used:

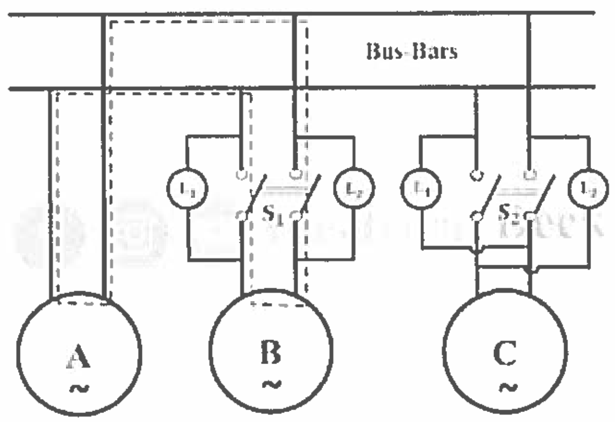
(i) Synchronization using incandescent lamp

(ii) Synchronization using synchroscope

Synchronization Of Alternator Using Incandescent Lamp

Dark Lamp Method (For Single-phase Alternators) :

Consider alternator B is to be connected to bus-bars, to which alternator A is already connected as shown in the figure below. The prime-mover of alternator B is brought up to its rated speed. The alternator is then excited and voltage is raised to that of bus-bars or alternator A voltage. If the frequencies of the alternators A and B are same and their terminal voltages are in phase opposition, no resultant voltage act across the lamps L1 and L2, and therefore these lamps remain dark.



If the frequencies of the alternators A and B are not equal, the current through the lamps and local series circuit (shown with dotted line) will be changing, resulting in the flickering of lamps.

The frequency will be changing, resulting in the flickering of lamps. The frequency of flickering is equal to (fA ~ fB). At this condition, lamps will glow up alternately. In the middle of the dark, the two voltages will be in-phase opposition with respect to the local circuit.

The speed of the alternator B is adjusted until the flickering of lamps is very slow. The voltage is also made equal to the incoming bus-bar voltage by changing field excitation. Now the switch S1 is closed in the middle of the dark period of the flickering lamps. Hence it is known as the dark lamp method.

14(a).Explain the principle of operation of a 3-phase synchronous motor with neat sketch (6M)

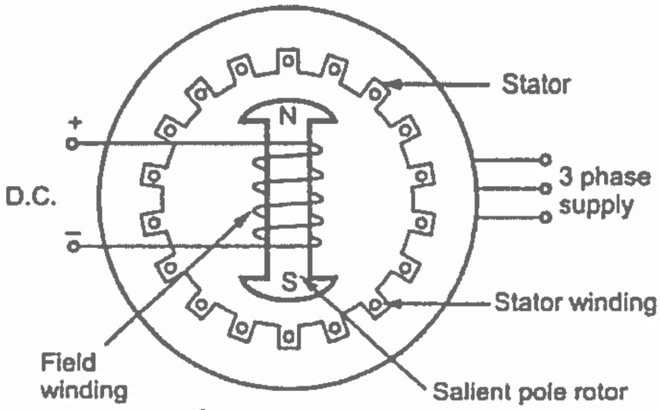
Synchronous Motor Working Principle

- Electric Motor is an electromechanical device which transforms electric energy into mechanical energy.
- According to their type of connection, electric motors are generally classified into the two types i.e single phase motor and three phase motor.
- A synchronous motor is a 3 phase motor and it closely resembles 3 phase alternator.
- 3 phase synchronous motor and 3 phase induction motor are most widely used AC motor.
- A synchronous motor is also called as doubly excited motor.

The synchronous motor consist of the two parts:

Stator: Stator is the armature winding. It consists of three phase star or delta connected winding and excited by 3 phase A.C supply.

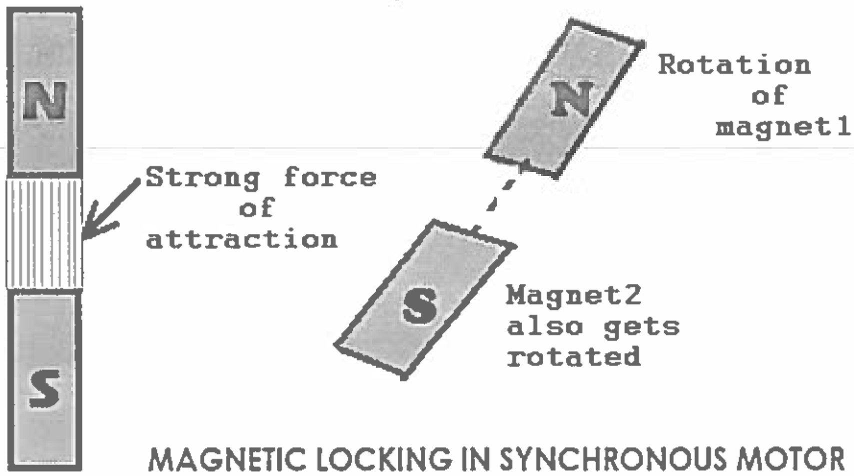
Rotor: Rotor is a field winding. The field winding is excited by the separate D.C supply through the slip ring. The construction of Rotor can be salient pole (projected pole) and non-salient pole (cylindrical pole) type.



SYNCHRONOUS MOTOR

Principle Of Working Of Synchronous Motor

- Synchronous motor work on the principle of magnetic locking.
- When two unlike strong unlike magnets poles are brought together, there exists a tremendous force of extraction between those two poles. In such condition, the two magnets are said to be magnetically locked.



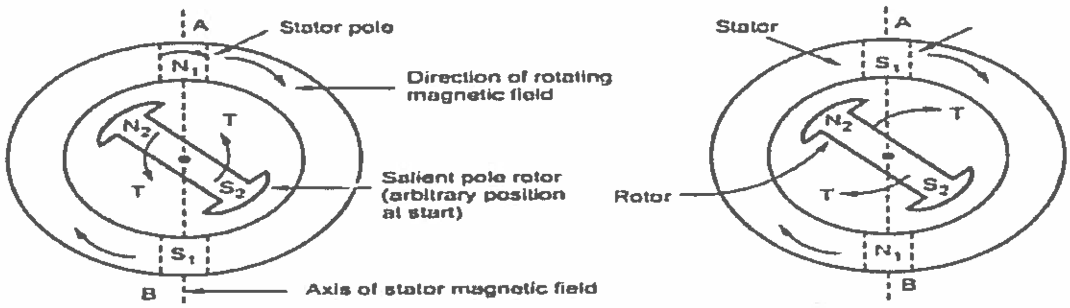
MAGNETIC LOCKING IN SYNCHRONOUS MOTOR

- If now one of the two magnets is rotated, the other magnets also rotate in the same direction with the same speed due to the strong force of attraction.
- This phenomenon is called as magnetic locking For magnetic locking condition, there must be two unlike poles and magnetic axes of this two poles must be brought very nearer to each other.
- Consider a synchronous motor whose stator is wound for 2 poles.
- The stator winding is excited with 3 phase A.C supply and rotor winding with D.C supply respectively. Thus two magnetic fields are produced in the synchronous motor.

- When the 3 phase winding is supplied by 3 phase A.C supply than the rotating magnetic field or flux is produced.
- This magnetic field or flux rotates in a space at a speed called synchronous speed.
- The rotating magnetic field or rotating flux has fixed relationship between, the number of poles, the frequency of a.c supply and the speed of rotation.
- The rotating magnetic field creates an effect which is similar to the physical rotation of magnets in space with a synchronous speed.
- So for rotating magnetic field

$$N_s = \frac{120f}{P} \text{ r.p.m.}$$

Where **f** = supply frequency **P** = Number of poles



- Suppose the stator poles are N1 and S1 which are rotating at a speed of Ns and the direction of rotation be clockwise.
- When the field winding on a rotor is excited by the D.C source, it produces the two stationary poles i.e N2 and S2.
- To establish the magnetic locking between the stator and rotor poles the, unlike poles N1 and S2 or N2 and S1 should be brought near to each other.
- As stator poles are rotating and due to magnetic locking the rotor poles will rotate in the same direction of rotating magnetic field as that of stator poles with the same speed Ns.
- Hence synchronous motor rotates at only one speed that is synchronous speed.
- The synchronous speed depends on the frequency therefore for constant supply frequency synchronous motor speed will be constant irrespective of the load changed.

14(b).Derive an expression for the power developed in an synchronous motor. (6M)

Power Developed by Motor :

The mechanical power developed / phase is,

Pm = Back emf * Armature current * Cosine of the angle between Eb and Ia

$= E_b I_a \cos(\alpha - \varphi)$ for lagging p.f

$= E_b I_a \cos(\alpha + \varphi)$ for leading p.f

The copper loss in a **synchronous motor** takes place in the armature windings.

Therefore,

Armature copper loss / phase = $I_a^2 R_a$

Total copper loss = $3 I_a^2 R_a$

By subtracting the **copper loss** from the power input, we obtain the mechanical power developed by a synchronous motor as,

$$P_m = P - P_{cu}$$

For three-phase,

$$P_m = \sqrt{3}\, I_L I_L \cos\varphi - 3 I_a^2 R_a$$

Power Output of the Motor :

To obtain the power output we subtract the iron, friction, and excitation losses from the power developed.

Therefore, Net output power, Pout = Pm - iron, friction, and excitation losses.

The above two stages can be shown diagrammatically called as Power Flow Diagram of a Synchronous Motor

The power developed in a synchronous motor as follows.

Motor Input Power, P

- Stator (Armature) copper loss Pcu
- Mechanical power developed, Pm
  - Iron, friction, and excitation losses
  - Output power, Pout

Net Power Developed by a Synchronous Motor :

The expression for power developed by the synchronous motor in terms of α, θ, V, E_b, and Z_s are as follows :

Let

- V = Supply voltage
- E_b = Back emf / phase
- α = Load angle
- θ = Internal or Impedance angle = Tan^{-1} (X_r / Z_s)
- I_a = Armature current / phase = E_r / Z_s
- $Z_s = R_a + J X_s$ = Synchronous impedance

Mechanical power developed / phase,

$$P_m = \frac{E_b V}{Z_s}\cos(\theta - \alpha) - \frac{E_b^2}{Z_s}\cos\theta$$

The armature resistance is neglected

If R_a is neglected, then $Z_s \approx X_s$ and θ = 90°. substituting these values in the above equation.

$$P_m = \frac{E_b V}{X_s}\cos(90 - \alpha) - \frac{E_b^2}{X_s}\cos 90^0$$

$$P_m = \frac{E_b V}{X_s} \sin\alpha$$

15(a). Explain any two starting methods of synchronous motor in detail (8M)

Methods of Starting of Synchronous Motor

The different methods used to start a synchronous motor are :

Using Pony Motors :

By using the small pony motors like a small induction motor, we can start the synchronous motor. This small induction motor is coupled to the rotor of the synchronous motor. The function of this induction motor is to bring the rotor of the synchronous motor to the synchronous speed.

Once the rotor attains the synchronous speed the pony motor is dis-coupled from the rotor. The synchronous motor continues to rotate at synchronous speed, by supplying d.c. excitation to the rotor through the slip-rings. One should remember that the motor used as the pony motor must have less number of poles than the synchronous motor used.

Using Small D.C. Machine :

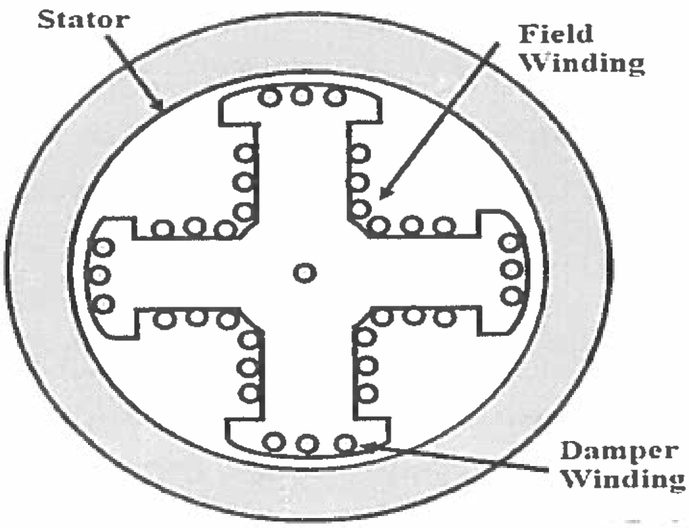
In the above method, we have a seen small induction motor to start the motor. Here we use d.c. motor instead of induction motor to bring the motor to synchronous motor.

Once the d.c. motor brings the rotor of the synchronous motor to synchronous speed. The motor starts acting as the **d.c. generator** and starts giving excitation to the field winding of the synchronous motor.

Using Damper Winding :

When a 3-phase supply is given to the synchronous motor it fails to start. In order to make it start copper bars circuited at both ends (similar to the squirrel cage rotor of an induction motor) are placed on the rotor, these bars or winding are known as 'Damper Winding'.

Now when the supply is given the field winding setups a rotating magnetic field. Due to the damper winding used, the rotor starts rotating as an induction motor i.e., less than the synchronous speed at starting. Once d.c. excitation is given to the field winding and the motor is then pulled into synchronism.

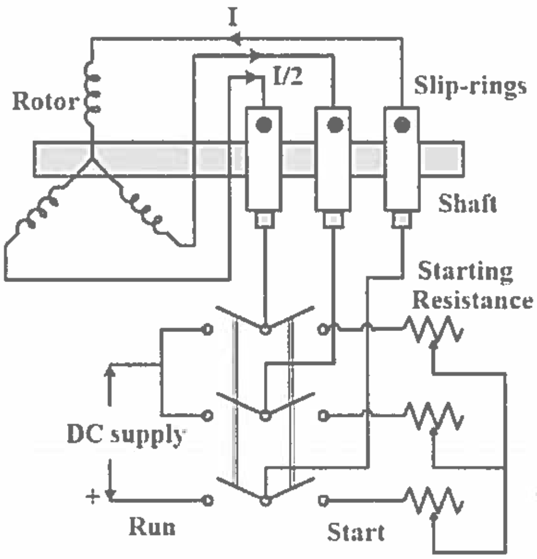


The damper winding is used to start the motor and hence can be used for starting purposes only. Because once the rotor rotates at synchronous speed the relative motion between the damper winding and rotating magnetic will be equal, and hence induced **emf** and current will be zero. The damper winding will be out of the circuit.

As a Slip Ring Induction Motor (Synchronous Induction Motor) :

In this method, an external rheostat is connected to the rotor through slip-rings. Here, ends of the damper winding are brought of the motor and connected either in star or delta. The rheostat is connected in series with the rotor. At starting high resistance is connected with the rotor to limited the current drawn by the motor. As the motor starts as a slip ring induction motor at starting, it draws large currents.

When the motor picks up its speed, resistance is gradually cut off from the rotor circuit. As the speed reaches near to synchronous speed, d.c. excitation is given to the rotor and it is pulled into synchronism.



The above figure shows the rheostat connected with the rotor circuit through slip-rings. From the figure as the dc supply is given current 'I' flows through the positive terminal, then it divides as 'I/2' through each phase at star point.

From these methods, damper winding is the most common method of starting a synchronous motor.

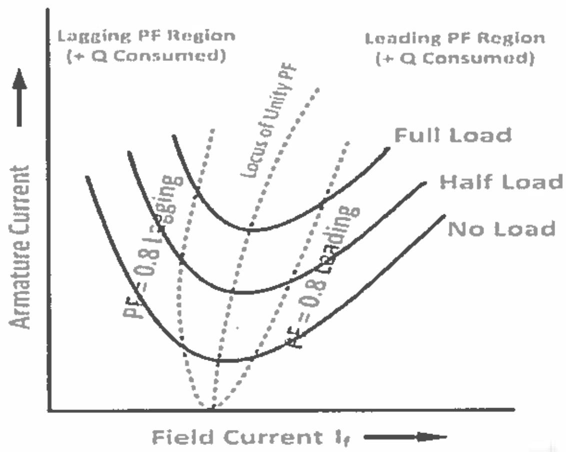
15(b). Illustrate the performance of a synchronous motor using V and inverted V curves. (4M)

V curve is a plot of the stator current versus field current for different constant loads. The Graph plotted between the armature current Ia and field current If at no load the curve is obtained known as V Curve. Since the shape of these curves is similar to the letter "V", thus they are called the V curve of a synchronous motor.

The power factor of the synchronous motor can be controlled by varying the field current If. As we know that the armature current Ia changes with the change in the field current If. Let us assume that the motor is running at NO load. If the field current is increased from this small value, the armature current Ia decreases until the armature current becomes minimum. At this minimum point, the motor is operating at a unity power factor. The motor operates at a lagging power factor until it reaches up to this point of operation.

If now, the field current is increased further, the armature current increases and the motor starts operating as a leading power factor. The graph drawn between armature current and field current is known as the V curve. If this procedure is repeated for various increased loads, a family of curves is obtained.

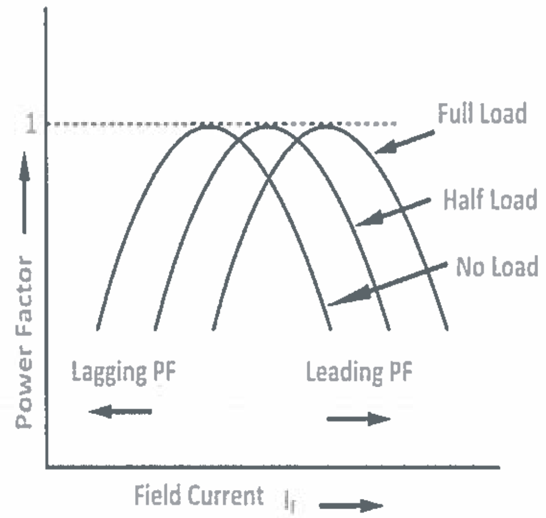
The V curves of a synchronous motor are shown below:



The point at which the unity power factor occurs is at the point where the armature current is minimum. The curve connecting the lowest points of all the V curves for various power levels is called the Unity Power Factor Compounding Curve. The compounding curves for 0.8 power factor lagging and 0.8 power factor leading are shown in the figure above by a red dotted line.

The loci of constant power factor points on the V curves are called Compounding Curves. It shows the manner in which the field current should be varied in order to maintain a constant power factor under changing load. Points on the right and left of the unity power factor corresponds to the over-excitation and leading current and under excitation and lagging current respectively.

The V curves are useful in adjusting the field current. Increasing the field current, If beyond the level for minimum armature current results in leading power factor. Similarly decreasing the field current below the minimum armature current results in a lagging power factor. It is seen that the field current for the unity power factor at full load is more than the field current for unity power factor at no load.The figure below shows the graph between power factor and field current at the different loads.



It is clear from the above figure that, if the synchronous motor at full load is operating at unity power factor, then removal of the shaft load causes the motor to operate at a leading power factor.

KHA Caselabs

NSRIT

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	**Program**	ECE			**Academic Year**	2021 - 2022
Course Code	20EC404	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	IV
Course	**Electromagnetic Waves & Transmission Lines**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define the terms Phase velocity and group velocity.	20EC404.1	L1
2	List any four applications of Smith Chart.	20EC404.2	L1
3	State Gauss's law and write its expression.	20EC404.3	L1
4	What is the significance of boundary conditions?	20EC404.4	L1
5	Define Skin effect and give its expression in medium parameter.	20EC404.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Derive the expression for characteristic impedance in terms of primary constants and angular frequency.	6M	20EC404.1	L3
6 (b)	For a transmission line G = 0.02, C = 10pF, L=0.2μH, R = 0.1Ω. Find the propagation constant and characteristic impedance at 1MHz.	6M	20EC404.1	L3
	OR			
7 (a)	Derive the condition for distortion less transmission line and also plot the open circuit short circuit wave forms of voltage and current at the receiving end.	8M	20EC404.1	L2
7 (b)	A lossless line has Z_0 =50Ω and β= 0.2π m^{-1} at f=60MHz. Find the distributed parameters L and C of the line.	4M	20EC404.1	L2
8 (a)	What is Smith Chart? Explain its significance.	8M	20EC404.2	L2
8 (b)	An arial of (200-j300)Ω is to be matched with 500Ω lines. The matching is to be done by means of a low loss line 600Ω stub line. Find the position and length of the stub line used if the operating wavelength is 20 m.	4M	20EC404.2	L3
	OR			
9 (a)	Characteristic impedance of a low loss transmission line is 90 Ω and it is terminated by another impedance of (130-j980)Ω. The wavelength of the line is 2.6m. Determine: i) VSWR, ii) Maximum impedance and iii) Minimum impedance.	8M	20EC404.2	L3
9 (b)	Write a brief note on the various impedance matching techniques.	4M	20EC404.2	L2
10 (a)	Point charges Q1 and Q2 are respectively located at (4, 0, -3) and (2, 0, 1) if Q2 = 4nc. Find Q1 such that the 'E' at (5, 0,6) has no z – component.	6M	20EC404.3	L3
10 (b)	Obtain the expression for the electric field due to finite length of the conductor along Z axis.	6M	20EC404.3	L3
	OR			
11 (a)	In a certain conducting region, $H=yz(x^2+y^2)a_x-y^2xza_y+4x^2y^2a_z$ mA/m. Determine J at (5,2,-3).	4M	20EC404.3	L2
11 (b)	Find an expression for the magnetic field produced by a straight current carrying conductor of finite length along Z axis.	8M	20EC404.3	L3
12 (a)	Discuss the boundary condition in static electric field at the interface between two perfect dielectric media.	6M	20EC404.4	L2
12 (b)	Given E=10 Sin (ωt-βz)ay V/m in free space, determine D B and H.	6M	20EC404.4	L3

	OR			
13 (a)	Define polarization and explain each of its types.	4M	20EC404.4	L2
13 (b)	Derive Maxwell's equations in Integral and Differential forms for time varying fields.	8M	20EC404.4	L3
14 (a)	Define uniform plane wave. Prove that uniform plane wave does not have field component in the direction of propagation.	6M	20EC404.5	L3
14 (b)	Discuss wave propagation in lossless media and in free space.	6M	20EC404.5	L3
	OR			
15 (a)	An EM wave travels in space with $E=100e^{j(0.866y+0.5z)}\ a_x$ v/m. Determine ω, λ, H.	6M	20EC404.5	L2
15 (b)	Obtain the expressions for Reflection and Transmission coefficients for Normal Incidence of Uniform Plane wave at Dielectric interface.	6M	20EC404.5	L3

ANSWER KEY AND SCHEME OF EVALUATION

Part A (Short Answer Questions 5 x 2 = 10 Marks)

1) Define the terms Phase velocity and group velocity.

Ans) Phase velocity:
It is defined as the velocity in which a signal of single frequency propagates along the length.
It is defined by 'up'.
where

$$\boxed{V_P = f\lambda}$$

$$V_P = f\frac{2\pi}{\beta}$$

$$\boxed{V_P = \frac{\omega}{\beta}}$$

Group Velocity: The velocity at which the envelope the complex wave propagates along a line is called 'Group Velocity'.
It is defined as the ratio of change in angular frequency to change in phase constant of the wave

$$\boxed{V_G = \frac{\omega_2-\omega_1}{\beta_2-\beta_1} = \frac{d\omega}{d\beta}}$$

2) List any four applications of Smith Chart.

Ans) Applications of Smith Chart: Smith Chart can be used for

For evaluating the rectangular components, or the magnitude and phase of an input impedance or admittance, voltage, current, and related transmission functions at all points along a transmission line, including:

- Complex voltage and current reflections coefficients
- Complex voltage and current transmission coefficents
- Power reflection and transmission coefficients
- Reflection Loss
- Return Loss
- Standing Wave Loss Factor
- Maximum and minimum of voltage and current, and SWR
- Shape, position, and phase distribution along voltage and current standing waves

3) State Gauss's law and write its expression.

Gauss's law states that the total electric flux ψ through any closed surface is equal to the total charge enclosed by the surface

Thus $\psi = Q_{enclosed}$ —— (1)

i.e. $\psi = \oint_S d\psi = \oint_S \bar{D}\cdot\bar{ds}$ = total charge enclosed

(or) $Q = \oint_S \bar{D}\cdot\bar{ds} = \int_v \rho_v \, dv$ —— (2)

By applying divergence theorem to the middle term in above

$$\oint \bar{D}\cdot\bar{ds} = \int (\nabla\cdot\bar{D})\,dv \text{ —— (3)}$$

comparing the two equations (2) & (3) yields

$$\boxed{\nabla\cdot\bar{D} = \rho_v}$$

4) **What is the significance of boundary conditions?**
Boundary conditions are practically essential for defining a problem and, at the same time, of primary importance in computational fluid dynamics. It is because the applicability of numerical methods and the resultant quality of computations can critically be decided on how those are numerically treated.

5) **Define Skin effect and give its expression in medium parameter.**
Skin effect is a tendency for alternating current (AC) to flow mostly near the outer surface of an electrical conductor, such as metal wire. The effect becomes more and more apparent as the frequency increases.

The skin depth is $\delta = \sqrt{(2\rho/\omega\mu)}$.

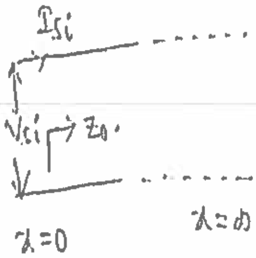
Part B (Long Answer Questions 5 x 12 = 60 Marks)

6 (a) Derive the expression for characteristic impedance in terms of primary constants and angular frequency. 6M

Expression for characteristic Impedance :- (Z_0)

Definition :- characteristic Impedance is defined as the ratio of applied Voltage to the current flowing through the infinite line.

It is denoted by Z_0.



$$Z_0 = \frac{V_{si}}{I_{si}} = \sqrt{\frac{R+j\omega L}{G+j\omega C}}$$

proof :- We know that, from Transmission line equation Concept.

Voltage equation is $-\frac{dV}{dx} = I(R+j\omega L) \rightarrow (1)$

For Infinite line, the propagation equations are

$V = V_{si} e^{-\gamma x} \rightarrow (2)$

$I = I_{si} e^{-\gamma x} \rightarrow (3)$

substitute eq (2) & (3) in eq (1)

$$\Rightarrow -\frac{d}{dx}\left(V_{si} e^{-\gamma x}\right) = I_{si} e^{-\gamma x} (R+j\omega L)$$

$$-V_{si} \cdot \cancel{e^{-\gamma x}} (-\gamma) = I_{si} \cdot \cancel{e^{-\gamma x}} (R+j\omega L)$$

$$V_{si}(\gamma) = I_{si}(R+j\omega L)$$

$$\Rightarrow \frac{V_{si}}{I_{si}} = \frac{R+j\omega L}{\gamma} = \frac{\cancel{R+j\omega L}\ \sqrt{R+j\omega L}}{\cancel{\sqrt{R+j\omega L}}\sqrt{G+j\omega C}}$$

$$\Rightarrow Z_0 = \frac{V_{si}}{I_{si}} = \frac{\sqrt{R+j\omega L}}{\sqrt{G+j\omega C}}$$

$\left(\because \gamma = \sqrt{(R+j\omega L)(G+j\omega C)}\right.$ = propagation constant)

$$\therefore \boxed{Z_0 = \frac{V_{si}}{I_{si}} = \sqrt{\frac{R+j\omega L}{G+j\omega C}}\ \Omega}$$

6 (b) For a transmission line G = 0.02, C = 10pF, L=0.2µH, R = 0.1Ω. Find the propagation constant and characteristic impedance at 1MHz. 6M

Sol : Given data

→ Give data & formulas 2 marks

→ propagation constant 2 marks

→ characteristic Impedance 2 marks

$G = 0.02\ \mho$

$C = 10pF = 10 \times 10^{-12} F$

$L = 0.2\ \mu H = 0.2 \times 10^{-6} H$

$R = 0.1\ \Omega$

$\gamma = \sqrt{(R + j\omega L)(G + j\omega C)}$

$\omega = 2\pi f = 2 \times 3.14 \times 10^6$

$= 6.28 \times 10^6$

$\gamma = \sqrt{(0.1 + j\, 6.28 \times \cancel{10^6} \times 0.2 \times \cancel{10^{-6}})(0.02 + j\, 6.28 \times 10^6 \times 10 \times 10^{-12})}$

$= \sqrt{(0.1 + j\, 6.28 \times 0.2)(0.02 + j}$

$= \sqrt{(1.921 \times 10^{-3} + 0.025\, i)}$

$Z_0 = \sqrt{\dfrac{R + j\omega L}{G + j\omega C}} = \sqrt{\dfrac{0.1 + j \times 6.28 \times 10^6 \times 0.2 \times 10^{-6}}{0.02 + j\, 6.28 \times 10^6 \times 10 \times 10^{-12}}}$

$Z_0 = \sqrt{79617.8 + 10^6\, i}$

7 (a) Derive the condition for distortion less transmission line and also plot the open circuit short circuit wave forms of voltage and current at the receiving end. 8M

Distortionless line ($R/L = G/C$)

- A distortionless line is one in which the attenuation constant α is frequency independent while is phase constant β is linearly dependent on frequency.
- A distortionless line results if the line parameters are such that

$$\boxed{\frac{R}{L} = \frac{G}{C}}$$

- Thus for distortionless line

$$\gamma = \sqrt{RG\left(1+\frac{j\omega L}{R}\right)\left(1+\frac{j\omega C}{G}\right)}$$

$$= \sqrt{RG}\left(1+\frac{j\omega C}{G}\right) = \alpha + j\beta$$

(or) $\alpha = \sqrt{RG}$ $\quad \beta = \omega\sqrt{LC}$

showing that α does not depend on frequency, where β is a linear function of frequency

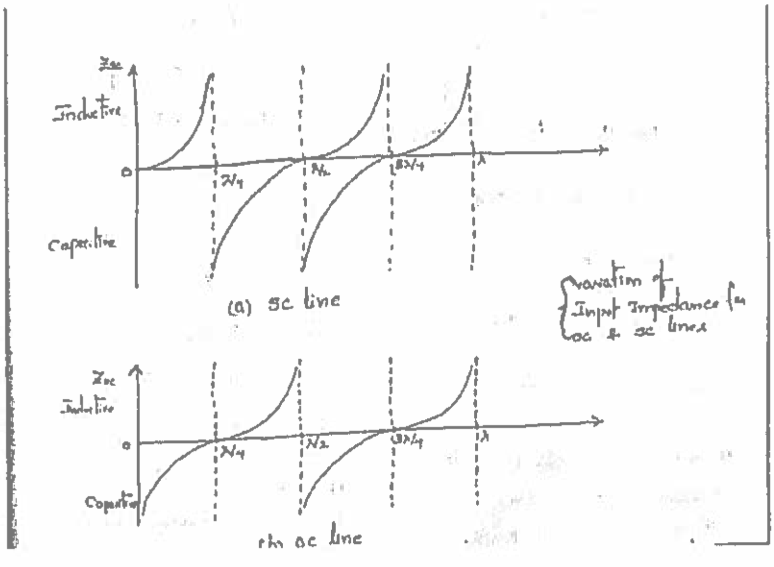
Also $\quad Z_0 = \sqrt{\dfrac{R(1+j\omega L/R)}{G(1+j\omega C/G)}} = \sqrt{\dfrac{R}{G}} = \sqrt{\dfrac{L}{C}} = R_0 + jX_0$

(or) $\quad R_0 = \sqrt{\dfrac{R}{G}} = \sqrt{\dfrac{L}{C}}, \quad X_0 = 0$

and $\quad v = \dfrac{\omega}{\beta} = \dfrac{1}{\sqrt{LC}} = f\lambda$

Distributed line parameters at High frequency

parameters	Coaxial line	Two-wire line
R (Ω/m)	$\frac{1}{2\pi\delta\sigma_c}\left[\frac{1}{a}+\frac{1}{b}\right]$	$\frac{1}{\pi a\delta\sigma_c}$
L (H/m)	$\frac{\mu}{2\pi}\ln\frac{b}{a}$	$\frac{\mu}{\pi}\cosh^{-1}\frac{d}{2a}$
G (S/m)	$\frac{2\pi\sigma}{\ln b/a}$	$\frac{\pi\sigma}{\cosh^{-1}\frac{d}{2a}}$
C (F/m)	$\ln\frac{b}{a}$	$\frac{\pi\varepsilon}{\cosh^{-1}\frac{d}{2a}}$



7 (b) A lossless line has Z 0 =50 and = 0.2 m -1 at f=60MHz. Find the distributed parameters L and C of the line. 4M

Sol Given data

→ Given data & formulae 1 marks

→ L value = 1.5 marks

C value = 1.5 marks

$$\Rightarrow Z_0 = 50\Omega = \sqrt{\frac{L}{C}} \quad \text{— (1)}$$

$$\beta = 0.2\pi \ m^{-1}$$

$$f = 60 \ MHz = 60 \times 10^6$$

$$\Rightarrow V = \frac{\omega}{\beta} = \frac{2\pi f}{\beta} = \frac{\cancel{2}\cancel{\pi} \times 60 \times 10^6}{\underset{0.1}{\cancel{0.2}} \times \cancel{\pi}} = \frac{1}{\sqrt{LC}} \rightarrow (2)$$

$$\text{from eq-(1)} \quad 2500 = \frac{L}{C} \quad \text{— (3)}$$

$$\text{from eq-(2)} \quad 60 \times 10^7 = \frac{1}{\sqrt{LC}} \quad \text{— (4)}$$

substitute eq-(3) in eq-(4), yields

$$60 \times 10^7 = \frac{1}{\sqrt{2500\, C \cdot C}}$$

$$60 \times 10^7 = \frac{1}{50 \times C}$$

$$C = \frac{1}{50 \times 60 \times 10^7}$$

$$C = 0.333 \ pF$$

$$L = 2500 \times 0.33 \ pF$$

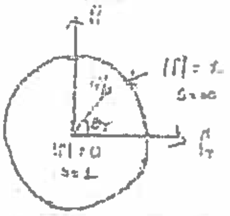
$$= 2500 \times 0.33 \times 10^{-12}$$

$$L = 8.25 \times 10^{-10} \ H$$

8 (a) What is Smith Chart? Explain its significance. 8M

Smith Chart:- It is a polar chart for calculating the transmission line characteristics.

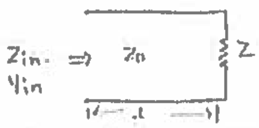
* It is a polar plot of the reflection coefficient in terms of normalized impedance $(r+jx)$.
* It is a graphical plot of normalized resistance and reactance in the reflection co-efficient plane.



* It consists of two sets of orthogonal circles, which represents the value of normalised impedence.
* The one set of circles represent resistive component "r" called "r" circles and other set of circles represent reactive component "x" called "x-circles".

Applications:-

1. It is used to find the input impedance (Z_i) & input admittance (Y_i) of a given transmission line of length "l".



we can easily find the input impedance and admittance of a transmission line of length "l" in Smith chart.

2. It is used to find the reflection coefficient (Γ)

we know that, to find the reflection coefficient (Γ) we want the "Z_0 and Z_L" values but by using smith chart we can find them easily.

i.e $\Gamma = \frac{Z_L - Z_0}{Z_L + Z_0}$

But in smith chart "Γ" is directly calculated.

3. It is used ...

we can easily find the input impedance and admittance of a transmission line of length "l" in Smith chart.

2. It is used to find the reflection coefficient (Γ)

we know that, to find the reflection coefficient (Γ) we want the "Z_0 and Z_L" values but by using smith chart we can find them easily.

i.e $\Gamma = \frac{Z_L - Z_0}{Z_L + Z_0}$

But in smith chart "Γ" is directly calculated.

3. It is used to find VSWR (voltage standing wave ratio) of the line.

As we know that $VSWR = \frac{1+|\Gamma|}{1-|\Gamma|}$

As the "Γ" is directly found in smith chart, so the VSWR is also calculated simply in smithchart.

(iv) We can find the V_{max}, V_{min}, I_{max} & I_{min} along the transmission line using smith chart.

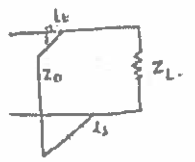
V_{max} = Maximum magnitude of voltage

V_{min} = minimum magnitude of voltage.

I_{max} = maximum magnitude of current

I_{min} = Minimum magnitude of current.

(v) It is used to find the length & locations of stubs in impedance matching.

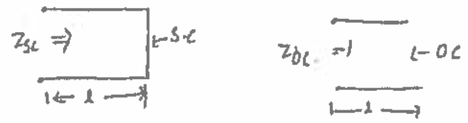


"l_t" = length of stub

"l_s" = location of stub

Irrespective of the mathematical calculation we can find the length & location of stubs very easily using Smith chart.

(vi) It is used to find the input impedance of short-circuit and open circuit line.



8 (b) An arial of (200-j300)Ω is to be matched with 500Ω lines. The matching is to be done by means of a low loss line 600Ω stub line. Find the position and length of the stub line used if the operating wavelength is 20 m. 4M

9 (a) Characteristic impedance of a low loss transmission line is 90 Ω and it is terminated by another impedance of (130-j980)Ω. The wavelength of the line is 2.6m. Determine: i) VSWR, ii) Maximum impedance and iii) Minimum impedance. 8M

Sol

Given data

$Z_0 = 90\Omega$

$Z_L = 130 - j980 \Omega$

$\lambda = 26 m$

$S = ?$

$Z_{max} = ?$

$Z_{min} = ?$

(1) Given data & formulae – 2 marks
(2) Γ – 2 marks
S – 2 marks
(3) Z_{max} – 1 mark
Z_{min} – 1 mark

$$\Gamma = \frac{Z_L - Z_0}{Z_L + Z_0}$$

$$\Gamma = 0.960 + 0.174j$$

i, $$S = \frac{1+|\Gamma|}{1-|\Gamma|}$$

$$= \frac{1+|0.960+0.174j|}{1-|0.960+0.174j|}$$

$$S = \frac{1+0.975}{1-0.975}$$

$$S = 81.8$$

$$Z_{max} = Z_L \times S$$
$$= (130-j980)\times 81.8$$

$$Z_{min} = \frac{Z_L}{S} = \frac{(130-j980)}{81.8}$$

9 (b) Write a brief note on the various impedance matching techniques. 4M

* Stub Matching

⇒ when a UHF line is terminated with a load Impedance which is not Equal to the characteristic Impedance of the line, mismatch occurs.

⇒ mismatch reduces Efficiency and increases power loss.

⇒ To avoid mismatching, it is necessary to add Impedance matching devices between the load and the line.

⇒ A quarter wave transformer can be used to achieve Impedance matching but we have to cut the line to insert

a transformer in between the line and the load.

Definition: A stub is a open (or) short circuited short length Transmission lines used as a matching device, which can be connected in parallel to the line at a distance from the load.

* This matching device is called stub matching.

⇒ Therefore the length of the short circuited stub is

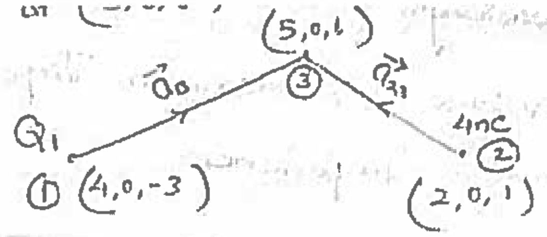
$$l_t = \frac{\lambda}{2\pi} \tan^{-1}\left(\frac{\sqrt{Z_R Z_0}}{Z_R - Z_0}\right) = \frac{\lambda}{2\pi} \tan^{-1} \frac{\sqrt{1-|K|^2}}{2|K|}$$

⇒ The location of the short-circuited stub is

$$l_s = \frac{1}{\beta} \tan^{-1} \sqrt{Z_0/Z_R} = \frac{\phi + \pi - \cos^{-1}|K|}{2\beta}$$

10 (a) Point charges Q1 and Q2 are respectively located at (4, 0, -3) and (2, 0, 1) if Q2 = 4nc. Find Q1 such that the 'E' at (5, 0,6) has no z –component. 6M

Sol



$\boxed{\begin{array}{l} E_{13} = 1 \text{ mark} \\ E_{23} = 1 \text{ mark} \\ E = 2 \text{ marks} \\ Q_1 = 2 \text{ marks} \end{array}}$

$$\vec{E}_{13} = \frac{Q_1}{4\pi\epsilon_0 R_{13}^2}\vec{a}_{13}$$

$$= \frac{Q_1}{4\pi\epsilon_0}\,\frac{1\vec{a}_x + 9\vec{a}_z}{(82)^{3/2}} \quad -①$$

$$\vec{E}_{23} = \frac{4\times10^{-9}}{4\pi\epsilon_0 R_{23}^2}\vec{a}_{23}$$

$$\vec{E}_{23} = \frac{4\times10^{-9}(3\vec{a}_x + 5\vec{a}_z)}{4\pi\epsilon_0 (9+25)^{3/2}}$$

$\vec{E}$ at pt ③ is $\vec{E} = \vec{E}_{13} + \vec{E}_{23}$

$$\vec{E} = \frac{Q_1(\vec{a}_x + 9\vec{a}_z)}{4\pi\epsilon_0 (82)^{3/2}} + \frac{4\times10^{-9}(3\vec{a}_x+5\vec{a}_z)}{4\pi\epsilon_0 (41)^{3/2}}$$

Given that z – component of $\vec{E}$ is Zero.

i.e $$\frac{Q_1\times 9\vec{a}_z}{4\pi\epsilon_0 (82)^{3/2}} + \frac{4\times10^{-9}\times 5\vec{a}_z}{4\pi\epsilon_0 (41)^{3/2}} = 0$$

$$\frac{Q_1\times 9\cancel{\vec{a}_z}}{\cancel{4\pi\epsilon_0}(82)^{3/2}} = -\frac{4\times10^{-9}\times 5\cancel{\vec{a}_z}}{\cancel{4\pi\epsilon_0}(41)^{3/2}}$$

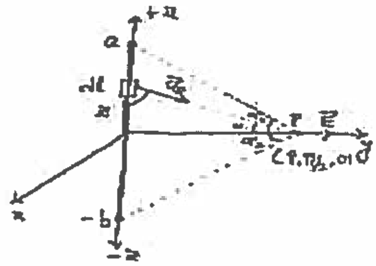
$$Q_1 = -\left(\frac{82}{41}\right)^{3/2}\frac{5}{9} = -(2)^{3/2}\cdot\frac{5}{9} \text{ Coulombs}$$

10 (b) Obtain the expression for the electric field due to finite length of the conductor along Z axis. 6M

Electric field due to a finite line charge

15

1. Consider a line of finite length l with uniform charge density ρ_L placed along the z-axis.
2. Let the end points of the line be at distances a and -b from the origin.
3. Consider a differential charge dq of length dl at a distance z from the origin, as shown in fig.
4. Then, $dq = \rho_L\, dl = \rho_L dz$
5. Let point p be on the y-axis, which is at a distance ρ from the line charge.



Field due to line charge of finite length.

6. Consider cylindrical coordinates.
7. The point of dq is given by coordinates $(0, \pi/2, z)$ and that of point p by $(\rho, \pi/2, 0)$
8. The distance vector between the points $(\rho, \pi/2, 0)$ and $(0, \pi/2, z)$ is

$$\vec{R} = (\rho - 0)\vec{a}_\rho + 0\,\vec{a}_\phi + (0 - z)\vec{a}_z$$

$$\vec{R} = \rho\,\vec{a}_\rho - z\,\vec{a}_z \qquad \text{——— (1)}$$

9. The unit vector is $\vec{a}_R = \dfrac{\vec{R}}{|\vec{R}|} = \dfrac{\rho\vec{a}_\rho - z\vec{a}_z}{\sqrt{\rho^2 + z^2}}$ ——— (2)
10. We know that the differential field strength is

$$\vec{dE} = \frac{\rho_L\, dl}{4\pi\epsilon R^2}\vec{a}_R = \frac{\rho_L\, dz}{4\pi\epsilon(\rho^2+z^2)^{3/2}}(\rho\vec{a}_\rho - z\vec{a}_z) \text{ — (3)}$$

$$= \frac{\rho_L}{4\pi\epsilon}\left[\frac{\rho\, dz\, \vec{a}_\rho}{(\rho^2+z^2)^{3/2}} - \frac{z\, dz\, \vec{a}_z}{(\rho^2+z^2)^{3/2}}\right]$$

The total electric field is in the radial direction from z-axis.

so $\vec{E} = \int_l \vec{dE} = \dfrac{\rho_L}{4\pi\epsilon}\displaystyle\int_{-b}^{a} \dfrac{\rho\,\vec{a}_\rho\, dz}{(\rho^2+z^2)^{3/2}}$ ——— (4)

→ To evaluate the integral, let α be the angle of $\vec{a}_R$ at point; also, let α_1 and α_2 be the angles of the end points a and -b of the line charge makes with point p respectively.

→ From fig, ~~we~~ $z = \rho\tan\alpha$, $a = \rho\tan\alpha_1$, and $-b = \rho\tan\alpha_2$

$$\alpha_1 = \tan^{-1}\left(\frac{a}{\rho}\right), \quad \alpha_2 = \tan^{-1}\left(\frac{-b}{\rho}\right)$$

$$dz = \rho\sec^2\alpha\, d\alpha$$

substituting the values in Eq-4, yields

$$\vec{E} = \frac{\rho_L}{4\pi\epsilon}\int_{\alpha_2}^{\alpha_1}\frac{\rho^2\sec^2\alpha\, d\alpha\,\vec{a}_\rho}{(\rho^2+\rho^2\tan^2\alpha)^{3/2}}$$

$$= \frac{\rho_L}{4\pi\epsilon}\int^{\alpha_1}\frac{\rho^2\sec^2\alpha\, d\alpha\,\vec{a}_\rho}{\rho^3\sec^3\alpha}$$

The total electric field is in the radial direction from z-axis.

$$\text{so} \quad \vec{E} = \int_L d\vec{E} = \frac{\rho_L}{4\pi\epsilon} \int_{-b}^{a} \frac{\rho\, \vec{a}_\rho\, dz}{(\rho^2 + z^2)^{3/2}} \quad \text{———} \; (4)$$

→ To evaluate the integral, let α be the angle of $\vec{a}_R$ at point. Also, let α_1 and α_2 be the angles of the end points a and $-b$ of the line charge makes with point p respectively

→ From fig, ~~z~~ $z = \rho \tan\alpha$, $a = \rho \tan\alpha_1$, and 29/5/1 $\tan\alpha_2$

$$\alpha_1 = \tan^{-1}\left(\frac{a}{\rho}\right), \quad \alpha_2 = \tan^{-1}\left(\frac{-b}{\rho}\right)$$

$$dz = \rho \sec^2\alpha \, d\alpha$$

substituting the values in Eq-4, yields

$$\vec{E} = \frac{\rho_L}{4\pi\epsilon} \int_{\alpha_2}^{\alpha_1} \frac{\rho^2 \sec^2\alpha \, d\alpha \, \vec{a}_\rho}{(\rho^2 + \rho^2 \tan^2\alpha)^{3/2}}$$

$$= \frac{\rho_L}{4\pi\epsilon} \int_{\alpha_2}^{\alpha_1} \frac{\rho^2 \sec^2\alpha \, d\alpha \, \vec{a}_\rho}{\rho^3 \sec^3\alpha}$$

$$= \frac{\rho_L}{4\pi\epsilon\rho} \int_{\alpha_2}^{\alpha_1} \cos\alpha \, d\alpha \, \vec{a}_\rho$$

$$\therefore \quad \vec{E} = \frac{\rho_L}{4\pi\epsilon\rho} \left[(\sin\alpha_1 - \sin\alpha_2)\, \vec{a}_\rho \right].$$

11 (a) In a certain conducting region, H=yz(x 2 +y 2)ax-y 2 xzay+4x 2 y 2 az mA/m. Determine J at (5,2,-3). 4M

11a $\vec{H} = yz(x^2+y^2)\vec{a_x} - y^2xz\vec{a_y} + 4x^2y^2\vec{a_z}$

Determine J at $(5,2,-3)$

Sol: we know Relation

$$\nabla \times \vec{H} = \vec{J}$$

Relation 1 marks
curl - 2 marks
Substitution of 1 mark
point

$$\nabla \times \vec{H} = \begin{vmatrix} \vec{a_x} & \vec{a_y} & \vec{a_z} \\ \frac{\delta}{\delta x} & \frac{\delta}{\delta y} & \frac{\delta}{\delta z} \\ yz(x^2+y^2) & -y^2xz & 4x^2y^2 \end{vmatrix}$$

$$\nabla \times \vec{H} = \vec{a_x}\left[\frac{\delta}{\delta y}(4x^2y^2) + \frac{\delta}{\delta z}(y^2xz)\right]$$
$$- \vec{a_y}\left[\frac{\delta}{\delta x}(4x^2y^2) - \frac{\delta}{\delta z}(yz(x^2+y^2))\right]$$
$$+ \vec{a_z}\left[\frac{\delta}{\delta x}(-y^2xz) - \frac{\delta}{\delta z}(yz(x^2+y^2))\right].$$

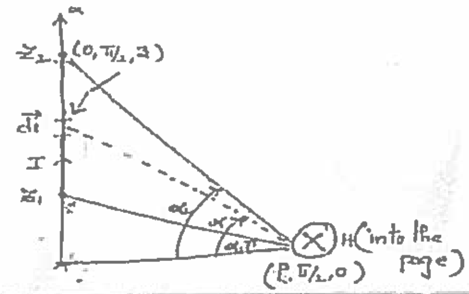
$$\nabla \times \vec{H} = \vec{a_x}\left[8x^2y + y^2x\right] - \vec{a_y}\left[8xy^2 - y(x^2+y^2)\right]$$
$$+ \vec{a_z}\left[-y^2z - y(x^2+y^2)\right] = \vec{J}$$

$$\vec{J}\Big|_{(5,2,-3)} = \vec{a_x}\left[8\times 25\times 2 + 4\times 5\right] - \vec{a_y}\left[8\times 5\times 4 - 2(25+4)\right]$$
$$+ \vec{a_z}\left[-4\times(-3) - 2\,(25+4)\right]$$

11 (b) **Find an expression for the magnetic field produced by a straight current carrying conductor of finite length along Z axis. 8M**

Sol

1. Consider a straight current carrying filamentary conductor of finite length $z_1 z_2$ as shown in figure.



2. Let the conductor is along the z-axis with its upper and lower ends, respectively, subtending angles α_2 and α_1 at P, the point at which $\vec{H}$ is to be determined.

3. Here current flows from z_1 to z_2

4. Let the contribution of $\vec{dH}$ at P due to element $\vec{dl}$

$$\vec{dH} = \frac{I\,\vec{dl} \times \vec{R}}{4\pi R^3} \quad \text{--- (1)}$$

But $\vec{dl} = dz\,\vec{a_z}$ and $\vec{R} = \rho\vec{a_\rho} - z\vec{a_z}$

$$\vec{dl} \times \vec{R} = \rho\, dz\, \vec{a_\phi} \quad \text{--- (2)}$$

Hence $$\vec{H} = \int_{z_1}^{z_2} \frac{I\rho\, dz}{4\pi[\rho^2+z^2]^{3/2}}\,\vec{a_\phi} \quad \text{--- (3)}$$

Let $z = \rho \tan\theta$ when $z = z_1$ then $\theta = \alpha_1$

$dz = \rho \sec^2\theta\, d\theta$ $z = z_2$ then $\theta = \alpha_2$

$$= \frac{I}{4\pi}\int_{\alpha_1}^{\alpha_2} \frac{\rho^2 \sec^2\theta\, d\theta}{\rho^3 \sec^3\theta}\,\vec{a_\phi}$$

$$= \frac{I}{4\pi\rho}\int_{\alpha_1}^{\alpha_2} \cos\theta\, d\theta\, \vec{a_\phi}$$

$$= \frac{I}{4\pi\rho}\, \sin\theta \Big|_{\alpha_1}^{\alpha_2} \vec{a_\phi}$$

(or) $$\vec{H} = \frac{I}{4\pi\rho}\left[\sin\alpha_2 - \sin\alpha_1\right]\vec{a_\phi} \quad \text{--- (4)}$$

$\vec{H}$ is always along the unit vector $\vec{a_\phi}$.

<u>Special case I</u>: when the conductor is semi-infinite (with respect to P) so that point z_1 is now at $(0,0,0)$ while z_2 is at $(0,0,\infty)$ and $\alpha_1 = 0$ and $\alpha_2 = 90^\circ$; then

Special case I : when the conductor is semi-Infinite (with respect to P) so that point z_1 is now at $(0,0,0)$ while z_2 is at $(0,0,\infty)$ and $\alpha_1 = 0$ and $\alpha_2 = 90°$; then eq-4 becomes

$$\vec{H} = \frac{I}{4\pi\rho}\vec{a}_\phi \quad \text{—— (5)}$$

Special case II when the conductor is infinite in length. For this case, point z_1 is at $(0,0,-\infty)$ while z_2 is at $(0,0,\infty)$; $\alpha_1 = -90°$, $\alpha_2 = 90°$ eq-(4) Reduces to

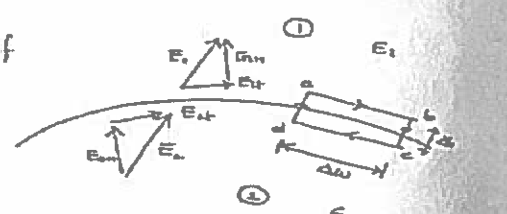
$$\vec{H} = \frac{I}{2\pi\rho}\vec{a}_\phi$$

12 (a) **Discuss the boundary condition in static electric field at the interface between two perfect dielectric media. 6M**

and $\oint \vec{D}\cdot d\vec{s} = Q_{enc}$ —— (2)

Dielectric - Dielectric Boundary Conditions

→ Consider the $\vec{E}$ field existing in a region that consists of two different dielectrics characterized by $\epsilon_1 = \epsilon_0\epsilon_{r1}$ and $\epsilon_2 = \epsilon_0\epsilon_{r2}$ as shown in a fig (a)



(a) Dielectric - dielectric boundary determining $E_{1t} = E_{2t}$

→ The fields $\vec{E}_1$ and $\vec{E}_2$ in media 1 and 2 respectively can be decomposed as

$$\vec{E}_1 = \vec{E}_{1t} + \vec{E}_{1n} \quad \text{—— (1)}$$
$$\vec{E}_2 = \vec{E}_{2t} + \vec{E}_{2n} \quad \text{—— (2)}$$

By applying $\oint \vec{E}\cdot d\vec{l} = 0$ to the closed path abcda of fig (a).

$$0 = E_{1t}\Delta w - E_{1n}\frac{\Delta h}{2} - E_{2n}\frac{\Delta h}{2} - E_{2t}\Delta w + E_{2n}\frac{\Delta h}{2} + E_{1n}\frac{\Delta h}{2} \quad \text{—— (3)}$$

As $\Delta h \to 0$, Eq-(3) becomes

$$\boxed{E_{1t} = E_{2t}} \quad \text{—— (4)}$$

→ Thus the tangential components of $\vec{E}$ are the same on the both sides of the boundary

⇒ Since $\vec{D} = \epsilon\vec{E} = \vec{D}_t + \vec{D}_n$, Eq-(3) can be written as

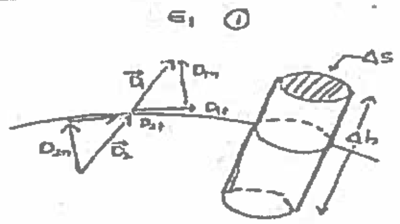
$$\frac{D_{1t}}{\epsilon_1} = E_{1t} = E_{2t} = \frac{D_{2t}}{\epsilon_2}$$

(or)

$$\boxed{\frac{D_{1t}}{\epsilon_1} = \frac{D_{2t}}{\epsilon_2}} \quad \text{—— (5)}$$

D_t undergoes some change across the Interface. Hence D_t is said to be discontinuous across the Interface.

*→ similarly, By applying $\oint \vec{D}\cdot d\vec{s} = Q$ to the pillbox (cylindrical Gaussian surface) of Fig (b).



→ The contribution due to the sides vanishes. As $\Delta h \to 0$

$$\Delta Q = \rho_s \Delta S = D_{1n}\Delta S - D_{2n}\Delta S$$

fig (b) determining $D_{1n} = D_{2n}$

(or) $\boxed{D_{1n} - D_{2n} = \rho_s}$ —— (6)

where ρ_s is the ~~free~~ surface charge density

12 (b) Given E=10 Sin (ωt-βz)ay V/m in free space, determine D B and H. 6M

12) b) Given $\vec{E} = 10 \sin(\omega t - \beta z)\,\vec{a_y}$ V/m in free space, determine $\vec{D}, \vec{B}$ and $\vec{H}$?

Sol Given data $\vec{E} = 10 \sin(\omega t - \beta z)\vec{a_y}$

(1) We know that

$$\vec{D} = \epsilon_0 \vec{E}$$

$$= 8.854 \times 10^{-12} \times 10 \sin(\omega t - \beta z)\vec{a_y}$$

(2) from faradays law.

$\vec{D}$ - 1 marks
Curl, $\vec{B}$ - 4 marks
$\vec{H}$ - 1 marks

$$\nabla \times \vec{E} = -\frac{\delta \vec{B}}{\delta t}$$

y component of $\vec{E}$ is present

$$-\frac{\delta \vec{B}}{\delta t} = \begin{vmatrix} \vec{a_x} & \vec{a_y} & \vec{a_z} \\ \frac{\delta}{\delta x} & \frac{\delta}{\delta y} & \frac{\delta}{\delta z} \\ 0 & E_y & 0 \end{vmatrix}$$

$$= \vec{a_x}\frac{\delta}{\delta x}E_y + \vec{a_z}\left(-\frac{\delta E_y}{\delta z}\right)$$

$$= 0 \quad +(-\vec{a_z})\,\epsilon_0 10\cos(\omega t - \beta z)(-\beta)$$

$$\frac{\delta \vec{B}}{\delta t} = -\epsilon_0 \beta 10 \cos(\omega t - \beta z)\,\vec{a_z}$$

$$\vec{B} = -\int \epsilon_0 \beta 10 \cos(\omega t - \beta z)\,\vec{a_z}\,dt$$

$$\vec{B} = -\frac{\beta \epsilon_0 10 \sin(\omega t - \beta z)}{\omega}\,\vec{a_z} \quad \text{Tesla}$$

$$\vec{B} = -\frac{\cancel{2\pi}}{\lambda} \cdot \frac{\epsilon_0}{\cancel{2\pi} f} 10 \sin(\omega t - \beta z) \quad \text{wb/m}^2$$

$$\vec{B} = -\frac{\epsilon_0}{3\times10^8} 10 \sin(\omega t - \beta z)\,\vec{a_z}$$

B (3) from relation $\vec{B} = \mu_0 \vec{H}$

$$\vec{H} = \frac{\vec{B}}{\mu_0}$$

$$\vec{H} = -\frac{\epsilon_0}{\mu_0} \times \frac{10 \sin(\omega t - \beta z)\vec{a_z}}{3\times10^8}$$

13 (a) Define polarization and explain each of its types. 4M

8 Polarization of a uniform plane wave

polarization of a uniform plane wave is defined as the time-varying behaviour of the electric field strength, $\vec{E}$ at a given fixed point in space.

There are three types of polarization

a) Linear polarization
b) Elliptical polarization
c) Circular polarization

a) Linear polarization: A wave is said to be linearly polarized, if the electric field strength remains along a straight line at a given point in space.

Linear polarization: It is divided into two types
(a) horizontal polarization
(b) vertical polarization

(11)

Consider a uniform plane wave travelling in the z-direction, with electric and magnetic fields lying in the x-y plane.

→ The electric field vector in time domain for linear polarization is

$$\vec{E}(z,t) = E_x e^{-\alpha z}\cos(\omega t-\beta z)\,\vec{a}_x + E_y e^{-\alpha z}\cos(\omega t-\beta z)\,\vec{a}_y$$
$$= E_x\vec{a}_x + E_y\vec{a}_y$$

→ Thus the electric field strength $\vec{E}$ has two components E_x and E_y.

→ If E_x is present and E_y is zero, then wave is said to be polarized in the x-direction.

→ Similarly, if E_y is present and E_x is zero, the wave is said to be polarized in the y direction. This phenomenon is called vertical polarization.

(a) Horizontal polarization   vertical polarization   polarized at angle θ

Circular polarization: If E_x and E_y have equal magnitude with 90° phase difference, then the locus of the resultant vector $\vec{E}$ is a circle.

→ Consider a uniform plane wave travelling in the z direction in a lossless medium $\alpha = 0$.

→ The Electric field vector in the time domain is

$$\vec{E}(z,t) = E_x\cos(\omega t-\beta z)\,\vec{a}_x + E_y\cos(\omega t-\beta z+\phi)\,\vec{a}_y \quad \text{—— (1)}$$

where the $\vec{E}_y$ component leads the $\vec{E}_x$ component by an angle ϕ

→ The electric field vector lies in the x-y plane. The vector $\vec{E}$ has two components E_x and E_y along the x- and y-axes respectively.

$$\therefore\ \vec{E}(z,t) = E_x\vec{a}_x + E_y\vec{a}_y \quad \text{—— (2)}$$

→ For circular polarization, E_x and E_y have equal magnitudes and $\phi = 90°$ phase difference

i.e. $E_1 = E_2 = E$ (say).

Elliptical polarization

→ If $\vec{E}_x$ and $\vec{E}_y$ have different magnitudes with a 90° phase difference, the resultant vector $\vec{E}$ is an ellipse, then the locus of the resultant vector is called an elliptical polarized wave.

→ The wave is called an elliptical polarized wave.

→ The wave is travelling in the z-direction.

Proof: Consider a uniform plane wave travelling in a lossless medium $\alpha = 0$

The electric field vector in time domain is

$$\vec{E}(z,t) = E_1\cos(\omega t-\beta z)\vec{a}_x + E_2\cos(\omega t-\beta z+\phi)\vec{a}_y$$

where the $\vec{E}_y$ component leads the E_x component by an angle ϕ.

The electric field vector lies in the x-y plane.

The vector $\vec{E}$ has two components $\vec{E}_x$ and $\vec{E}_y$ along the x- and y-axes respectively.

13 (b) Derive Maxwell's equations in Integral and Differential forms for time varying fields. 8M

Gauss's law states that the total Electric flux ψ through any closed surface is equal to the total charge enclosed by that surface

Thus $\quad \psi = Q_{enclosed}$ —— (1)

i.e $\quad \psi = \oint_S d\psi = \oint \vec{D}\cdot\vec{ds}$ = total charge enclosed.

(or) $\quad Q = \oint_S \vec{D}\cdot\vec{ds} = \int_V \rho_v \, dv$ — (2)

By applying divergence theorem to the middle term in above equation

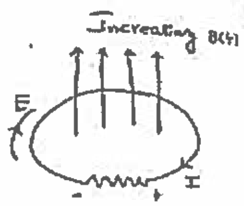
$$\oint \vec{D}\cdot\vec{ds} = \int (\nabla\cdot\vec{D})\, dv \quad \text{— (3)}$$

Comparing the Two equations (2) & (3) yields

$$\boxed{\nabla\cdot\vec{D} = \rho_v}$$

<u>Transformer EMFs</u>

1. Consider a stationary conducting loop is in a time varying magnetic $\vec{B}$ field



Induced emf due to a stationary loop in a time varying B field

$V_{emf} = \oint \vec{E}\cdot\vec{dl} = -\int \frac{\delta\vec{B}}{\delta t}\cdot\vec{ds}$ — (1)

2. This emf Induced by the time-varying current (producing the time-varying $\vec{B}$ field) in a stationary loop is often referred to as a transformer emf in power analysis since it is due to transformer action.

3. By applying stoke's theorem to the middle term in eq-(1)

We get $\quad \boxed{\int_S (\nabla\times\vec{E})\cdot\vec{ds} = -\int_S \frac{\delta\vec{B}}{\delta t}\cdot\vec{ds}}$ — (2)

For the two Integrals to be equal, their Integrand must be equal, that is

$$\boxed{\nabla\times\vec{E} = -\frac{\delta\vec{B}}{\delta t}} \text{ —— (3)}$$

→ This is one of the maxwell's equation for time-varying fields

1. For static EM fields $\nabla \times \vec{H} = \vec{J}$ — (1)
2. But the divergence of the curl of any vector field is identically zero
   Hence $\nabla \cdot (\nabla \times \vec{H}) = 0 = \nabla \cdot \vec{J}$ — (2)
3. But according to continuity equation
$$\nabla \cdot \vec{J} = -\frac{\delta \rho_v}{\delta t} \neq 0 \quad \text{— (3)}$$
4. Eq-(2) and (3) are incompatible for time-varying conditions. We must modify Eq-(1) to agree with Eq-(3)
5. We add a term to Eq-(1), so it becomes
$$\nabla \times \vec{H} = \vec{J} + J_d \quad \text{— (4)}$$
   where J_d has to be defined
6. The divergence of the curl of any vector is zero
   Hence $\nabla \cdot (\nabla \times \vec{H}) = 0 = \nabla \cdot (\vec{J} + J_d)$
$$= \nabla \cdot \vec{J} + \nabla \cdot J_d \quad \text{— (5)}$$
$$\nabla \cdot \vec{J_d} = -\nabla \cdot \vec{J} = +\dot{\rho}_v$$
   So. $\nabla \cdot J = \dot{\rho}_v = \frac{\delta \rho_v}{\delta t} = \frac{\delta}{\delta t}(\nabla \cdot D) = \nabla \cdot \frac{\delta \vec{D}}{\delta t}$
   (or) $\boxed{J_d = \frac{\delta \vec{D}}{\delta t}}$ — (6)
7. Substituting Eq-(6) in Eq-(4) results in
$$\nabla \times \vec{H} = \vec{J} + \frac{\delta \vec{D}}{\delta t}$$
   This is Maxwell's Equation (Based on Ampere's Circuit law) for time varying fields

__Fourth Maxwell's Equation__

An Isolated magnetic charge does not exist

* The total flux through a closed surface in a magnetic field is equal to zero.

   i.e $\boxed{\oint \vec{B} \cdot \vec{ds} = 0}$ — (1)

* Above equation is called law of conservation of magnetic field (or) Gauss's law for magnetostatics
* Here magnetostatic field is not conservative, magnetic flux is constant.
* By applying divergence theorem to Eq-(1), we get.
$$\oint_S \vec{B} \cdot \vec{ds} = \int_V \nabla \cdot \vec{B}\, dv = 0$$
   (or) $\boxed{\nabla \cdot \vec{B} = 0}$

This Equation is called fourth Maxwell's Equations.

14 (a) Define uniform plane wave. Prove that uniform plane wave does not have field component in the direction of propagation. 6M

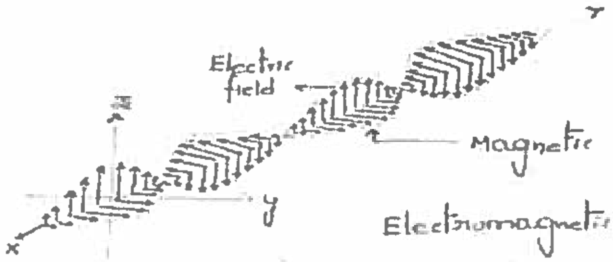
Uniform plane wave equation

Definition : An EM wave propagating in x- direction is said to be a uniform plane if its E and H are Independent of y and z- directions.

→ A uniform plane wave propagating in x-direction has no x- components of E and H, that is $E_x = 0$, $H_x = 0$

→ The electric and magnetic fields of an EM wave are always perpendicular to each other.

→ A typical wave is shown in fig



Electromagnetic wave

proof :

The plane wave equation in free space is given by

$$\nabla^2 \vec{E} = \mu_0 \epsilon_0 \ddot{\vec{E}}$$

that is

$$\frac{\delta^2 \vec{E}}{\delta x^2} + \frac{\delta^2 \vec{E}}{\delta y^2} + \frac{\delta^2 \vec{E}}{\delta z^2} = \mu_0 \epsilon_0 \frac{\delta^2 \vec{E}}{\delta t^2}$$

As per the definition of uniform plane wave,

$$E \neq f(z),\ E \neq f(y)$$

$$\frac{\delta^2 \vec{E}}{\delta z^2} = 0,\ \frac{\delta^2 \vec{E}}{\delta y^2} = 0$$

Hence the wave equation becomes, ③

$$\frac{\delta^2 \vec{E}}{\delta x^2} = \mu_0 \epsilon_0 \frac{\delta^2 \vec{E}}{\delta t^2}$$

that is

$$\frac{\delta^2 E_x}{\delta x^2} \vec{a}_x + \frac{\delta^2 E_y}{\delta x^2} \vec{a}_y + \frac{\delta^2 E_z}{\delta z^2} \vec{a}_z = \mu_0 \epsilon_0 \left[\frac{\delta^2 E_x}{\delta t^2} \vec{a}_x + \frac{\delta^2 E_y}{\delta t^2} \ldots\right.$$

Equating the respective components on both sides, we get

$$\frac{\delta^2 E_x}{\delta x^2}\bar{a}_x + \frac{\delta^2 E_y}{\delta x^2}\bar{a}_y + \frac{\delta^2 E_z}{\delta z^2}\bar{a}_z = \mu_0\epsilon_0\left[\frac{\delta^2 E_x}{\delta t^2}\bar{a}_x + \frac{\delta^2 E_y}{\delta t^2}\bar{a}_y + \frac{\delta^2 E_z}{\delta t^2}\bar{a}_z\right]$$

Equating the respective components on both sides, we get

$$\frac{\delta^2 E_x}{\delta x^2} = \mu_0\epsilon_0 \frac{\delta^2 E_x}{\delta t^2} \quad \text{--- } 1(a)$$

$$\frac{\delta^2 E_y}{\delta x^2} = \mu_0\epsilon_0 \frac{\delta^2 E_y}{\delta t^2} \quad \text{--- } 1(b)$$

$$\frac{\delta^2 E_z}{\delta x^2} = \mu_0\epsilon_0 \frac{\delta^2 E_z}{\delta t^2} \quad \text{--- } 1(c)$$

We know that $\nabla\cdot\vec{D} = 0$ [free space] [as $\rho_v = 0$]

(or) $\nabla\cdot\epsilon_0\vec{E} = 0$

that is $\nabla\cdot\vec{E} = 0$

So $\frac{\delta E_x}{\delta x} + \frac{\delta E_y}{\delta y} + \frac{\delta E_z}{\delta z} = 0$

As $\frac{\delta E_y}{\delta y} = 0$, $\frac{\delta E_z}{\delta z} = 0$, we have

$$\frac{\delta E_x}{\delta x} = 0 \quad \text{--- } (2)$$

Substituting Eq-(2) in Eq-1(c), we get

$$\frac{\delta^2 E_x}{\delta t^2} = 0$$

This means that E_x shold have one of the following solutions

1. $E_x = 0$
2. $E_x = a$
3. E_x increases uniformly with time

If $E_x = a$ constant and $E_x = kt$, it will not be a part of wave motion.

Therefore $\boxed{E_x = 0}$

Similarly $\boxed{H_x = 0}$

→ This means that the components of Electric and magnetic fields of a uniform plane wave in the direction of propagation are zero.

14 (b) Discuss wave propagation in lossless media and in free space. 6M

Wave propagation in Lossless Medium

The wave equation is $\nabla^2 \vec{E} = -\omega^2 \mu \epsilon \vec{E}$

that is $$\frac{\delta^2 \vec{E}}{\delta x^2} = -\omega^2 \mu \epsilon \vec{E}$$

(or) $$\frac{\delta^2 \vec{E}}{\delta x^2} = -\beta^2 \vec{E}$$

where $\beta = \omega\sqrt{\mu \epsilon}$

The y-component of $\vec{E}$ may be written as

$$E_y = A e^{-j\beta x} + B e^{j\beta x}$$

where A and B are arbitrary complex constants.

Then $$\tilde{E}_y(x,t) = Re\{E_y(x) e^{j\omega t}\}$$
$$= Re\{A e^{j(\omega t - \beta x)} + B e^{j(\omega t + \beta x)}\}$$

If A and B are real, it becomes

$$\tilde{E}_y(x,t) = A \cos(\omega t - \beta x) + B \cos(\omega t + \beta x)$$

⇒ This is the sum of two waves. They travel in opposite direction.

⇒ If A = B, the wave combine together and form a standing wave. Such waves do not progress

The wave equation in free space is

$$\nabla^2 \vec{E} = \mu_0 \epsilon_0 \ddot{\vec{E}}$$

Applying the conditions of uniform plane wave equation. the above equation becomes

$$\nabla^2 \vec{E} = \frac{\delta^2 \vec{E}}{\delta x^2} \qquad [\text{as } E \neq f(y),\ E = f(z)]$$

$$\frac{\delta^2 \vec{E}}{\delta x^2} = \mu_0 \epsilon_0 \frac{\delta^2 \vec{E}}{\delta t^2} \text{ --- } (1)$$

Equating the respective components on either side and $E_x = 0$, we have

$$\frac{\delta^2 E_y}{\delta x^2} = \mu_0 \epsilon_0 \frac{\delta^2 E_y}{\delta t^2}$$

$$\frac{\delta^2 E_z}{\delta x^2} = \mu_0 \epsilon_0 \frac{\delta^2 E_z}{\delta t^2}$$

Eq-(1) has a general solution given by

$$\vec{E} = f_1(x - v_0 t) + f_2(x + v_0 t)$$

where f_1 and f_2 are fuctions of $(x - v_0 t)$ and $(x + v_0 t)$ respectively

$$v_0 = \frac{1}{\sqrt{\mu_0 \epsilon_0}} = \text{velocity of propagation}$$

x is the direction of propagation of the wave

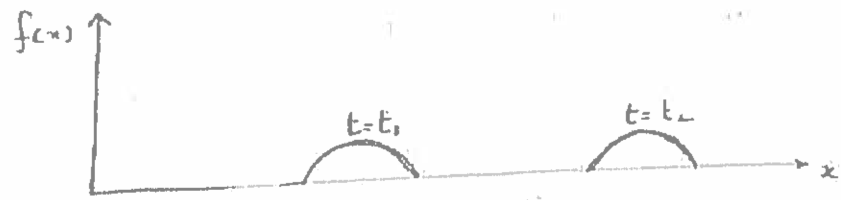
$f_1(x - v_0 t)$ represents a forward wave &

$f_2(x + v_0 t)$ represents a reflected wave

⇒ This reflected wave is present when there is a conductor which acts as a reflector. otherwise, it is absent. As we are considering free space propagation. $\vec{E}$ will be $f_1(x - v_0 t)$ only. That is

$$\boxed{\vec{E} = f(x - v_0 t)}$$

⇒ This the solution of uniform plane wave equation in free space. The behaviour is represented typically



15 (a) An EM wave travels in space with E=100e j(0,866y+0.5z) a x v/m. Determine , λ, H. 6M

$$E = 100 e^{(0.866y + 0.5z)} a_x \ V/m$$

$$\boxed{\omega = 0.866}$$

$$\beta = 0.5$$

$$\frac{2\pi}{\lambda} = 0.5$$

$$\lambda = \frac{2\pi}{0.5}$$

$$\boxed{\lambda = 4\pi}$$

$$\nabla \times E = -\frac{\partial B}{\partial t}$$

$$\begin{vmatrix} \vec{a_x} & \vec{a_y} & \vec{a_z} \\ \frac{\partial}{\partial x} & \frac{\partial}{\partial y} & \frac{\partial}{\partial z} \\ E & 0 & 0 \end{vmatrix} = -\frac{\partial \vec{B}}{\partial t}$$

$$-\vec{a_y}\frac{\partial}{\partial z}(E) + \vec{a_z}\frac{\partial E}{\partial y} = -\frac{\partial B}{\partial t}$$

$$-\vec{a_y}\, 0.5 \vec{E} = -\frac{\partial B}{\partial t}$$

$$\frac{\partial B}{\partial t} = 2E \ a_y$$

$$\vec{B} = 2\int E \, dt \ \vec{a_y}$$

$$\vec{B} = 2\int 100 \ e^{(0.866y + 0.5z)} \vec{a_y} \ V/m$$

$$\vec{B} = \frac{2 \times 100}{0.866} e^{(0.866y + 0.5z)} \vec{a_y}$$

$$\vec{B} = \frac{200}{.866} \ 230 \ e^{(0.866y + 0.5z)} \vec{a_y}$$

$$\boxed{\vec{H} = \frac{230}{\mu} e^{(0.866y + 0.5z)} \vec{a_y}}$$

15 (b) Obtain the expressions for Reflection and Transmission coefficients for Normal Incidence of Uniform Plane wave at Dielectric interface. 6M

- When an EM wave is incident normally on the surface of a dielectric, reflection and transmission takes place.
- For a perfect dielectric, $\sigma = 0$. Hence, there is no loss or no absorption of energy in it.

Reflection coefficient: It is defined as the ratio of reflected wave and incident wave.

$$\text{Reflection coefficient} = \frac{\text{reflected wave}}{\text{incident wave}}$$

$$\text{Reflection coefficient for } E, \ \Gamma_E = \frac{E_r}{E_i}$$

where E_r = reflected electric field
E_i = incident electric field
H_r = reflected magnetic field
H_i = incident magnetic field.

Transmission Coefficient: It is defined as the ratio of transmitted wave and incident wave

$$\text{Transmission coefficient} = \frac{\text{transmitted wave}}{\text{incident wave}}$$

$$\text{Transmission coefficient of } E = T_E = \frac{E_t}{E_i}$$

$$\text{Transmission coefficient for } H \text{ is}, \ T_H = \frac{H_t}{H_i}$$

Expression for reflection coefficient are

$$\Gamma_E = \frac{E_r}{E_i} = \frac{\eta_2 - \eta_1}{\eta_2 + \eta_1} \qquad \Gamma_H = \frac{\eta_1 - \eta_2}{\eta_1 + \eta_2}$$

Expression for Transmission coefficient

$$T_E = \frac{E_t}{E_i} = \frac{2\eta_2}{\eta_1 + \eta_2} \qquad T_H = \frac{2\eta_1}{\eta_1 + \eta_2}$$

where η_1 and η_2 are intrinsic Impedances of medium 1 and 2 respectively.

Susatton

NSRIT

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	Program	CSE/CSE(AI&ML)/CS(DS)			Academic Year	2021 - 2022
Course Code	20CS302	Test Duration	3 Hrs.	Max. Marks	70	Semester	IV
Course	Operating Systems						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define operating system.	20CS404.1	L1
2	Define process.	20CS404.2	L1
3	What is safe state?	20CS404.3	L1
4	Define Thrashing.	20CS404.4	L1
5	List any two file attributes.	20CS404.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	What are the functionalities of Operating Systems? Explain in detail	6M	20CS404.1	L2
6 (b)	Explain the following Operating Systems concepts a. Multi - Programming b. Multi -Tasking	6M	20CS404.1	L2
	OR			
7	What is system call? Explain various system calls in detail	12M	20CS404.1	L2
8	Consider the following four processes, with the length of the CPU burst time given in milliseconds, time slice = 3 ms Process / Arrival Time(ms) / Burst Time (ms) P_1 / 1 / 6 P_2 / 1 / 5 P_3 / 2 / 5 P_4 / 2 / 3 Compute Average Waiting Time for a given process using FCFS, SJF and RR Algorithms.	12M	20CS404.2	L3
	OR			
9 (a)	Explain about the different types of Schedulers in detail.	6M	20CS404.2	L2
9 (b)	Define Process, explain different Process states with a neat diagram.	6M	20CS404.2	L2
10 (a)	Explain how dinning philosopher's problem is solved using Semaphores with an example.	6M	20CS404.3	L3
10 (b)	Explain the necessary condition for deadlock.	6M	20CS404.3	L2
	OR			
11	Consider the following page reference string : 7, 2, 3, 1, 2, 5, 3, 4, 6, 7, 7, 1, 0, 5, 4, 6, 2, 3, 0, 1 Assume demand paging with three frames, how many page faults would occur for the following page replacement Algorithms 1. LRU replacement 2. FIFO replacement	12M	20CS404.3	L3
12 (a)	Write Short notes on Segmentation.	6M	20CS404.4	L1
12 (b)	Discuss about Demand paging.	6M	20CS404.4	L2
	OR			
13	Suppose that a disk drive has 5000 cylinders numbered 0 to 4999. The drive is currently serving a request at cylinder 143. The queue of pending requests in FIFO order 86,1470,913,1774,948,1509, 1022, 1750, 130 starting from current head position. Determine the total distance that disk arm moves to satisfy all the pending request for FCFS, SSTF, SCAN, C-SCAN disk scheduling algorithm.	12M	20CS404.4	L2

14(a)	Discuss about different file access methods.	6M	20CS404.5	L2
14(b)	Explain in detail about allocation methods.	6M	20CS404.5	L2
	OR			
15	Explain in detail about file system structure and implementation.	12M	20CS404.5	L2

N S RAJU INSTITUTE OF TECHNOLOGY
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SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	**Program**	CSE/CSE(AI&ML)/CS(DS)			**Academic Year**	2021 - 2022
Course Code	20CS404	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	IV
Course	Operating Systems						

Questions (1 through 5)

Define operating system.

An Operating System (OS) is an interface between a computer user and computer hardware. An operating system is a software which performs all the basic tasks like file management, memory management, process management, handling input and output, and controlling peripheral devices such as disk drives and printers.

Define process.

A program under execution is called process.

What is safe state?

A state is safe if the system can allocate resources to each process(up to its maximum requirement) in some order and still avoid a deadlock.

Define Thrashing.

Thrashing is a condition or a situation when the system is spending a major portion of its time servicing the page faults, but the actual processing done is very negligible.

List any two file attributes.

Name, identifier, type, location, size etc.

Questions (6 through 15)

6 a. What are the functionalities of Operating Systems? Explain in detail - 6M

- Program Execution: The Operating System is responsible for the execution of all types of programs whether it be user programs or system programs. The Operating System utilizes various resources available for the efficient running of all types of functionalities.
- Handling Input/Output Operations: The Operating System is responsible for handling all sorts of inputs, i.e, from the keyboard, mouse, desktop, etc. The Operating System does all interfacing in the most appropriate manner regarding all kinds of Inputs and Outputs.
  For example, there is a difference in the nature of all types of peripheral devices such as mice or keyboards, the Operating System is responsible for handling data between them.
- Manipulation of File System: The Operating System is responsible for making decisions regarding the storage of all types of data or files, i.e, floppy disk/hard disk/pen drive, etc. The Operating System decides how the data should be manipulated and stored.
- Error Detection and Handling: The Operating System is responsible for the detection of any type of error or bugs that can occur while any task. The well-secured OS sometimes also acts as a countermeasure for preventing any sort of breach to the Computer System from any external source and probably handling them.
- Resource Allocation: The Operating System ensures the proper use of all the resources available by deciding which resource to be used by whom for how much time. All the decisions are taken by the Operating System.
- Accounting: The Operating System tracks an account of all the functionalities taking place in the computer system at a time. All the details such as the types of errors that occurred are recorded by the Operating System.
- Information and Resource Protection: The Operating System is responsible for using all the information and resources available on the machine in the most protected way. The Operating System must foil an attempt from

any external resource to hamper any sort of data or information.

All these services are ensured by the Operating System for the convenience of the users to make the programming task easier. All different kinds of Operating systems more or less provide the same services.

6B. Explain the following Operating Systems concepts

a. Multi - Programming – 3M

- In a modern computing system, there are usually several concurrent application processes which want to execute. Now it is the responsibility of the Operating System to manage all the processes effectively and efficiently.
- One of the most important aspects of an Operating System is to multi program.

  In a computer system, there are multiple processes waiting to be executed, i.e. they are waiting when the CPU will be allocated to them and they begin their execution. These processes are also known as jobs. Now the main memory is too small to accommodate all of these processes or jobs into it. Thus, these processes are initially kept in an area called job pool. This job pool consists of all those processes awaiting allocation of main memory and CPU. CPU selects one job out of all these waiting jobs, brings it from the job pool to main memory and starts executing it. The processor executes one job until it is interrupted by some external factor or it goes for an I/O task.

b. Multi -Tasking 3M

As the name itself suggests, multi tasking refers to execution of multiple tasks (say processes, programs, threads etc.) at a time. In the modern operating systems, we are able to play MP3 music, edit documents in Microsoft Word, surf the Google Chrome all simultaneously, this is accomplished by means of multi tasking. Multitasking is a logical extension of multi programming. The major way in which multitasking differs from multi programming is that multi programming works solely on the concept of context switching whereas multitasking is based on time sharing alongside the concept of context switching.

7. What is system call? Explain various system calls in detail - 12M

System calls provide an interface to the services made available by an operating system. These calls are generally available as functions written in C and C++ (high level language)

1. Example of using System calls:

Writing a simple program to read data from one file and copy them to another file. The first input that the program will need is the names of the two files: the input file and the output file.

One approach is to pass the names of the two files as part of the command—for example, the **UNIX cp command:**

cp in.txt out.txt

This command copies the input file in.txt to the output file out.txt.

A second approach is for the program to ask the user for the names. In an interactive system, this approach will require a sequence of system calls, first to write a prompting message on the screen and then to read from the keyboard the characters that define the two files. Fig. 1.1 shows the sequence of system calls to be called for achieving the same in an interactive system.

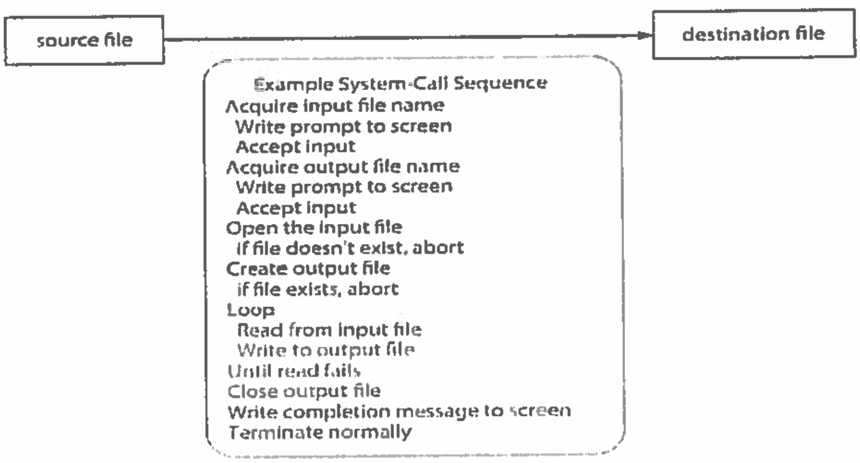


Figure 1.1 Example of how system calls are used

Application Programming Interface

The above example illustrates as you can see, even simple programs may make heavy use of the operating system. Typically, application developers design programs according to an pplication programming interface (API). The API specifies a set of functions that are available to an application programmer, including the parameters that are passed to each function and the return values the programmer can expect.

Three of the most common APIs available to application programmers are the **Windows API for Windows systems, the POSIX API for POSIX-based systems (which include virtually all versions of UNIX, Linux, and macOS), and the Java API for programs that run on the Java virtual machine.** A programmer accesses an API via a library of code provided by the operating system.

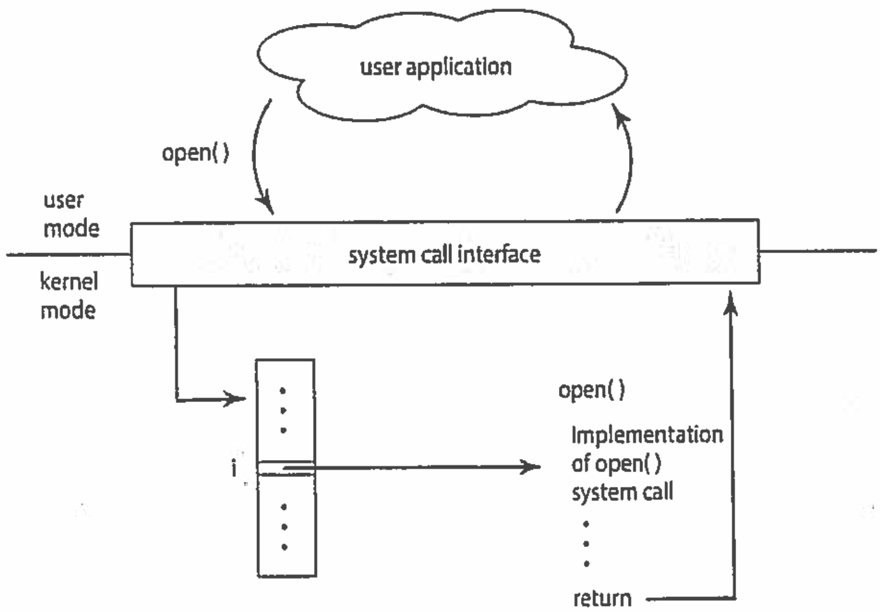


Figure 1.2 The handling of a user application invoking the open() system call

system-call interface as shown in figure 1.2 that serves as the link to system calls made available by the operating system. The system-call interface intercepts function calls in the API and invokes the necessary system calls within the operating system. The relationship among an API, the system-call interface, and the operating system is shown in Figure 1.2, which illustrates how the operating system handles a user application invoking the open() system call.

Types of System Calls:

System calls can be grouped roughly into six major categories: **process control, file management, device management, information maintenance, communications,** and **protection.** Figure 1.3 summarizes the types of system calls normally provided by an
operating system.

Process control system calls help to create, terminate a process, load and execute a process, to obtain process attributes and allocate & deallocate memory .

File management system calls help the user to create a file, delete a file, open a file for reading/writing/appending, get file attributes etc.

Device management: A process may need several resources to execute—main memory, disk drives, access to files, and so on. If the resources are available, they can be granted, and control can be returned to the user process. Otherwise, the process will have to wait until sufficient resources are available. All the resources are treated as devices.

- Process control
  - create process, terminate process
  - load, execute
  - get process attributes, set process attributes
  - wait event, signal event
  - allocate and free memory
- File management
  - create file, delete file
  - open, close
  - read, write, reposition
  - get file attributes, set file attributes
- Device management
  - request device, release device
  - read, write, reposition
  - get device attributes, set device attributes
  - logically attach or detach devices
- Information maintenance
  - get time or date, set time or date
  - get system data, set system data
  - get process, file, or device attributes
  - set process, file, or device attributes
- Communications
  - create, delete communication connection
  - send, receive messages
  - transfer status information
  - attach or detach remote devices
- Protection
  - get file permissions
  - set file permissions

8. Consider the following four processes, with the length of the CPU burst time given in milliseconds, time slice = 3 ms (12M – EACH 4M)

Process	Arrival Time(ms)	Burst Time (ms)
P_1	1	6
P_2	1	5
P_3	2	5
P_4	2	3

Compute Average Waiting Time for a given process using FCFS, SJF and RR Algorithms.

RR :-

P_1	P_2	P_3	P_4	P_1	P_2	P_3

1 4 7 10 13 [illegible]

Waiting time [illegible] $(4-1)+(16-7) = 12$

Avg. waiting time

P_3 = $(7-2)+(18-10)$ = [illegible]

Avg waiting time = $\frac{9+12+14+[illegible]}{4}$ = [illegible] 10.[illegible]

FCFS, SJF, RR [illegible]

Process	Arrival Time	Burst Time
P_1	1	6
P_2	1	5
P_3	1	5
P_4	2	3

FCFS

P_1	P_2	P_3	P_4
1	7	12	17	20

Waiting Time :-

$P_1 = 0$; $P_2 = 6$; $P_3 = 11$; $P_4 = 15$

$$\text{Avg. waiting Time} = \frac{0+6+11+15}{4} = \frac{32}{4} = 8 \text{ ms}$$

SJF

P_2	P_4	P_3	P_1
1	6	09	14	20

Waiting Time: $P_1 = 13$; $P_2 = 0$; $P_3 = 08$, $P_4 = 4$

$$\text{Avg. waiting time} = \frac{13+0+8+4}{4} = \frac{25}{4} = 6.25 \text{ ms}$$

9A. Explain about the different types of Schedulers in detail. – 6M

Process Scheduling handles the selection of a process for the processor on the basis of a scheduling algorithm and also the removal of a process from the processor. It is an important part of multiprogramming operating system.

There are many scheduling queues that are used in process scheduling. When the processes enter the system, they are put into the job queue. The processes that are ready to execute in the main memory are kept in the ready queue. The processes that are waiting for the I/O device are kept in the I/O device queue.

The different schedulers that are used for process scheduling are –

Long Term Scheduler

The job scheduler or long-term scheduler selects processes from the storage pool in the secondary memory and loads them into the ready queue in the main memory for execution.

The long-term scheduler controls the degree of multiprogramming. It must select a careful mixture of I/O bound and CPU bound processes to yield optimum system throughput. If it selects too many CPU bound processes then the I/O devices are idle and if it selects too many I/O bound processes then the processor has nothing to do.

The job of the long-term scheduler is very important and directly affects the system for a long time.

Short Term Scheduler

The short-term scheduler selects one of the processes from the ready queue and schedules them for execution. A scheduling algorithm is used to decide which process will be scheduled for execution next.

The short-term scheduler executes much more frequently than the long-term scheduler as a process may execute only for a few milliseconds.

The choices of the short term scheduler are very important. If it selects a process with a long burst time, then all the processes after that will have to wait for a long time in the ready queue. This is known as starvation and it may happen if a wrong decision is made by the short-term scheduler.

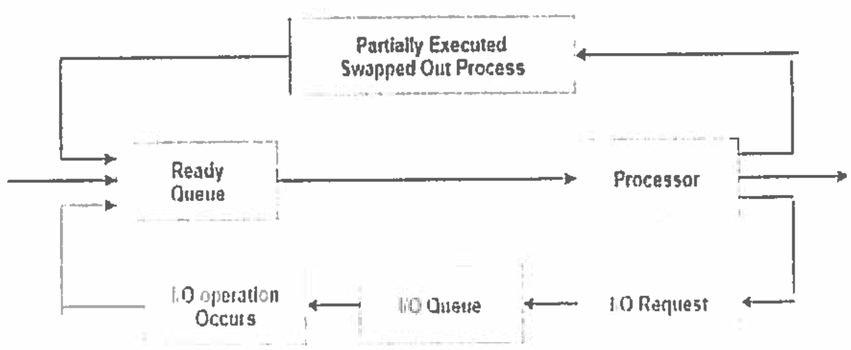
A diagram that demonstrates long-term and short-term schedulers is given as follows –

Medium Term Scheduler

The medium-term scheduler swaps out a process from main memory. It can again swap in the process later from the point it stopped executing. This can also be called as suspending and resuming the process.

This is helpful in reducing the degree of multiprogramming. Swapping is also useful to improve the mix of I/O bound and CPU bound processes in the memory.

A diagram that demonstrates medium-term scheduling is given as follows –



9B. Define Process, explain different Process states with a neat diagram – 6M

A Process Control Block is a data structure maintained by the Operating System for every process. The PCB is identified by an integer process ID (PID). A PCB keeps all the information needed to keep track of a process as listed below in the table –

S.N.	Information & Description
1	Process State The current state of the process i.e., whether it is ready, running, waiting, or whatever.
2	Process privileges This is required to allow/disallow access to system resources.
3	Process ID Unique identification for each of the process in the operating system.
4	Pointer A pointer to parent process.
5	Program Counter Program Counter is a pointer to the address of the next instruction to be executed for this process.
6	CPU registers Various CPU registers where process need to be stored for execution for running state.
7	CPU Scheduling Information Process priority and other scheduling information which is required to schedule the process.
8	Memory management information This includes the information of page table, memory limits, Segment table depending on memory used by the operating system.
9	Accounting information This includes the amount of CPU used for process execution, time limits, execution ID etc.
10	IO status information This includes a list of I/O devices allocated to the process.

The architecture of a PCB is completely dependent on Operating System and may contain different information in different operating systems. Here is a simplified diagram of a PCB –

Process ID
State
Pointer
Priority
Program counter
CPU registers
I/O information
Accounting information
etc....

The PCB is maintained for a process throughout its lifetime, and is deleted once the process terminates.

10A. Explain how dinning philosopher's problem is solved using Semaphores with an example. – 6M

A solution of the Dining Philosophers Problem is to use a semaphore to represent a chopstick. A chopstick can be picked up by executing a wait operation on the semaphore and released by executing a signal semaphore.

The structure of the chopstick is shown below –

```
semaphore chopstick [5];
```

Initially the elements of the chopstick are initialized to 1 as the chopsticks are on the table and not picked up by a philosopher.

The structure of a random philosopher i is given as follows –

```
do {
    wait(chopstick[i]);
    wait(chopstick[(i+1) % 5]);
     . . .
    // eat
     . . .
    signal(chopstick[i]);
    signal(chopstick[(i+1) % 5]);
     . . .
    // think
     . . .
}while (TRUE);
```

In the above structure, first wait operation is performed on chopstick[i] and chopstick[(i+1) % 5]. This means that the philosopher i has picked up the chopsticks on his sides. Then the eating function is performed.

After that, signal operation is performed on chopstick[i] and chopstick[(i+1) % 5]. This means that the philosopher i has eaten and put down the chopsticks on his sides. Then the philosopher goes back to thinking.

10B. Explain the necessary condition for deadlock – 6M

There are four conditions that are necessary to achieve deadlock:

1. **Mutual Exclusion** - At least one resource must be held in a non-sharable mode; If any other process requests this resource, then that process must wait for the resource to be released.
2. **Hold and Wait** - A process must be simultaneously holding at least one resource and waiting for at least one resource that is currently being held by some other process.
3. **No preemption** - Once a process is holding a resource (i.e. once its request has been granted), then that resource cannot be taken away from that process until the process voluntarily releases it.
4. **Circular Wait** - A set of processes { P0, P1, P2, . . ., PN } must exist such that every P[i] is waiting for P[(i + 1) % (N + 1)]. (Note that this condition implies the hold-and-wait condition, but it is easier to deal with the conditions if the four are considered separately.

11. Consider the following page reference string : - 12M (EACH 6M)
7, 2, 3, 1, 2, 5, 3, 4, 6, 7, 7, 1, 0, 5, 4, 6, 2, 3, 0, 1
Assume demand paging with three frames, how many page faults would occur for the following page replacement Algorithms
1. LRU replacement 2. FIFO replacement

LRU:

7	2	3	1	2	5	3	4	6	7	7
7	7	7	1		1	3	3	3	7	
	2	2	2		2	2	4	4	4	
		3	3		5	5	5	6	6	

1	0	5	4	6	2	3	0	1
7	7	5	5	5	2	2	2	1
1	1	1	4	4	4	3	3	3
6	0	0	0	6	6	6	0	0

No. of Misses = 18

FIFO

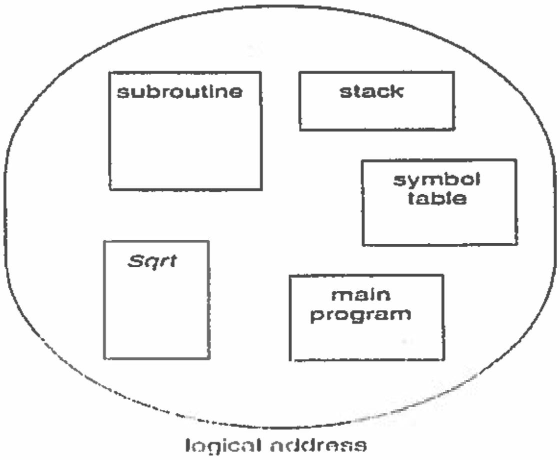
7	2	3	1	2	5	3	4	6	7	7
7	7	7	1		1	3	3	3	7	
	2	2	2		2	2	4	4	4	
		3	3		5	5	5	6	6	

1	0	5	4	6	2	3	0	1
7	7	5	5	5	2	2	2	1
1	1	1	4	4	4	3	3	3
6	0	0	0	6	6	6	0	0

No. of Misses = 17

12A. Write Short notes on Segmentation – 6M

- Most users (programmers) do not think of their programs as existing in one continuous linear address space.
- Rather they tend to think of their memory in multiple ***segments***, each dedicated to a particular use, such as code, data, the stack, the heap, etc.
- Memory ***segmentation*** supports this view by providing addresses with a segment number (mapped to a segment base address) and an offset from the beginning of that segment.
- For example, a C compiler might generate 5 segments for the user code, library code, global (static) variables, the stack, and the heap, as shown in below Figure.



Programmer's view of a program.

Segmentation Hardware

- A ***segment table*** maps segment-offset addresses to physical addresses, and simultaneously checks for invalid addresses, using a system similar to the page tables and relocation base registers discussed previously. (Note that at this point in the discussion of segmentation, each segment is kept in contiguous memory and may be of different sizes, but that segmentation can also be combined with paging as we shall see shortly.)

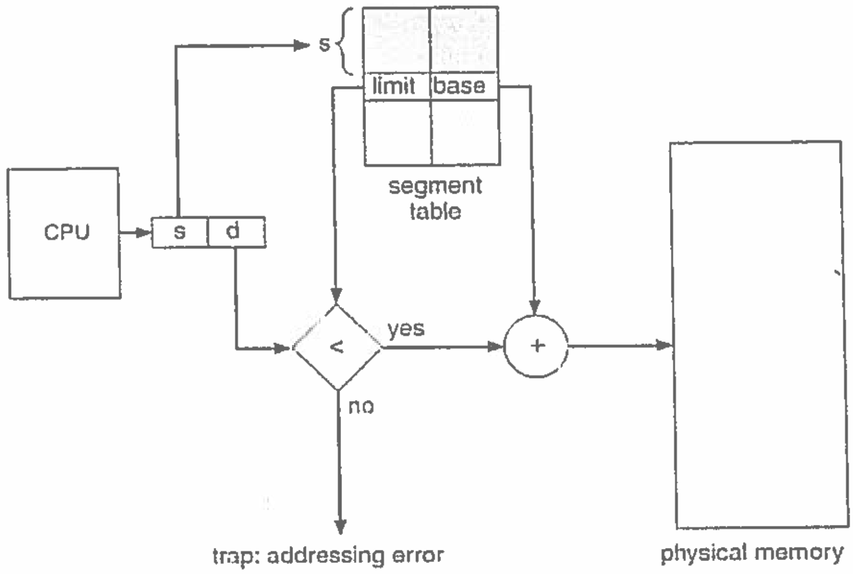


Figure Segmentation hardware

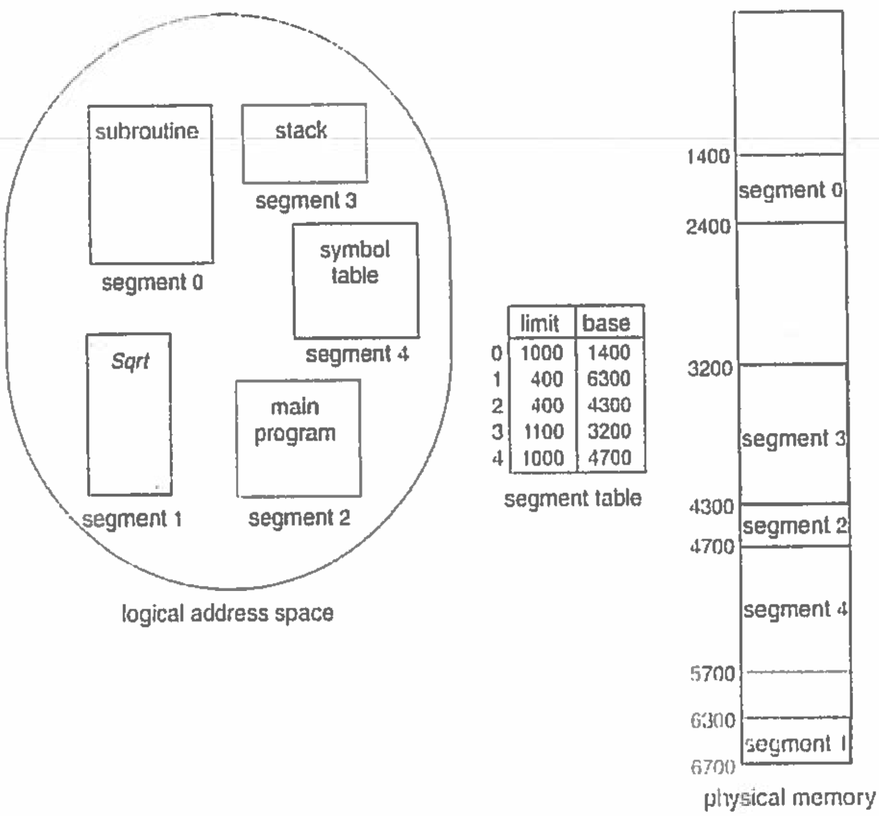


Figure Example of segmentation

12 B. Discuss about Demand paging. – 6M

The basic idea behind ***demand paging*** is that when a process is swapped in, its pages are not swapped in all at once. Rather they are swapped in only when the process needs them. (on demand.) This is termed a ***lazy swapper***, although a ***pager*** is a more accurate term.

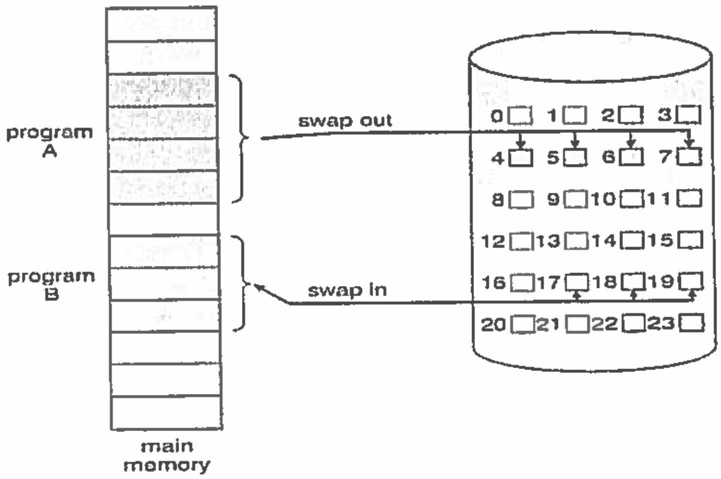


Figure Transfer of a paged memory to contiguous disk space

13. Suppose that a disk drive has 5000 cylinders numbered 0 to 4999. The drive is currently serving a request at cylinder 143. The queue of pending requests in FIFO order 86,1470,913,1774,948,1509, 1022, 1750, 130 starting from current head position.
Determine the total distance that disk arm moves to satisfy all the pending request for FCFS, SSTF, SCAN, C-SCAN disk scheduling algorithm. – 12m (each 3M)

[illegible]

→ 86, 1470, 913, 1774, 948, 1509, 1022, 1750, 130

FCFS : (143 − 86) + (1470 − 86) + (1470 − 913) + (1774 − 913) + [illegible]
(1509 − 948) + (1509 − 1022) + (1750 − 1022) + (1750 − 130)
= 57 + 1384 + 557 + 861 + 826 + 561 + 427 + 728 + 1620

SSTF : (143 − 130) + (130 − 86) + (913 − 86) + (948 − 913) + (1022 − 948) +
(1470 − 1022) + (1509 − 1470) + (1750 − 1509) + (1774 − 1750)

C-SCAN : (143 − 86) + (130 − 86) + (913 − 130) + (948 − 913) + (1022 − 948) +
(1509 − 1022) + (1750 − 1509) + (1774 − 1750)

SCAN : (143 − 86) + (86 − 0) + (130 − 0) + (913 − 130) + (948 − 913) + (1022 − 948)
+ (1509 − 1022) + (1750 − 1509) + (1774 − 1750)

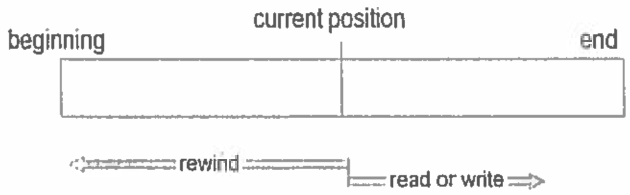
14a. Discuss about different file access methods. – 6M

Files store information. When it is used, this information must be accessed and read into computer memory. The information in the file can be accessed in several ways.

Sequential Access

The simplest access method is **sequential access.** Information in the file is processed in order, one record after the other. This mode of access is by far the most common; for example, editors and compilers usually access files in this fashion.

Reads and writes make up the bulk of the operations on a file. A read operation— read next()— reads the next portion of the file and automaticallyadvances a file pointer, which tracks the I/O location. Similarly, the write operation— write next()— appends to the end of the file and advances to the end of the newly written material (the new end of file). Such a file can be resetto the beginning, and on some systems, a program may be able to skip forward or backward *n* records for some integer *n* —perhaps only for *n* = 1. Sequentialaccess, which is depicted in Figure 13.4, is based on a tape model of a file andworks as well on sequential-access devices as it does on random-access ones.



Direct Access

Another method is **direct access** (or **relative access**). Here, a file is made upof fixed-length **logical records** that allow programs to read and write recordsrapidly in no particular order. The direct-access method is based on a disk model of a file, since disks allow random access to any file block. For direct access, the file is viewed as a numbered sequence of blocks or records. Thus, we may read block 14, then read block 53, and then write block 7. There are norestrictions on the order of reading or writing for a direct-access file.

For the direct-access method, the file operations must be modified to include the block number as a parameter. Thus, we have read(n), where *n* is the block number, rather than read next(), and write(n) rather than write next(). An alternative approach is to retain read next() and write next() and to add an operation position file(n) where *n* is the block number. Then, to effect a read(n), we would position file(n) andthen read next().

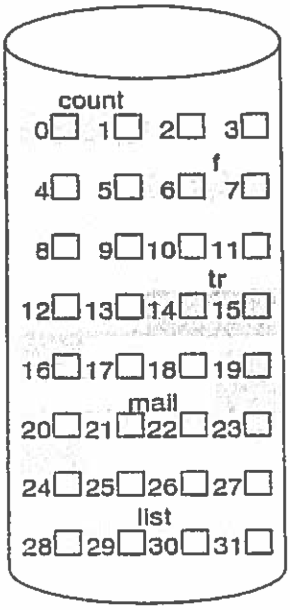
14b. Explain in detail about allocation methods – 6M

Allocation Methods

- There are three major methods of storing files on disks: contiguous, linked, and indexed.

Contiguous Allocation

- ***Contiguous Allocation*** requires that all blocks of a file be kept together contiguously.
- Performance is very fast, because reading successive blocks of the same file generally requires no movement of the disk heads, or at most one small step to the next adjacent cylinder.

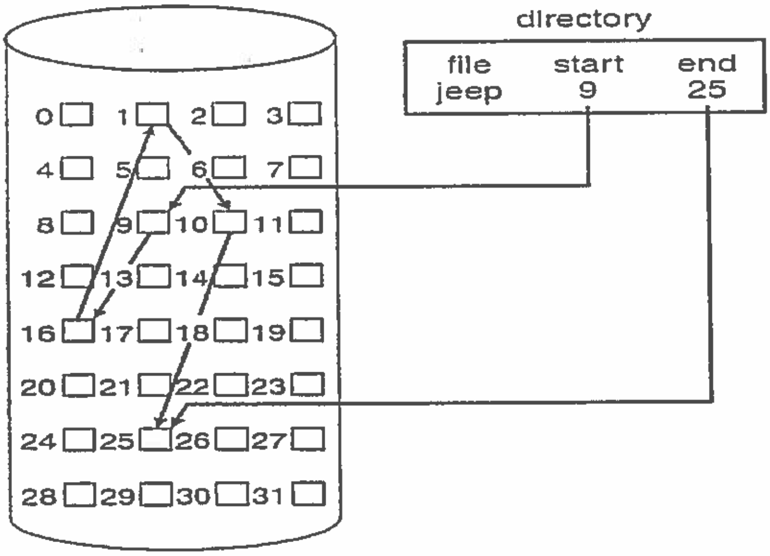


directory

file	start	length
count	0	2
tr	14	3
mail	19	6
list	28	4
f	6	2

Linked Allocation

- Disk files can be stored as linked lists, with the expense of the storage space consumed by each link. (E.g. a block may be 508 bytes instead of 512.)
- Linked allocation involves no external fragmentation, does not require pre-known file sizes, and allows files to grow dynamically at any time.

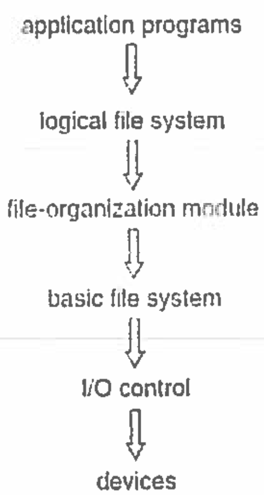


15 Explain in detail about file system structure and implementation.

File-System Structure – 6M

- Hard disks have two important properties that make them suitable for secondary storage of files in file systems: (1) Blocks of data can be rewritten in place, and (2) they are direct access, allowing any block of data to be accessed with only (relatively) minor movements of the disk heads and rotational latency. (See Chapter 12)
- Disks are usually accessed in physical blocks, rather than a byte at a time. Block sizes may range from 512 bytes to 4K or larger.
- File systems organize storage on disk drives, and can be viewed as a layered design:

- At the lowest layer are the physical devices, consisting of the magnetic media, motors & controls, and the electronics connected to them and controlling them. Modern disk put more and more of the electronic controls directly on the disk drive itself, leaving relatively little work for the disk controller card to perform.
- ***I/O Control*** consists of ***device drivers***, special software programs (often written in assembly) which communicate with the devices by reading and writing special codes directly to and from memory addresses corresponding to the controller card's registers. Each controller card (device) on a system has a different set of addresses (registers, a.k.a. ***ports***) that it listens to, and a unique set of command codes and results codes that it understands.
- The ***basic file system*** level works directly with the device drivers in terms of retrieving and storing raw blocks of data, without any consideration for what is in each block. Depending on the system, blocks may be referred to with a single block number, (e.g. block # 234234), or with head-sector-cylinder combinations.
- The ***file organization module*** knows about files and their logical blocks, and how they map to physical blocks on the disk. In addition to translating from logical to physical blocks, the file organization module also maintains the list of free blocks, and allocates free blocks to files as needed.
- The ***logical file system*** deals with all of the meta data associated with a file (UID, GID, mode, dates, etc), i.e. everything about the file except the data itself. This level manages the directory structure and the mapping of file names to ***file control blocks, FCBs***, which contain all of the meta data as well as block number information for finding the data on the disk.



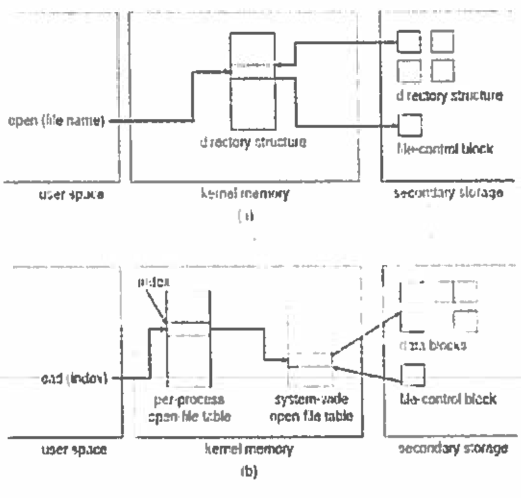


Figure 14.3 In-memory file-system structures. (a) File open. (b) File read.

File-System Implementation - 6M

Overview

File systems store several important data structures on the disk:

- A ***boot-control block***, (per volume) a.k.a. the ***boot block*** in UNIX or the ***partition boot sector*** in Windows contains information about how to boot the system off of this disk. This will generally be the first sector of the volume if there is a bootable system loaded on that volume, or the block will be left vacant otherwise.
- A ***volume control block,*** (per volume) a.k.a. the ***master file table*** in UNIX or the ***superblock*** in Windows, which contains information such as the partition table, number of blocks on each filesystem, and pointers to free blocks and free FCB blocks.
- A directory structure (per file system), containing file names and pointers to corresponding FCBs. UNIX uses inode numbers, and NTFS uses a ***master file table.***
- The ***File Control Block, FCB,*** (per file) containing details about ownership, size, permissions, dates, etc.

UNIX stores this information in inodes, and NTFS in the master file table as a relational database structure.

file permissions
file dates (create, access, write)
file owner, group, ACL
file size
file data blocks or pointers to file data blocks

- There are also several key data structures stored in memory:
    - An in-memory mount table.
    - An in-memory directory cache of recently accessed directory information.
    - ***A system-wide open file table***, containing a copy of the FCB for every currently open file in the system, as well as some other related information.
    - ***A per-process open file table***, containing a pointer to the system open file table as well as some other information. (For example the current file position pointer may be either here or in the system file table, depending on the implementation and whether the file is being shared or not.)
    - Figure 14.3 illustrates some of the interactions of file system components when files are created and/or used

8m

11 b)

Project scheduling with limited resources:

1. man power
2. equipment
3. money.

Resources levelling: (to minimize) the peak requirement and smooth out period to period variation.

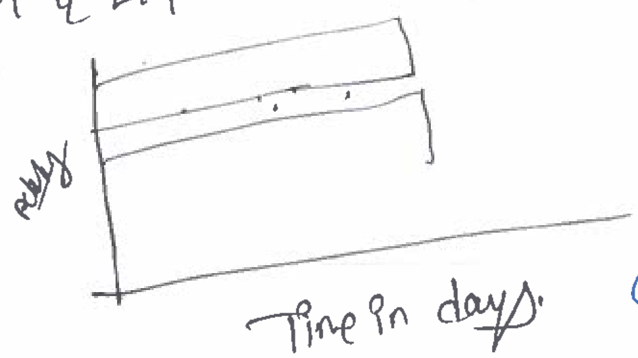
2. Resource allocation: To adjust the non-critical activities such that the resource requirement in each period is with in the available range.

Resource levelling: consider the following problem of project scheduling to obtain a schedule which will minimize the peak man power requirement and smooth out period to period variation of man power requirement

steps: → Draw the network diagram

→ Find the critical path.

→ The EST & LFT for each event provided.



G. Dung

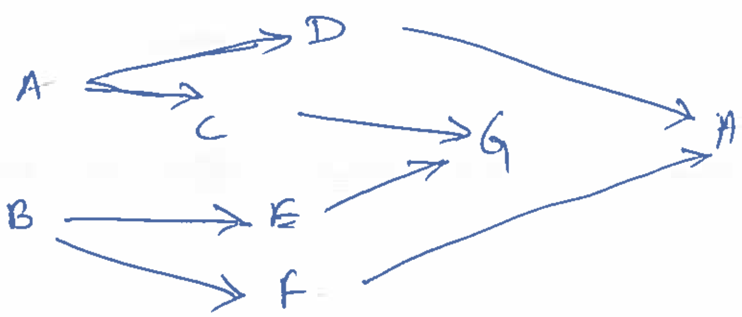
10 a 8m

procedure:

1. Network diagram to be draw
2. cost slope need to find

$$\text{Cost slope} = \frac{\text{crash cost} - \text{Normal cost}}{\text{Normal time} - \text{crash time}}$$

3. paths need to find (critical paths).



critical paths = A - D - H
A - C - G
B - E - G
B - F - H.

Need to find time ~~used~~ for total work and have to crash by given.

Computations for TE and TL

$$T_E^{j} = T_E^{i} + t_E^{ij}$$

$$T_L^{i} = T_L^{j} - T_E^{ij}$$

FLOAT :-

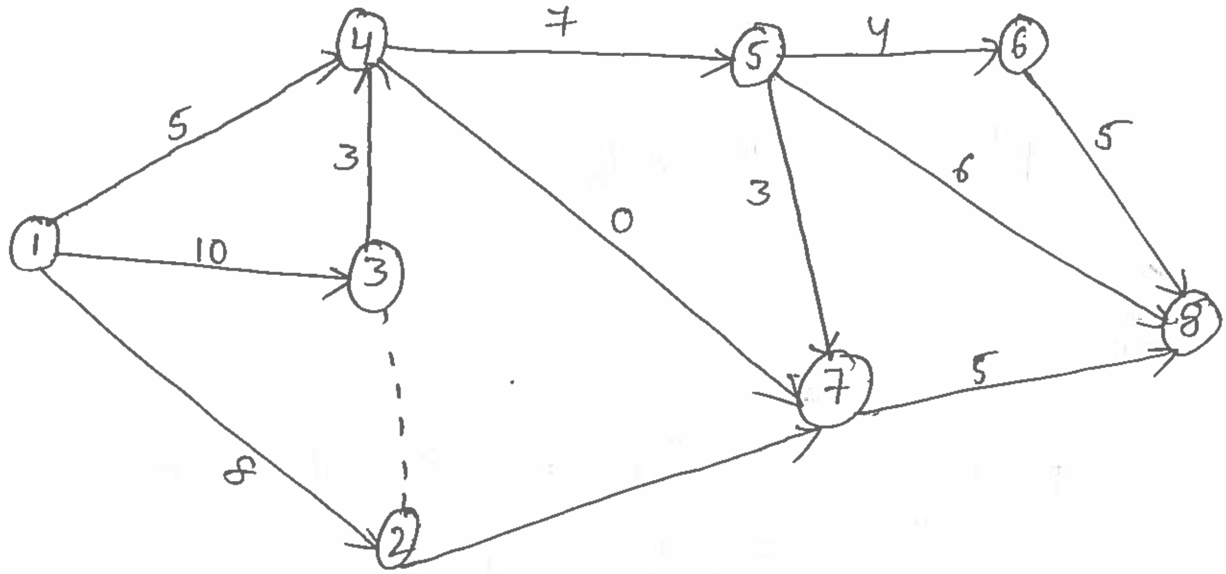
TOTAL FLOAT = time available − time required
(TF)

$$= \left(T_L^{j} - T_E^{i}\right) - t^{ij}$$

Free FLOAT :-

$$FT = \left(T_L^{j} - T_E^{i}\right) - t^{ij}$$

(8 b) Sol



Network Diagram

Earliest Expected time :- (TE)

Latest allowable occurrence time (TL)

$$\underset{(\text{Successor Event})}{TE} = TE(\text{Predecessor event}) + t_E(\text{activity})$$

$$TE^j = TE^i + t_E^{ij}.$$

$$TE^j = (T_E^i + t_E^{ij})_{max}.$$

Latest allowable occurrence time (TL) :-

$$T_L^i = (T_L^j - t_E^{ij})_{min}.$$

ANSWER KEY AND SCHEME OF EVALUATION

Step 5: Monitoring and Reviewing Risks

Monitoring and reviewing of risks is a continuous process. Managers need to keep checking the likelihood of risks occurring. They must also regularly follow up on their risk prevention strategies. This step is important because risks are inevitable and they never remain static.

6m 14b

There are many causes of an accident on a construction site. The top causes of construction worker deaths on the job were falls, followed by struck by object, electrocution , and caught-in/between. These "Fatal Four" were responsible for nearly three out of five of the construction worker deaths. Human Error. One of the most critical causes of construction accidents is that of human.
Failure to Follow Safety Precautions. In order to stay safe on jobsites, construction worker
Failure to Identify Unsafe Conditions. General contractors and site managers should always
Failure to Wear Proper Safety Protection. Failure to wear safety protection is one of the root

6m 15a

Safety risk management encompasses the assessment and mitigation of safety risks. The objective of safety risk management is to assess the risks associated with identified hazards and to develop and implement effective and appropriate mitigations.

- Provide employees options to work from home.
- Social gatherings at the workplace are disallowed.
- Observe good personal hygiene.
- Minimise need for physical touchpoints.
- Step up cleaning of workplace premises through the following.
- Provide cleaning and disinfecting agents at the following areas.
- Record proximity data on phones

6m 15b

According to the National Safety Council, an effective safety management program should:

- Reduce the risk of workplace incidents, injuries, and fatalities through data-driven measurements and improvements
- Involve people from different parts of the organization to make safety a shared responsibility
- Be well organized and structured to ensure consistent growth and performance
- Be proactive, preventive and integrated into the culture of the entire organization

ANSWER KEY AND SCHEME OF EVALUATION

like **sand, gravel, water and cement** and then transported to concrete construction site ready to be poured for use

6m 13a

Earthmoving Equipment: Types of machines commonly used and their applications in construction.

- 1. Excavators. These heavy machines consist of a base cabin and a long arm with a bucket attachment at its end. They use a hydraulic system to ...
- 2. Backhoe Loaders. Backhoe loaders or backhoes are tyre mounted machines with a shovel at the front and a bucket attached to a jointed arm at the ...
- 3. Bulldozers. Considered to be one of the most heavy-duty machines that can be spotted at a construction site, one of the most common applications of ...
- 4. Skid-Steer Loaders. As the name suggests, these are tyre mounted small-sized machines which can skid on their own axis

6m 13b

Earth compaction equipment is used to decrease the porosity of earth and to increase density and strength of the earth. Compaction of the earth is done by rolling, kneading, ramming, tamping, vibrating etc. There are types of compacting equipment. Rolling equipment, ramming equipment and vibrating equipment.

Compaction of the earth is done by **rolling, kneading, ramming, tamping, vibrating** etc. There are types of compacting equipment. Rolling equipment, ramming equipment and vibrating equipment.

6m 14a

Risk management basically means the identification and mitigation of losses. It is a systematic process by which an organization identifies, analyzes, prepares and reduces losses.

Step 1: Establishing the Context

Before dealing with risks, managers must be able to understand and identify them clearly. In order to do this, they first need to comprehend the context in which the risks arise.

Step 2: Identifying the Loss

After understanding the context, managers should list down all possible risks that may arise. This will depend on the nature of the organization's business, its environment, etc. For example, a company manufacturing chemicals may face the risk of leakage from its production units.

Step 3: Analysing and Evaluating Risks

Every organization faces several kinds of risks but the chances of them occurring differ in every case. Managers should analyze each possible risk individually and evaluate the chances of it happening. This is because they have to accord more importance to serious risks than less serious ones.

Step 4: Treating the Risks

After identifying and analyzing risks, managers next have to treat them. This process can include avoiding risks altogether. Alternatively, it is also possible to reduce the possible impact of a risk.

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SONTYAM, ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

6m 12a

Excavation based on material

Topsoil Excavation

This excavation type is particularly used to remove the topsoil which is the uppermost level of soil in the surface of earth that is normally no more 12 inches. The excavation method gets out decaying materials, soil, vegetations that have the possibility of making the soil compressible as well as infit to bear structural loads. The topsoil is also removed because it has high moisture content.

Muck Excavation

Muck, combination of soil and water, usually causes problems in the construction sites since it makes the ground unsuitable as well as unstable for building. The good thing is once the much is excavated, it is possible to either move it to another area or spread it out to dry, then reuse it.

Earth Excavation

This process is the removal of soil under the surface topsoil, the digging is deepened based on the type of project. The removal of these various layers enables the construction firm to lay foundation for various construction types such as building drainage ditches, constructing bridges among other engineering projects.

Excavation based on Task

Drainage Excavation

The purpose of drainage excavation is to alter the flow of water from specific areas; hence the procedure must be accurate and done correctly to ensure the drainage remain unblocked and run freely. Drainage is connected to carrying water away from certain areas which comprises of drainage for runoffs, agricultural drainage, storm drains trenches and so on. The drainage is used to direct water from particular areas such as infrastructure, habitation along with agriculture.

Bridge Excavation

These three types of bridges suspension, beam as well as the arch require a sturdy foundation to build on.The bridge excavation process should be well informed to support the weight of the bridge above. The purpose of bridge excavation is to remove materials that might curtail the construction of the foundation, bridge footing, abutments as well as substructures of the bridge.

6m 12b

Hoisting equipment is used in a variety of areas to support processing and handling throughout a facility:

- **Assembly:** Moving products through production processes
- **Positioning:** Securing a component for additional work
- **Transportation:** Loading finished products onto open trailers or railcars
- **Staging:** Holding work-in-process for additional production processes
- **Storage:** Transporting heavy items to and from storage areas
- **Warehousing:** Moving large, heavy products to and from docks

A concrete plant or batching plant is very important equipment for the concrete construction. With the help of concrete batching plant, concrete used for the construction is produced by proper mixing of all the ingredients

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ANSWER KEY AND SCHEME OF EVALUATION

Outline the Project Plan

A work breakdown structure (WBS) in construction is a hierarchical way of organizing a building project. The WBS is a single document that divides the project deliverables into manageable chunks known as work packages.

4m 9a

CPM	PERT
Focus on cost optimization	Focus on time control
Used in projects with predictable activities such as construction	Used in projects with unpredictable work such as R&D
Deterministic treatment of time – single outcome	Probabilistic treatment of time – uses three estimates of duration
Can show activities that can be sacrificed if needed to execute project because it distinguishes between critical and noncritical activities	PERT does not provide this information because it does not distinguish between critical and noncritical activities
Enables "crashing," a shortening of the duration of specific tasks while incurring the least extra expenses	Crashing does not apply

4m 10 a

Direct Costs and Indirect Costs

- Direct Costs. Direct costs can be defined as costs which can be accurately traced to a cost object with little effort.
- Indirect Costs. Costs which cannot be accurately attributed to specific cost objects are called indirect costs. These...
- Example. Classify the above costs as direct or indirect.

4/6m 11a

You Need Effective Resource Allocation. Save money: Effective resource allocation leads to no waste of money. It lets you know the performance of team members in a project. Hence it can Boost productivity, Improve time management. Improve staff morale:. Predict the future project plan: More items.

Resource smoothing is used to settle certain resources and time but with a diverse approach than resource levelling. It aims to finish the project or activity in a given time period while preventing any unnecessary resource demands. Time is an important constraint here and therefore no activity can be delayed more than their float project by comparing the progress reports with the project plan to measure the performance of the project activities. If any deviation is found from the already defined plan corrective measures are made. The first option of action should always be to bring the project back to the original plan. If that cannot happen, the team should record variations from the original plan and record and publish modifications to the plan. all through this step, project sponsors, and other key stakeholders are kept informed about the project's status as per the agreed rate and format of communication. The plan should be updated and available on a regular basis. Status reports should always highlight the probable end point in terms of cost, schedule, and quality of deliverables. Each project deliverable produced should be reviewed for quality and measured against the acceptance criteria. When deliverables have been produced and the customer has agreed on the final solution, the project is said to be ready for closure.

6m 7b.a

Initiation Phase of Construction Project

We have to create and evaluate the project in order to determine if it is feasible and if it should be undertaken, at the beginning of the project. Here the project objective or need is identified, this can be a business problem or opportunity. A suitable response to the need is documented in a business case with recommended solution options. A feasibility study is conducted to examine whether each option clearly identifies the project objective and a final recommended solution is determined. Many questions related to the issues of feasibility i.e. "can we do the project?" and justification like "should we do the project?" are mentioned and faced. When a solution is approved, a project is initiated to implement the approved solution. For this, a project manager is appointed. At this stage, the major deliverables and the participating work groups are identified. This is the time when the project team begins to take shape. Approval is then required by the project manager to move onto the detailed planning phase. Implementation is generally considered by team members as when the project starts. In a phase-controlled project, project team members are only minimally involved prior to the implementation phase. At this point, the scope should be approved and the project is starting in earnest.

~~6m~~ 4m 8a.

- The Project Planning Process

1. Identify Project Stakeholders Start your project planning process by identifying the stakeholders of your project. ...
2. Identify Project Goals and Objectives A projects goals and objectives depend on the needs of the project stakeholders. ...
3. Identify Project Deliverables Project deliverables are the tangible products that are produced or provided as a result of the project. ...
4. Create the Project Schedule In traditional project management, the project schedule lists all activities and deliverables with their intended start and end dates, and thus provides a timeline ...
5. Create Supporting Plans Your project plan needs to include all the information necessary to manage, monitor, and complete the project successfully. ...
6.

ANSWER KEY AND SCHEME OF EVALUATION

The project manager should keep track of the project at each step and make sure that the assigned tasks are being delivered as expected. Also, he/ she should make sure to communicate with all the team members to track their progress individually and likewise provide input to better their performances if required.

5. Motivate the Team

Setting up the team, assigning tasks, and creating a deadline is not the only thing that needs to be done; a good project manager should make sure to acknowledge the hard work of all the members of the team and accordingly provide rewards and compensation to keep them motivated. Not all the members are equally driven to achieve the goal; hence it is necessary to communicate with them individually to understand what they require to accomplish their tasks and help them with that so they can align their own growth with the project's growth and put their best efforts.

6. Budget Management

A budget of a project is usually fixed right at the start of the project with the project manager. What a project manager should ensure then is that the budget is enough to get the project done and there are no budget issues introduced in the middle of the project as it progresses, because it will become difficult to solve these issues then and the project would run a risk of costing more than it is bringing in. So, the budget, once decided for the project, should also be known to the team members as it will help them have an idea as to how much work hours they have to put in to successfully complete the task knowing the hours budget.

7. Risk Management

This is a very important skill for the project manager, one in which he should be adept. A project manager should be competent enough to identify the risks the project might potentially run into and be prepared with alternatives to tackle them. They should be ready with ways to handle unexpected scenarios taking into consideration that the project will be still delivered on time without bearing a lot of expenses on other solutions.

6m 7a.a

Planning Phase of Construction Project

The planning phase involves further development of the project in detail to meet the project's objective. The team identifies all of the work to be done. The project's tasks and resource requirements are identified, along with the strategy for producing them. In a broader sense identification of each activity as well as their resource allocation is also carried out. A project plan outlining the activities, tasks, dependencies, and timeframes is created. The project manager is the one who coordinates the preparation of a project budget by providing cost estimates for the labour, equipment, and materials costs. This is mainly carried out by project scheduling software like MS project or PRIMAVERA. This scheduling charts would help us to track the stages of our project as time passes. This is also referred to as "scope management." The budget of the project already estimated is used to monitor and control cost expenditures during project implementation. Finally, we require a document to show the quality plan, providing quality targets, assurance, and control measures, along with an acceptance plan, listing the criteria to be met to gain customer acceptance. At this point, the project would have been planned in detail and is ready to be executed.

Execution Phase of Construction Project

This is the implementation phase, where the project plan is put into motion and the work of the project is performed practically on site. It is essential to maintain control and communicate as needed during each implementation stages. Progress should be continuously monitored and appropriate adjustments are made and recorded as variances from the original plan. A project manager is the one who spends most of the time in this step. Throughout the project implementation, people carry out the tasks, and progress information is being reported through regular project team meetings. The project manager uses this information to preserve control over the direction of the

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ANSWER KEY AND SCHEME OF EVALUATION

Short answers

1 Define project planning process.
2 Differentiate between Activity on arrow network diagrams and activity on node network diagrams.
3 Define resource levelling and allocation.
4 What is Trenching?
5 Define safety.

2m 1a. Project planning is a procedural step in project management, where required documentation is created to ensure successful project completion. Documentation includes all actions required to define, prepare, integrate and coordinate additional plans

2m 2a. In activity on arrow diagrams, arrows are used to show activities. In activity on arrow diagrams, nodes are called events. First node is always the "START" event and last node is the "END" event. On the other hand, in activity on node diagrams, activities are shown on the node

2m 3a. Resource levelling is a project management technique that involves resolving overallocation or scheduling conflicts to ensure a project can be completed with the available resources. Resource allocation, also known as resource scheduling, recognizes and assigns resources for a specific period to various activities.

2m 4a. Trenching is a construction method that involves digging a trench in the ground to install, maintain or inspect pipes, conduits or cables underground. When the installation of the pipe, conduit or cable is completed, the trench is backfilled and disturbed grounds are returned to their original state.

2m 5a. Construction safety involves any safety procedure that is related to the construction industry or construction sites. Construction safety aims to ensure that a construction site or the industry as a whole is not the cause of immediate danger to the public around a construction site

Long answers

12m 6a.

1. Requirement Gathering

The project manager needs to make sure that the requirements of the project are correctly gathered from the source/ client, as this is the first and main step of any project. If the requirements are correctly known, only then can the project manager select resources to work on the project according to its need.

2. Organizing Team and Dividing Tasks

After gathering the information on the project and understanding what needs to be done, the project manager must organize the whole team and divide the tasks between them according to their skills and requirement so that they can help in achieving the goal of the project with their full potential. Here, the project manager's organizational skills are tested, as any fault in rightly organizing the team would lead to failure of the project.

3. Creating a Timeline for the Project

The most important thing to keep in mind for a project manager is to create a timeline for the project and adhere to it. It is very important for the whole team to know what tasks are expected to be completed from their end and at what time. Hence, according to the timeline of the project, the team members would keep the right pace in finishing their tasks and ensure that everything is delivered on time.

4. Monitoring the Project and Providing Inputs when Required

	production equipment.			
	OR			
13 (a)	What are the equipments used for earth work excavation? Compare and contrast among various Excavating and Earth Moving Equipments.	6M	20CE405.4	L2
13 (b)	Define compaction and explain in detail the different types of equipments used in Earthwork compaction.	6M	20CE405.4	L2
14 (a)	What is risk management and explain the steps in risk management.	6M	20CE405.5	L1
14 (b)	Discuss the causes of accidents on various sites.	6M	20CE405.5	L2
	OR			
15 (a)	What is safety management and what are the measures and safety policies to be adopted?	6M	20CE405.5	L1
15 (b)	Determine the safety parameters for safety management.	6M	20CE405.5	L2

Activity	Duration in Days			Immediate predecessor
	t_o	t_m	t_p	
A	8	10	12	--
B	6	7	9	--
C	3	3	4	--
D	10	20	30	A
E	6	7	8	C
F	9	10	11	B,D,E
G	6	7	10	B,D,E
H	14	15	16	F
I	10	11	13	F
J	6	7	8	G,H
K	4	7	8	I,H
L	1	2	4	G,H

10 (a) Define direct and indirect cost? Discuss about time-cost trade-off? 4M 20CE405.3 L2

10 (b) A Construction company has been awarded a contract to construct a flyover in a city with a completion time period of 18 months. The major activities in the project and the relationships among them, the normal and crash durations, and the corresponding normal and crash costs are given in the table below. 8M 20CE405.3 L2

Activity	Immediate predecessors	Duration (Months)		Cost (in Rs.)	
		Normal	Crash	Normal	Crash
A	---	6	4	24,000	34,000
B	---	4	3	12,000	22,000
C	A	5	3	20,000	28,000
D	A	7	4	29,000	47,000
E	B	6	5	26,000	34,000
F	B	8	5	34,000	52,000
G	C, E	10	6	27,000	47,000
H	D, F	9	7	34,000	48,000

OR

11 (a) What are the objectives of resource allocation? Explain the steps involved for doing resource smoothing? 4M 20CE405.3 L2

11 (b) In a small construction project, there are nine activities. The duration of each activity and the labour required to them are given in the table. The project must be completed in 27 days. Nevertheless, the contractor wishes to carry out some resource levelling/smoothening in order that there are no excessive peaks or troughs in his labour schedule. Prepare labour schedule based on early start and late start of activities, and by visual inspection indicate the adjustment you would make in activities in order to perform resource scheduling. 8M 20CE405.3 L2

Activities	Duration (days)	Resource required (Labour)
1-2	4	2
2-3	6	3
2-5	9	4
2-4	2	4
3-4	3	3
3-7	8	3
5-6	10	2
6-7	4	2
4-7	2	1

12 (a) Explain in detail various methods of excavation-based on material used and the purpose. 6M 20CE405.4 L1

12 (b) Explain in detail about Hoisting equipment and Aggregate and concrete 6M 20CE405.4 L2

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	Program	Civil Engineering			Academic Year	2021 - 2022
Course Code	20CE405	Test Duration	3 Hrs.	Max. Marks	70	Semester	IV
Course	CONSTRUCTION PROJECT MANAGEMENT						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define project planning process.	20CE405.1	L1
2	Differentiate between Activity on arrow network diagrams and activity on node network diagrams.	20CE405.2	L2
3	Define resource levelling and allocation.	20CE405.3	L1
4	What is Trenching?	20CE405.4	L1
5	Define safety management.	20CE405.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Explain the role of each constituent of the construction team? List the Project life Cycle Phases and stages in construction.	12M	20CE405.1	L2
	OR			
7 (a)	Explain the planning and execution phrases in project management.	6M	20CE405.1	L2
7 (b)	Explain the initiation and implementation phrases in project management.	6M	20CE405.1	L1
8 (a)	Explain project planning process and also discuss the work breakdown structure.	4 M	20CE405.2	L2
8 (b)	The Following table lists the activities, durations, and their sequence of operation for a construction project. Prepare the network and compute in a table their Early Start time (EST), Early Finish time (EFT), Late start time (LST), Late finish times (LFT). Determine the critical path and find the total float for all the activities? (see table below)	8 M	20CE405.2	L2

Activity	Duration(days)
1-2	8
1-3	10
1-4	5
2-7	6
3-4	3
4-5	7
4-7	0
5-6	4
5-7	3
5-8	6
6-8	5
7-8	5

OR

No.	Questions	Marks	Learning Outcome (s)	DoK
9 (a)	Differentiate between CPM and PERT Models for project scheduling.	4M	20CE405.2	L2
9 (b)	A Construction company engaged in undertaking small projects has recently been awarded a construction project. The project activities and estimated time for their completion are listed in the table below along with the information on immediate predecessors. i. Construct a network for the project. ii. Determine the critical path and the project completion time. iii. What is the probability of completing the project in the completion time you have arrived at? iv. Determine the time interval within which the probability of completion of the project will be 90%	8M	20CE405.2	L3

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Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	**Program**	Mechanical Engineering	**Academic Year**	2021 - 2022
Course Code	20ME405	**Test Duration**	3 Hrs. Max. Marks 70	**Semester**	IV
Course	**IC Engines and Gas Turbines**				

Part A (Short Answer Questions 5x 2= 10 Marks)

No.	Questions (1 through 2)	Learning Outcome (s)	DoK
1	Define Mean Effective Pressure and Compression Ratio	20ME405.1	L1
2	Draw Actual port Timing Diagram for two Stroke Engine.	20ME405.2	L1
3	What are Different Ignition systems being used for SI Engine?	20ME405.3	L1
4	What is the Chemical Composition of Liquefied Petroleum Gas?	20ME405.4	L1
5	List any three applications of pulse jet engines?	20ME405.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 10)	Marks	Learning Outcome (s)	DoK
6 (a)	Draw the Dual Cycle P-V and T-S Diagram; Find the Efficiency in terms of Compression Ratio.	6M	20ME405.1	L2
6 (b)	In an Air Standard Diesel Cycle, the Compression ratio is 16, at the beginning isentropic compression, the temperature is 15 ^{0}C and the pressure is 0.1 MPa. Heat is added until the Temperature at the end of constant pressure process is 1480 ^{0}C .Calculate the following. (i) The cut-off ratio (ii) The heat supplied for Kg of air	6M	20ME405.1	L3
	OR			
7 (a)	Draw the Diesel Cycle P-V and T-S Diagram; Find the Efficiency in terms of Compression Ratio.	7M	20ME405.1	L2
7 (b)	Explain (i) Time loss Factor (ii) Heat Loss Factor (iii) Volumetric Efficiency.	5M	20ME405.1	L2
8 (a)	Describe the working principle of the Four stroke CI Engine. Mention the typical values of Valve timing diagram for four stroke CI Engine	6M	20ME405.2	L2
8 (b)	Draw a labeled sketch showing the circuit diagram of Battery Ignition system and discuss its working principles.	6M	20ME405.2	L2
	OR			
9 (a)	Classify the IC engines.	5M	20ME405.2	L2
9 (b)	Draw a labeled sketch showing the circuit diagram of Magneto Ignition system and Discuss its working principles	7M	20ME405.2	L2

10 (a)	What is the significance of heat balance sheet? Discuss the procedure to draw heat balance sheet for CI engine?	6M	20ME405.3	L2
10 (b)	What is wilaan's line? How do you measure frictional power using this?	6M	20ME405.3	L2
	OR			
11 (a)	Explain the Combustion Stages of SI Engine	6M	20ME405.3	L2
11 (b)	Explain knocking, properties and its effects in CI engine.	6M	20ME405.3	L2
12 (a)	Explain Different Categories of CI Emissions. Also explain various factors effecting exhaust emission.	7M	20ME405.4	L2
12 (b)	What are the Different Gaseous fuels and their Limitations?	5M	20ME405.4	L2
	OR			
13 (a)	What is the use of LPG, hydrogen and natural gas in SI Engine?	6M	20ME405.4	L2
13 (b)	What is Cetane number? What is the role of Cetane number in the performance of engine?	6M	20ME405.4	L2
14 (a)	What are the different rocket propulsion systems? Brief the working differences between the propeller-jet, turbojet and turbo-prop	6M	20ME405.5	L3
14 (b)	A turbo-jet engine flying at a speed of 960 km/h consumes air at the rate of 54.5 kg/s calculate i). Exit velocity of the jet when the enthalpy change for the nozzle is 200 KJ/kg and velocity coefficient is 0.97. ii).fuel flow rate in kg/s when air fuel ratio is 75:1 iii). Thrust specific fuel consumption iv). Propulsive power v). Propulsive Efficiency.	6M	20ME405.5	L2
	OR			
15 (a)	With a neat diagram explain the working of rocket engine.	5M	20ME405.5	L2
15 (b)	Draw the Brayton Cycle P-V and T-S Diagram; find the Efficiency in terms of Compression Ratio.	7M	20ME405.5	L2

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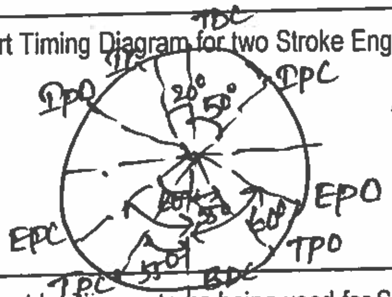
ANSWER KEY AND SCHEME OF EVALUATION

Degree	B.Tech (U.G.)			Year	II	Academic Year	2021 - 2022
Course Code	20ME405	Test Duration	3 Hrs	Max. Marks	70	Semester	IV
Course	IC Engines and Gas Turbines						

Part A

No.	Answers	Marks
1.	Define Mean Effective Pressure and Compression Ratio The mean effective pressure (MEP) is **a quantity relating to the operation of a reciprocating engine and is a measure of an engine's capacity to do work that is independent of engine displacement**. The **compression ratio** (CR) is defined as the ratio of theTotal volume of the cylinder and its Swept Volume	Definition -2M
2.	Draw Actual port Timing Diagram for two Stroke Engine.  IPO – Inlet port Open IPC – Inlet port Close EPO – Exhaust port Open TPC – Transfer port Close EPC – Exhaust port Close IS – Ignition Start	Diagram -2M
3	What are Different Ignition systems being used for SI Engine? 1.Battery Ignition System 2. Magneto Ignition System 3. Transistor Ignition System 4.Electronic Ignition System	Definition -2M
4.	What is the Chemical Composition of Liquefied Petroleum Gas? The Chemical Composition components of liquefied petroleum gas (LPG) are **propane, butane, propylene, butylene, and isobutane**.	Composition- 2M

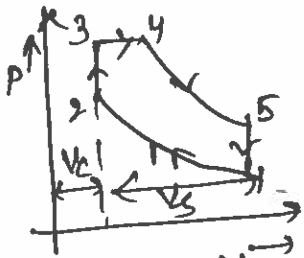
5.	List any three applications of pulse jet engines?

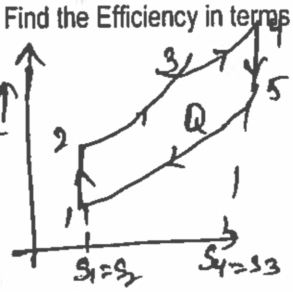
Applications of Pulse Jet Engines

1. The pulse jet engines are used for domestic purposes.
   Ex:- Microoven, iron boxes
2. This is widely used in Industrial dryness.

PART - B

6.(A) Draw the Dual Cycle P-V and T-S Diagram; Find the Efficiency in terms of Compression Ratio. [8M]





This Cycle consists of following process
(1) Two reversible adiabatic
(2) Two const. volume
(3) One const. pressure

<u>process 2-3</u> :- Const. volume heat addition process

$$Q_{s_1} = mC_v(T_3 - T_2)$$

During the process
$V_2 = V_3$

<u>process 3-4</u> :-

Const. pressure process

$$Q_{s_2} = mC_p(T_4 - T_3)$$

<u>process 4-5</u> :- During the process,
Expands P_4 to P_5
dec. T_4 to T_5

<u>process 5-1</u> :-

Const. volume heat rejection process

$$Q_R = mC_v(T_5 - T_1)$$

$$Q_s = Q_{s_1} + Q_{s_2}$$

$$= mC_v(T_3 - T_2) + mC_p(T_4 - T_3)$$

$$\eta = \frac{W}{Q_s} = \frac{Q_s - Q_R}{Q_s}$$

$$\eta = 1 - \frac{(T_5 - T_1)}{(T_3 - T_2) + \gamma(T_4 - T_3)} \quad \left[\frac{C_p}{C_v} = \gamma\right] \text{ --- } (1)$$

Compression ratio, $r = \frac{V_1}{V_2}$

pressure ratio, $k = \frac{P_3}{P_2}$

Cut-off ratio, $\rho = \frac{V_4}{V_3}$

Expansion ratio, $\frac{V_5}{V_4} = \frac{V_1}{V_4} = \frac{V_1}{V_2} \times \frac{V_3}{V_4} = \frac{r}{\rho}$

process 1-2 :-

$$\frac{T_2}{T_1} = (r)^{\gamma-1} \Rightarrow T_2 = T_1 (r)^{\gamma-1}$$

process 2-3 :-

$$\frac{P_2}{T_2} = \frac{P_3}{T_3}$$

$$T_3 = \frac{P_3}{P_2} T_2 = k\, T_1 (r)^{\gamma-1}$$

process 3-4 :-

$$\frac{V_3}{T_3} = \frac{V_4}{T_4}$$

$$T_4 = \frac{V_4}{V_3} T_3 = \rho k\, T_1 (r)^{\gamma-1}$$

process 4-5 :-

$$\frac{T_4}{T_5} = \left(\frac{V_5}{V_4}\right)^{\gamma-1} = \left(\frac{r}{\rho}\right)^{\gamma-1}$$

$$T_5 = \frac{T_4}{(r/\rho)^{\gamma-1}} = \frac{T_4\, \rho^{\gamma-1}}{(r)^{\gamma-1}} = \frac{T_1 (r)^{\gamma-1} k \rho \rho^{\gamma-1}}{(r)^{\gamma-1}}$$

$$T_5 = T_1 k \rho^{\gamma}$$

Sub T_2, T_3, T_4, T_5 in Eq (1)

$$\eta = 1 - \frac{T_1 k \rho^{\gamma} - T_1}{\left[T_1 (r)^{\gamma-1} k - T_1 (r)^{\gamma-1} + \gamma \left[T_1 (r)^{\gamma-1} k \rho - T_1 (r)^{\gamma-1} k\right]\right]}$$

$$= 1 - \frac{T_1 \left[k \rho^{\gamma} - 1\right]}{T_1 (r)^{\gamma-1} \left[(k-1) + \gamma k (\rho - 1)\right]}$$

$$\boxed{\eta_{Dual} = 1 - \frac{1}{(r)^{\gamma-1}} \left[\frac{k\rho^{\gamma} - 1}{(k-1) + \gamma k (\rho - 1)}\right]}$$

6 (b) Given Data [6]

$r = 16$

$T_1 = 15+273\ K$

$P_1 = 0.1\ MPa$

$T_4 = 1480°C$

$$\frac{T_2}{T_1} = \left(\frac{V_1}{V_2}\right)^{\gamma-1} = (16)^{(1.4-1)}(15+273)$$

$T_2 = 873.05\ K$

$$\frac{T_2}{T_1} = \left(\frac{P_2}{P_1}\right)^{\frac{\gamma-1}{\gamma}}$$

$$P_2 = \left(\frac{T_2}{T_1}\right)^{\frac{\gamma}{\gamma-1}} \times P_1$$

$$= \left(\frac{873}{288}\right)^{\frac{1.4}{0.4}} \times 0.1\times10^6$$

$P_3 = P_2$

Consider process 2-3

$Q_s = mc_p(T_3 - T_2)$

$$\frac{V_2}{T_2} = \frac{V_3}{T_3}$$

Cut-off ratio, $\rho = \frac{V_3}{V_2}$

$Q_s = mc_p(T_3 - T_2)$

$Q_s = 1\times 1.005(T_3 - 873.05)$

$$\boxed{Q_s = 1400\ kJ/kg}$$

Given Data [2M]

Cut-off [4M]

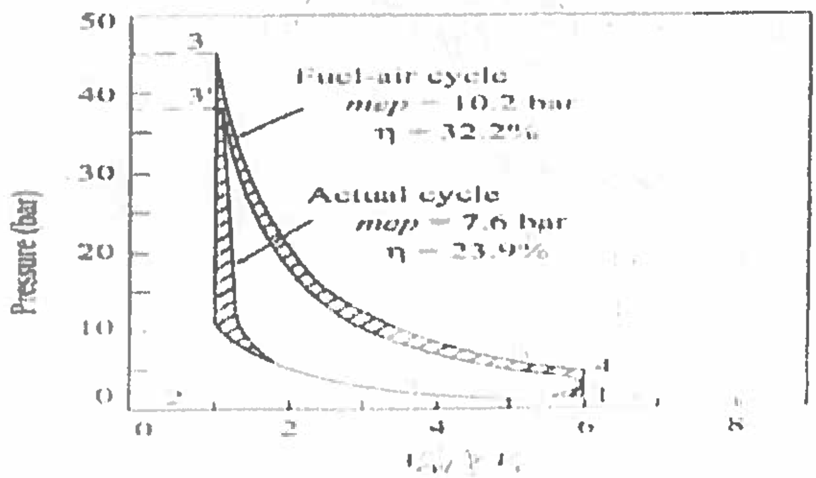
Q_s [2M]

7 (b)

Time Loss Factor

Time loss factor: loss due to time required for mixing of fuel and air and also for combustion.

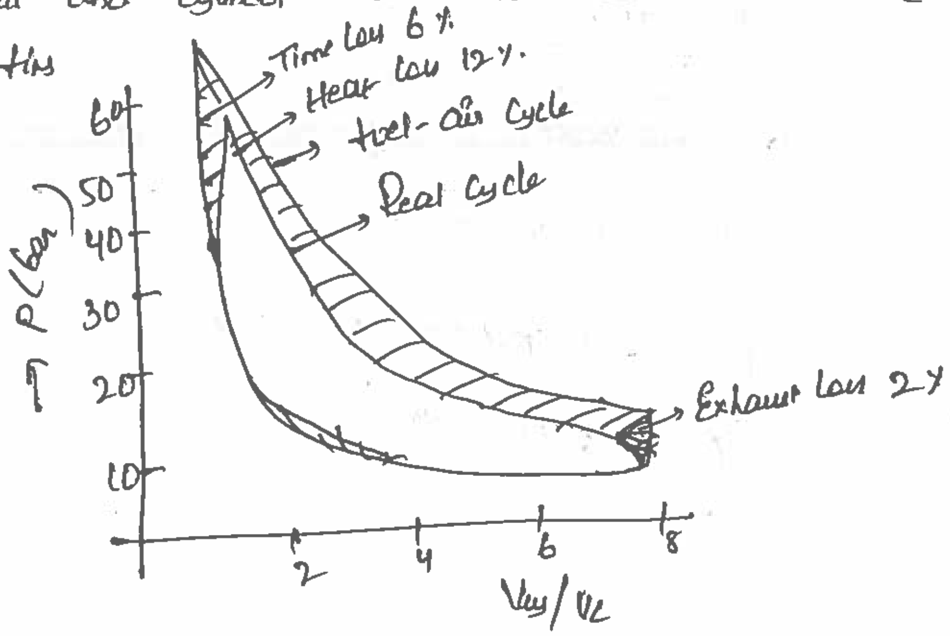
[5M]



In air standard cycles the heat addition is assumed to be an instantaneous process, where as in

Heat Loss Factor:-

During Combustion process and the Expansion Stroke the heat flows from Cylinder gases through the cylinder wall and cylinder head into the water Jacket (or) Cooling fins



Volumetric Efficiency :-

It is a breathing ability of the Engine and ratio of Volume of air actually inducted at ambient condition to swept volume.

The Volumetric Efficiency is affected by many Variables

(i) The density of fresh charge
(ii) The Exhaust gas in clearance volume
(iii) The design of intake and Exhaust value
(iv) Timing of intake & Exhaust value

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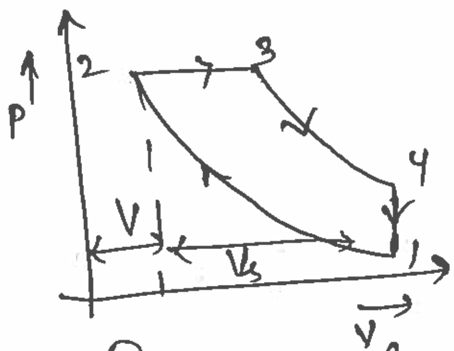
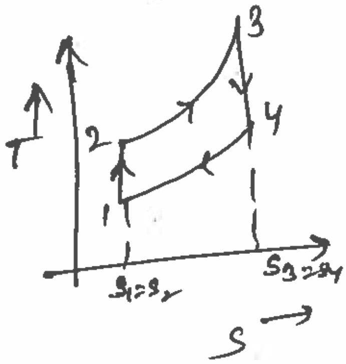
7 (a) [7M]





It consists of
(i) Two reversible adiabatic
(ii) One const Volume
(iii) One const pressure

Process 2-3 :- Heat addition process

$$Q_s = mC_p (T_3 - T_2)$$

Process 4-1 :- Const. Volume heat rejection

$$Q_R = mC_v (T_4 - T_1)$$

$$\eta = 1 - \frac{Q_R}{Q_s}$$

$$= 1 - \frac{m c_v (T_4 - T_1)}{m c_p (T_3 - T_2)}$$

$$\eta_{Dies} = 1 - \frac{(T_4 - T_1)}{\gamma (T_3 - T_2)} \quad \left[\frac{C_p}{C_v} = \gamma\right] \quad \text{—— } (1)$$

$$r = \frac{V_1}{V_2}, \quad \rho = \frac{V_3}{V_2}, \quad \text{Expansion} = \frac{r}{\rho}$$

process 1-2 :

$$\frac{V}{T} = C$$

$$\frac{T_3}{T_2} = \frac{V_3}{V_2} = \rho$$

$$T_3 = T_2 \rho = T_1 (r)^{\gamma-1} \rho$$

$$T_3 = T_1 (r)^{\gamma-1} \rho$$

process 1-2 :

$$\frac{T_2}{T_1} = \left(\frac{V_1}{V_2}\right)^{\gamma-1} = (r)^{\gamma-1}$$

$$T_2 = T_1 (r)^{\gamma-1}$$

process 3-4 :

$$\frac{T_3}{T_4} = \left(\frac{V_4}{V_3}\right)^{\gamma-1} = \left(\frac{r}{\rho}\right)^{\gamma-1}$$

$$T_4 = \frac{T_3}{(r/\rho)^{\gamma-1}} = \frac{T_1 \cancel{(r)^{\gamma-1}} \rho \, \rho^{\gamma-1}}{\cancel{(r)^{\gamma-1}}}$$

$$\boxed{T_4 = T_1 \rho^{\gamma}}$$

Sub T_2, T_3, T_4 Value in Eq (1)

$$\eta_{Diesel} = 1 - \frac{1}{\gamma}\left[\frac{T_1 \rho^{\gamma} - T_1}{T_1 (r)^{\gamma-1} \rho - T_1 (r)^{\gamma-1}}\right]$$

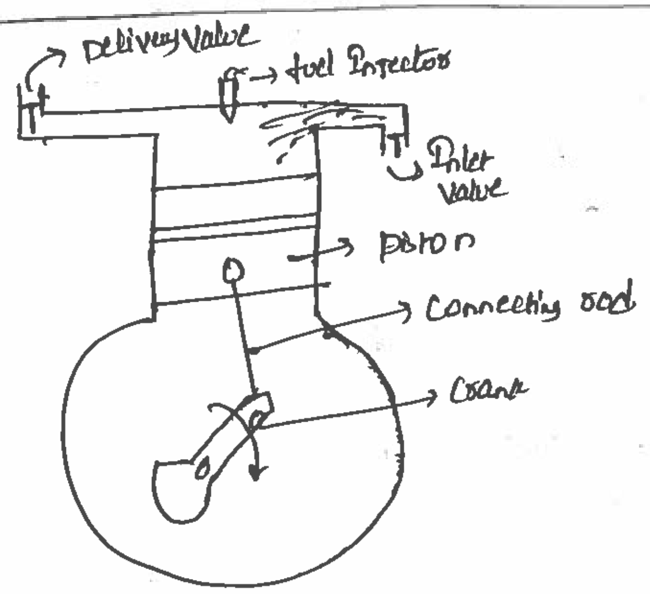
$$= 1 - \frac{1}{\gamma}\left[\frac{T_1 (\rho^{\gamma} - 1)}{T_1 \, r^{\gamma-1} (\rho - 1)}\right]$$

$$\boxed{\eta_{Diesel} = 1 - \frac{1}{\gamma (r)^{\gamma-1}}\left[\frac{\rho^{\gamma} - 1}{\rho - 1}\right]}$$

8(a)

Working Of Four Stroke Cycle (Diesel) CI Engine :-

[6M]



① Suction Stroke :-

Piston moves from TDC to BCC. The inlet Value Open Condition, the fresh Air is admitted inside the Cylinder through inlet Value.

② Compression Stroke :-

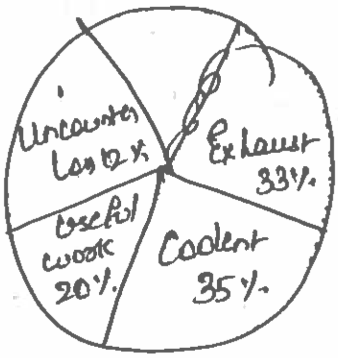
Both inlet and Exhaust Value are Closed. piston moves from BDC to TDC. Compression ration Varies from 12 to 18. pressure Compression is 3500 to 4000 kN/m^2.

③ power Stroke :-

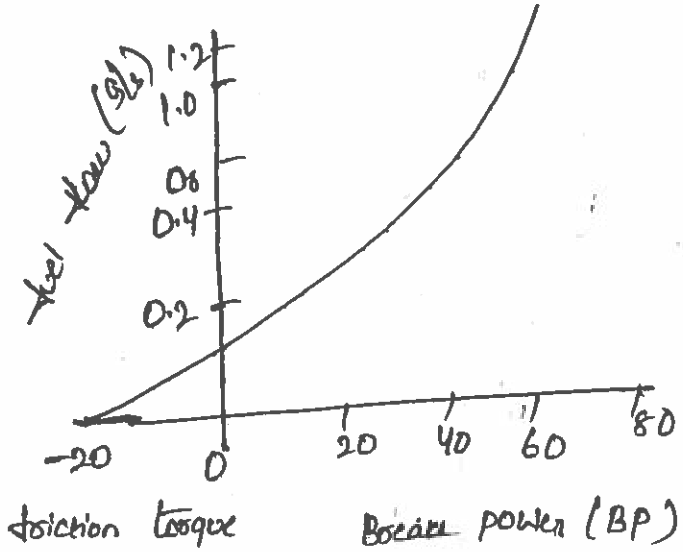
Both Value are in Closed position. The injector Opens Just before begining Of third Stroke. Ignition takes place Automatically. It pushes piston rod. Thus produce power Stroke.

④ Exhaust Stroke :-

Exhaust Value Open. It blows Out the burnt gases from the Cylinder. Thus One cycle Operation is Completed and repeated again in the Same manner.

8 b	Sketch – 2M It Consists Of battery, Ignition Coil, Condenser, Contact breaker, distributor and Spark plug.	[6M]
	(or)	
9 a	(a) Types Of Ignitions (b) Cycle Of Operation (c) Engine Cycle per Stroke (d) Types Of fuels used (e) method Of cooling (f) no. Of Cylinder (g) Valve location (h) field of Application	[5M]
9 b	Sketch – 2M It Consists Of rotating magnet assembly driven by Engine and fixed armature. The armature Of primary and secondary winding. The primary Circuit Consists Of primary winding Condenser and Contact breaker	[7M]
10 a	 Heat Input from fuel 100% 1. Only a part Of Energy is transfer into useful work 2. The rest Of its Either wasted (or) used for turbo components.	[6M]

10 (b)



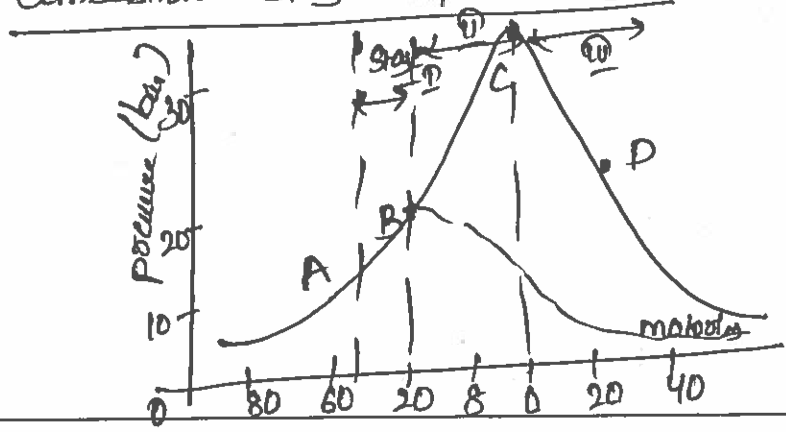
[6M]

A graph connecting fuel consumption and B.P at const. speed is drawn and it extrapolated on the negative axis of B.P, taken as frictional power of engine speed.

At start, engine does not develop any power i.e., bp = 0 it consumes certain amount of fuel. The energy spent to over come friction. Hence, the extrapolated negative intercept to mechanical friction, pumping and blow by and whole is termed as frictional losses

11 (a)

Combustion Stages of SI Engine :-

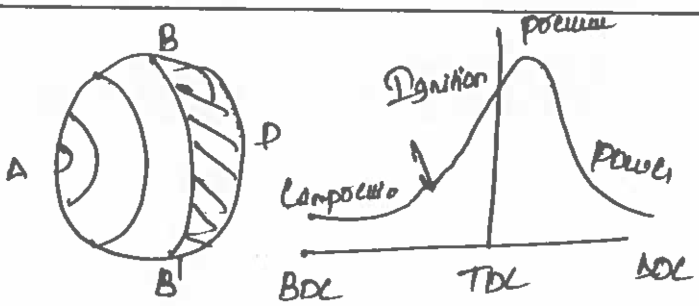


Stage - I → Ignition lag

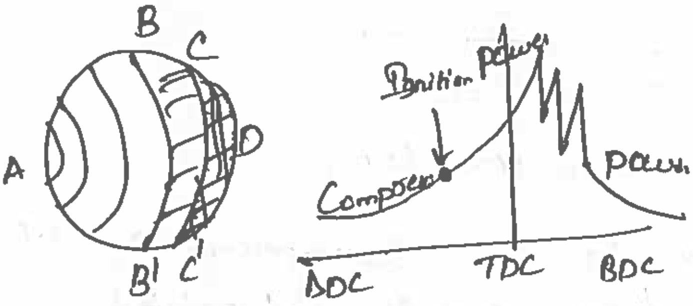
Stage - II (B→C) → Flame spread through combustion

Stage - III (C→D) → maximum pressure reached.

11 b



[6M]



If the temp of unburnt mixture exceeds the self-ignition temperature of the fuel and remains at (or) above temp during the period of preflame reaction, spontaneous ignition occurs the phenomenon called knock

Effect of Engine :-

(1) Density Factors
(2) Mass of Inducted Charge
(3) Inlet temp of mixture
(4) Temp of Combustion Chamber
(5) Power Output

---

12 a

CI Emission

Various factors Effecting Exhaust Emission [7M]

b Different gaseous fuels [5M]

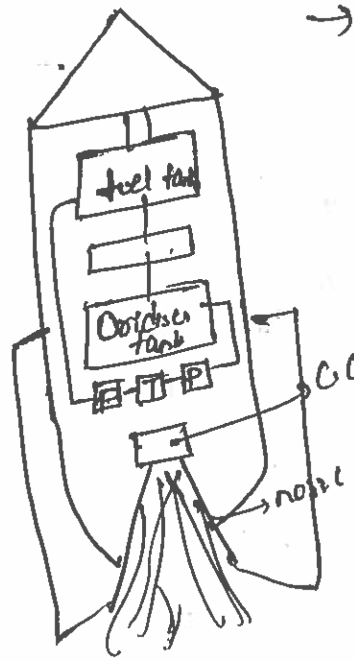
---

13 a

Uses of LPG, Hydrogen, natural gas [6M]

Cetane no. definition [6M]

role

14 (a)	Types of Rocket : (1) Solid - propellent (2) Liquid propellent (3) Hybrid - propellent Any three differences Propeller | Turbojet | Turbo-prop 1. Increase thrust in forward motion | Good performance at supersonic speed | Good performance at subsonic speed	[6M]
14 (b)	Given data - 2M Finding Exit Velocity - 1M fuel flow rate - 2M propulsive power - 1M	[6M]
	(OR)	
15 (a)	Rocket Engine :-  → The fuel and Oxygen are supplied by pumps to the C.C where the fuel is ignited by Electrical means. → The pumps are driven with the help of Steam turbine → The gases Coming from nozzle produce thrust Rocket moves forward motion	[5M]

Application :

(1) Long range artillery

(2) For satellite

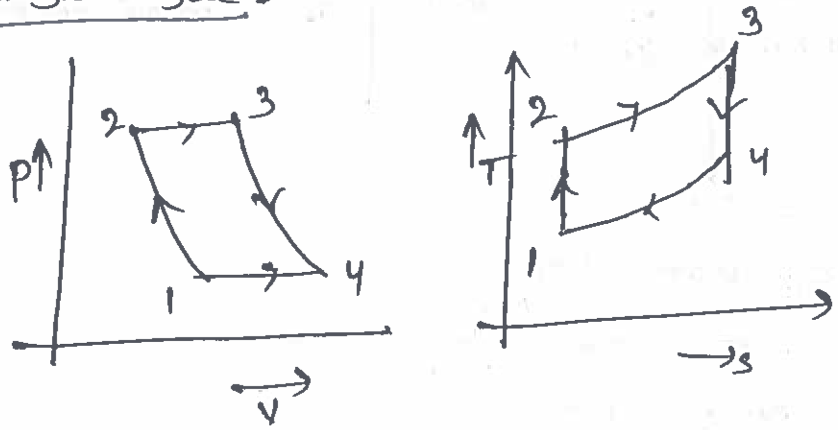
(3) Research

(4) Signalling and fire work

OR

15 (b) Brayton Cycle :

[7M]



(1) Two reversible adiabatic process

(2) Two const pressure process

$$W_c = mc_p(T_2 - T_3)$$

$$Q_s = mc_p(T_3 - T_2)$$

$$W_T = mc_p(T_3 - T_4)$$

$$Q_R = mc_p(T_4 - T_1)$$

$$\eta_{Brayton} = \frac{W}{Q_s} = 1 - \frac{m\cancel{c_p}(T_4 - T_1)}{m\cancel{c_p}(T_3 - T_2)} = \frac{T_4 - T_1}{T_3 - T_2}$$

$$\gamma = \frac{V_1}{V_2} = \frac{V_4}{V_3} ; \quad \rho_p = \frac{P_2}{P_1} = \frac{P_3}{P_4}$$

$$\eta_{Brayton} = 1 - \frac{1}{(\gamma)^{\gamma - 1}}$$

HOD

23/6

23-06-2022

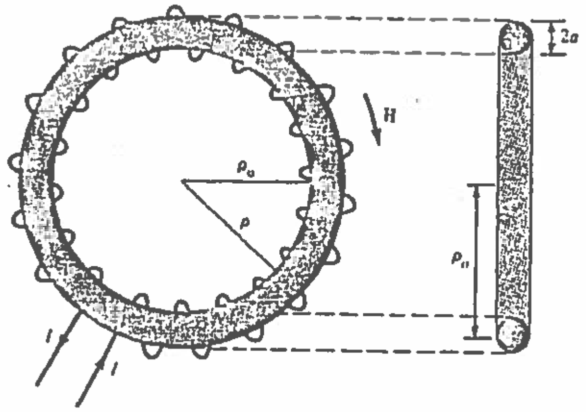
Semester End Regular Examination, June, 2022

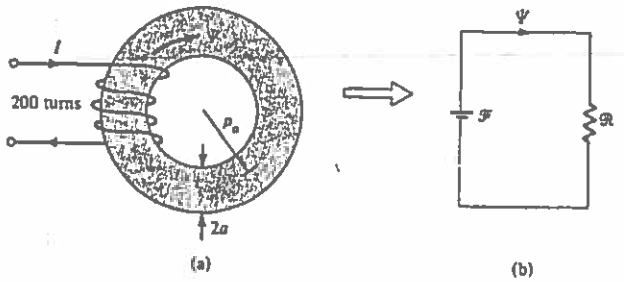
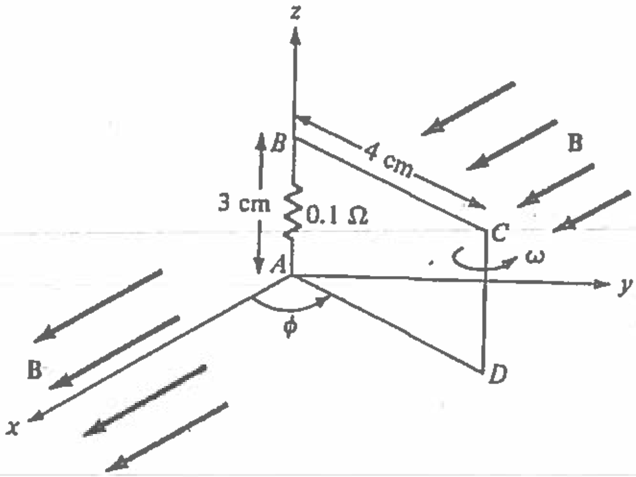
Degree	B. Tech. (U. G.)	Program	EEE			Academic Year	2021 - 2022
Course Code	20EE405	Test Duration	3 Hrs.	Max. Marks	70	Semester	IV
Course	Electro Magnetic Field Theory						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define Divergence, Gradient and Curl.	20EE405.1	L1
2	Express the energy stored in the capacitor.	20EE405.2	L1
3	Give the relationship between magnetic flux and magnetic flux density.	20EE405.3	L2
4	Express Lorentz force equation.	20EE405.4	L1
5	What is time varying field?	20EE405.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6(a)	State and prove stokes's theorem.	6M	20EE405.1	L2
6(b)	Derive the expression for Maxwell's second equation.	6M	20EE405.1	L2
	OR			
7	Derive the expression for electric field due to an infinite sheet of charge in the xy-plane with uniform charge density ρ_s.	12M	20EE405.1	L2
8 (a)	Three point charges - 1 nC, 4 nC, and 3 nC are located at (0, 0, 0), (0, 0, 1), and (1, 0, 0), respectively. Find the energy in the system.	6M	20EE405.2	L2
8 (b)	List any six properties of materials in electric field.	6M	20EE405.2	L2
	OR			
9 (a)	Derive the expression for capacitance of coaxial cable.	6M	20EE405.2	L2
9 (b)	Explain Dielectric-Dielectric boundary conditions.	6M	20EE405.2	L2
10	Toroid whose dimensions are shown in Figure1 has N turns and carries current I. Determine H inside and outside the toroid.  Figure 1	12M	20EE405.3	L2
	OR			
11	Derive Maxwell's third equation, MFI due to an infinite sheet of current carrying conductor.	12M	20EE405.3	L2
12	Derive an expression for force between two straight long and parallel current carrying conductor.	12M	20EE405.4	L2
	OR			
13 (a)	Calculate the self-inductance per unit length of an infinitely long solenoid.	8M	20EE405.4	L2
13 (b)	The toroidal core of Figure 2 has ρo = 10 cm and a circular cross section with a = 1 cm. If the core is made of steel ($\mu = 1000\mu_o$) and has	4M	20EE405.4	L2

	a coil with 200 turns, calculate the amount of current that will produce a flux of 0.5 mWb in the core.  Figure 2			
14	The loop shown in Figure is inside a uniform magnetic field $B = 50a_x$ mWb/m². If side *DC* of the loop cuts the flux lines at the frequency of 50 Hz and the loop lies in the yz-plane at time $t = 0$, find (a) The induced emf at $t = 1$ ms (b) The induced current at $t = 3$ ms  Figure 3	12M	20EE405.5	L2
	OR			
15 (a)	Explain statically induced emf.	6M	20EE405.5	L2
15 (b)	Express the integral form of Faraday's law.	6M	20EE405.5	L2

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

EMFT SEMESTER EXAM KEY

Part-A

Sl.No	Question	Marks
1	**Define Divergence, Gradient and Curl** Divergence: The divergence theorem states that the total outward flux of a vector field A through the closed surface S is the same as the volume integral of the divergence of A. In mathematical form $\oint_S A \bullet ds = \int_v (\nabla \bullet A) dv$ Curl: Curl can be defined as an axial vector whose magnitude is the maximum circulation of A per unit area as the area tends to zero and whose direction is the normal to the area, when the area is oriented so as to make the circulation maximum. Gradient (or) Potential Gradient: A potential gradient is the rate of change of the potential with respect to displacement, i.e. spatial derivative, or gradient. This quantity frequently occurs in equations of physical processes because it leads to some form of flux. In electrical engineering it refers specifically to electric potential gradient, which is equal to the electric field.	2M
2	**Express the energy storage equation for capacitor** Capacitance is the ability of a body to store an electrical charge. A material with a large capacitance holds more electric at a given voltage, than one with low capacitance.	2M

	$$\varepsilon = \varepsilon_0 \varepsilon_r$$ $$\boxed{C = \frac{\varepsilon_0 \varepsilon_r A}{d} \quad F}$$ Where *C* is the capacitance, in farads; *A* is the area of overlap of the two plates, in square meters; ε_r is the relative static permittivity (sometimes called the dielectric constant) of the material between the plates (for a vacuum, $\varepsilon_r = 1$); ε_0 is the electric constant ($\varepsilon_0 \approx 8.854 \times 10^{-12}$ F.m^{-1}); and *d* is the separation between the plates, in meters;	
3	Give the relationship between magnetic flux and magnetic flux density. The magnetic flux density B is similar to the electric flux density D. As D = $\varepsilon_0 E$ in free space, the magnetic flux density B is related to the magnetic field intensity H according to $$B = \mu_0 H$$ Where, μ_0 is a constant known as the *permeability of free space.* The constant is in henrys/meter (H/m) and has the value of $$\mu_0 = 4\pi.10^{-7} \quad H/m$$ The magnetic flux through a surface *S* is given by $$\varphi = \int_s B.ds$$	2M
4	Express Lorentz force equation. The flow of an electric current down a conducting wire is ultimately due to the motion of electrically charged particles (in most cases, electrons) through the conducting medium. It seems reasonable, therefore, that the force exerted on the wire when it is placed in a magnetic field is really the resultant of the forces exerted on these moving charges $$m\frac{d\mathbf{v}}{dt} = q\,\mathbf{E} + q\,\mathbf{v} \times \mathbf{B},$$	2M
5	What is time varying field? Stationary Charges are produced by electrostatic fields which produce steady currents which in return is a cause to produce magneto static fields, there by time varying currents are produced by electro-magnetic fields	2M

Part-B

Sl.No	Question	Marks
6(a)	**State and Prove Stokes theorem** Stoke's theorem states that the circulation of a vector field A around a closed path L is equal to the surface integral of the curl of A over the open surface S bounded by L provided that A and delta A are continuous on S .in mathematical terms it can be $$\oint_c A \bullet dl = \iint (\nabla \times A) \bullet ds$$ **PROOF OF STOKE'S THEOREM:** Consider $$\oint_c A \bullet dl = \sum_{i=1}^{N} \oint_{c_i} A \bullet dl_i = \sum_{i=1}^{N} ds_i \left(\frac{\oint_{c_i} A \bullet dl_i}{ds_i}\right)$$ Observe what happens to the right hand side as N is made enormous and d_{si} shrink. The quantity in the parentheses becomes $$(\nabla \times A) \bullet a_i$$ whereas is the unit vector normal to the ith patch So we have on the right the sum, over all the patches that make up the entire surface S spanning C, of the product "patch area times normal component of (Curl of A)". This is nothing but the surface integral over S , of the vector curl A $$\sum_{i=1}^{N} ds_i \left(\frac{\oint_{c} A \bullet dl_i}{ds_i}\right) = \sum_{i=1}^{N} ds_i (\nabla \times A) \bullet a_i = \int_s (\nabla \times A) \bullet ds$$ It relates the line integral of a vector to the surface integral of the curl of the vector. $$\oint_c A \bullet dl = \int_s (\nabla \times A) \bullet ds$$	6M

6(b)	Derive the expression for Maxwell's Second equation.	6M

Maxwell Equation - II

St:- The closed Surface Integral of magnetic flux density is always equal to scalar magnetic flux enclosed within the surface of any shape and ~~#~~ or size. Laying in any medium mathematially it is represented as. has magnetic $\boxed{\int\oint_s B\cdot ds = \phi_{enclosed}}$ — (1)

flux cannot be enclosed within a closed Surface

it can be written as $\boxed{\int\oint B\cdot ds = 0}$ — (2).

By using a divergence theorm

$$\int\oint B\cdot ds = \iint\oint_v \nabla v\cdot dV$$

In this case it is written as $\iint\oint_v \nabla B\cdot dv$

$$\nabla B = 0$$

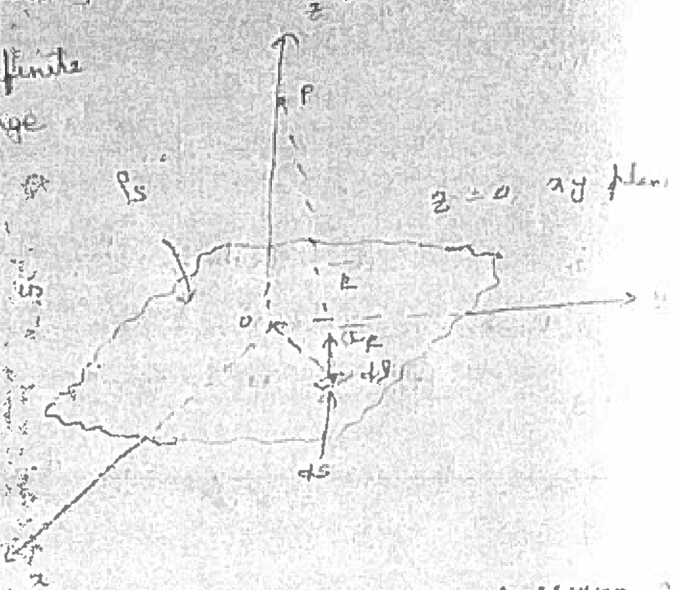
Since $\boxed{B = \mu H}$

it is written as $\nabla H = 0$

	(OR)	
7	**Derive the expression for electric field due to an infinite sheet of charge in the xy-plane with uniform charge density ϱ_s.**	12M

Electric field Intensity due to infinite sheet of charge

Consider an infinite sheet of charge having uniform ρ_s charge density ρ_s c/m² placed in xy plane



Point P at which $\bar{E}$ is to be calculated on z-axis

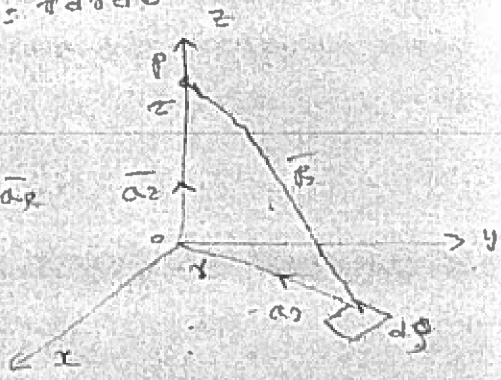
Consider differential surface area dS carrying charge dQ. Hence dS normal direction to dS is z direction is normal direction to z

$$= r\,dr\,d\phi$$

and $dQ = \rho_s\, dS = \rho_s\, r\,dr\,d\phi$

$$d\bar{E} = \frac{dQ}{4\pi\epsilon_0 R^2}\,\bar{a}_R$$

$$= \frac{\rho_s\, r\,dr\,d\phi}{4\pi\epsilon_0 R^2}\,\bar{a}_R$$



Distance vector R has two components as

(i) radial component r along $\bar{a}_r \rightarrow -r\,\bar{a}_r$

(ii) component z along $\bar{a}_z \rightarrow z\,\bar{a}_z$

with these two components $\bar{R}$ can be obtained from differential area towards point P

$$\bar{R} = -r\bar{a}_r + z\bar{a}_z$$

$$|\bar{R}| = \sqrt{r^2+z^2}$$

$$\bar{a}_R = \frac{\bar{R}}{|\bar{R}|} = \frac{-r\bar{a}_r + z\bar{a}_z}{\sqrt{r^2+z^2}}$$

$$d\bar{E} = \frac{\rho_s \cdot r\,dr \cdot d\phi}{4\pi\epsilon_0\left(\sqrt{r^2+z^2}\right)^2}\left[\frac{-r\bar{a}_r + z\bar{a}_z}{\sqrt{r^2+z^2}}\right]$$

For infinite sheet in xy plane, r varies from 0 to ∞ while ϕ from 0 to 2π

By neglecting a_r component:

$$\bar{E} = \int_{\phi=0}^{2\pi}\int_{r=0}^{\infty} d\bar{E} = \int_0^{2\pi}\int_0^{\infty} \frac{\rho_s\, r\,dr\,d\phi}{4\pi\epsilon_0(r^2+z^2)^{3/2}}\,(z\,a_z)$$

Consider

$$r^2+z^2 = u^2$$

for $r=0$, $u=z$

hence $2r\,dr = 2u\,du$

$r=\infty$, $u=\infty$

$$\therefore \bar{E} = \int_0^{2\pi}\int_{u=z}^{\infty} \frac{\rho_s}{4\pi\epsilon_0}\frac{u\,du}{(u^2)^{3/2}}\,d\phi\, z\,\bar{a}_z$$

$$= \int_0^{2\pi}\int_{u=z}^{\infty} \frac{\rho_s}{4\pi\epsilon_0}\frac{du}{u^2}\,d\phi\, z\,\bar{a}_z$$

$$= \int_0^{2\pi} \frac{\rho_s}{4\pi\epsilon_0}\,d\phi\, z\,\bar{a}_z\left[-\frac{1}{u}\right]_z^{\infty}$$

$$= \frac{\rho_s}{4\pi\epsilon_0}\left[\phi\right]_0^{2\pi}(z\,a_z)\left[-\frac{1}{\infty} - \left[-\frac{1}{z}\right]\right]$$

	$= \dfrac{\rho_s}{4\pi\epsilon_0}(2\pi)\, a_z$ $E = \dfrac{\rho_s}{2\epsilon_0} a_z \ \ v/m$ As a_z is direction normal to differential surface area ds considered. $E = \dfrac{\rho_s}{2\epsilon_0} \bar{a}_n \ \ v/m.$ [For above xy plane] $a_n = -a_z$ $E = \dfrac{-\rho_s}{2\epsilon_0} a_z \ \ v/m$ [For below xy plane]	
8(a)	**Three point charges - 1 nC, 4 nC, and 3 nC are located at (0, 0, 0), (0, 0, 1), and (1, 0, 0), respectively. Find the energy in the system.** Total energy contained in the system is: $W = W_1 + W_2 + W_3$ $= 0 + Q_2V_{21} + Q_3(V_{31} + V_{32})$ $= Q_2 \cdot \dfrac{Q_1}{4\pi\epsilon_0 \lvert(0,0,1)-(0,0,0)\rvert} + \dfrac{Q_3}{4\pi\epsilon_0}\left[\dfrac{Q_1}{\lvert(1,0,0,)-(0,0,0)\rvert} + \dfrac{Q_2}{\lvert(1,0,0,)-(0,0,1)\rvert}\right]$ $= \dfrac{1}{4\pi\epsilon_0}\left(Q_1Q_2 + Q_1Q_3 + \dfrac{Q_2Q_3}{\sqrt{2}}\right)$ $= \dfrac{1}{4\pi \times \dfrac{10^{-9}}{36\pi}}\left(-4 - 3 + \dfrac{12}{\sqrt{2}}\right) \times 10^{-18}$ $= 9\left(\dfrac{12}{\sqrt{2}} - 7\right) \text{nJ} = 13.37 \text{ nJ}$	6M

8(b)	Illustrate the properties of materials in electric field	6M

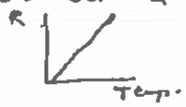
Properties of Materials

Materials

- Conductors.
  (Allows the free flow of $\bar{e}$
  Ex :- Copper, Iron, Silver, gold; Aluminiu
- Semiconductors
  Partially allows the flow of $\bar{e}$
  Ex :- Ge, Si
- Insulators:
  which offers high resistive path for flow of $\bar{e}$
  Ex :- wood, plastic, rubber.

Conductors:

* In conductors electric field is zero. as electric field is zero the charge density $\rho=0$. It is a +ve temp coefficient.



Semi Conductors :-

* At 0°c it acts as a Insulator.
* At room temp. it acts as Semiconductor (or) conducts
* It has -ve temp coefficient.

Insulations :

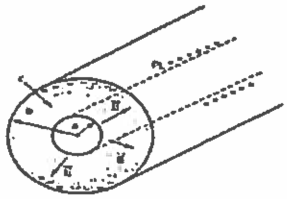
* High Resistance is offered.
* Will have high dielectric strength.
* It has high thermal strength
* Has low permetivity and low thermal expansion.

(OR)

9(a) Derive the expression for capacitance of coaxial cable. **8M**

Consider a co-axial cable with "a" and "b" being inner and outer radius, and "l" being the length.

Inner conductor carries a charge of $+\rho_L$ and outer conductor with $-\rho_L$.



$Q = \rho_L \times L$

Electric field intensity is given by

$$\bar{E} = \frac{\rho_L}{2\pi\varepsilon r}\bar{a}_r$$

E is directed from inner to outer conductor. The potential difference is work-done in moving a unit charge against E from b to a

$$V = -\int_{-}^{+} \bar{E} \cdot d\bar{L} = -\int_{r=b}^{r=a} \frac{\rho_L}{2\pi\varepsilon r}\bar{a}_r \cdot dr\,\bar{a}_r$$

$$= -\frac{\rho_L}{2\pi\varepsilon}\left[\ln r\right]_b^a = -\frac{\rho_L}{2\pi\varepsilon}\ln\left[\frac{a}{b}\right]$$

$$V = \frac{\rho_L}{2\pi\varepsilon}\ln\left[\frac{b}{a}\right] \text{ V}$$

the capacitance is given by,

$$C = \frac{Q}{V} = \frac{\rho_L \times L}{\frac{\rho_L}{2\pi\varepsilon}\ln\left[\frac{b}{a}\right]}$$

$$\boxed{C = \frac{2\pi\varepsilon L}{\ln\left[\frac{b}{a}\right]} \text{ F}}$$

9(b)	Explain Dielectric-Dielectric boundary conditions	6M

→ Dielectric to Dielectric boundary condition ①

The boundary conditions explain relations b/w vector fields at dielectric interface only. [not at points above or below]

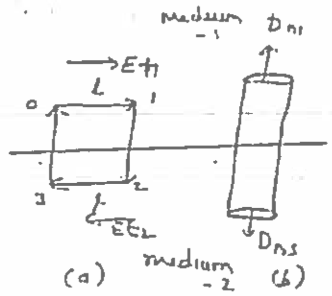
Let permittivity of 2 mediums be ϵ_1 & ϵ_2

$\epsilon_1 = \epsilon_0 \epsilon_{r1}$ & $\epsilon_2 = \epsilon_0 \epsilon_{r2}$

$\bar{D}_1, \bar{D}_2$ are flux densities $\bar{E}_1, \bar{E}_2$ are EFI

Consider a box, which encloses 2 mediums & ΔS be surface area & assume it is charge free

According to Gauss law, flux coming out of charged body is equal to charge enclosed by that body



$D_{n1} \Delta S - D_{n2} \Delta S = q$

but $q = \rho_s \cdot \Delta S$

$\therefore D_{n1} \Delta S - D_{n2} \Delta S = \rho_s \Delta S$

$D_{n1} - D_{n2} = \rho_s$

As assumed earlier that it is charge free

$\rho_s = 0$

$D_{n1} - D_{n2} = 0$

$D_{n1} = D_{n2}$

As $D = \epsilon E$

$D_{n1} = \epsilon_1 E_{n1}$

$D_{n2} = \epsilon_2 E_{n2}$

$$\frac{D_{n1}}{D_{n2}} = \frac{\epsilon_1 E_{n1}}{\epsilon_2 E_{n2}} = 1 \Rightarrow \frac{E_{n1}}{E_{n2}} = \frac{\epsilon_2}{\epsilon_1} = \frac{\epsilon_{r2}}{\epsilon_{r1}}$$

Consider the rectangular closed path "0 1 2 3 0"
workdone in moving a unit charge around closed path
is zero. Thus

$$\oint E \cdot dl = 0 \Rightarrow \text{Applying limit}$$

$$\oint_0^1 E \cdot dl + \oint_1^2 E \cdot dl + \oint_2^3 E \cdot dl + \oint_3^0 E \cdot dl = 0$$

workdone in moving the charge from 1 to 2; 3 to 0 is zero

$$\oint_0^1 E \cdot dl + \oint_2^3 E \cdot dl = 0 \Rightarrow E_{t1} \cdot \Delta L - E_{t2} \Delta L = 0$$

$$\boxed{E_{t1} = E_{t2}}$$

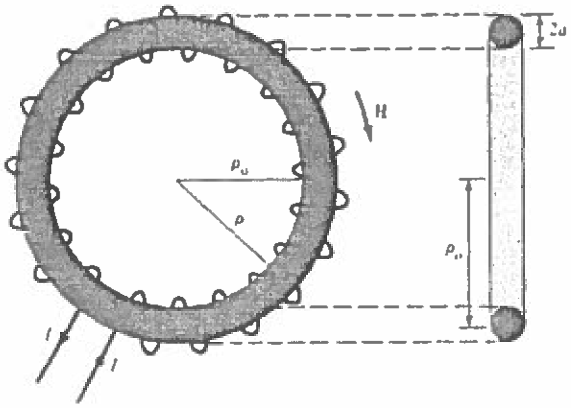
As we know flux density $\bar{D} = \epsilon \bar{E}$

$$D_{t1} = \epsilon_1 E_{t1} \quad D_{t2} = \epsilon_2 E_{t2}$$

D_{t1} & D_{t2} are tangential component of flux Density

$$\frac{E_{t1}}{E_{t2}} \Rightarrow \frac{D_{t1}}{\epsilon_1} = \frac{D_{t2}}{\epsilon_2}$$

$$\boxed{\frac{D_{t1}}{D_{t2}} = \frac{\epsilon_1}{\epsilon_2} = \frac{\epsilon_{r1}}{\epsilon_{r2}}}$$

10	Toroid whose dimensions are shown in Figure1 has N turns and carries current I. Determine H inside and outside the toroid.  We apply Ampère's circuit law to the Amperian path, which is a circle of radius ρ shown dashed in Figure Since N wires cut through this path each carrying current I, the net current enclosed by the Amperian path is NI. Hence, $$\oint H \cdot dl = I_{enc} \rightarrow H \cdot 2\pi\rho = NI$$ or $$H = \frac{NI}{2\pi\rho}, \text{ for } \rho_o - a < \rho < \rho_o - a$$ where ρ_o is the mean radius of the toroid as shown in Figure An approximate value of H is $$H_{approx} = \frac{NI}{2\pi\rho_o} = \frac{NI}{\ell}$$ Notice that this is the same as the formula obtained for H for points well inside a very long solenoid ($\ell \gg a$). Thus a straight solenoid may be regarded as a special toroidal coil for which $\rho_o \rightarrow \infty$. Outside the toroid, the current enclosed by an Amperian path is $NI - NI = 0$ and hence $H = 0$.	12M
	(OR)	
11	**Derive Maxwell's third equation, MFI due to an infinite sheet of current carrying conductor.** Maxwell's Third Equation: Statement: It states that closed integral of Magneto Motive Force is equal to surface integral of current density taken	12M

over the surface enclosed by closed path.

In point form it is expressed as

$$\nabla \times \bar{H} = \bar{J}$$

Maxwell third equation is derived from Ampere's circuit law which states that line integral of magnetic field taken around any closed path is equal to current enclosed by that path. Mathematically,

$$\oint \bar{H} \cdot \bar{dl} = I.$$

Also

$$I = \int_S \bar{J} \cdot \bar{ds}$$

by substituting in the above

$$\oint \bar{H} \cdot \bar{dl} = \int \bar{J} \cdot \bar{ds}$$

Now applying stokes theorem to above

L.H.S.

$$\oint \bar{H} \cdot \bar{dl} = \int_S (\nabla \times \bar{H}) \cdot \bar{ds}$$

$$= \int \bar{J} \cdot \bar{ds}$$

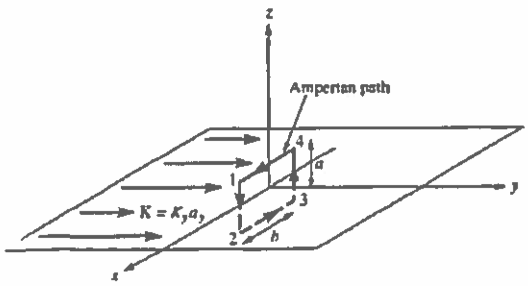
So $\Rightarrow$ $$\int_S (\nabla \times \bar{H}) \, ds = \int \bar{J} \cdot \bar{ds}$$

From above

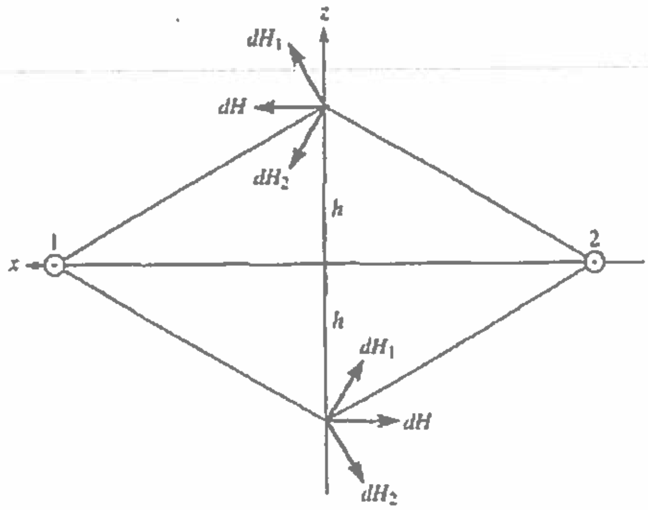
$$\nabla \times \bar{H} = \bar{J}$$

MFI due to infinite sheet of current

Consider an infinite current sheet in the $z = 0$ plane.



Application of Ampere's law to an infinite sheet: closed path 1-2-3-4-1,



Symmetrical pair of current filaments with current along a_y.

If the sheet has a uniform current density $K = K_y a_y$ A/m as shown in figure, applying Ampere's law to the rectangular closed path (Amperian path) gives

$$\oint H.dl = I_{enc} = K_y.b$$

To evaluate the integral, we first need to have an idea of what H is like. To achieve this, we regard the infinite sheet as comprising of filaments; As evident in Figure above the resultant dH has only an x-component. Also, H on one side of the sheet is the negative of that on the other side. Due to the infinite extent of the sheet, the sheet can be regarded as consisting of such filamentary pairs so that the characteristics of H for a pair are the same for the infinite current sheets, that is,

$$H = \begin{cases} H_o a_x & ;z > 0 \\ -H_o a_x & ;z < 0 \end{cases}$$

$$\oint H.dl = \left(\int_1^2 + \int_2^3 + \int_3^4 + \int_4^1 \right) H.dl$$

For path 1-2 $d\overline{L} = dz\overline{a_x}$,

Fro path 3-4 $d\overline{L} = dz\overline{a_x}$

But $\overline{H}$ is in x direction while $\bar{a}_x \cdot \bar{a}_z = 0$.
Hence along the paths 1-2 and 3-4, the integral $\oint \overline{H} \cdot dL = 0$.

Consider path 2-3 along which $d\bar{L} = dx\,\bar{a}_x$.

$$\therefore \quad \int_2^3 H \cdot dL = \int_2^3 (-H_x\,\bar{a}_x)\cdot(dx\,\bar{a}_x) = H_x \int_2^3 dx = b\,H_x$$

Consider path 4-1 along which $d\bar{L} = dx\,\bar{a}_x$ and it is in the region $z > 0$ hence $\bar{H} = H_x\,\bar{a}_x$.

$$\oint H.dl = 0(-a) + (H_x)(b) + 0.a + H_x.b =_2 H_o b$$

Now equating the above equation

$$2\,b\,H_x = K_y\,b$$

$$H_x = \frac{1}{2}K_y$$

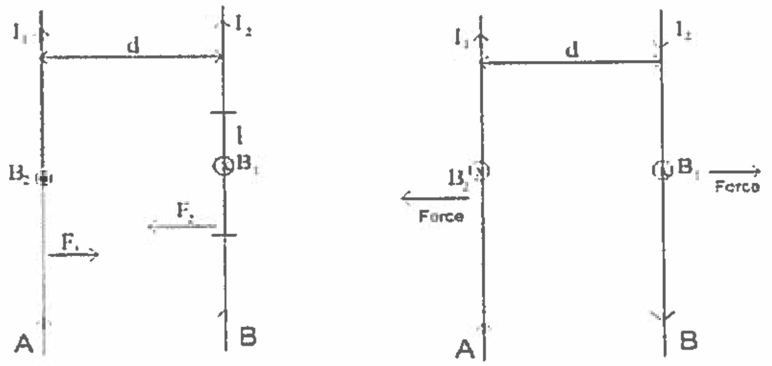
$$H = \begin{cases} \frac{1}{2}K_y.a_x & ;z > 0 \\ -\frac{1}{2}K_y.a_x & ;z < 0 \end{cases}$$

In general, for an infinite sheet of current density K A/m.

$$H = \frac{1}{2}K \times a_n$$

12 **Derive an expression for force between two straight long and parallel current carrying conductor.** 12M

It is experimentally established fact that two current carrying conductors attract each other when the current is in same direction and repel each other when the current are in opposite direction Figure below shows two long parallel wires separated by distance d and carrying currents $I_1$and I_2

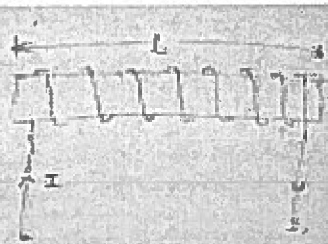


(A) (B)

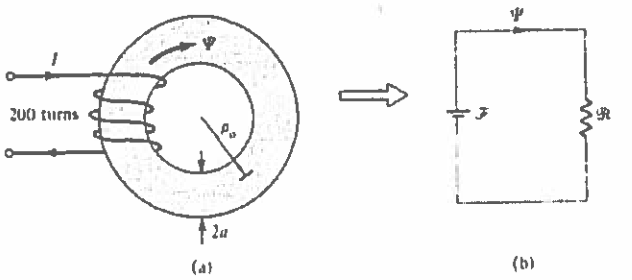
Consider fig wire A will produce a field B_1 at all nearby points .The magnitude of B_1 due to current I_1 at a distance d i.e. on wire b is

$B_1 = \mu_0 I_1 / 2\pi d$

	According to the right hand rule the direction of B_1 is in downward as shown in figure. Consider length l of wire B and the force experienced by it will be (I_2lXB) whose magnitude is $F_2 = I_2lB = \frac{\mu_0}{2\pi}\frac{lI_1I_2}{d}$ Direction of F_2 can be determined using vector rule .F_2 Lies in the plane of the wires and points to the left From figure we see that direction of force is towards A if I_2 is in same direction as I_1 fig and is away from A if I_2 is flowing opposite to I_1 Force per unit length of wire B is $\frac{F_2}{l} = \frac{\mu_0}{2\pi}\frac{I_1I_2}{d}$ Similarly force per unit length of A due to current in B is $\frac{F_1}{l} = \frac{\mu_0}{2\pi}\frac{I_1I_2}{d}$ and is directed opposite to the force on B due to A. Thus the force on either conductor is proportional to the product of the current We can now make a conclusion that the conductors attract each other if the currents are in the same direction and repel each other if currents are in opposite direction.	
(OR)		
13(a)	**Calculate the self-inductance per unit length of an infinitely long solenoid.** Inductance of Solenoid → Consider a solenoid of N turns with current flowing be 'I' amps and cross sectional area of 'A' Field intensity inside a solenoid be $H = \frac{NI}{l}$ A/m Total flux linkage $= N\phi$ $= N(B)(A) = N(\mu H)A$ $= \mu N\left[\frac{NI}{l}\right]A = \frac{N^2\mu IA}{l}$ Inductance $L = \frac{\text{Total flux linkage}}{\text{Total current}}$ $L = \frac{\mu N^2 I A}{l(I)}$ $\boxed{L = \frac{\mu N^2 A}{l}}$ H	8M



13(b) The toroidal core of Figure 2 has ρ_o = 10 cm and a circular cross section with a = 1 cm. If the core is made of steel (μ = 1000μ_o) and has a coil with 200 turns, calculate the amount of current that will produce a flux of 0.5 mWb in the core. **4M**



Since ρ_o is large compared with a,

$$B = \frac{\mu N I}{\ell} = \frac{\mu_o \mu_r N I}{2\pi\rho_o}$$

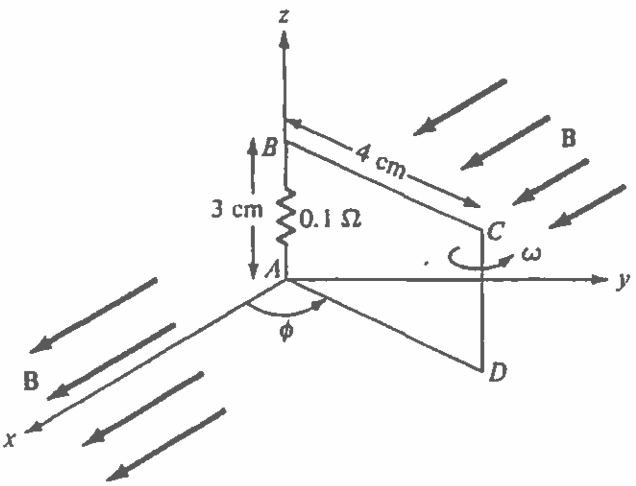
Hence

$$\Psi = BS = \frac{\mu_o \mu_r N I \pi a^2}{2\pi\rho_o}$$

$$I = \frac{2\rho_o \Psi}{\mu_o \mu_r N a^2} = \frac{2(10\times10^{-2})(0.5\times10^{-3})}{4\pi\times10^{-7}(1000)(200)(1\times10^{-4})} = \frac{100}{8\pi} = 3.979 A$$

14 The loop shown in Figure is inside a uniform magnetic field B = 50a_x mWb/m². If side DC of the loop cuts the flux lines at the frequency of 50 Hz and the loop lies in the yz-plane at time t = 0, find **12M**

(a) The induced emf at t = 1 ms
(b) The induced current at t = 3 ms



(a) Since the B field is time invariant, the induced emf is motional, that is

$$V_{emf} = \int_L (u \times B) \cdot dl$$

where

$$dl = dl_{DC} = dz a_z, \quad u = \frac{dl'}{dt} = \frac{\rho d\phi}{dt} a_\phi = \rho\omega a_\phi$$

$$\rho = AD = 4cm, \quad \omega = 2\pi f = 100\pi$$

Because u and dl are in cylindrical coordinates, we transform B into cylindrical coordinates by using eq. (2.9):

$$a_x = \cos\phi a_\rho - \sin\phi a_\phi$$
$$a_y = \sin\phi a_\rho - \cos\phi a_\phi$$
$$a_z = a_z$$

$$B = B_o a_x = B_o(\cos\phi a_\rho - \sin\phi a_\phi)$$

where $B_o = 0.05$. Hence

$$u \times B = \begin{vmatrix} a_\rho & a_\phi & a_z \\ 0 & \rho\omega & 0 \\ B_o\cos\phi & -B_o\sin\phi & 0 \end{vmatrix} = -\rho\omega B_o \cos\phi a_z$$

$$(u \times B) \cdot dl = -\rho\omega B_o \cos\phi dz$$

$$= -0.04(100\pi)(0.05)\cos\phi dz = -0.2\pi\cos\phi dz$$

$$V_{emf} = \int_{z=0}^{0.03} -0.2\pi\cos\phi dz = -6\pi\cos\phi mV$$

To determine ϕ, recall that

$$\omega = \frac{d\phi}{dt} \rightarrow \phi = \omega t + C_o$$

where C_o is an integration constant. At $t = 0$, $\phi = \pi/2$ because the loop is in the yz-plane at that time, $C_o = \pi/2$. Hence

$$\phi = \omega t + \frac{\pi}{2}$$

and

$$V_{emf} = -6\pi\cos\left(\omega t + \frac{\pi}{2}\right) = 6\pi\sin(100\pi t) mV$$

At $t = 1ms$, $V_{emf} = 6\pi\sin(0.1\pi) = 5.825mV$

(b) The current induced is

$$i = \frac{V_{emf}}{R} = 60\pi \sin(100\pi t) mA$$

At $t = 3ms$

$$i = 60\pi \sin(0.3\pi) mA = 0.1525 A$$

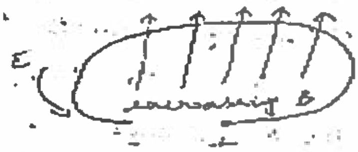
(OR)

15(a) Explain statically induced emf.

→ Statically Induced EMF:

This is a condition in which closed path is stationary & magnetic field B is varying with time.

The closed circuit in which EMF is induced is stationary & magnetic flux is varying sinusoidally with time



$$\oint \bar{E}\cdot d\bar{l} = -\int_S \frac{\partial B}{\partial t}\cdot d\bar{s}$$

Magnetic flux density is only quantity varying with time

Using partial derivative to define relationship as B may be changing with co-ordinates as well as time

Using stokes theorem

$$\int_S (\nabla\times E)\cdot ds = -\int_S \frac{\partial B}{\partial t}\cdot ds$$

Assuming both surfaces takes are identical

$$\therefore (\nabla\times E)\cdot ds = -\frac{\partial \bar{B}}{\partial t}\cdot ds$$

$$\boxed{\nabla\times E = \frac{-\partial \bar{B}}{\partial t}}$$

If $\bar{B}$ is not varying with time

$$\oint \bar{E}\cdot d\bar{l} = 0 \quad \& \quad \nabla\times\bar{E} = 0$$

15(b)	Express the integral form of Faraday's law	
	Integral Form: The closed line integral of EFI around loop of conductor is equal to surface integral of rate of change of flux density with respect to time, over loop area. i.e., $e = \oint \bar{E} \cdot \bar{dl} = \frac{-d}{dt} \int_s \bar{B} \cdot \bar{ds}$ **Derivation:** According to Faradays law $e = \oint \bar{E} \cdot \bar{dl}$ magnetic flux passing through specified area $\phi = \int_s \bar{B} \cdot \bar{ds}$ $e = \frac{-d}{dt} \int_s \bar{B} \cdot \bar{ds}$ From above $e = \oint \bar{E} \cdot \bar{dl} = \frac{-d}{dt} \int \bar{B} \cdot \bar{ds}$	

Prepared by

ABRR

(Mr. A. B. R. Rao)

NSRIT

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	**Program**	ECE			**Academic Year**	2021 - 2022
Course Code	20EC405	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	IV
Course	**Electronic Circuit Analysis**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Draw the small signal high frequency CE model of a transistor and list its elements.	20EC405.1	L1
2	Express the current gain for Darlington pair.	20EC405.2	L1
3	Identify any three advantages of negative feedback amplifier.	20EC405.3	L1
4	List any four types of oscillator.	20EC405.4	L1
5	Identify the factors that influences on the selectivity of a single tuned amplifier.	20EC405.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Derive expression for the CE short circuit current gain A_I as a function of frequency.	9M	20EC405.1	L2
6 (b)	Draw Hybrid - π model for a transistor in the CB configuration.	3M	20EC405.1	L2
	OR			
7 (a)	State and explain Miller's theorem.	8M	20EC405.1	L2
7 (b)	Draw Hybrid - π model for a transistor in the CE configuration	4M	20EC405.1	L2
8 (a)	Explain three types of coupling methods used in multistage amplifiers	8M	20EC405.2	L2
8 (b)	Draw the circuit diagram of cascade (Two stage RC coupled) amplifier with and without biasing circuit. Also mention the advantages.	4M	20EC405.2	L3
	OR			
9	Draw and explain Darlington emitter follower configurations with respect to i. current gain ii. input impedance iii. voltage gain iv. output impedance and compare with emitter follower.	12M	20EC405.2	L2
10	Draw the circuit for voltage shunt amplifier and justify the type of feedback. Also derive the expressions for AV, β, input and output resistance with feedback	12M	20EC405.3	L2
	OR			
11	Draw the circuit for Voltage series feedback amplifier and derive the expressions for A_f and β for the circuit. Also mention the advantages.	12M	20EC405.3	L3
12	Derive the expression frequency of oscillation and condition for sustained oscillations of a FET based RC Phase shift oscillator.	12M	20EC405.4	L3
	OR			
13	Describe the operation of Hartley oscillator circuit using bipolar junction transistor with necessary diagrams.	12M	20EC405.4	L3
14 (a)	Describe the operation of class B push pull amplifier and also	9M	20EC405.5	L2

	explain how the crossover distortion is minimized?			
14 (b)	Identify the effects of Harmonic distortions in power amplifiers.	3M	20EC405.5	L2
	OR			
15 (a)	With a neat diagram show how to cascade tuned (staggered) amplifier and explain briefly.	8M	20EC405.5	L2
15 (b)	Describe the features of single tuned amplifier.	4M	20EC405.5	L2

N S RAJU INSTITUTE OF TECHNOLOGY

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ANSWER KEY AND SCHEME OF EVALUATION

Degree	B. Tech. (U. G.)	**Program**	ECE			**Academic Year**	2021 - 2022
Course Code	20EC405	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	IV
Course	**Electronic Circuit Analysis**						

1	Circuit diagram of small signal high frequency CE model of a transistor and list its elements.	2M
2	Expression for current gain for Darlington pair.	2M
3	Identification of three advantages of negative feedback amplifier.	2M
4	List of four types of oscillators	2M
5	Identify of factors that influences on the selectivity of a single tuned amplifier.	2M
6 (a)	Circuit diagram Derivation of expression for the CE short circuit current gain A_I	3M 6M
6 (b)	Circuit diagram for Hybrid - π model for a transistor in the CB configuration	3M
7 (a)	Statement for Miller's theorem Explanation of Miller's theorem.	3M 5M
7 (b)	Circuit diagram for Hybrid - π model for a transistor in the CE configuration	4M
8 (a)	Three circuit diagrams Explanation of three types of coupling methods used in multistage amplifiers	3M 5M
8 (b)	Circuit diagram of cascade (Two stage RC coupled) amplifier with Without biasing circuit Advantages.	3M 1M
9	Circuit diagram and explanation of Darlington emitter follower i. current gain ii. input impedance iii. voltage gain iv. output impedance Comparison with emitter follower.	2M 2M 2M 2M 2M 2M
10	Circuit for voltage shunt feedback amplifier Justification of the type of feedback Derivation for A_V and β Expression for input and output resistance with feedback	2M 1M 6M 3M
11	Circuit for Voltage series feedback amplifier derivation for A_f and β Advantages	3M 7M 2M

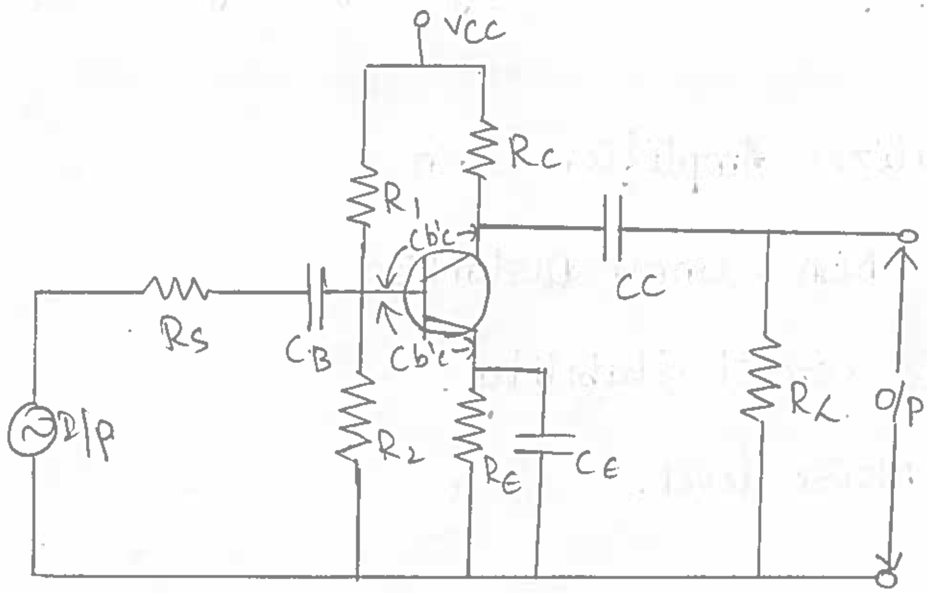
12	Circuit diagram for a FET based RC Phase shift oscillator Derivation for frequency of oscillation condition for sustained oscillations	2M 7M 3M
13	Circuit diagrams of Hartley oscillator using bipolar junction transistor Operation of Hartley oscillator Derivation for frequency of oscillations	3M 3M 6M
14 (a)	Circuit diagram Operation of class B push pull amplifier Explanation of crossover distortion minimization	3M 4M 2M
14 (b)	Identification of effects of Harmonic distortions in power amplifiers.	3M
15 (a)	Circuit diagram of cascade (staggered) tuned amplifier Explanation	3M 5M
15 (b)	Four features of single tuned amplifier	4M

IV/II-II SEMESTER END REGULAR EXAMINATION, JUNE, 2022

COURSE CODE: 20EC405 (ELECTRONIC CIRCUIT ANALYSIS)

KEY

Draw the Small Signal high frequency CE Model of a transistor and list its Elements.



where, R_C = Collector Resistance

R_E = Emitter Resistance

C_C = Collector Capacitance

R_L = Load Resistance

C_E = Emitter Capacitance

2. Express the Current gain for Darlington pair.

$$A_{i_1} = \frac{1+h_{fe}}{1+h_{oe}h_{fe}R_E}$$

$$A_{i_2} = 1+h_{fe}$$

$$A_i = A_{i_1} \times A_{i_2}$$

$$= \frac{1+h_{fe}}{1+h_{oe}h_{fe}R_E} \times (1+h_{fe})$$

$$A_i = \frac{(1+h_{fe})}{1+h_{oe}h_{fe}R_E}$$

3) Identify any three Advantages of negative feedback amplifier.

(1) It Stabilizes Amplifier Gain

(2) Reduces Non-linear Distortion.

(3) Increases Circuit Stability.

(4) Reduces Noise level.

4) List Out four types of Oscillator.

(1) Hartley Oscillator

(2) Colpitts Oscillator

(3) Wein Bridge Oscillator

(4) RC phase Shift Oscillator.

(5) Crystal Oscillator.

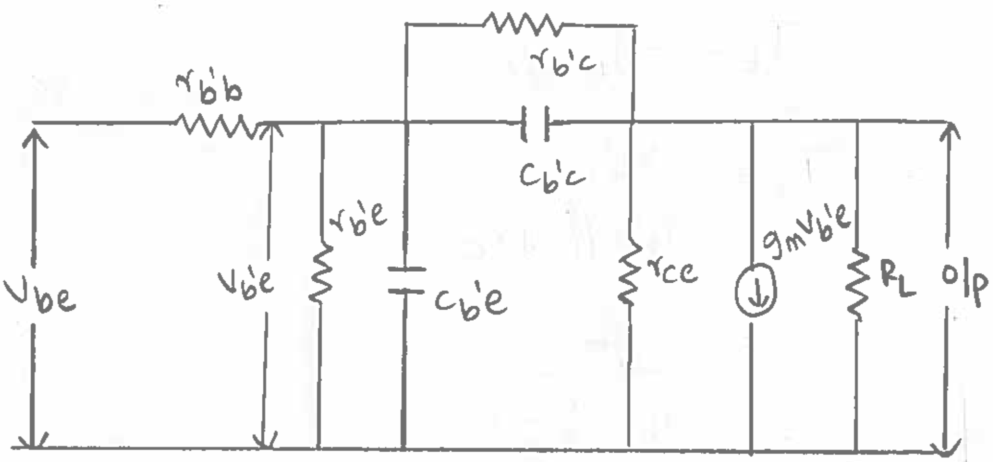
5) Identify the factors that Influences On the Selectivity of a Single tuned amplifier.

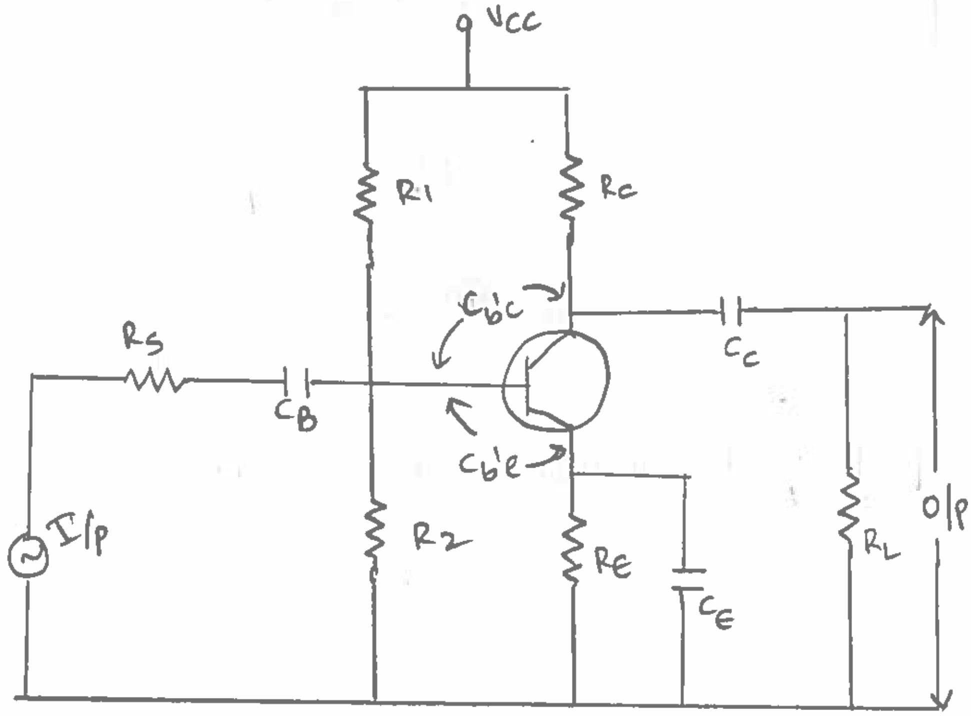
(1) To Select the desired Carrier frequency and to amplify the allowed bandwidth around this

Selected Carrier frequency.

(2) To obtain higher gain and improved discriminating Property, a tuned circuit with higher q-factor is employed.

2) Derive expression for the CE Short circuit Current gain A_I as a function of frequency.

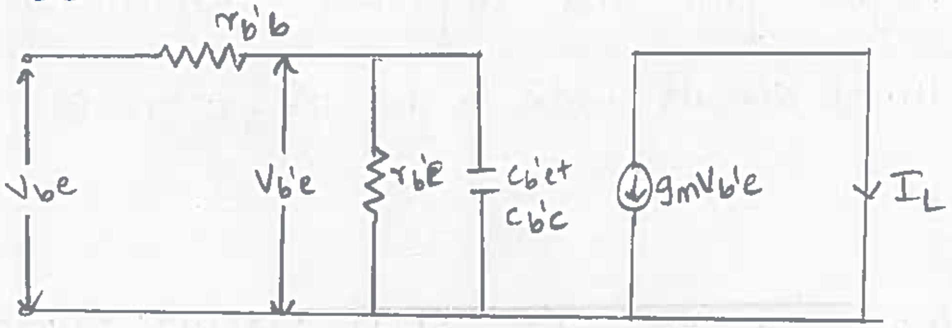




If Output is Shorted R_L is Zero, $C_{b'c}$ is in Series with $C_{b'e}$ and Parallel to $r_{b'e}$ and R_{ce}

is Zero.

(iii) $r_{b'c} >> r_{b'e}$



Current gain $A_i = \frac{I_L}{I_i}$

$$I_L = -g_m V_{b'e}$$

$$I_i = \frac{V_{b'e}}{r_{b'e} // -jX_C}$$

$$= \frac{V_{b'e}}{r_{b'e} // \frac{-j}{\omega(C_{b'e}+C_{b'c})}}$$

$$= \frac{V_{b'e}}{\frac{r_{b'e} \cdot -j/\omega(C_{b'e}+C_{b'c})}{r_{b'e} + \frac{(-j)}{2\pi f(C_{b'e}+C_{b'c})}}}$$

Now, multiply and divide denominators by $r_{b'e}$.

$$= \frac{V_{b'e}}{\frac{\frac{-j\, r_{b'e}}{2\pi f(C_{b'e}+C_{b'c})} \times \frac{r_{b'e}}{r_{b'e}}}{r_{b'e} - \frac{j}{2\pi f(C_{b'e}+C_{b'c})} \times \frac{r_{b'e}}{r_{b'e}}}}$$

$$= \cfrac{V_{b'e}}{\cfrac{\cfrac{-j(r_{b'e})^2}{2\pi f r_{b'e}(C_{b'e}+C_{b'c})}}{r_{b'e} - \cfrac{j r_{b'e}}{2\pi f (r_{b'e}(C_{b'e}+C_{b'c}))}}}$$

Consider $f_\beta = \dfrac{1}{2\pi\, r_{b'e}(C_{b'e}+C_{b'c})}$

$$I_i = \cfrac{V_{b'e}}{\cfrac{-j(r_{b'e})^2 f_\beta/f}{r_{b'e} - \cfrac{j r_{b'e} f_\beta}{f}}}$$

$$= \cfrac{V_{b'e}}{\cfrac{\cfrac{-j(r_{b'e})^2 \cdot f_\beta}{\cancel{f}}}{\cfrac{f r_{b'e} - j r_{b'e} f_\beta}{\cancel{f}}}} = \cfrac{V_{b'e}}{\cfrac{-j(r_{b'e})^{\cancel{2}} f_\beta}{\cancel{r_{b'e}}(f - f_\beta j)}}$$

$$I_i = \cfrac{V_{b'e}}{\cfrac{-j r_{b'e} f_\beta}{f - j f_\beta}}$$

$$A_I = -\cfrac{g_m \cancel{V_{b'e}}}{\cfrac{\cancel{V_{b'e}}}{\cfrac{-j r_{b'e} f_\beta}{f - j f_\beta}}} = \frac{j g_m r_{b'e} f_\beta}{f - j f_\beta}$$

Divide the Numerator & Denominator by J

$$A_I = \frac{\cancel{j} g_m r_{b'e} f_\beta / \cancel{j}}{f - j f_\beta / j}$$

$$= \frac{g_m r_{b'e} f_\beta}{\left(\frac{f}{j} - f_\beta\right)} = \frac{g_m r_{b'e} f_\beta}{-jf - f_\beta}$$

$$= \frac{g_m r_{b'e} f_\beta}{-(jf + f_\beta)}$$

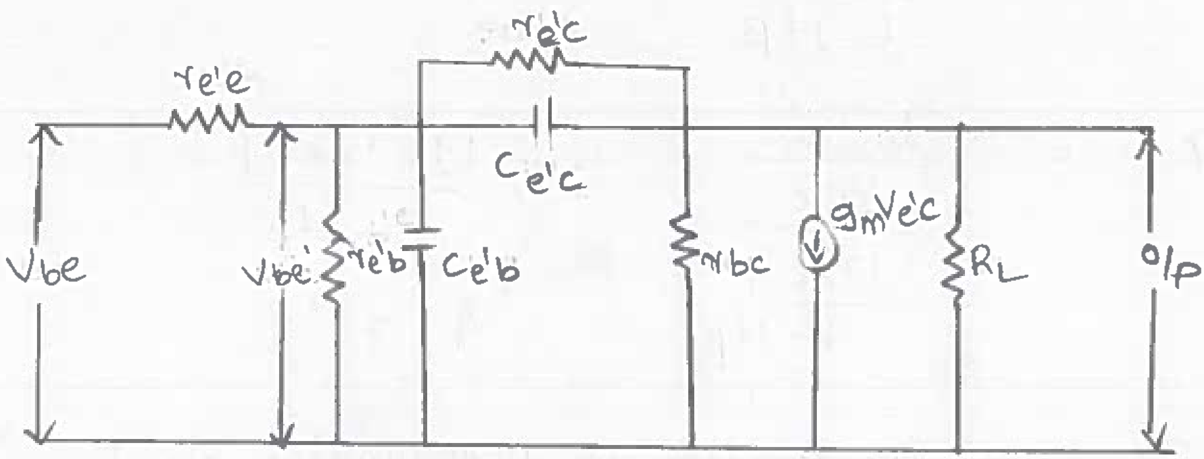
Divide the Numerator & Denomenator with "$-f_\beta$"

$$A_I = \frac{\frac{g_m r_{b'e} \cancel{f_\beta}}{-\cancel{f_\beta}}}{\frac{\cancel{-}(jf + f_\beta)}{\cancel{-}f_\beta}} = \frac{-g_m r_{b'e}}{j\frac{f}{f_\beta} + 1}$$

$[\because g_m r_{b'e} = h_{fe}]$

$$A_I = \frac{-h_{fe}}{1 + j\left(\frac{f}{f_\beta}\right)}$$

6) B) Draw Hybrid - π Model for a transistor in the CB Configuration.



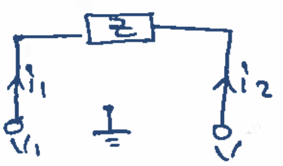
7a) State and explain miller's theorem:-

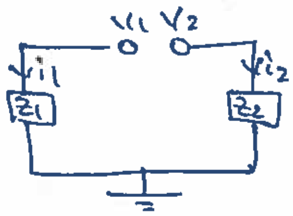
Definition :- It states that if an impedance "Z" is connected between input and output terminals of a network which provides a voltage gain "A". An equivalent circuit that gives the same effect can be drawn by removing Z and connecting an impedance $Z_i = \frac{Z}{1-A}$ across the input and $Z_o = \frac{ZA}{A-1}$ across the output.

if $Z = R$, $R_1 = \frac{R}{1-A}$

$R_2 = \frac{RA}{A-1}$

Miller's theorem:-





$$i_1 = \frac{V_1 - V_2}{Z} \rightarrow (1)$$

$$i_2 = \frac{V_2 - V_1}{Z} \rightarrow (2)$$

$$i_1 = \frac{V_1 - 0}{Z_1}$$

$$i_2 = \frac{V_2 - 0}{Z_2}$$

equating eqn (1) & (2)

$$\Rightarrow \frac{V_1 - V_2}{Z} = \frac{V_1 - 0}{Z_1}$$

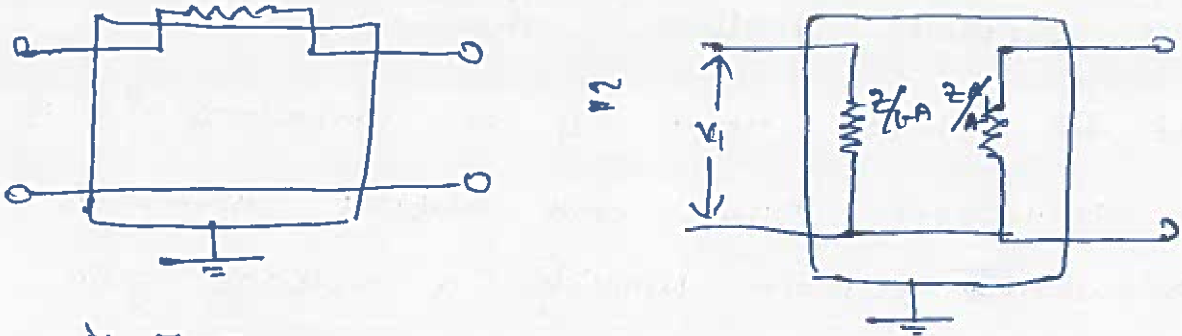
$$\Rightarrow Z_1 = \frac{ZV_1}{V_1 - V_2} \Rightarrow \frac{V_1 Z}{V_1\left(1 - \frac{V_2}{V_1}\right)}$$

$$\Rightarrow Z_1 = \frac{Z}{1-A}$$

equating eqn (2) & (4)

$$\frac{V_2 - V_1}{Z} = \frac{V_2 - 0}{Z}$$

$$\frac{V_2 - V_1}{Z} = \frac{V_2}{Z}$$



If $Z = C$

1. $\frac{1}{j\omega C_1} = \frac{1}{j\omega C(1-A)}$

$\frac{1}{C_1} = \frac{1}{C(1-A)}$

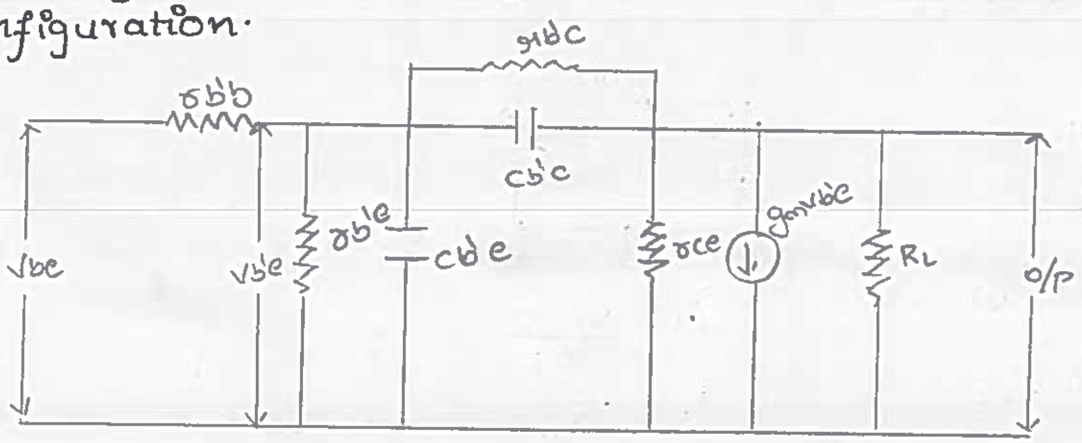
$C_1 = C(1-A)$

2. $\frac{1}{j\omega C_2} = \frac{A}{j\omega C(A-1)}$

$\frac{1}{C_1} = \frac{A}{C(A-1)}$

$C_2 = \frac{C(A-1)}{A}$

~~7)~~ 8) Draw Hybrid-π model for a transistor in the CE configuration.



$r_{b'b}$ = resistance between acutal base and vitual base

$r_{b'c}$ = resistance between vitual base and collector base

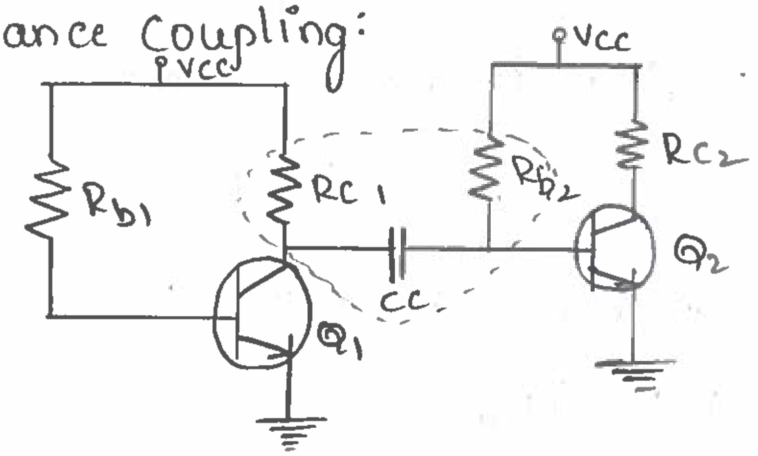
$C_{b'c}$ = capacitance between vitual base and collector terminal

r_{ce} = resistance between collector and emitter terminal

V_{be} = voltage between base and emitter terminal

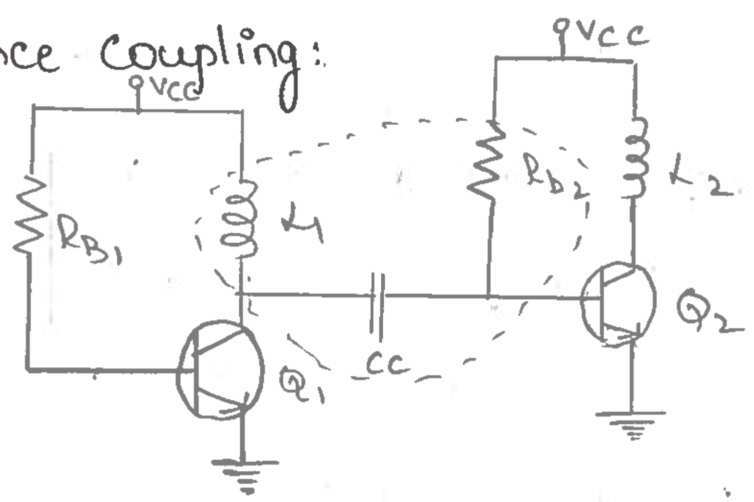
8(a) Explain ~~these~~ three types of coupling Methods used in multistage amplifiers.

1) Resistance and Capacitance Coupling / Resistive Capacitance Coupling:



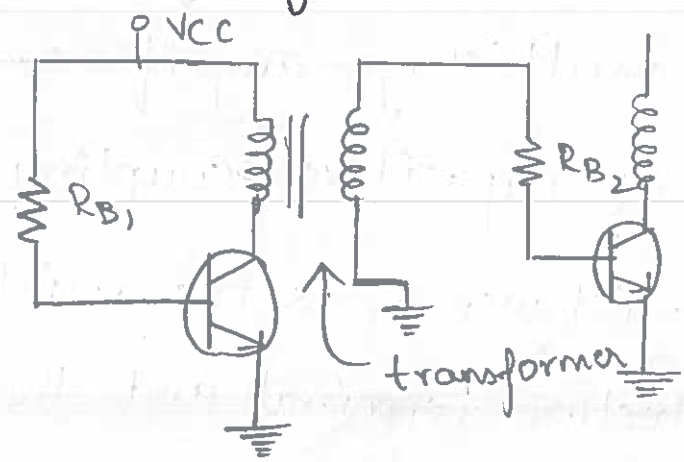
Rc Coupling is the most widely used method of coupling in multistage amplifiers. It is an application of capacitive Coupling. In this case the Resistance R is the resistor connected at the collector terminal and the capacitor c is connected in between the amplifiers.

2) Impedence Coupling:



Impedance Coupling is very similar to RC Coupling. The difference is the use of impedence device to replace the load resistor of the first stage. Impedance Coupled transistor amplifier. The formula shows that for inductive reactance to be large, either inductance (or) frequency or both must be high.

3) Transformer Coupling:



Transformer Coupling is frequently used to step up transmission line signals. voltage signals amplified in this way are not constrained by local supply voltages, so the amplifier's rated current rather than its voltage swing usually limits the power delivered to the load.

8(b) Draw the circuit diagram of Cascade (Two stage RC Coupled) amplifier with and without biasing circuit. Also mention the advantages.

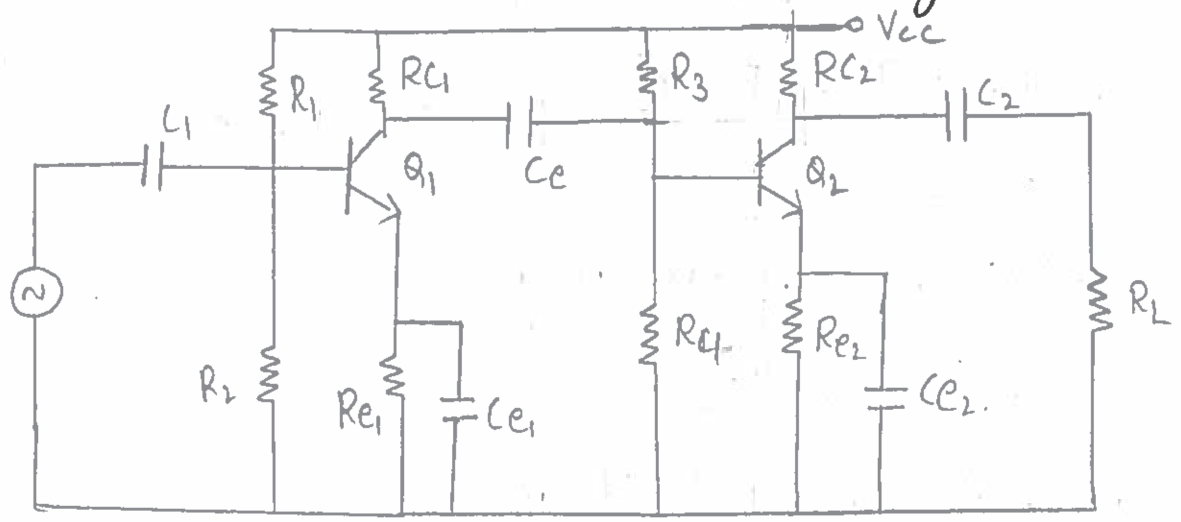


Fig:- RC coupled two stage Amplifier

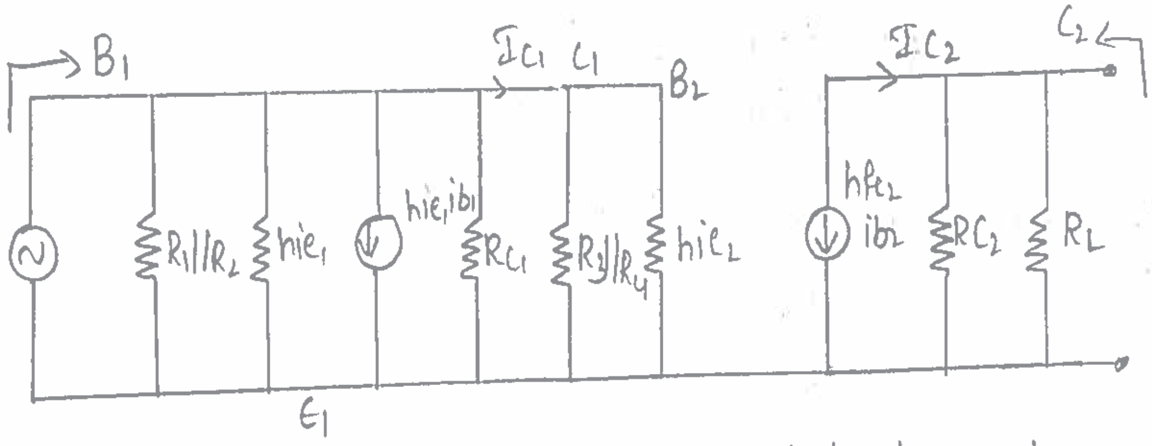


Fig:- Hybrid model of RC coupled two stage Amplifier

Input Resistance of 1st Stage

$$R_i = R_{B_1} \| \beta_1 r_{e_1}' \qquad (\because R_{B_1} >> \beta_1 r_{e_1}')$$

$$= \beta_1 r_{e_1}'$$

Output Resistance of 1st Stage:

$$R_{o_1} = R_{c_1} \| R_{i_2}$$

Input resistance of 2nd Stage:

$$R_{i_2} = R_{B_2} \| \beta_2 r_{e_2}' \quad (\because R_{B_2} \gg \beta_2 r_{e_2}')$$

$$= \beta_2 r_{e_2}'$$

Output resistance of 2nd Stage:

$$R_{o_2} = R_{o_2} \| R_L$$

voltage gain at 1st Stage:

$$V_1 = \beta \frac{R_{o_1}}{R_{i_1}}$$

$$= \cancel{\beta} \frac{R_{o_1}}{\cancel{\beta_1} r_{e_1}'} = \frac{R_{o_1}}{r_{e_1}'}$$

voltage gain at 2nd Stage:

$$V_2 = \beta \frac{R_{o_2}}{R_{i_2}}$$

$$= \cancel{\beta} \frac{R_{o_2}}{\cancel{\beta_2} r_{e_2}'} = \frac{R_{o_2}}{r_{e_2}'}$$

Over all voltage gain $A_V = V_1 \times V_2$

$$A_V = \frac{R_{o_1}}{r_{e_1}'} \times \frac{R_{o_2}}{r_{e_2}'}$$

Both the transistors are identicals then;

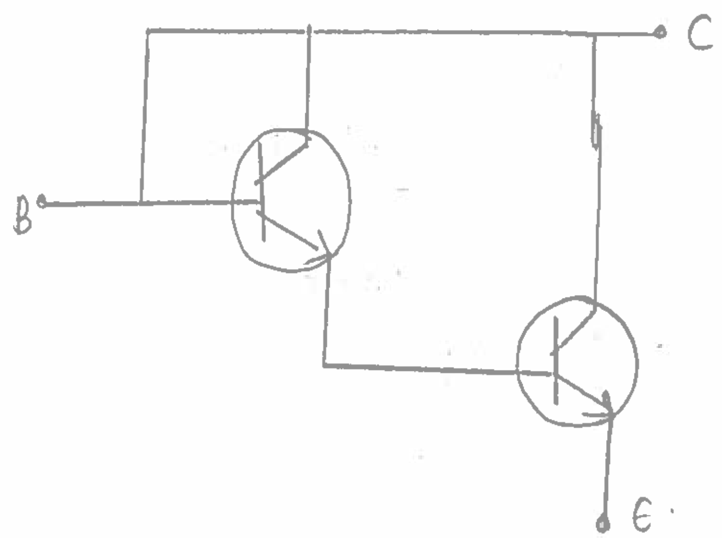
$$r_{e_1}' = r_{e_2}'$$

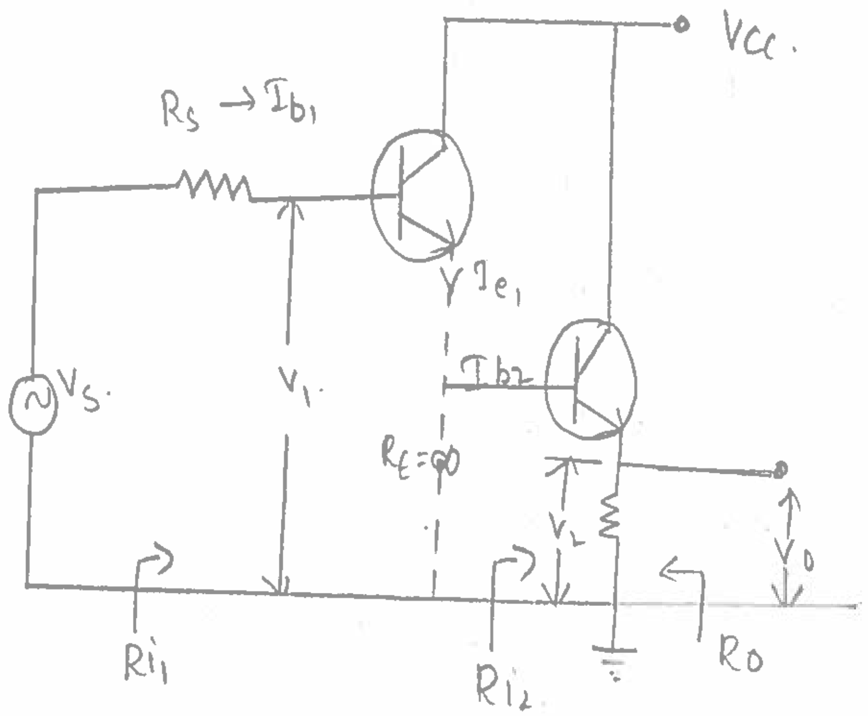
$$\therefore A_V = \frac{R_{o_1} \times R_{o_2}}{(r_{e_1}')^2}$$

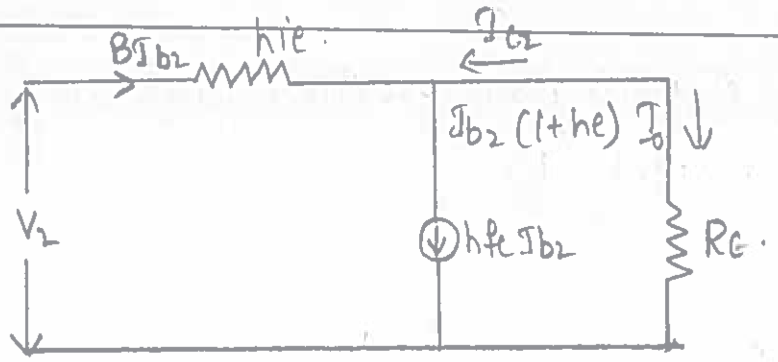
9) Draw and explain Darlington emitter follower con-
-figuration with respect to

(i) Current gain
(ii) input Impedance
(iii) voltage gain
(iv) Output impedence

and Compare with emitter follower.







Approximate Analysis.

$$I_{e_1} = I_{b_2}$$

1) current gain

$$A_{i_2} = \frac{I_0}{I_b} = \frac{-I_{e_2}}{I_b} = \frac{I_{b_2} + h_{fe} I_{b_2}}{I_{b_2}}$$

$$= \frac{\cancel{I_{b_2}}(1+h_{fe})}{\cancel{I_{b_2}}}$$

$$= 1 + h_{fe}$$

2) Input resistance 2nd stage

$$R_{i_2} = \frac{V_2}{I_{b_2}}$$

apply kvl to outer loop

$$V_2 - h_{ie} I_{b_2} - I_0 R_E = 0$$

$$V_2 = h_{ie} I_{b_2} + I_0 R_E$$

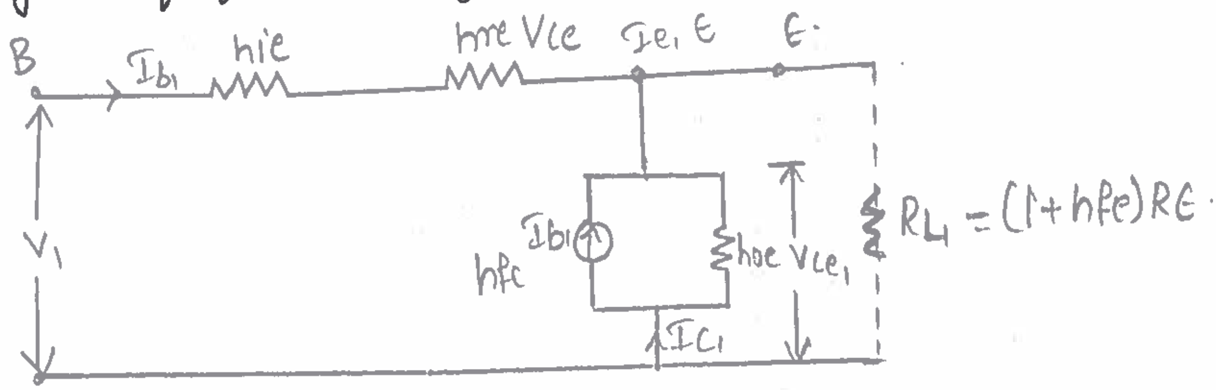
$$\frac{V_2}{I_{b_2}} = h_{ie} + \frac{I_0}{I_{b_2}} R_E$$

$$R_{i_2} = h_{ie} + A_{i_2} R_E$$

$$= h_{ie} + (1+h_{fe}) R_E$$

$$\therefore R_{i_2} = (1+h_{fe}) R_E \quad (\because h_{ie} << (1+h_{fe}) R_E)$$

Analysis of first stage:



1) Current gain $A_{i_1} = \dfrac{I_{b_2}}{I_{b_1}}$

$$= \frac{I_{e_1}}{I_{b_1}}$$

$$I_{e_1} = -(I_{b_1} + I_{C_1})$$

$$I_{C_1} = h_{fe} I_{b_1} + h_{oe}(-I_{b_2} R_{L_1})$$

$$= h_{fe} I_{b_1} + h_{oe}(-I_{b_2} R_{L_1})$$

$$= h_{fe} I_{b_1} + h_{oe} I_{e_1} R_{L_1}$$

Sub the value of I_{C_1} to the eq ①

$$I_{e_1} = -(I_{b_1} + h_{fe} I_{b_1} + h_{oe} I_{e_1} R_{L_1})$$

$$= -I_{b_1} - h_{fe} I_{b_1} - h_{oe} I_{e_1} R_{L_1}$$

$$I_{e_1} + h_{oe} I_{e_1} R_L = -I_{b_1}(1+h_{fe})$$

$$I_{e_1}(1 + h_{oe} R_{L_1}) = -I_{b_1}(1+h_{fe})$$

$$\therefore \frac{I_{e_1}}{I_{b_1}} = \frac{1+h_{fe}}{1+h_{oe} R_{L_1}}$$

$$A_{i_1} = \frac{1+h_{fe}}{1+h_{oe}(1+h_{fe})R_E}$$

$$= \frac{1+h_{fe}}{1+h_{oe} h_{fe} R_E} \quad (\because h_{fe} << 1)$$

Input Resistance of first Stage:

$$v_i - I_{b_1} h_{ie} - h_{re} V_{ce_1} + V_{ce_1} = 0$$

$$v_i = I_{b_1} h_{ie} - V_{ce_1}$$

$h_{re} V_{c_1}$ is negligible since h_{re} is in the order of

$$h_{re} = 2.5 \times 10^{-4}$$

$$v_i = I_{b_1} h_{ie} - (-I_{b_2} R_{L_1})$$

$$= I_{b_1} h_{ie} + I_{b_2} R_L$$

$$v_i = I_{b_1} h_{ie} + I_{b_2} (1 + h_{fe}) R_E$$

$$\frac{v_i}{I_{b_1}} = h_{ie} + \frac{I_{b_2}}{I_{b_1}} (1 + h_{fe}) R_E$$

$$R_{i_1} = h_{ie} + A_{i_1} (1 + h_{fe}) R_E$$

$$R_{i_1} = h_{ie} + \frac{(1 + h_{fe})(1 + h_{fe}) R_E}{1 + h_{oe} h_{fe} R_E}$$

$$R_i = h_{ie} + \frac{(1 + h_{fe})^2 R_E}{1 + h_{oe} h_{fe} R_E}$$

$$R_i = \frac{(1 + h_{fe})^2 R_E}{1 + h_{oe} h_{fe} R_E} \qquad \left[h_{fe} << \frac{(1 + h_{fe})^2 R_E}{1 + h_{oe} h_{fe} R_E} \right]$$

$\Rightarrow$ **Overall Current gain**

$$A_i = A_{i_1} \times A_{i_2}$$

$$= \frac{1 + h_{fe}}{1 + h_{oe} h_{fe} R_E} \times 1 + h_{fe}$$

$$A_i = \frac{(1 + h_{fe})^2}{1 + h_{oe} h_{fe} R_E}$$

Overall voltage gain: (A_v)

$$A_v = \frac{A_i R_L}{R_i}$$

from cc configuration

$$R_i = h_{ie} + h_{re} A_i R_L$$

$$1 - A_v = 1 - \frac{A_i R_L}{R_i}$$

$$1 - A_v = \frac{R_i - A_i R_L}{R_i}$$

$$= \frac{h_{ic_1} + h_{re_1} A_i R_L - A_i R_L}{R_i}$$

$$= \frac{h_{ic_1} + (1) \cancel{A_i R_L} - \cancel{A_i R_L}}{R_i}$$

$$h_{rc} = (1 - h_{re}) \approx 1$$

$$A_{v_1} = 1 - \frac{h_{ie_1}}{R_i} \quad (\because h_{ic} = h_{ie})$$

$$A_v = A_{v_1} \times A_{v_2}$$

$$= \left(1 - \frac{h_{ie_1}}{R_{i_1}}\right)\left(1 - \frac{h_{ie_2}}{R_{i_2}}\right)$$

$$= 1 - \frac{h_{ie}}{R_{i_2}} - \frac{h_{ie_1}}{R_1} + \frac{h_{ie_1} h_{ie_2}}{R_{i_1} R_{i_2}}$$

$R_{i_1} >> R_{i_2}$ then neglect the terms 3 & 4

$$A_v = 1 - \frac{h_{ie}}{R_{i_2}}$$

Output Impedence (Ro_2):

$$R_o = \frac{1}{y_o}$$

from CC amplifier.

$$y_{o_1} = h_{oc} - \frac{h_{fe} h_{re}}{h_{ic} R \cdot S}$$

$$h_{oc} = h_{oe}$$

$$h_{fc} = (1+h_{fe})$$

$$h_{rc} = 1 - h_{re} \approx 1$$

$$h_{ic} = h_{ie}$$

$$y_{o_1} = h_{oe} - \frac{-((1+h_{fe}))}{h_{ie}+R_s}$$

$$= h_{oe} + \frac{1+h_{fe}}{h_{ie}+R_s}$$

$$y_{o_1} = \frac{1+h_{fe}}{h_{ie}+R_s} \quad \left(\because h_{oe} << \frac{1+h_{fe}}{h_{ie}+R_{S_1}}\right)$$

$$Ro_1 = \frac{1}{y_{o_1}}$$

$$= \frac{h_{ie}+R_s}{1+h_{fe}}$$

$$Ro_1 = R_{S_2}$$

$$Ro_2 = \frac{R_{S_2} + h_{ie_2}}{1+h_{fe}}$$

$$= \frac{\left(\frac{h_{ie_1}+R_s}{1+h_{fe}}\right) + h_{ie_2}}{1+h_{fe}}$$

$$= \frac{h_{ie_1} + R_s}{(1+h_{fe})^2} + \frac{h_{ie_2}}{(1+h_{fe})}$$

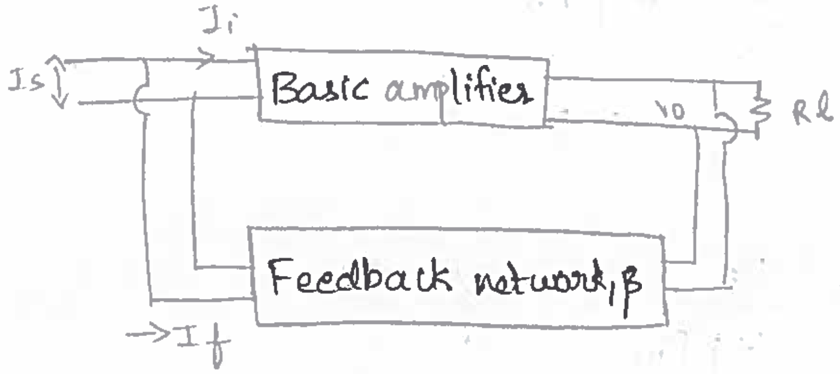
$$= \frac{\cancel{(1+h_{fe})} h_{ie_2}}{(1+h_{fe})^{\cancel{2}}} + \frac{R_s}{(1+h_{fe})^2} + \frac{h_{ie_2}}{1+h_{fe}}$$

$$= \frac{h_{ie_2}}{1+h_{fe}} + \frac{R_s}{(1+h_{fe})^2} + \frac{h_{ie_2}}{1+h_{fe}}$$

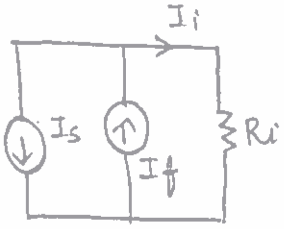
$$= 2\left(\frac{h_{ie_2}}{1+h_{fe}}\right) + \frac{R_s}{(1+h_{fe})^2} //$$

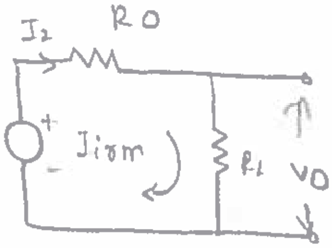
16.

Voltage shunt feedback Amplifier



Input impedance





$$I_s = I_i + I_f$$

$$I_i = I_s - I_f$$

$$\beta = \frac{I_f}{V_o} \qquad I_f = \beta V_o$$

$$V_o = \frac{r_m I_i R_o}{R_o + R_L}$$

Consider $R_m = \dfrac{r_m R_o}{R_o + R_L}$

then $V_o = R_m I_i$

$$I_s = I_i + I_f$$

$$= I_i + \beta V_o$$

$$= I_i + (\beta R_m) I_i$$

$$I_s = I_i [1 + \beta R_m]$$

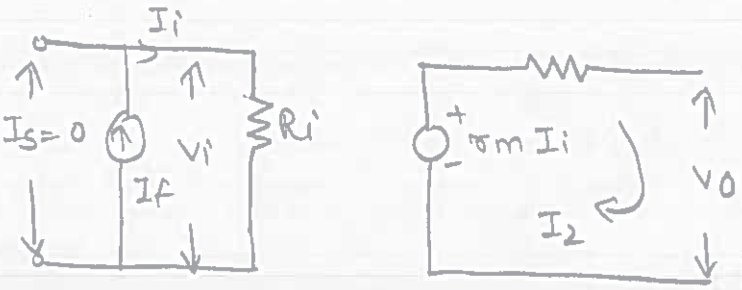
Now divide LHS and RHS by V_i

$$\therefore \frac{I_s}{V_i} = \frac{I_i}{V_i}(1+\beta R_m)$$

$$\frac{1}{R_{if}} = \frac{1}{R_i}(1+\beta R_m)$$

$$R_{if} = \frac{R_i}{1+\beta R_m}$$

Output impedance :-



$$I_i = I_s - I_f$$

$$= 0 - I_f$$

$$I_i = -I_f$$

$$\beta = \frac{I_f}{V_o} \Rightarrow I_f = \beta V_o$$

$$I_i = -\beta V_o$$

$$I_2 = \frac{V_o - r_m I_i}{R_o}$$

$$= \frac{V_o - r_m(-\beta V_o)}{R_o}$$

$$= \frac{V_o + r_m \beta V_o}{R_o}$$

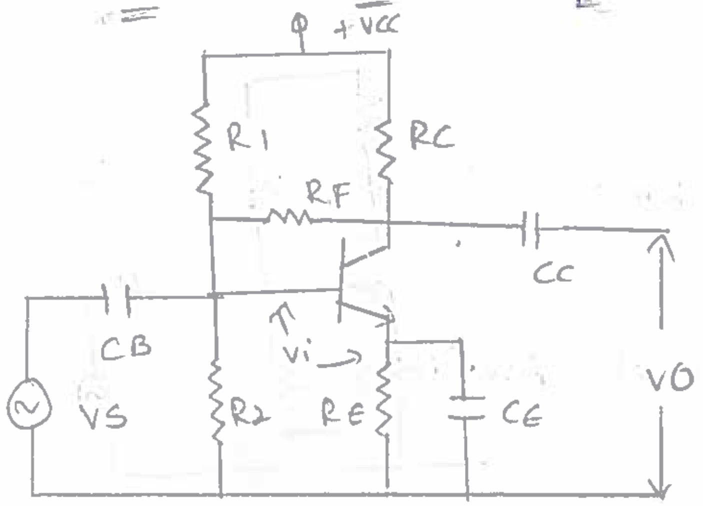
$$I_2 = \frac{V_o(1+\beta r_m)}{R_o}$$

$$\frac{I_2}{V_0} = \frac{(1+\beta r_m)}{R_0}$$

$$\frac{1}{R_{of}} = \frac{(1+\beta r_m)}{R_0}$$

$$R_{of} = \frac{R_0}{1+\beta r_m}$$

Voltage shunt feedback Amplifier



$$R_{if} = \frac{R_i}{1+\beta R_m}$$

$$R_{of} = \frac{R_o}{1+\beta R_m}$$

$$I_F = \frac{V_i - V_o}{R_F}$$

$$I_F R_F = V_i - V_o$$

$$I_f R_f = -V_o$$

$$\beta = \frac{I_f}{V_o} = \frac{\cancel{I_f}}{-\cancel{I_f} R_f}$$

$$\beta = \frac{-1}{R_f}$$

when $A\beta >> 1$

$$A_f = \frac{A}{1+A\beta}$$

$$= \frac{\cancel{A}}{\cancel{A}\beta} = \frac{1}{\beta}$$

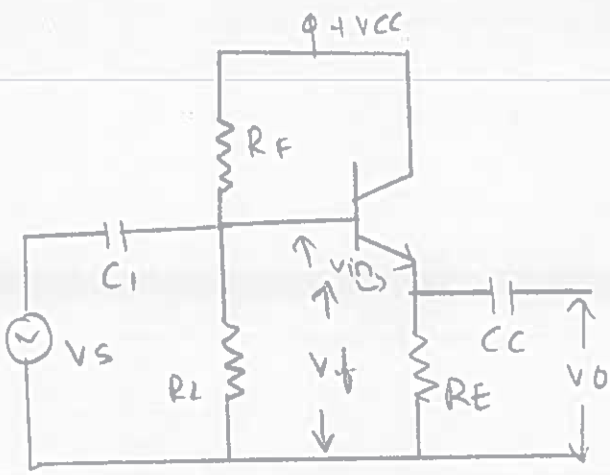
substitute β in above equation

$$\therefore A_f = \frac{1}{\frac{-1}{R_f}} = -R_f$$

$$\therefore A_f = -R_f$$

$$A_f = \frac{V_o}{I_s} = R_m = \frac{1}{\beta} = -R_F$$

11. Voltage Series Feedback Amplifier

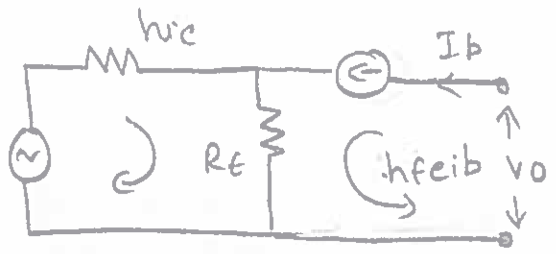


voltage series feedback amplifier is called emitter follower.

input impedance $R_{if} = R_i(1+A_v\beta)$

output impedance $R_{of} = \frac{R_o}{1+A_v\beta}$

H- parameter for the normal voltage series feedback amplifier is given by



$$A_f = \frac{V_o}{V_s} \rightarrow (1)$$

$$V_o = (I_b + h_{fe}\, i_b)R_E$$

$$V_o = I_b\,(1+h_{fe})R_E \rightarrow (2)$$

$$V_s = I_b h_{ie} + (I_b + h_{fe} I_b)R_E$$

$$V_s = I_b[h_{ie} + (1+h_{fe})R_E] \rightarrow (3)$$

substitute eq (2) & (3) in eq (1)

$$A_f = \frac{I_b\,(1+h_{fe})R_E}{I_b\,[h_{ie} + (1+h_{fe})R_E]}$$

$$A_f = \frac{(1+h_{fe})R_E}{h_{ie} + (1+h_{fe})R_E}$$

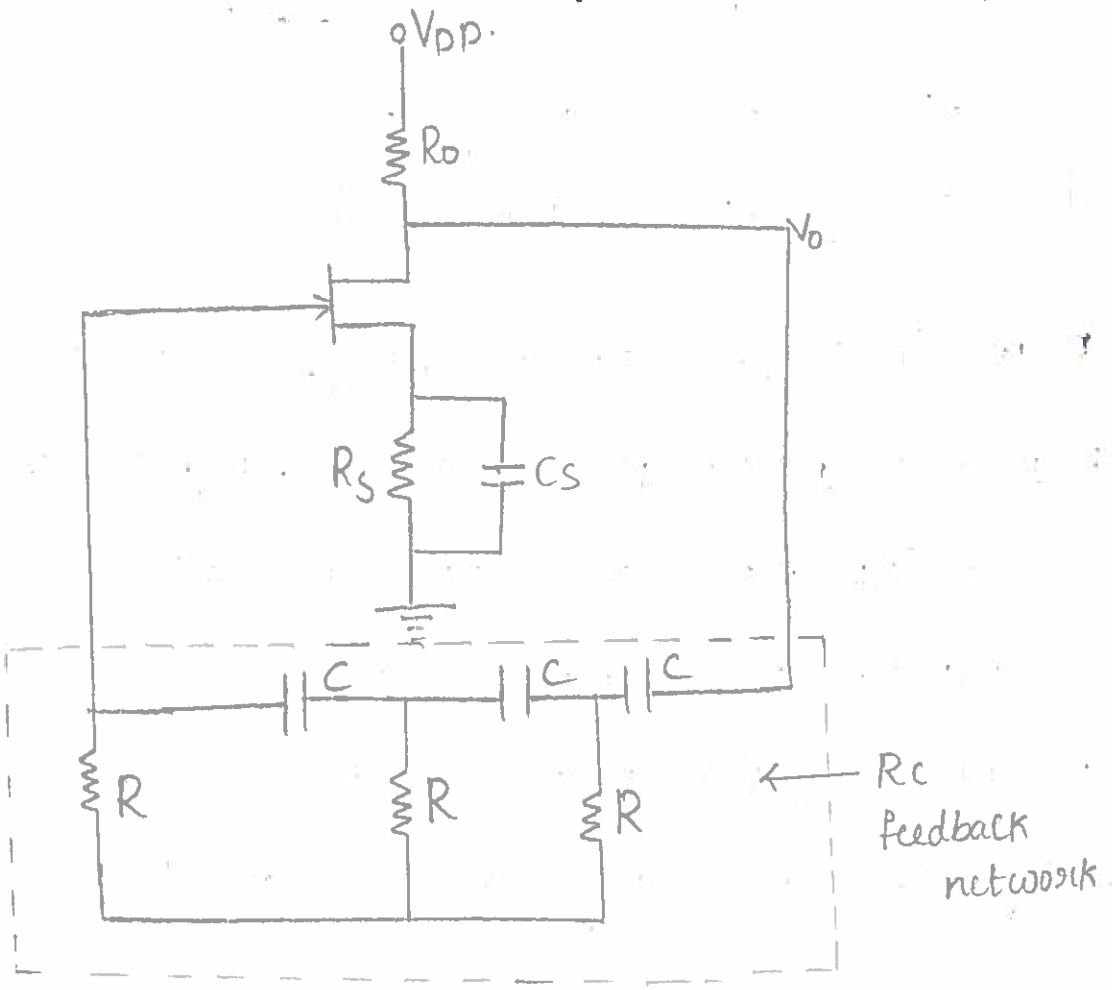
$$= \frac{\cancel{(1+h_{fe})R_E}}{\cancel{(1+h_{fe})R_E}}$$

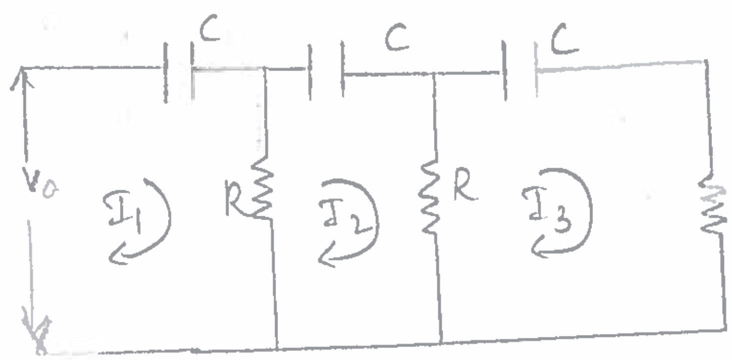
$[\because h_{ie} << 1+h_{fe}]$

$$\therefore A_f \cong 1$$

12) Derive the expression frequency of oscillation and condition for sustained oscillations of a FET based RC phase shift oscillator.

A:- <u>RC phase shift oscillator using FET</u>:-





Apply KVL for 3 loops

$$RI_1 + \frac{1}{j\omega C} I_1 - RI_2 + 0 = V_0$$

$$\left(R+\frac{1}{j\omega C}\right)I_1 - RI_2 = V_0 \rightarrow ①$$

$$-RI_1 + \left(2R+\frac{1}{j\omega C}\right)I_2 - RI_3 = 0 \rightarrow ②$$

$$0 - RI_2 + \left(2R+\frac{1}{j\omega C}\right)I_3 = 0 \rightarrow ③$$

$$\begin{bmatrix} I_1 \\ I_2 \\ I_3 \end{bmatrix} \begin{bmatrix} R-jX_C & -R & 0 \\ -R & 2R-jX_C & -R \\ 0 & -R & 2R-jX_C \end{bmatrix} = \begin{bmatrix} 0 \\ 0 \\ 0 \end{bmatrix}$$

$$\Delta = R-jX_c\left[((2R-jXc)(2R-jX_C)-R^2\right]+R\left[-R(2R-jXc)-0\right]+0=0$$

$$= R-jXc\left[4R^2-j2RXc-j2RXc-Xc^2-R^2\right]+R\left[-2R^2+jRXc\right]=0$$

$$= R-jXc\left[3R^2-j4RXc-Xc^2\right]-2R^3+jR^2Xc=0$$

$$= 3R^3-j4R^2Xc-RXc^2-j3R^2Xc-4RXc^2+jXc^3-2R^3+jR^2Xc=0$$

$$\Delta = R^3-5RXc^2-j6R^2Xc+jXc^3=0 \rightarrow ④$$

Equating the imaginary terms equals to zero.

$$-6R^2Xc+Xc^3=0$$

$$Xc^{\cancel{3}} = 6R^2\cancel{Xc}$$

$$Xc^2 = 6R^2$$

$$Xc = \sqrt{6}R$$

$$\frac{1}{\omega C} = \sqrt{6}R$$

$$\omega = \frac{1}{\sqrt{6}RC}$$

$$2\pi f = \frac{1}{\sqrt{6}RC}$$

$$\boxed{\therefore f = \frac{1}{2\pi\sqrt{6}RC}}$$

$$\beta = \frac{R^3}{R^3-5R(6R^2)}$$

$$\beta = \frac{R^3}{R^3-30R^3}$$

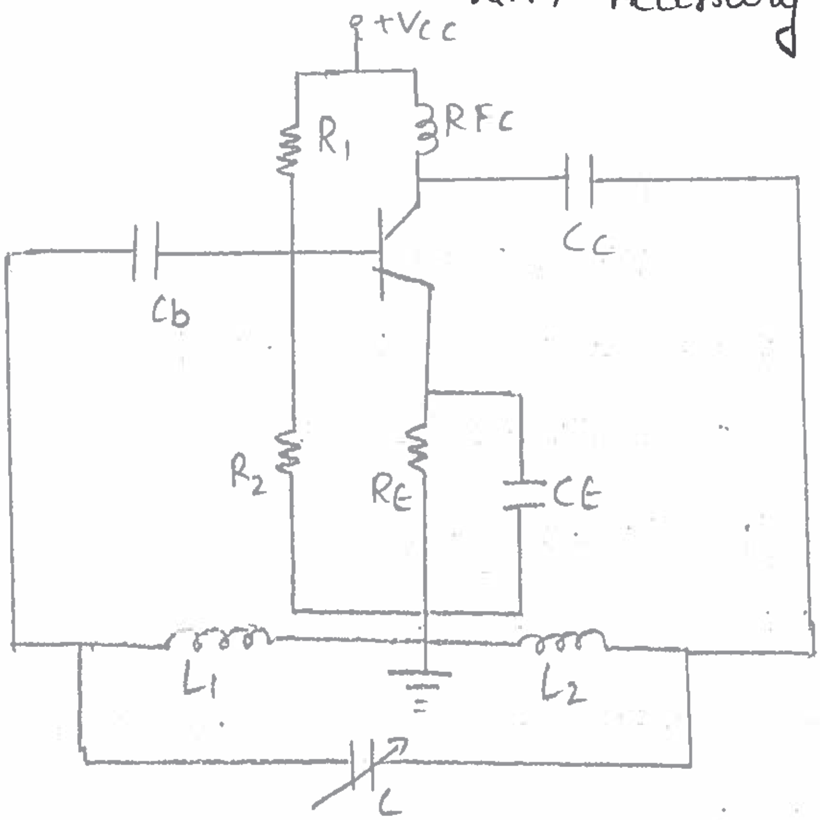
$$= \frac{\cancel{R^3}}{\cancel{R^3}(1-30)} = \frac{1}{-29}$$

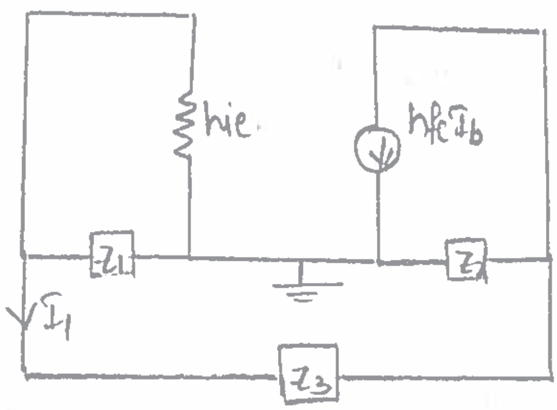
$$|A\beta| \geq 1$$

$$|A| = \frac{1}{\beta} = \frac{1}{1/29}$$

$$= 29$$

13) Describe the operation of Hartley oscillator circuit using bipolar junction transistor with necessary diagrams.





1) $\frac{1}{Z_1'} = \frac{1}{Z_1} + \frac{1}{hie}$

$= \frac{Z_1 + hie}{Z_1 hie}$

$Z_1' = \frac{Z_1 hie}{Z_1 + hie}$

$Z_1' + Z_3 = \frac{Z_1 hie + Z_1 Z_3 + Z_3 hie}{Z_1 + hie}$

The load impedance Z_L is the parallel combination of $z_1' + z_3 \parallel z_2$

$$\frac{1}{Z_L} = \frac{1}{Z_2} + \frac{1}{z_1' + z_3}$$

$$= \frac{1}{Z_2} + \frac{z_1 + h_{ie}}{z_1 h_{ie} + z_1 z_3 + z_3 h_{ie}}$$

$$\frac{1}{Z_L} = \frac{h_{ie}(z_1 + z_3) + z_1 z_3 + z_2 z_1 + z_2 h_{ie}}{z_2 [h_{ie}(z_1 + z_3)] + z_1 z_3}$$

$$Z_L = \frac{z_2 [h_{ie}(z_1 + z_3) + z_1 z_3]}{h_{ie}(z_1 + z_3) + z_2) + z_1 z_3 + z_1 z_2}$$

The Voltage gain Without feedback is given as

$$A = \frac{-h_{fe} Z_L}{h_{ie}}$$

The feedback fraction can be calculated as the output voltage between third & second terminals is given as

$$V_0 = (z_1' + z_3) I_1 = \left[\frac{z_1 h_{ie}}{z_1 + h_{ie}} + z_3 \right] I_1$$

$$= \left[\frac{z_1 h_{ie} + z_1 z_3 + z_3 h_{ie}}{z_1 + h_{ie}} \right] I_1$$

The Voltage feedback to the input terminals 1 and 2 is given by

$$V_{fb} = z_1' I_1$$

$$= \left[\frac{z_1 h_{ie}}{z_1 + h_{ie}} \right] I_1$$

$$\beta = \frac{V_{fb}}{V_b} = \frac{\left[\frac{Z_1 h_{ie}}{Z_1+h_{ie}}\right] I_1}{\left[h_{ie}(Z_1+Z_3)+Z_1Z_3 / Z_1+h_{ie}\right] I_1}$$

$$\beta = \frac{Z_1 h_{ie}}{h_{ie}(Z_1+Z_3)+Z_1Z_3}$$

Apply the criterion of oscillator.

$$A\beta = 1$$

$$-\frac{h_{fe}Z_3}{h_{ie}} \times \frac{h_{ie}Z_1}{h_{ie}(Z_1+Z_3)+Z_1Z_3} = 1$$

$$Z_1' + Z_3 = \frac{Z_1 h_{ie} + Z_3 Z_1 + Z_3 h_{ie}}{Z_1 + h_{ie}}$$

The load impedance Z_L is the parallel combination of $Z_1' + Z_3$ parallel to Z_1.

$$\frac{1}{Z_1} = \frac{1}{\cancel{Z}_3} + \frac{1}{Z_1'+Z_3} = \frac{1}{Z_2} + \frac{Z_1'+h_{ie}}{h_{ie}(Z_1+Z_3)+Z_1Z_3}$$

$$Z_1 = \frac{Z_1\left[h_{ie}(Z_1+Z_3)+Z_1Z_3\right]}{h_{ie}[Z_1+Z_2+Z_3]+Z_1Z_3+Z_1Z_2} \rightarrow (1)$$

The voltage gain without feedback is given as

$$A = \frac{h_{fe}Z_L}{h_{ie}}$$

The feedback function can be calculated the output voltage between 2^{nd} & 3^{rd} terminal is given as

$$h_{ie}\, j\omega\left[L_1+L_2+2m-\frac{1}{\omega^2 C}\right] - \omega^2\left[L_1+L_2+L_1M\right]+\left[L_1M+M^2\right]$$

$$(1+h_{fe}) + \frac{L_1}{C} + \frac{M}{C} = 0$$

$$h_{ie}j\omega\left[L_1+L_2+2M-\frac{1}{\omega^2 C}\right]-\omega^2\left[(L_1+M)(L_2+M)(1+h_{fe})+\frac{L_1+M}{C}\right]=0$$

$$h_{ie}j\omega\left[L_1+L_2+2M-\frac{1}{\omega^2 C}\right]-\omega^2(L_1+M)\left[(L_2+M)(1+h_{fe})-\frac{1}{\omega^2 C}\right]=0 \rightarrow ①$$

Now equating Imaginary part to zero

where $\left[L_1+L_2+2M-\frac{1}{\omega^2 C}\right]=0$

$$L_1+L_2+2M=\frac{1}{\omega^2 C}$$

$$\omega^2 C=\frac{1}{\sqrt{(L_1+L_2+2M)C}}$$

$$2\pi f=\frac{1}{\sqrt{(L_1+L_2+2M)C}}$$

$$f=\frac{1}{2\pi\sqrt{(L_1+L_2+2M)C}}$$

Equating the real part of eq ① to zero

$$(L_2+M)(1+h_{fe})-\frac{1}{\omega^2 C}=0$$

$$\frac{1}{\omega^2}=(L_2+M)(1+h_{fe})$$

$$h_{fe}=\frac{1}{\omega^2 C(L_2+M)}$$

$$=\frac{1}{L_2 M}\cdot\frac{1}{L_1+L_2+2M}$$

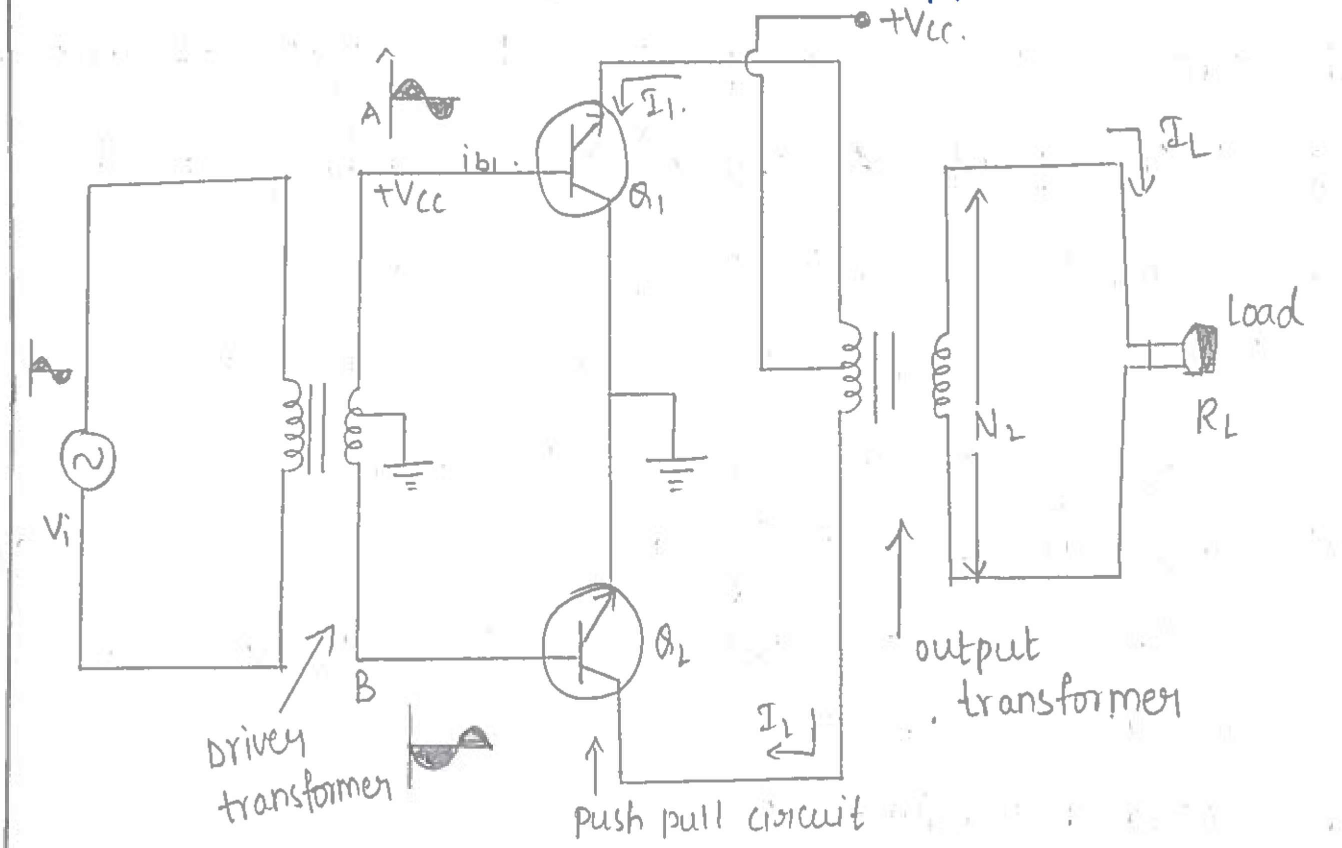
$$=\frac{L_1+M}{L_2+M}+\frac{L_2+M}{L_2+M}$$

$$\cancel{1}+h_{fe}=\frac{L_1+M}{L_2+M}+\cancel{1}$$

$$h_{fe}=\frac{L_1+M}{L_2+M}$$

14(a) Describe the operation of class B Push pull amplifier and also explain how the crossover distortion is minimized?

A:- The push pull circuit requires two transformers, one as input transformer called driver transformer and the other to connect the load called output transformer. The input signal is applied to the primary of the driver transformer. Both the transformers are centre tapped transformer.



In the circuit, both Q_1 & Q_2 transistors are of n-p-n type. The circuit can use both Q_1 & Q_2 of p-n-p type. In such a case, the only change is that the supply voltage must be $-V_{cc}$, then basic circuit remains the same. Generally the circuit using n-p-n transistor is used. Both the transistors are in common emitter configuration.

The driver transformer drives the circuit. The Input signal is applied to the primary of the driver transformer. The centre tap on the secondary of the driver transformer is grounded. The centre tap on the primary of the output transformer is connected to the supply voltage +Vcc. With respect to the centre tap, for a positive half cycle of input signal, the point A shown on secondary of the driver transformer will be positive. While the point B will be negative. Thus, the voltage in the two halves of the secondary of the driver transformer will be equal but while opposite polarity. Hence the input signals applied to the base of the transistors Q_1 and Q_2 will be 180° out of phase.

Cross over distortion:-

Cross over distortion is a type of distortion which is caused by switching between devices driving a load. It is most commonly seen in complementary, class-B amplifier stages, although it is seen in other circuits also.

A simple way to eliminate crossover distortion in a class B amplifier is to add two small voltages sources to the circuit to bias both the transistors at a point slightly above their cut-off point.

14b) Identify the effects of Harmonic distortions in power amplifiers.

Ans. • Harmonic distortion can have detrimental effects on electrical equipment.

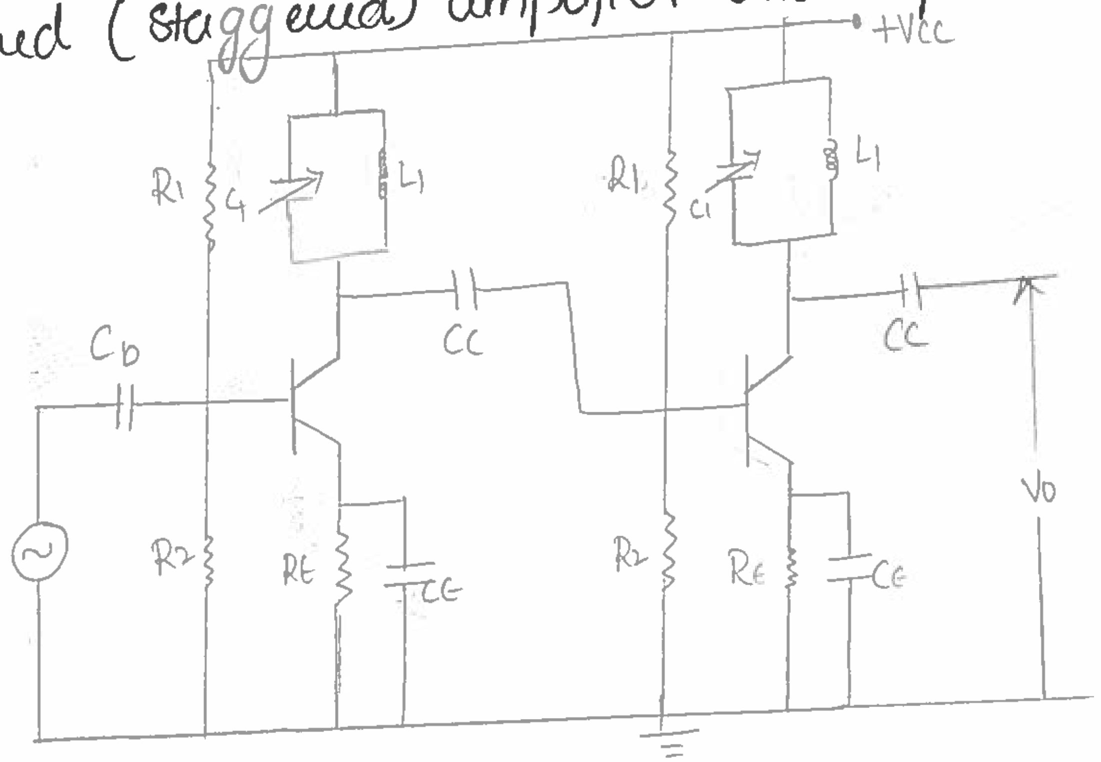
• Unwanted distortion can increase the current in power systems which results in higher temperature in neutral conductors and distortion transformer.

• Harmonics are caused by power electronic equipment VFDs, electronically commutated (EC) motors rectifiers, computers, LED lights, EV charger.

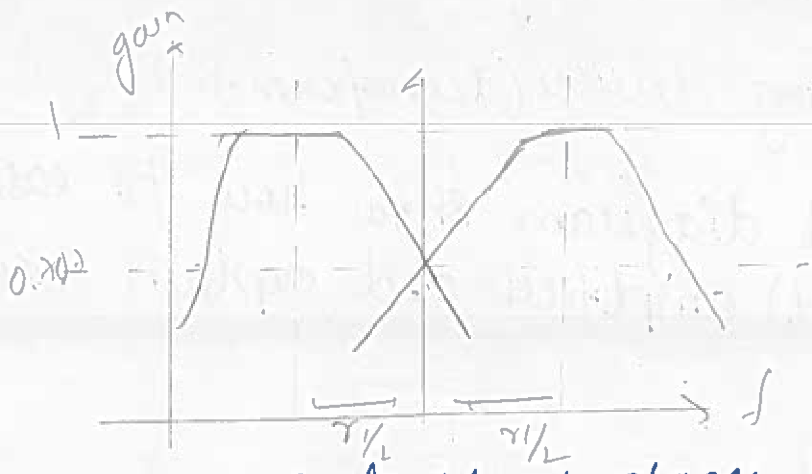
• Harmonic causes arcing devices (welders, arc furnances, fluorescent lights, etc.)

• Iron saturating devices (transformer)

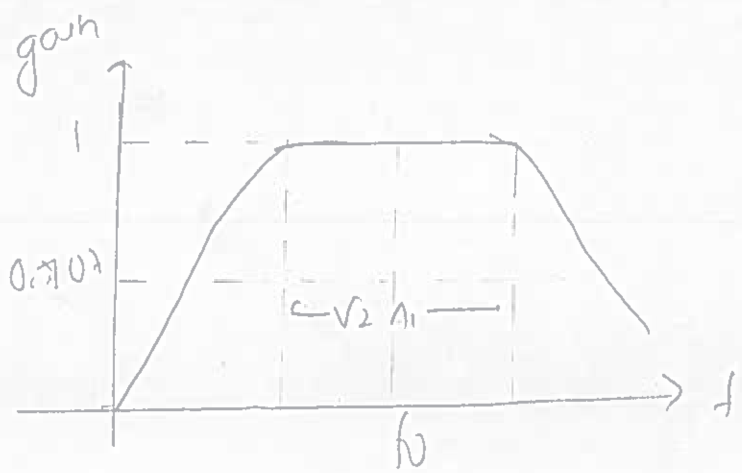
15a) With a neat diagram show how to cascade tuned (staggered) amplifier and explain briefly.



- The back to back connections of two single tuned amplifiers is the staggered tuned amplifier, i.e
- The two single tuned amplifier having certain band width are taken and their resonant frequencies are adjusted, that they are separated by an equal amount of bandwidth to each stage.
- Since, the resonant frequencies are, displayed (or) staggered, they are known as staggered tuned amplifiers



- This is the individual stage is shown above



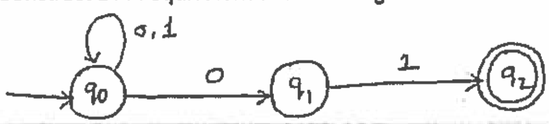
Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	Program	CSE		Academic Year	2021 - 2022
Course Code	20CS405	Test Duration	3 Hrs.	Max. Marks 70	Semester	IV
Course	Theory of Computation					

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Differentiate between DFA and NFA.	20CS405.1	L1
2	What is Pumping Lemma?	20CS405.2	L1
3	State Halting Problems.	20CS405.3	L1
4	What is the role of the Syntax Analyzer?	20CS405.4	L1
5	Define Local Optimization.	20CS405.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Construct a DFA for accepting the set of all strings with three consecutive 0's.	6M	20CS405.1	L2
6 (b)	Construct DFA equivalent to the NFA given below. 	6M	20CS405.1	L3
	OR			
7 (a)	Prove that a language L is accepted by some ε–NFA if and only if L is accepted by some DFA.	6M	20CS405.1	L2
7 (b)	Consider the following ε–NFA. Compute the ε–closure of each state and find its equivalent DFA.	6M	20CS405.1	L3

	ε	A	b	C
p	{q}	{p}	Φ	Φ
q	{r}	ϕ	{q}	Φ
*r	Φ	ϕ	ϕ	{r}

No.	Questions	Marks	Learning Outcome (s)	DoK
8 (a)	Give a detailed description of ambiguity in Context-free grammar.	6M	20CS405.2	L2
8 (b)	Let G be a grammar s->OB/1A, A->O/OS/1AA, B->1/1S/OBB. For the string 00110101 find its leftmost derivation and derivation tree.	6M	20CS405.2	L3
	OR			
9 (a)	Explain the design of Push Down Automata.	6M	20CS405.2	L2
9 (b)	Construct a equivalent grammar G in CNF for the grammar G1 where G1=({S,A,B},{a,b},{S->bA/aB,A->bAA/aS/a, B->aBB/bS/b},S)	6M	20CS405.2	L3
10 (a)	Explain the Basic Turing Machine model and explain in one move. What are the actions that take place in TM?	12M	20CS405.3	L2
	OR			
11 (a)	Explain turing machine with model and design turing machine for $a^nb^n/n\geq1$	12M	20CS405.3	L2
12 (a)	Explain the role of Lexical Analysis with an example.	6M	20CS405.4	L2
12 (b)	Explain the Context-free grammar writing a grammar.	6M	20CS405.4	L2
	OR			

13 (a)	Describe LR Parsing with an example.	6M	20CS405.4	L2
13 (b)	Explain LALR Parsers in detail	6M	20CS405.4	L2
14 (a)	Explain the generation variants of Syntax tree three address code.	6M	20CS405.5	L2
14(b)	Explain the Back patching with an example.	6M	20CS405.5	L2
	OR			
15 (a)	Describe the Loop Optimization in detail.	6M	20CS405.5	L2
15 (b)	Explain the DAG representation of Basic Blocks.	6M	20CS405.5	L2

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM, ANADAPURAM, VISAKHAPATNAM-531173

ANSWER KEY AND SCHEME OF EVALUATION

Degree	B. Tech. (U. G.)	**Program**	CSE			**Academic Year**	2021 - 2022
Course Code	20CS405	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	IV
Course	**Theory of Computation**						

1. Differentiate between DFA and NFA

An NFA is a Nondeterministic Finite Automaton. Nondeterministic means it can transition to, and be in, multiple states at once (i.e. for some given input). A DFA is a Deterministic Finite Automaton. Deterministic means that it can only be in, and transition to, one state at a time (i.e. for some given input)

2. What do you mean by pumping lemma?

In the theory of formal languages, the pumping lemma may refer to: Pumping lemma for regular languages, the fact that all sufficiently long strings in such a language has a substring that can be repeated arbitrarily many times, usually used to prove that certain languages are not regular.

3. State Halting Problems

The halting problem is a decision problem about properties of computer programs on a fixed Turing-complete model of computation, i.e., all programs that can be written in some given programming language that is general enough to be equivalent to a Turing machine.

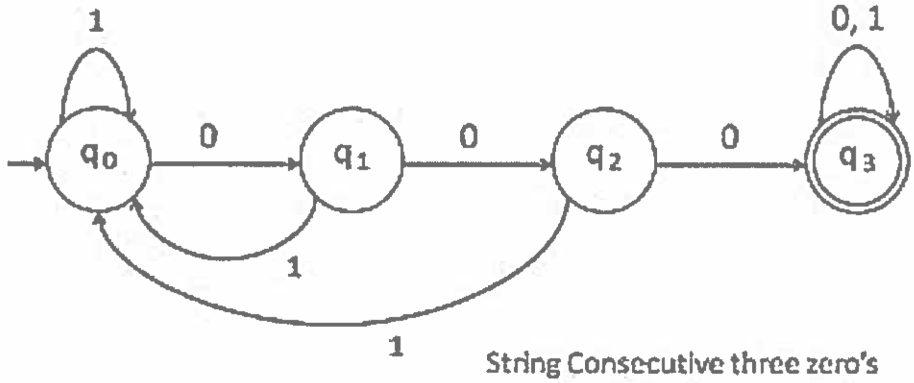
4. What is the role of the Syntax Analyzer?

A syntax analyzer or parser takes the input from a lexical analyzer in the form of token streams. The parser analyzes the source code (token stream) against the production rules to detect any errors in the code. The output of this phase is a parse tree.

5. Define Local Optimization.

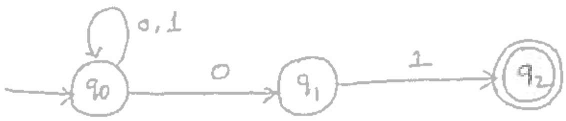
Optimization is a program transformation technique, which tries to improve the code by making it consume less resource (i.e. CPU, Memory) and deliver high speed. In optimization, high-level general programming constructs are replaced by very efficient low-level programming codes.

6A. Construct a DFA for accepting the set of all strings with three consecutive 0's.



String Consecutive three zero's

6B. Construct DFA equivalent to the NFA given below.



State / Alphabet	0	1
→q0	q0	q0, q1
q1	–	*q2
*q2	–	–

<u>Step-01:</u>

Let Q' be a new set of states of the Deterministic Finite Automata (DFA).

Let T' be a new transition table of the DFA.

<u>Step-02:</u>

Add transitions of start state q0 to the transition table T'.

State / Alphabet	0	1
→q0	q0	{q0, q1}

<u>Step-03:</u>

New state present in state Q' is {q0, q1}.

Add transitions for set of states {q0, q1} to the transition table T'.

State / Alphabet	0	1
→q0	q0	{q0, q1}
{q0, q1}	q0	{q0, q1, q2}

Step-04:

New state present in state Q' is {q0, q1, q2}.

Add transitions for set of states {q0, q1, q2} to the transition table T'.

State / Alphabet	0	1
→q0	q0	{q0, q1}
{q0, q1}	q0	{q0, q1, q2}
{q0, q1, q2}	q0	{q0, q1, q2}

Step-05:

Since no new states are left to be added in the transition table T', so we stop.

States containing q2 as its component are treated as final states of the DFA.

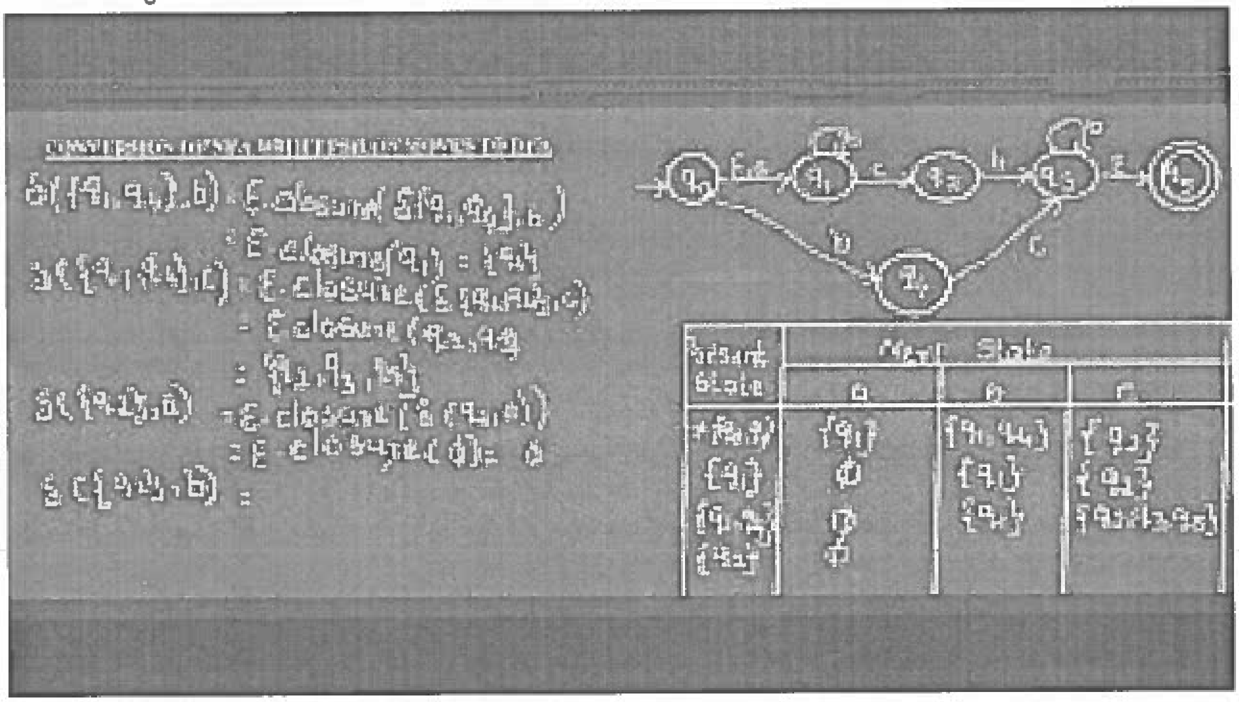
Finally, Transition table for Deterministic Finite Automata (DFA) is-

State / Alphabet	0	1
→q0	q0	{q0, q1}
{q0, q1}	q0	*{q0, q1, q2}

*{q0, q1, q2}	q0	*{q0, q1, q2}

7A. Prove that a language L is accepted by some ε–NFA if and only if L is accepted by some DFA.

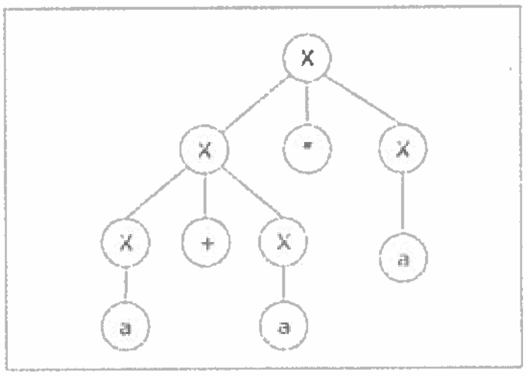
1. Step 1 – Consider M={Q, Σ, δ,q0,F) is NFA with ε. We have to convert this NFA with ε to equivalent DFA denoted by. ...
2. Step 2 – We will obtain δ transition on [p1,p2,p3,... pn] for each input. ...
3. Step 3 – The state obtained [p1,p2,p3,... pn] ∈ Q0 . ...
4. The DFA diagram is as follows



7B.CONVERSION ONLY THROUGH EPSILON CLOSURE

8A. Give a detailed description of ambiguity in Context-free grammar.

If a context free grammar G has more than one derivation tree for some string w □ L(G), it is called an ambiguous grammar. There exist multiple right-most or left-most derivations for some string generated from that grammar.



8B. Let G be a grammar s->OB/1A, A->O/OS/1AA, B->1/1S/OBB. For the string 00110101 find its leftmost derivation and derivation tree.

00110101

A) leftmost derivation

S → 0B
= 00BB
= 001B
= 0011S
= 00110B
= 00110B
= 0011010S
= 0011010B
= 00110101

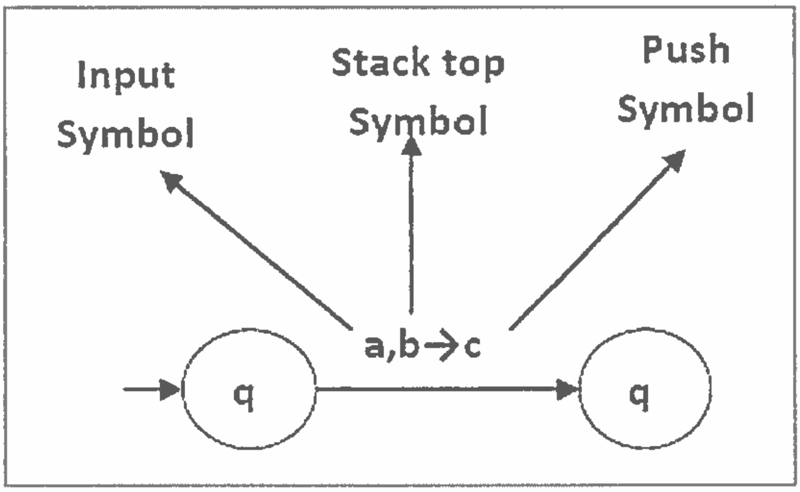
B) Rightmost derivation

S → 0B
= 00BB
= 00B1S
= 00B10B
= 00B101S
= 00B1010B
= 00B10101
= 00110101

9A

Pushdown Automata is a finite automata with extra memory called stack which helps Pushdown automata to recognize Context Free Languages. A Pushdown Automata (PDA) can be defined as : Q is the set of states. ∑is the set of input symbols. Γ is the set of pushdown symbols (which can be pushed and popped from stack)

A pushdown automaton is a way to implement a context-free grammar in a similar way we design DFA for a regular grammar. A DFA can remember a finite amount of information, but a PDA can remember an infinite amount of information. a stack with infinite size



9B Construct a equivalent grammar G in CNF for the grammar G1 where G1=({S,A,B},{a,b},{S->bA/aB,A->bAA/aS/a, B->aBB/bS/b},S)

- nto CNF. Since we use the symbols A and B in this grammar already, let us call the new nonterminals we need to incorporate to achieve the form of CNF, X (for a) and Y (for b).

- The grammar becomes:

$S \rightarrow YA$	$B \rightarrow XBB$
$S \rightarrow XB$	$B \rightarrow YS$
$A \rightarrow YAA$	$B \rightarrow b$
$A \rightarrow XS$	$X \rightarrow a$
$A \rightarrow a$	$Y \rightarrow b$

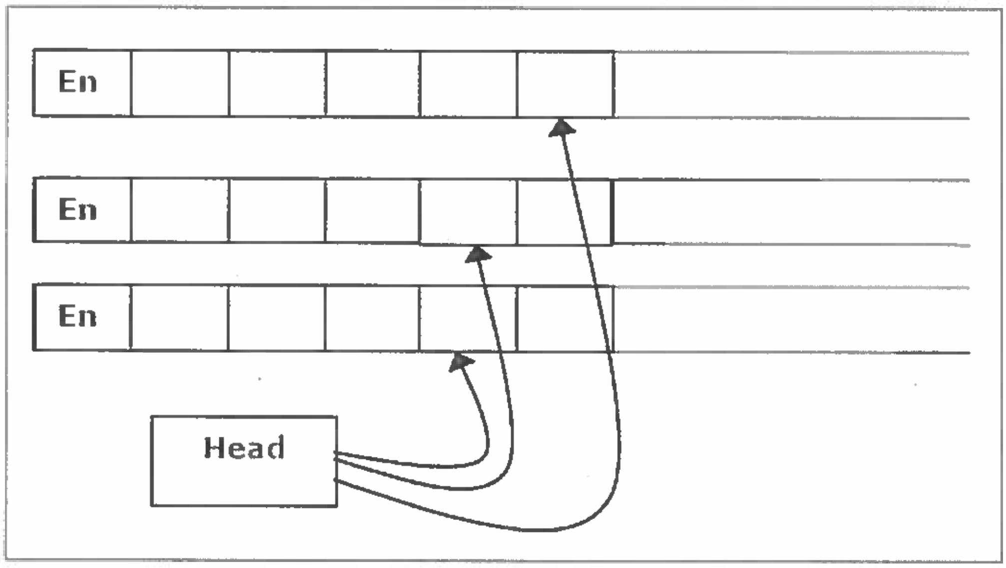
- Notice that we have left well enough alone in two instances:
- $A \rightarrow a$ and $B \rightarrow b$
- We need to simplify only two productions:
- $A \rightarrow YAA$ becomes $A \rightarrow YR_1$; $R_1 \rightarrow AA$ and
  $B \rightarrow XBB$ becomes $B \rightarrow XR_2$; $R_2 \rightarrow BB$
- The CFG has now become:
- $S \rightarrow YA \mid XB$
  $A \rightarrow YR \mid XS \mid a$
  $B \rightarrow XR_2 \mid YS \mid b$
  $X \rightarrow a$
  $Y \rightarrow b$
  $R_1 \rightarrow AA$
  $R_2 \rightarrow BB$

- which is in CNF. This is one of the more obscure grammars for the language EQUAL.

10A. Explain the Basic Turing Machine model and explain in one move. What are the actions that take place in TM?

δ is a transition function which maps Q × T → Q × T × {L,R}. Depending on its present state and present tape alphabet (pointed by head pointer), it will move to new state, change the tape symbol (may or may not) and move head pointer to either left or right. q0 is the initial state. F is the set of final states.

Each head can move independently of the other heads. Initially the input is on tape 1 and others are blank. At first, the first tape is occupied by the input and the other tapes are kept blank. Next, the machine reads consecutive symbols under its heads and the TM prints a symbol on each tape and moves its heads.



11A. Explain turing machine with model and design turing machine for $a^n b^n / n \geq 1$

A Turing machine can be formally described as seven tuples

(Q,X, Σ, δ,q0,B,F)

Where,

- Q is a finite set of states
- X is the tape alphabet
- Σ is the input alphabet
- δ is a transition function: δ:QxX→QxXx{left shift, right shift}
- q0 is the initial state
- B is the blank symbol
- F is the final state.

A Turing Machine (TM) is a mathematical model which consists of an infinite length tape divided into cells on which input is given. It consists of a head which reads the input tape. A state register stores the state of the Turing machine.

After reading an input symbol, it is replaced with another symbol, its internal state is changed, and it moves from one cell to the right or left. If the TM reaches the final state, the input string is accepted, otherwise rejected.

The Turing machine has a read/write head. So we can write on the tape.

Now, let us construct a Turing machine which accepts equal number of a's and b's,

The language it is generated is L ={ $a^n b^n$ | n>=1}, the strings that are accepted by the given language is –

L= {ab, aabb, aaabbb, aaaabbbb,.........}

Example

Consider n=3 so, a^3b^3 , the tape looks like –

a	a	a	b	b	b	B	B	....
q0								

B= blank

We need to convert every 'a' as X and every 'b' as Y. If the Turing machine contains an equal number of X and Y then it reaches the final state.

Step 1 – Consider the initial state as q0. This state replace 'a' as X and move to right, now state changes for q0 toq1, so the transition function is –

δ(q0, a) = (q1,X,R)

X	a	a	b	b	b	B	B	....
	q1							

Step 2 – Move right until you see the blank symbol.

δ(q1, a) = (q1,a,R)

δ(q1, b) = (q1,b,R)

After reaching the blank symbol B, move left and change the state to q2, because we need to change the last 'b' to Y.

δ(q1, B) = (q2,B,L) //1st iteration

δ(q1, Y) = (q2,Y,L) // remaining iterations

X	a	a	b	b	b	B	B	....
						q2		

Step 3 – When we see the symbol 'b', replace it as Y and change the state to q3 and move left.

δ(q2, B) = (q3,Y,L)

X	a	a	b	b	Y	B	B	....
				q3				

Step 4 – Move to left until reach the symbol X.

δ(q3, a) = (q3,a,L)

δ(q3, b) = (q3,b,L)

When we reach X move right and change the state as q0, and the next iteration is started.

After replacing every 'a' and 'b' as X and Y by changing the states to q0 to q4, we get the following –

δ(q0, Y) = (q4,Y,N)

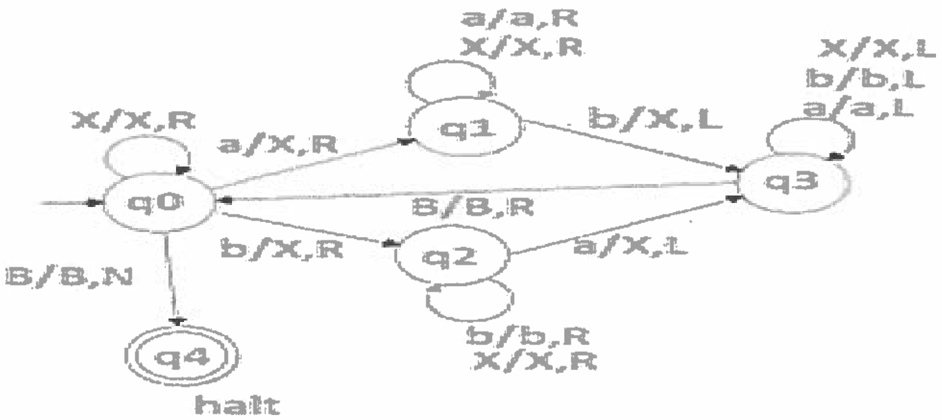
N represents No movement.

X	X	X	Y	Y	Y	B	B	....

q4 is the final state and q0 is the initial state of the Turing Machine, the intermediate states are q1, q2, q3.

Transition diagram

The transition diagram for Turing Machine is as follows –

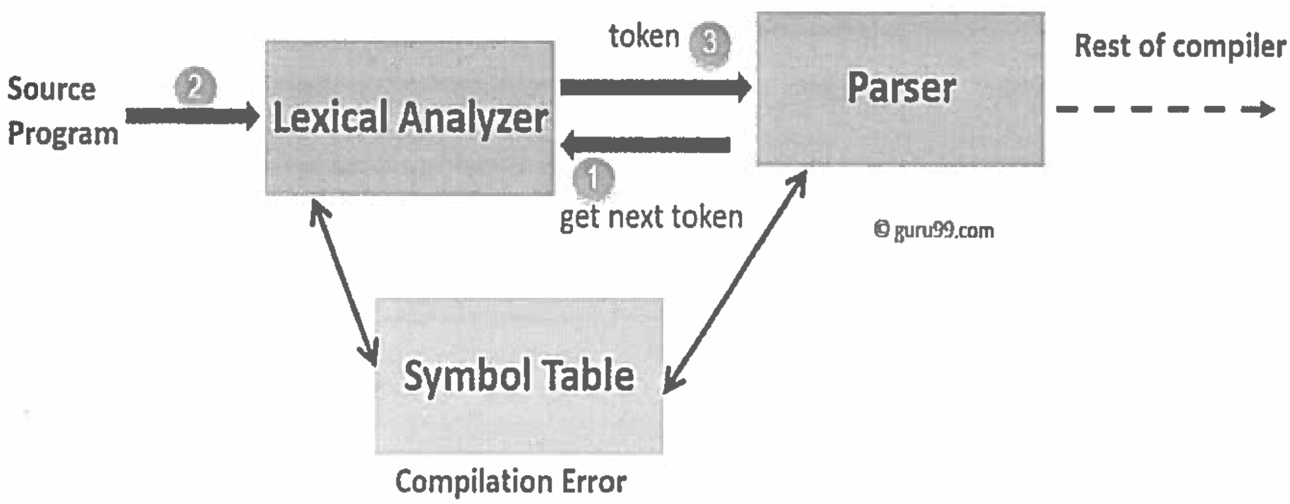


12A. Explain the role of Lexical Analysis with an example.

The lexical analysis is the first phase of the compiler where a lexical analyser operate as an interface between the source code and the rest of the phases of a compiler. It reads the input characters of the source program, groups them into lexemes, and produces a sequence of tokens for each lexeme.

The first step of compilation, called lexical analysis, is to convert the input from a simple sequence of characters into a list of tokens of different kinds, such as numerical and string constants, variable identifiers, and programming language keywords. The purpose of lex is to generate lexical analyzers.

Lexical Analysis is the very first phase in the compiler designing. A Lexer takes the modified source code which is written in the form of sentences . In other words, it helps you to convert a sequence of characters into a sequence of tokens. The lexical analyzer breaks this syntax into a series of tokens.



12B.

A formal grammar is "context free" if its production rules can be applied regardless of the context of a nonterminal. No matter which symbols surround it, the single nonterminal on the left hand side can always be replaced by the right hand side. This is what distinguishes it from a context-sensitive grammar.

CFG stands for context-free grammar. It is is a formal grammar which is used to generate all possible patterns of strings in a given formal language. Context-free grammar G can be defined by four tuples as: G = (V, T, P, S)

Context-Free Grammar Example

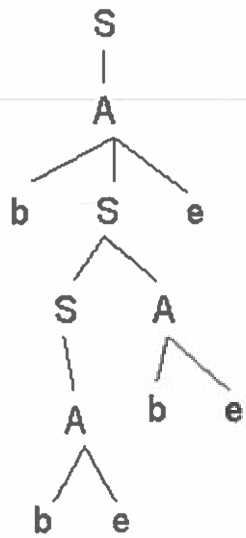
- S -> SA
  S -> A
  A -> bSe
  A -> be

Example: b and e matched as parentheses

derivation of bbebee

S
A
bSe
bSAe
bAAe
bbeAe
bbebee

hierarchical parse tree



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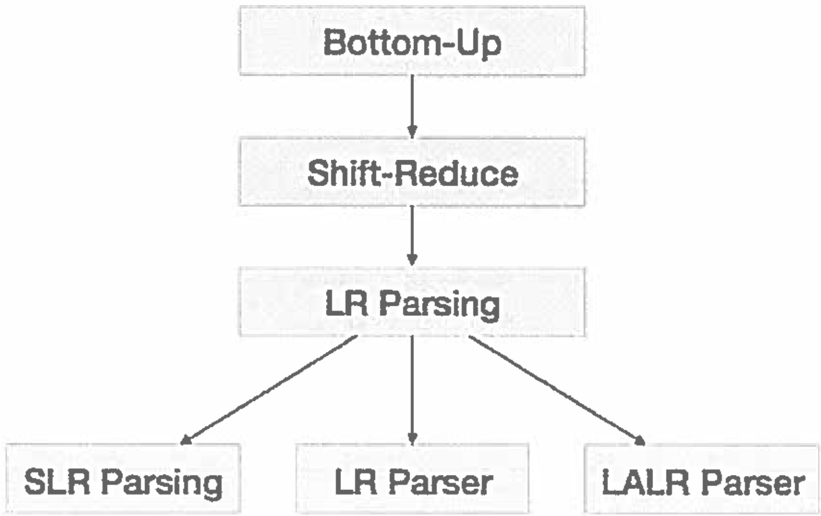
13A. Describe LR Parsing with an example.

LR parser : LR parser is a bottom-up parser for context-free grammar that is very generally used by computer programming language compiler and other associated tools. LR parser reads their input from left to right and produces a right-most derivation.

LR Parser. The LR parser is a non-recursive, shift-reduce, bottom-up parser. It uses a wide class of context-free grammar which makes it the most efficient syntax analysis technique.



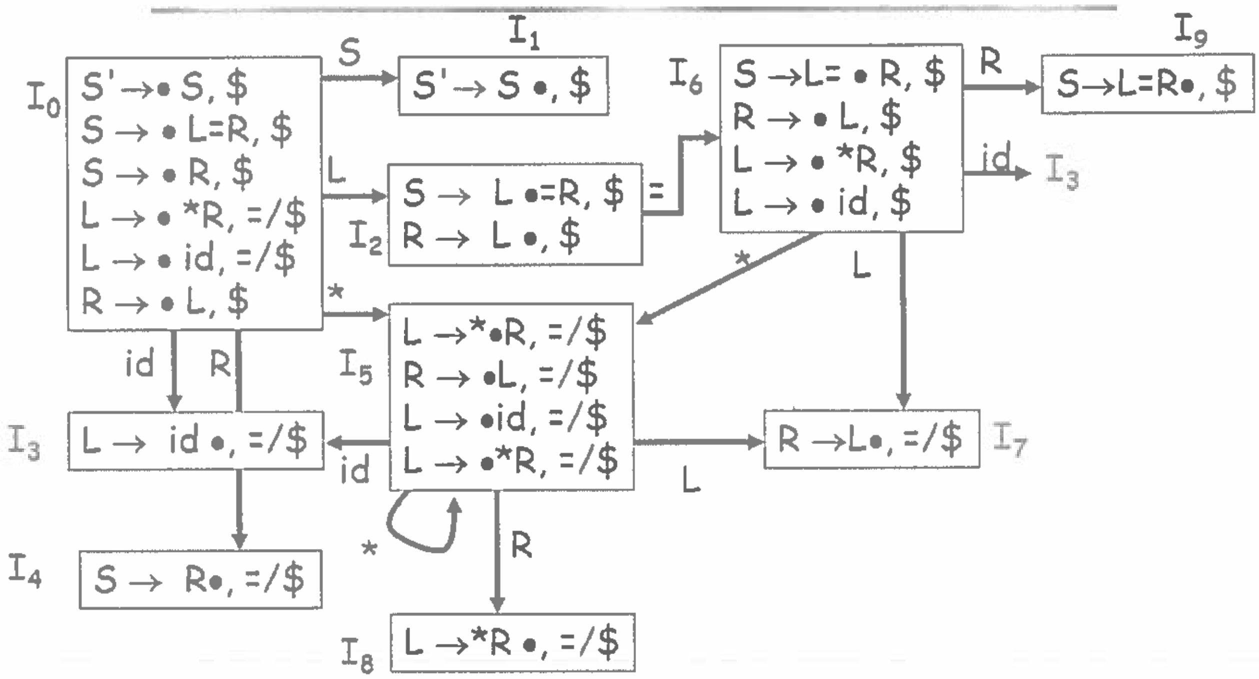
parsing (also known as syntax analysis) can be defined as a process of analyzing a text which contains a sequence of tokens, to determine its grammatical structure with respect to a given grammar.

13B.

In computer science, an LALR parser or Look-Ahead LR parser is a simplified version of a canonical LR parser, to parse a text according to a set of production rules specified by a formal grammar for a computer language.

LALR Parser is lookahead LR parser. It is the most powerful parser which can handle large classes of grammar. The size of CLR parsing table is quite large as compared to other parsing table. LALR reduces the size of this table. LALR works similar to CLR.

LALR(1) parsing



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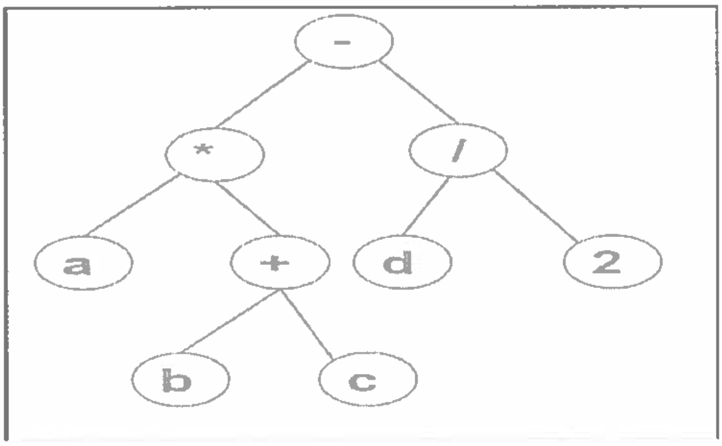
14A. Explain the generation variants of Syntax tree three address code.

A syntax tree basically has two variants which are described below:

- Directed Acyclic Graphs for Expressions (DAG)
- The Value-Number Method for Constructing DAGs.

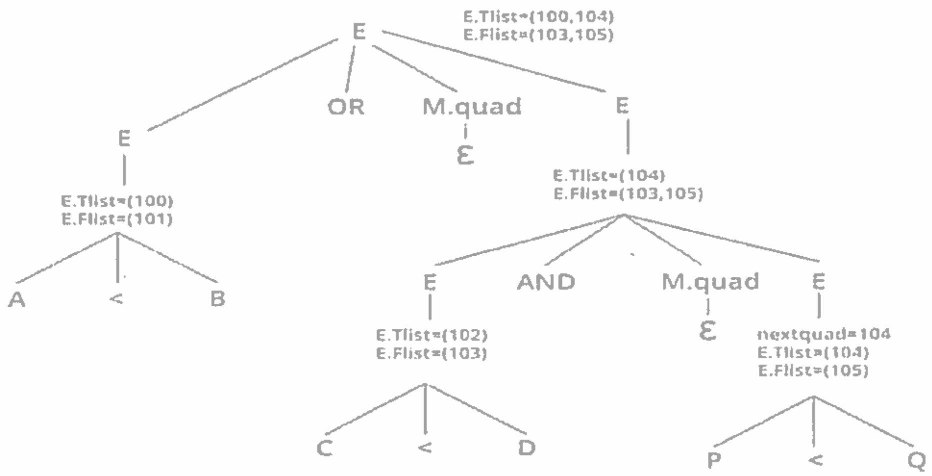


A syntax tree's nodes can all be performed as data with numerous fields. One element of the node for an operator identifies the operator, while the remaining field contains a pointer to the operand nodes. The operator is also known as the node's label. The nodes of the syntax tree for expressions with binary operators are created using the following functions. Each function returns a reference to the node that was most recently created.

1. mknode (op, left, right): It creates an operator node with the name op and two fields, containing left and right pointers.
2. mkleaf (id, entry): It creates an identifier node with the label id and the entry field, which is a reference to the identifier's symbol table entry.
3. mkleaf (num, val): It creates a number node with the name num and a field containing the number's value, val. Make a syntax tree for the expression a 4 + c, for example. p1, p2,..., p5 are pointers to the symbol table entries for identifiers 'a' and 'c', respectively, in this sequence.

14B. Explain the Back patching with an example

Backpatching is basically a process of fulfilling unspecified information. This information is of labels. It basically uses the appropriate semantic actions during the process of code generation. It may indicate the address of the Label in goto statements while producing TACs for the given expressions



15A. Describe the Loop Optimization in detail.

Loop Optimization is the process of increasing execution speed and reducing the overheads associated with loops. It plays an important role in improving cache performance and making effective use of parallel processing capabilities. Most execution time of a scientific program is spent on loops

1. oop Optimization Techniques
2. Loop Optimization in Compiler Design.
3. Intermediate Code Generation in Compiler Design.
4. Three address code in Compiler.
5. Compiler Design | Detection of a Loop in Three Address Code.
6. Code Optimization in Compiler Design.
7. Peephole Optimization in Compiler Design.

   For loop optimization the following three techniques are important:

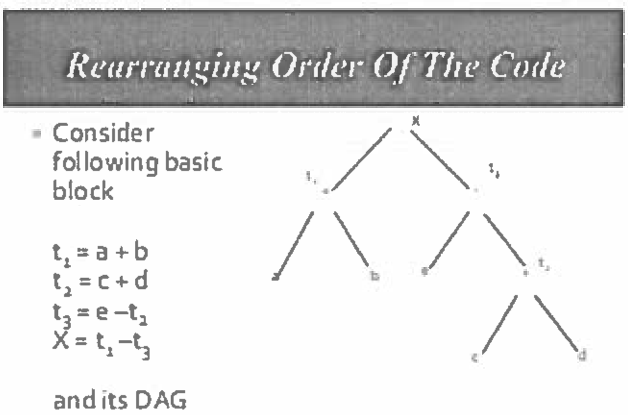
   8. Code motion.
   9. Induction-variable elimination.
   10. Strength reduction.

15B. Explain the DAG representation of Basic Blocks.

A DAG for basic block is a directed acyclic graph with the following labels on nodes: The leaves of graph are labeled by unique identifier and that identifier can be variable names or constants. Interior nodes of the graph is labeled by an operator symbol.

The Directed Acyclic Graph (DAG) is used to represent the structure of basic blocks, to visualize the flow of values between basic blocks, and to provide optimization techniques in the basic block.

Directed acyclic graph(DAG) is a useful data structure for implementing transformations on basic blocks. DAG is used in. · Determining the common sub-expressions. · Determining which names are used inside the block and computed outside the block



NSRIT

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	**Program**	Civil Engineering (Honors)	**Academic Year**	2021 - 2022
Course Code	20CEH02	**Test Duration**	3 Hrs. Max. Marks 70	**Semester**	IV
Course	**Energy Efficient Buildings**				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Classify the different types of Energy Sources.	20CEH02.1	L1
2	Define the term "Green House Effect".	20CEH02.2	L1
3	Differentiate between active and passive solar building.	20CEH02.3	L1
4	Interpret the importance of Green cover.	20CEH02.4	L1
5	List out the different classes of biopolymers.	20CEH02.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Write about non-conventional energy sources.	6M	20CEH02.1	L2
6 (b)	Describe the need for Energy Conservation system.	6M	20CEH02.1	L2
	OR			
7 (a)	Explain the classification of quality and concentration of energy from various energy sources.	6M	20CEH02.1	L2
7 (b)	Outline the various forms of energy and energy scenario in India.	6M	20CEH02.1	L2
8 (a)	Outline the concepts involved in Green Energy systems.	6M	20CEH02.2	L2
8 (b)	Comment on various rating systems for the assessment of sustainability.	6M	20CEH02.2	L2
	OR			
9 (a)	Explain the impacts of greenhouse gas emission process.	6M	20CEH02.2	L2
9 (b)	Enumerate the adoptive process and the agreements related to energy and sustainability.	6M	20CEH02.2	L2
10 (a)	Explain in detail the different ways of Utilization of Solar energy in buildings.	6M	20CEH02.3	L2
10 (b)	Elucidate the role of designing buildings related to the climatic conditions.	6M	20CEH02.3	L2
	OR			
11 (a)	Elucidate the ways of water Utilization in Buildings through natural sources.	6M	20CEH02.3	L2
11 (b)	Describe the differences of passive cooling and heating process in buildings.	6M	20CEH02.3	L2
12 (a)	Explain in brief about the techniques adopted for the management of Sullage Water and Sewage water for better sustainability.	6M	20CEH02.4	L2
12 (b)	Explain the guidelines and techniques related to green belt development.	6M	20CEH02.4	L2
	OR			
13 (a)	Write a note on water recycling and energy conservation.	6M	20CEH02.4	L2
13 (b)	Enumerate the importance of energy Approaches to Water Management.	6M	20CEH02.4	L2
14 (a)	Explain the techniques involved in preparation of Nano particles.	6M	20CEH02.5	L2
14 (b)	Enumerate the importance of Bio nanocomposites for sustainable future.	6M	20CEH02.5	L2
	OR			
15 (a)	Explain the sources and preparation of biopolymers.	6M	20CEH02.5	L2
15 (b)	Elucidate the concept of hybrid systems of thermal comfort. State	6M	20CEH02.5	L2

its outstanding features.

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(AUTONOMOUS)

SONTYAM, ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

PART-A

1. Classify the different types of Energy Sources.

   Ans)Sources of Energy
   - Solar Energy.
   - Wind Energy.
   - Biomass and Biofuels.
   - Water and geothermal

2. Define the term "Green House Effect".

   Ans)
   - The greenhouse effect is a warming of Earth's surface and the air above it.
   - It is caused by gases in the air that trap energy from the Sun. These heat-trapping gases are called greenhouse gases.
   - The most common greenhouse gases are water vapor, carbon dioxide, and methane.

3. Differentiate between active and passive solar building.

   Ans)
   - Active systems have devices to convert the sun's energy into a more usable form, such as hot water or electricity.
   - Passive systems are structures whose design, placement, or materials optimize the use of heat or light directly from the sun.

4. Interpret the importance of Green cover.

   Ans)
   - Greenery in our living environment benefits more than just our health and well-being.
   - It also facilitates water management and promotes biodiversity in built-up areas, and can help reduce the effects of noise pollution.
   - Greenery also helps to raise the property value of homes and offices

5. List out the different classes of biopolymers.

   Ans) There are three main classes of biopolymers, classified according to the monomers used and the structure of the biopolymer formed.

   They are:
   - Polynucleotides
   - Polypeptides
   - Polysaccharides

PART-B

6. a) Write about non-conventional energy sources.

   Ans) Non-conventional energy sources of energy are also known as a renewable source of energy
   - They are mainly used for household purposes

- These are not responsible for the cause of pollution
- Examples of non-conventional sources of energy include solar energy, bioenergy, tidal energy and wind energy.

Solar Energy

Solar Energy is produced by sunlight. The photovoltaic cells are exposed to sunlight based on the form of electricity that needs to be produced. The energy is utilized for cooking and distillation of water.

Wind Energy

Wind energy is generated by harnessing the power of wind and mostly used in operating water pumps for irrigation purposes. India stands as the second-largest country in the generation of wind power.

Tidal Energy

Tidal energy is generated by exploiting the tidal waves of the sea. This source is yet to be tapped due to the lack of cost-effective technology.

6. b) Describe the need for Energy Conservation system.

Ans) Energy conservation is the practice of reducing the consumption of energy by living organisms. Energy conservation is an idea and practice that focuses on saving our natural resources, especially those resources which are available in a limited amount. Non-renewable sources of energy are those that are consumed at a rate faster than that at which they are replenished.

The following are the importance of energy conservation:

- Energy conservation is necessary because it reduces the cost of consumption of energy. For example, when we reduce the use of electricity when not in use, then the cost per unit of energy also reduces. By using less electricity at home and using more energy-efficient appliances, we can reduce our electricity bills. That is how the conservation of electricity works.
- Energy conservation helps in reducing the use of natural resources of energy like fossil fuels. For example, more amount of coal and petroleum is used to heat water and generate electricity in thermal power plants. If we save electrical energy, we save our natural resources, which are consumed in producing electrical energy.
- Energy conservation reduces the waste which is released into the environment. It reduces unwanted carbon emissions into the atmosphere. For example, the burning of fossil fuels produces energy, and in this process, a lot of harmful gases are emitted into the air. It causes air pollution. Burning less amount of fuel reduces the unwanted contamination of the air.
- Energy conservation helps in improving the quality of life. It also helps in reducing global warming and other pollutants

7. a) Explain the classification of quality and concentration of energy from various energy sources.

Ans) classification of quality and concentration of energy from various energy sources
ENERGY QUALITY is of 3 types:
1) Energy quality in physical-chemical science (direct energy transformations):
i) Constant energy form, but variable energy flow
ii) Variable energy form, but constant energy flow

2) Energy quality in ecological physical chemistry (direct and indirect energy transformations):
i) Constant energy form and constant energy flow
ii) Variable energy form and variable energy flow
3) Energy quality in biophysical economics (indirect energy transformations)
ENERGY SOURCE classification :

1) Petroleum & Their Products
2) Bio Fuels
3) Natural Gas
4) Coal
5) Biomass
6) Hydrogen
7) Nuclear Energy
8) Solar Energy
9) Hydroelectric
10) Wind Power
11) Geothermal
12) Others

7. a) Outline the various forms of energy and energy scenario in India.

Ans) Energy is one of the major inputs for the economic development of any country. In the case of the developing countries, the energy sector assumes a critical importance in view of the everincreasing energy needs requiring huge investments to meet them.

Energy can be classified into several types based on the following criteria:

- Primary and Secondary energy
- Commercial and Non commercial energy
- Renewable and Non-Renewable energy

Indian Energy Scenario:

- Coal dominates the energy mix in India, contributing to 55% of the total primary energy production.
- Over the years, there has been a marked increase in the share of natural gas in primary energy production from 10% in 1994 to 13% in 1999.
- There has been a decline in the share of oil in primary energy production from 20% to 17% during the same period.

8. a) Outline the concepts involved in Green Energy systems.

Ans) Green energy is any energy type that is generated from natural resources, such as sunlight, wind or water.

Green energy is important for the environment as it replaces the negative effects of fossil fuels with more environmentally-friendly alternatives. Derived from natural resources, green energy is also often renewable and clean, meaning that they emit no or few greenhouse gases and are often readily available.

Types:

- Solar energy from the sun.
- Geothermal energy from heat inside the earth.
- Wind energy.
- Biomass from plants.
- Hydropower from flowing water

8. b) Comment on various rating systems for the assessment of sustainability.

   Ans) Various rating systems for the assessment of sustainability based on most popular, influential and technically advanced rating tools available are

   - BREEAM (Building Research Establishment's Environmental Assessment Method) is the leading and most widely used environmental assessment method for buildings. It was developed in the UK in 1990 and is the building environmental assessment method with the longest track record
   - LEED (Leadership in Energy and Environmental Design) Green Building Rating System, developed by the U.S. Green Building Council (USGBC) in 1998, provides a suite of standards for environmentally sustainable construction. Since its inception in 1998, LEED has grown to encompass more than 14,000 projects in the US and 30 countries covering 99 billion m^2 of development area
   - CASBEE (Comprehensive Assessment System for Building Environmental Efficiency) was developed in Japan in 2001. There are 4 basic versions of CASBEE which correspond to the individual stages of the building's lifecycle (Pre-design, New Construction, Existing buildings and Renovation)
   - GREEN STAR is a voluntary environmental rating system for buildings in Australia. It was launched in 2003 by the Green Building Council of Australia. The system considers a broad range of sustainable issues while also considering occupant health and productivity, and cost savings
   - HK-BEAM was developed 1996 in Hong Kong by the BEAM Society. It aims at promoting voluntary initiatives to measure, improve and label the environmental performance of buildings on environmental sustainability.

9. a) Explain the impacts of greenhouse gas emission process.

   Ans)The greenhouse effect is a process that occurs when gases in Earth's atmosphere trap the Sun's heat. This process makes Earth much warmer than it would be without an atmosphere. The greenhouse effect is one of the things that makes Earth a comfortable place to live.

   Impacts of greenhouse gas emission process are:

   - Causing more frequent and/or intense extreme weather events, including heat waves, hurricanes, droughts,and floods.
   - Exacerbating precipitation extremes, making wet regions wetter and dry regions drier.
   - Raising sea levels due to melting glaciers and sea ice and an increase in ocean temperatures (warmer water expands, which can contribute to sea level rise).
   - Altering ecosystems and natural habitat, shifting the geographic ranges, seasonal activities, migration patterns, and abundance of land, freshwater, and marine species

9. b) Enumerate the adoptive process and the agreements related to energy and sustainability.

10. a) Explain in detail the different ways of Utilization of Solar energy in buildings.

    Ans) **Solar power**

    - Solar energy is a renewable source of energy that is gaining ground because of the benefits it offers.
    - In India, sunlight is available in abundance and there is technology available to harness this energy and convert it into electric power.

- Solar power panels serve the purpose of absorbing solar energy and converting it to electric power through the photovoltaic (PV) effect.
- Most homes have a roof or a backyard which can be utilized to install solar panels and produce electricity
- A home solar system must provide enough electric energy to fulfil all the power requirements of a home.
- It should also be capable of providing AC power as traditionally all homes use AC power to operate lighting systems, gadgets, appliances and equipment such as computers, refrigerators, mixers, fans, air conditioners, TVs and music systems.

10. b) Elucidate the role of designing buildings related to the climatic conditions.

Ans)

- The purpose of climatic design is to facilitate an increase in the energy efficiency of buildings.
- Thermal design improves the living and working environment for occupants through ecologically sustainable means.
- It also seeks to reduce the effect on public health by adverse climatic conditions

Buildings designed for climate:

- Climatic design is practiced throughout the world and has been shown to produce buildings with low energy costs, reduced maintenance, and superior comfort.
- Some of the design features are outlined below:
- Utilising climatic factors may not require mechanical heating or cooling.
- Homes that are passively designed take advantage of natural energy flows to maintain thermal comfort. (Well-designed envelopes maximise cooling air movement and exclude sun in summer, trap and store heat from the sun in winter and minimise heat loss to the external environment.)
- Building envelope is a term used to describe the roof, walls, windows, floors and internal walls of a home.
- Maximise the thermal comfort and minimise the need for energy reliant heating and cooling appliances to achieve accepted levels of thermal comfort.

11. a) Elucidate the ways of water Utilization in Buildings through natural sources.

Ans) As access to fresh water continues to be a source of worry in many areas of the world (including India), water efficiency strategies in green building practices have become paramount to both new and existing construction efforts.

- Green building mentions a building structure that is designed to be environmental-friendly and makes nominal and efficient use of natural resources.
- Such buildings are resource-efficient and eco-friendly during its entire lifespan starting from its construction to demolition. A Green building design largely emphasises on making effectual use of natural resources like water, energy, etc. while reducing several bad effects on the environment and the occupant's health during its use. The 5 main gears of green buildings are:

1. Site And Design Efficiency
2. Reduced Energy Usage
3. Reduced Water Consumption
4. Environmentally Safe Construction Materials

5. Better Air Quality
   Considering water efficiency in Green Buildings, today several technologies are being used rainwater harvesting, recycling and reuse of grey water, low-flow fixtures, sensors etc.

   Water efficiency measures in residential and commercial buildings can greatly reduce water waste, yielding lower sewage volumes, reduced energy use, and bring in financial benefits too.

11. b) Describe the differences of passive cooling and heating process in buildings.
    Ans) Passive cooling
    - Passive cooling systems are designed to use natural means to transfer heat from buildings, including convection/ventilation, evaporation, radiation, and conduction.
    - However, the most important element in both passive and conventional cooling design is to prevent heat from entering the building in the first place.
    - Cooling conservation techniques involve building surface colors, insulation, special window glazings, overhangs and orientation,and numerous other architectural/engineering features.

    Passive Solar Heating
    Passive heating systems contain the five basic components of all solar systems
    Typical passive realizations of these components are
    - Collector: windows, walls, and floors
    - Storage: walls and floors, large interior masses (often, these are integrated with thecollector absorption function)
    - Distribution system: radiation, free convection, simple circulation fans
    - Controls: movable window insulation, vents both to other inside spaces or to ambient
    - Backup system: any non-solar heating system

12. a) Explain in brief about the techniques adopted for the management of Sullage Water and Sewage water for better sustainability.
    Ans) Four common ways to treat wastewater include physical water treatment, biological water treatment, chemical treatment, and sludge treatment.

    Physical Water Treatment
    In this stage, physical methods are used for cleaning the wastewater. Processes like screening, sedimentation and skimming are used to remove the solids. No chemicals are involved in this process.
    One of the main techniques of physical wastewater treatment includes sedimentation, which is a process of suspending the insoluble/heavy particles from the wastewater. Once the insoluble material settles down at the bottom, you can separate the pure water

    Biological Water Treatment
    This uses various biological processes to break down the organic matter present in wastewater, such as soap, human waste, oils and food. Microorganisms metabolize organic matter in the wastewater in biological treatment. It can be divided into three categories:
    - Aerobic processes: Bacteria decomposes the organic matter and converts it into carbon dioxide that can be used by plants. Oxygen is used in this process.
    - Anaerobic processes: Here, fermentation is used for fermenting the waste at a specific temperature. Oxygen is not used in anaerobic process.

- Composting: A type of aerobic process where wastewater is treated by mixing it with sawdust or other carbon sources.

Chemical Water Treatment

As the name suggests, this treatment involves the use of chemicals in water. Chlorine, an oxidizing chemical, is commonly used to kill bacteria which decomposes water by adding contaminants to it. Another oxidizing agent used for purifying the wastewater is ozone. Neutralization is a technique where an acid or base is added to bring the water to its natural pH of 7. Chemicals prevent the bacteria from reproducing in water, thus making the water pure.

Sludge Treatment

This is a solid-liquid separation process where the least possible residual moisture is required in the solid phase and the lowest possible solid particle residues are required in the separated liquid phase.

12. b) Explain the guidelines and techniques related to green belt development.

Ans)

Green belts are planned open spaces safeguarded from developmental activities such as construction of buildings, factories, dams, etc.

Safeguarded in the sense that no infrastructural development will be allowed on such designated areas and these areas will only be used for growing vegetation cover on it. Green belts in and around urban and industrial areas are important to the ecological health of any given region.

Following are the key points that all industries need to follow while moving ahead with the establishment of manufacturing/processing unit in certain areas. These are;

- No forest land shall be converted into non-forest activity for the sustenance of the industry (Reference: Forest Conservation Act, 1980).
- No prime agricultural land shall be converted into industrial site.
- Within the acquired site the industry must locate itself at the lowest location to remain obscured from general sight.
- Land acquired shall be sufficiently large to provide space for appropriate treatment of waste water still left for treatment after maximum possible reuse and recycle.The green belt shall be 1/2 km wide around the battery limit of the industry.
- The green belt between two adjoining large scale industries shall be one kilometer.
- Enough space should be provided for storage of solid wastes so that these could be available for possible reuse.
- Lay out and form of the industry that may come up in the area must conform to the landscape of the area without affecting the scenic features of that place.
- Associated township of the industry must be created at a space having physiographic barrier between the industry and the township.
- Each industry is required to maintain three ambient air quality measuring stations within 120 degree angle between stations.

13. a) Write a note on water recycling and energy conservation.

Ans) Water Recycling

- Water reuse is **the method of recycling treated wastewater for beneficial purposes**, such as agricultural and landscape irrigation, industrial processes, toilet flushing, and groundwater replenishing (EPA, 2004).

- One of the key advantages of recycling water is to protect water resources by reducing water pollution discharges and the need for water to be removed from natural habits.
- Wastewater recycling helps to preserve aquatic life and biodiversity by reducing polluting discharges into surface water.
- These discharges have major economic repercussions and negative impacts on the environment since they make any kind of normal activity impossible in the contaminated zone for a long time afterwards

Energy Conservation

- Energy conservation is the practice of reducing the consumption of energy by living organisms.
- Energy conservation is an idea and practice that focuses on saving our natural resources, especially those resources which are available in a limited amount.
- Non-renewable sources of energy are those that are consumed at a rate faster than that at which they are replenished.

13. b) Enumerate the importance of energy Approaches to Water Management.

Ans)

- Energy is of primary importance for water management and developments. The water infrastructures solely rely on energy throughout its value chain, groundwater extraction, transportation, purification, distillation, distribution, collection and wastewater management and treatment.
- Energy does not only play an important role in the functioning of water infrastructures, but also in the operational costs.

Improved energy and water services are a necessary input for achieving most MDGs. These are some examples:

- Improved water and energy services reduce the burden on women and young girls who often spend several hours each day collecting water and gathering biomass for cooking thus free up time for their participation in education and income generation activities. The provision of cleaner water and energy services is also linked to improvements in the health, micro-enterprise activity, and agricultural productivity of women.
- The lack of availability and access to basic water and energy services impedes individuals and communities from achieving greater levels of well-being and benefitting from opportunities for social and economic development. This is particularly true for the most poor and vulnerable segments of the population, such as women and children. Investing in water and energy services will lead to increased levels of human health, reduced levels of poverty and indigence, and increased opportunities for education and employment, resulting in overall national economic development.
- In many poor countries, biomass accounts for 90% of household energy consumption. Hence, ecosystem services not only sustain energy supply in low-income countries, but they are also critically affected by the predominant choice of energy carrier and aggregate consumption levels. Water security and ecosystems have a reciprocal relationship necessary for the enhancement of both and thereby conserving energy.

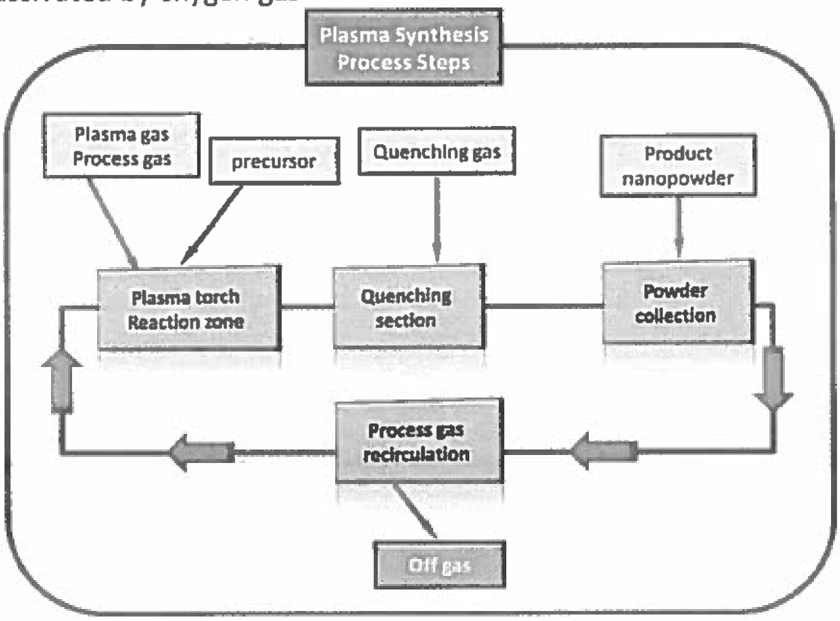
14. a) Explain the techniques involved in preparation of Nano particles.

Ans) Various methods of preparation of nanoparticles have been developed and they are suitable for synthesis of nanoparticles in different sizes and shapes. They are:

- Plasma method
- Chemical vapor deposition
- Microwave irradiation
- Pulsed laser method
- Sonochemical reduction
- Gamma radiation.

Plasma Method

- Plasma method is another method that is used to produce nanoparticles. The plasma is generated by radio frequency (RF) heating coils.
- The initial metal is enclosed in a pestle and the pestle is enclosed in an evacuated chamber.
- The metal is then heated above its evaporation point by high voltage RF coils wrapped around the evacuated chamber.
- The gas that is used in the procedure is Helium (He), which forms a high-temperature plasma in the region of the coils after flowing into the system.
- The metal vapor nucleates on the helium gas atoms and diffuses up to a cold collector rod, this is where nanoparticles are collected and they are passivated by oxygen gas



14. b) Enumerate the importance of Bio nanocomposites for sustainable future.

Ans)

- Nanocomposite is a multiphase solid material where one of the phases has one, two or three dimensions of less than 100 nanometers (nm) or structures having nano-scale repeat distances between the different phases that make up the material
- Polymer nanocomposites are the future for the global packaging industry.
- Once production and materials cost are less, companies will be using this technology to increase their product's stability and survivability through the supply chain to deliver higher quality to their customers while saving money.
- A range of polymeric nanocomposites are used for biomedical applications such as tissue engineering, drug delivery, cellular therapies.

- Due to unique interactions between polymer and nanoparticles, a range of property combinations can be engineered to mimic native tissue structure and properties

15. a) Explain the sources and preparation of biopolymers.

Ans)Sources of biopolymers:

Biopolymers are naturally occurring materials formed during the life cycles of green plants, animals, bacteria and fungi.

Biopolymers include animal protein- based biopolymers such as wool, silk, gelatin and collagen and polysaccharides such as cellulose, starch, carbohydrate polymers produced by bacteria and fungi.

- Biopolymers are produced from living matter which consists of monomeric units with linearly or branched like structured molecules.
- The monomeric unit refers to the molecules containing nucleic acids of nucleotides, saccharides or amino acids obtained from protein sources.
- Biopolymers are largely preferred over the conventional polymers for their renewability, eco-friendliness, bioavailability and biodegradability.
- They are also called natural biodegradable polymers as they have more economic value which are directly obtained from the environment

Methods of preparation

- Biopolymers can be produced either by fermentation or by polymerization of monomers.
- Biopolymers that are produced using microorganisms with specific carbon, nitrogen, minerals and salts as sources by the process of fermentation are termed as microbial biopolymers.
- The mechanism behind the production of these microbial biopolymers are mainly due to their defense mechanism or storage material.
- Another method of preparation is by chemical polymerization of monomeric units that can be degraded by microorganisms, enzymes or by natural resources.

15. b) Elucidate the concept of hybrid systems of thermal comfort. State its outstanding features.

Ans)

Hybrid ventilation combines mechanical ventilation with passive ventilation to optimize IAQ, thermal comfort and energy conservation. Ventilation's purpose is to provide acceptable indoor air quality and thermal comfort. Most residential buildings ventilate using both passive ventilation and mechanical ventilation

Hybrid Ventilation

- Hybrid ventilation uses both passive ventilation and mechanical ventilation in one system.
- Advanced hybrid ventilation, which includes intelligent controls, can achieve a balance between indoor air quality, thermal comfort, energy consumption, and electric peak load.
- Hybrid ventilation can reduce both electricity consumption and electricity peak demand. It provides more options for local control and minimizing fan energy when the outdoor climactic conditions are favorable

Adaptive Thermal Comfort (ATC)

- Occupants who enjoy more control of their indoor environment may tolerate a wider range of the indoor temperature.
- Adaptive thermal comfort is an idea based on an occupant's connection to the outdoors and control over their environment.

- This connection and control allow occupants to adapt to a wider range of thermal conditions than normal.
- The greatest opportunity for energy and power savings may come from hybrid ventilation's ability to provide adaptive comfort.
- Research indicates that building occupants appreciate some level of control of their thermal comfort and indoor air quality.
- The more transparent, simple and responsive the ventilation system, the better the occupant feels through adaptive comfort.
- Adaptive comfort requires educated occupants and means for them to control their environment.
- Adaptive thermal comfort can reduce the energy consumption of heating, cooling, and ventilation systems in residential buildings.
- ATC includes the following strategies.
    - Provide education about comfort and allow a sense of control
    - Open and closing windows and blinds
    - Don and shed clothing to affect comfort
    - Make only necessary adjustments to HVAC system

NSRIT

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	**Program**	EEE (Honors)			**Academic Year**	2021 - 2022
Course Code	20EEH01	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	IV
Course	**Smart Grid**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define the term smart grid and mention its components.	20EEH01.1	L1
2	Indicate the role of SCADA in smart grid.	20EEH01.2	L2
3	Identify the features of smart substation.	20EEH01.3	L2
4	List any two components of AMI used in smart grid.	20EEH01.4	L1
5	Define the term cloud computing and its need in smart grid operation.	20EEH01.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Describe the different opportunities and Barriers of Smart Grid in India.	12M	20EEH01.1	L2
	OR			
7 (a)	Compare the features of conventional & Smart grid technologies.	8M	20EEH01.1	L2
7 (b)	Explain the challenges and issues of smart grid implementation.	4M	20EEH01.1	L2
8	Describe the substation automation system involved in smart grid.	12M	20EEH01.2	L2
	OR			
9	Describe the power quality issues of grid connected renewable energy sources and solutions.	12M	20EEH01.2	L2
10	Explain the Outage Management System (OMS) used in the distribution networks.	12M	20EEH01.3	L2
	OR			
11	Describe the concept of distribution management system.	12M	20EEH01.3	L2
12	Compare the role of conventional metering and smart metering while involved in demand side management applications.	12M	20EEH01.4	L2
	OR			
13	Describe the concept of Advanced Metering infrastructure (AMI) in smart grid.	12M	20EEH01.4	L2
14	Explain the important role of Local Area Network and PLC in the smart grid systems.	12M	20EEH01.5	L2
	OR			
15 (a)	Illustrate the role of Broadband over power line (BPI) in the Sma grid operation.	8M	20EEH01.5	L2
15 (b)	Explain the important features of house area network (HAN) used in smart grid.	4M	20EEH01.5	L2

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(AUTONOMOUS)
SONTYAM, ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

Smart Grid

20EE H01

1. Definition – 1m
   Components – 1m
2. Definition – 1m
   Role of SCADA – 1m
3. Any two features – 2m
4. Two components – 2m
5. Definition – 1m
   Explanation – 1m
6. Opportunities – Any two – 4m
   Barriers – 5 numbers – 8m
7. (a) Any four comparisons – 2m each – 8m
   (b) Any four issues – 1m each – 4m.
8. Explanation – 8m.
   Figures – 4m.

9. Explanation of any 3 issues – 4m each – 12m.

10. Explanation — 6m

Figure — 3m + Figure Description — 3m.

12.. Any 6 comparisons — 2m each – 12m.

12. Explanation — 6m

Figure – 3m + Figure Description – 3m.

13. Explanation – 6m.

Figure – 3m + Figure Description – 3m.

14. Explanation of LAN – 6m.

" " PLC — 6m.

15(a) Explanation + Definition ~~Figure~~ ~~(Example)~~ — 4m each.

(b) Any four features — 4m.

1. Define the term smart grid and mention its components.
   Smart Grid: A Smart Grid is an electricity network that can intelligently integrate the actions of all users connected to it – generators, consumers and those that do both – in order to efficiently deliver sustainable, economic and secure electricity supplies.
2. Indicate the role of SCADA in smart grid.
   Supervisory control and data acquisition (SCADA) systems are extensively used for monitoring and controlling geographically distributed processes in a variety of industries. The basic requirement of any automation system is the availability of data from the field, and the SCADA system brings in the required data to the energy control center for further processing and necessary control activity.
3. Identify the features of smart substation.
   Fully equipped with advanced digital technologies
   Autonomous
   Coordination and communication capability
   Self-healing capability
4. List any two components of AMI used in smart grid
   Data Collection Unit
   Communication Network
5. Define the term cloud computing and its need in smart grid operation.
   Cloud computing is the on-demand availability of computer system resources, especially data storage (cloud storage) and computing power, without direct active management by the user. The deployment of CC in SG applications has multiple standpoints which can be grouped as organizational, technical, economic, and political standpoints.
   Part – B
6. Describe the different opportunities and Barriers of Smart Grid in India.
   Policy and regulation
   The current policy and regulatory frameworks were typically designed to deal with the existing networks and utilities. To some extent the existing model has encouraged competition in generation and supply of power but is unable to promote clean energy supplies. With the move towards smart grids, the prevailing policy and regulatory frameworks must evolve in order to encourage incentives for investment. The new frameworks will need to match the interests of the consumers with the utilities and suppliers to ensure that the societal goals are achieved at the lowest cost to the consumers
   Business Scenario:
   High capital and operating costs – Capital and operating costs include large fixed costs linked to the chronic communications network. Hardware costs do not cause in significant growths in economies of scale and software integration possess a significant delivery and integration risks.
   Benefits are constrained by the regulatory framework – When calculating the benefits, organizations tend to be conservative in what they can gather as cash benefits to the shareholders. For example, in many cases, line losses are considered to be put on to the customer and as a result any drop in losses would have no net impact on the utility shareholder. The smart grid benefits case may begin on a positive note but, as misaligned policy and regulatory incentives are factored in, the investment becomes less attractive. Therefore regulators are required to place such policies and regulations in place which could provide benefits both to the utilities and the

consumers. Therefore the first factor to be considered is to provide incentives to the utilities in order to remove inefficiencies from the system. They should be aptly remunerated for the line losses on their networks

Technology maturity and delivery risk

Technology is one of the essential constituents of Smart Grid which include a broad range of hardware, software, and communication technologies. In some cases, the technology is well developed; however, in many areas the technologies are still at a very initial stage of development and are yet to be developed to a significant level. As the technologies advances, it will reduce the delivery risk; but till then risk factor have to be included in the business situation.

On the software and data management side, the major challenge is to overcome the integration of the entire hardware system and to manage high volume of data. With multiple software providers come multiple data formats and the need for complex data models. In addition, the proliferation of data puts stresses on the data management architecture that are much similar to the telecommunications industry than the utilities industry. Many of these issues are currently being addressed in pilots such as Smart Grid task force and, as a consequence, the delivery risk will reduce as standards will be set up.

Lack of awareness

Consumer's level of understanding about how power is delivered to their homes is often low.

- Consumers should be made aware about their energy consumption pattern at home, offices...etc.
- Policy makers and regulators must be very clear about the future prospects of Smart Grids.
- Utilities need to focus on the overall capabilities of Smart Grids rather than mere implementation of smart meters. They need to consider a more holistic view.

Access to affordable capital

Funds are one of the major roadblocks in implementation of Smart Grid. Policy makers and regulators have to make more conducive rules and regulations in order to attract more and more private players. Furthermore the risk associated with Smart Grid is more; but in long run it is expected that risk-return profile will be closer to the current situation as new policy framework will be in place and risk will be optimally shared across the value chain.

Skills and knowledge

As the utilities will move towards Smart Grid, there will be a demand for a new skill sets to bridge the gap and to have to develop new skills in analytics, data management and decision support. To address this issue, a cadre of engineers and managers will need to be trained to manage the transition. This transition will require investment of both time and money from both government and private players to support education programs that will help in building managers and engineers for tomorrow. To bring such a change utility have to think hard about how they can manage the transition in order to avoid over burdening of staff with change.

Cyber security and data privacy

With the transition from analogous to digital electricity infrastructure comes the challenge of communication security and data management; as digital networks are more prone to malicious attacks from software hackers, security becomes the key issue to be addressed.

7. (a)Compare the features of conventional & Smart grid technologies

Feature/component	Conventional network	Smart Grid
Communications	None or one-way, typically not real-time	Two-way, real-time
Customer interaction	Limited	Extensive
Metering	Electromechanical	Digital (enabling real-time pricing and net metering)
Operation and maintenance	Manual equipment checks	Remote monitoring, predictive, time-based maintenance
Generation	Centralized	Centralized and distributed
Power flow control	Limited	Comprehensive, automated
Reliability	Prone to failures and cascading outages, essentially reactive	Automated, proactive protection, prevents outages before they start
Restoration following disturbance	Manual	Self-healing
Topology of distribution networks	Radial, generally one-way power flow	Network, multiple power flow pathways

(b) Explain the challenges and issues of smart grid implementation

Technical challenges

Inadequacies in grid infra structure

In developing countries like India, the grid infrastructure is still evolving. The existing grid network is inadequate to accommodate the upcoming needs of clean energy and distributed generation which may throw several challenges in design, erection, operation and maintenance. Besides focusing on SG, there is also a need to address issues of existing grid infra structure. In India, several electrical parts of country are unevenly connected to national grid in order to optimally evacuate large wind farms or solar parks which otherwise demand for installation of entire infrastructure. In this context, it is good to learn that Government of India is taking all possible and positive measures to overcome the Grid operation and connectivity problems through its working arms Central/State Transmission Utilities and National/Regional/State Load Dispatch Centre.

Cyber security

Connecting grid to cyber network triggers numerous vulnerabilities in the system and regrettably we are unaware about them. Recognizing and eliminating such loopholes before any security breach happens is very essential. Mainly three objectives of cyber security in SG have been addressed and discussed in Kappagantu et al. (2015a), i.e. availability, integrity and confidentiality. Availability refers to reliable and timely access to database and other information; Integrity includes protection from improper format/modification/destruction of information; Confidentiality refers to security of information from unauthorized access. Cyber security is one of the substantial issues for operation, since any single loophole has a potential threat to turn into disaster for utilities and individuals involved with grid. Well known cyber threats are hackers, zero day, malware, etc. Provision of any security feature alone is not sufficient enough to tackle such threat of logic bomb on grid. Infact smart grid has a multilayer structure and each layer demands for specific security concerns. There is no silver bullet for cyber threats but it mandates the development of advanced techniques for tackling the ever-evolving sophisticated cyber threats

Storage concerns

SG incorporates renewables for bulk power as well as distributed power generation. As the power generation from renewables is not uniform i.e., intermittent and variable, they may demand storage. Battery, the most common storage device, has very short life span of 4–5 years. Other storage technologies like flywheels, thermal storage, hydrogen storage, etc. have their respective varying concerns. Pumped storage technique, which is in regions of US, China, Japan, India and Norway, have efficiencies in the range of 70–85%. The problem with pumped storage techniques is that, it requires large areas as reservoirs which are normally available in mountain side only. For significant growth of SG, this option requires to move away from Pumped Storage in the mountain ranges. Research on its hybrid system with offshore wind is underway. In few regions of Germany storing compressed air in underground storage is in practice too, which can be used for electricity generation when needed. Although efficient, the complexities of storage facility become a hurdle for this technology. Flywheel is capable in absorbing energy in few seconds and delivering back quickly. Researchers found that Flywheels are very useful for supporting grid frequency for few seconds but they are not stable for longer duration. The most common technique for electricity storage is batteries and among them lead-acid batteries are the most popular. Portability is their advantage but low energy density, weight and size are the concerns for innovators to research. Further, risk of shortage of raw material for batteries is also a serious issue. Research on increasing efficiency and reducing cost of storage technologies is going on, but still battery storage technologies are expensive. Advanced Lead Acid Batteries, Flow Batteries and Lithium Ion Batteries are the options being tried in SG project in India for large scale storage purpose. At Puducherry, as per REAP provisions, the two commonly available configurations are rooftop solar PV system with and without battery back-up Figs. 1 and 2 which are very user friendly and needs no technical expertise.

Data management

SG infuse power network with enormous quantum of meters, sensors and controllers. The data from these units and from other sources like weather forecast, security cameras, etc. enhance the capability of operators. Through accurate analysis of data, a breakdown or damage could be avoided before occurrence. Further this big data could be utilized for system operation, alarms, forecasting demand, generation, price, etc. The data so collected is really big in volume, for example from employing smart meter that enables reading at each 15 min instead of once in a month increases the data almost 3000 times. Voluminous data from these devices is not only difficult for collection and storage but also poses critical challenges in retrieval and handling. Database management is a vital issue in SG. High volume of data may slow down the process of data collection, analysis and report generation. Apart from developing technology to manage the data, defining standards and protocols are of utmost importance and also necessity. Cloud based technologies may help in big data handling and analysis

Communication issues

We have a wide range of communication technologies for deployment in SG but they all have their own limitations. One technology has limited bandwidth while the second operates in limited distance, third has higher data loss and other has limited success in underground installations. Thus, despite numerous advantages, communication technology for SG still lacks a fool proof solution. Communication protocols are not well defined in SG network. Few technologies of this category are GSM, GPRS, PLCC, 3G, ZigBee, Broad band over PLC, etc. GSM and GPRS have coverage range of upto 10 km but they lack in data rates. 3G requires costlier spectrum, whereas ZigBee is limited by coverage range of 30–50 m only. Wired communication like power line communication overcomes the issues of wireless communication but face the problem of interferences. Optical fiber is fast and secure but is very expensive too. Router based RF technology with a canopy may solve the problems to some extent but it again lacks the history of proven in situ performance success stories besides the economy issues.

Stability concerns

SG is supposed to incorporate distributed generation (Renewables) and micro grids (MGs) on a large scale. The distributed generation causes bidirectional power flow. Renewables have various advantages over conventional and nuclear energy sources but high penetration of renewables and MGs would raise stability concerns like:

• Angular stability due to lower overall system inertia.
• Voltage stability due to lower power sharing support.
• Low-frequency power oscillation.
• Worsening of SG transients profile during MG islanding.
• Inability to serve as system reserve.

Energy management and electric vehicle

Using electric vehicle (EV) as storage prospective is on proposal. Research for efficient utilization of electric vehicle during periods of peak hour is going on. Batteries of EV can be charged in off peak period and can be used as source during peak periods (Software, 2012). Few basic controls in managing energy through EV include:

• Flow of power from
    - vehicle to grid (V2G),
    - grid to vehicle (G2V),
    - vehicle to vehicle (V2V).
• Reactive power control.
• DC link voltage control.
• Grid voltage support.

All these controls are not well defined yet and are still evolving. Development of standards for these is also on the anvil.

Socio-economic challenges

Socio-economic scenario plays a vital role in implementation and success of any technology. A technology becomes irrelevant if it fails to attract the investors or users, leading to failure of pilot projects, rejection of new technology, etc. Sometimes such issues may arise as a result of some economic or technological and some sometimes due to lack of appropriate awareness among stakeholders. Following are some discussions on few major issues in this regard

High capital investment

With high initial investment involved, SG is beneficial from economical perceptive that is realized on a long-term basis besides several technical advantages that it offers. Due to this initial capital investment in SG technology that appears to be high and the prevailing Indian conditions with inadequate financial health of Indian power utilities, SG deployment in India poses a major concern. Hence, awareness programs and incentives are essential to encourage utilities, organizations and individuals; to understand the SG benefits and cost burden. Merger of government, utilities and other sectors for sharing the initial burden could be a way out to build the business model

Stakeholder's engagement

New technology, high capital investment, lack of accurate information, etc. leads to negative perception of stake- holders that can derail even the SG project despite the highest potential benefits it offers. Advocate of smart grid should identify SG benefits to induce faith factor in stakeholders.

System operation aspects

Different operational aspects such as billing, tariff structure and operational strategies are some factors of utilities that may influence the SG deployment. They mainly depend on policies, participants, mindset of consumer, perception of operator, and state of supporting elements, etc. which vary with time. Defining any unified guidelines for operation of system is irrelevant and a flexible approach is needed from place to place for SG deployment.

Lack of awareness

Educating people about SG is much essential for its acceptance. To induce the faith for acceptance, society needs to be aware of SG. Myths create hurdles for any technology. Along with SG installation, utilities also need to focus on consumer awareness programs to teach about power delivery system and role of SG in building economy and efficiency. Consumers are also to be informed about economic and environmental benefits of the technology. Policy makers and regulators also need appropriate awareness; they must be clear about present and future scenario of the technology. Scrutinizing and feedback of awareness activities may also be considered

Privacy

Inadequacy in vigilance of huge data handling poses a risk of potential consumer privacy. Safety and security of consumers' information is of utmost concern. Breach of privacy of consumers' information may occur as consequence of any cyber threat or lack of proper policy as well. To maintain faith of consumers, their privacy must be kept intact through cyber security as well as tough regulations. Hence complete assurance to maintain consumer's trust is required for acceptance of SG like technology.

Fear of obsolescence

Very recently, user of smartphones, computers, etc. have witnessed the rapid growth of technology. Consumers are well aware of how fast these new technologies are becoming obsolete despite the additional benefits they bring forth. Further consumers are also aware that the higher costs associated with these new technologies eventually comes to their shoulders only. Such experience from IT and communication industry could become a road block for SG, if not addressed appropriately.

Fear of electricity charge increase

Because of awareness paucity, consumers apprehend about the rate increase of electricity charges due to SG deploy- ment. They believe that because of integration of new technology, associated with other factors the tariff would increase eventually. Consumers are also not well aware about new tariff approaches pursued by the utilities and government

New tariff

New tariff scheme like as real time use, time of use, critical time pricing, etc. have proven advantages in efficiency from operator's perceptive but each and every consumer may have his own opinion about it. At present, consumers who are comfortable with existing scheme are not in general accepting the new scheme willingly. The low tariff at off peak loads is attracting some consumers but consumers who are liberal in their usage of electricity are in opposition of tariff hike in peak periods

Radio frequency (RF) signal and health issues

Few consumers, medical groups and NGOs have registered their concerns about RF signals transmitted from the SG devices and their impact on the health. However, no accurate data is available in this regard, either in favor or against of such claims. A detailed research and awareness initiatives are required to deal with such issues.

Miscellaneous

Regulation and policies

Power theft

Work force

Co-ordination

8. Describe the substation automation system involved in smart grid.

Substation automation involves the deployment of substation and feeder operating functions and applications ranging from supervisory control and data acquisition (SCADA) and alarm processing, to integrated volt-var control in order to optimize the management of capital assets and enhance operation and maintenance (O&M) efficiencies with minimal human intervention.

Smart devices for substation automation are:

a) IEDs
b) Instrument transformers with digital interface
c) Intelligent breaker
d) Merging units (MUs)

IEDs facilitate the exchange of both operational and nonoperational data. Operational data, also called supervisory control and data acquisition (SCADA) data, are instantaneous values of power system analog and status points such as volts, amps, MW, MVAR, circuit breaker status, and switch position. These data are time critical and are used to monitor and control the power system (e.g., opening circuit breakers, changing tap settings, equipment failure indication, etc.). Nonoperational data consist of files and waveforms such as event summaries, oscillographic event reports, or sequential events records, in addition to SCADA-like points (e.g., status and analog points) that have logical state or a numerical value.

Instrument transformers with digital interface

New instrument transformers are available in the market. They can be directly linked to a merging unit, which in turn sends digital data over a network to the protection and metering devices and can eliminate the hardwiring to a large extent.

The new instrument transformers use capacitive, optical, and Rogowski techniques to capture the voltage and current from the field, thus making the systems robust, smaller, and reliable.

Nonconventional instrument transformers with digital interfaces based on IEC 61850-9-2 (process bus).

Intelligent breaker

Intelligent breaker has a digital interface that can access digital data from a local area network (LAN) and take action accordingly. It can also transmit back information, especially status changes and other data, through the LAN. An intelligent breaker has a controller inside which can be programmed to make appropriate decisions as per the system conditions.

Merging units (MUs)

The interface of the instrument transformers (both conventional and nonconventional) with different types of substation protection, control, monitoring, and recording equipment is through a device called a merging unit.

This is defined in IEC 61850-9-1 as follows: "Merging unit: interface unit that accepts multiple analogue CT/VT and binary inputs and produces multiple time synchronized serial unidirectional multi-drop digital point to point outputs to provide data communication via the logical interfaces 4 and 5."

Levels of automation in a substation:

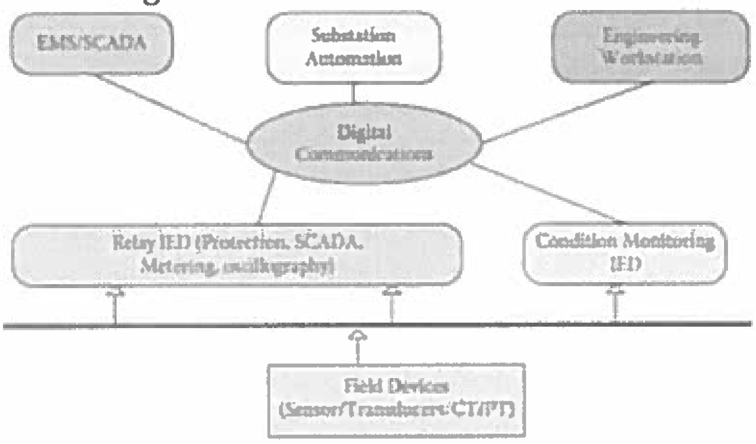
The first level is IED implementation where different IEDs are installed in the substation.

	Utility Enterprise
Level III	Substation Automation functions
Level II	IED Integration
Level I	IED Implementation
	Power system equipment (transformers, breakers, etc.)

The second level is IED integration, utilizing the two-way communications of the IED. IEDs from different vendors and with different functionalities have to be integrated to form a cohesive protection, monitoring, and control system, which also performs a number of other functions like waveform recording and metering.

Once the IEDs are integrated, a number of substation automation applications, the third level, can be run to effectively monitor and control the substation and associated feeder and customer automation functions in the power system.

Integration of functions in a substation.



9. Describe the power quality issues of grid connected renewable energy sources and solutions

10.Explain the Outage Management System (OMS) used in the distribution networks

An outage management system is a critical subsystem, where the distribution network is brought back from a state of emergency to normal state, in a minimum time frame, with disturbance to the least number of customers. Outages are sustained interruptions in the power supply to the customers. OMS includes functions such as trouble call management, outage analysis, crew management, and reliability reporting. Outages can be classified as unplanned and planned.

Unplanned outages

Outages in a distribution system can occur when a fuse or recloser or a circuit breaker operates to clear a fault and the customers located downstream lose power. This may be due to the sudden failure of a component such as transformers, insulators, and so on. The information about the failure is available to the control center via trouble calls from the customers and switch status changes from SCADA, and also the maintenance crew may detect the fault or outage. The OMS will work differently on systems, depending on the level of automation of the distribution system and also by the number of customers served by a distribution transformer. In automated systems, the outage of a component will be known to the SCADA system before any trouble call comes from customers. Especially with an automated metering infrastructure in place, the outage event will be reported to the DMS within a matter of seconds. The OMS program can continuously process and analyze incoming SCADA messages and the trouble calls to locate the fault or outage and the loss of power to customers. The system can also work out the time required to clear the contingency and inform the customers accordingly. Interactive voice response(IVR) systems generally permit trouble call entry into the OMS without human intervention, and the OMS can inform the customers about the outage status already inferred from the SCADA AMI, provide a restoration schedule, and also call back the customers later to verify the supply availability.

Planned outage

Planned outages are scheduled by the utility for routine maintenance or replacement of equipment. Customers are generally informed in advance about these outages. Planned outage can also be due to the load management algorithm implemented by the operator to maintain the load within the incoming supply limits. Planned outages are also handled by the OMS and crew management, and informing the customers in advance has to be completed before the planned outage.

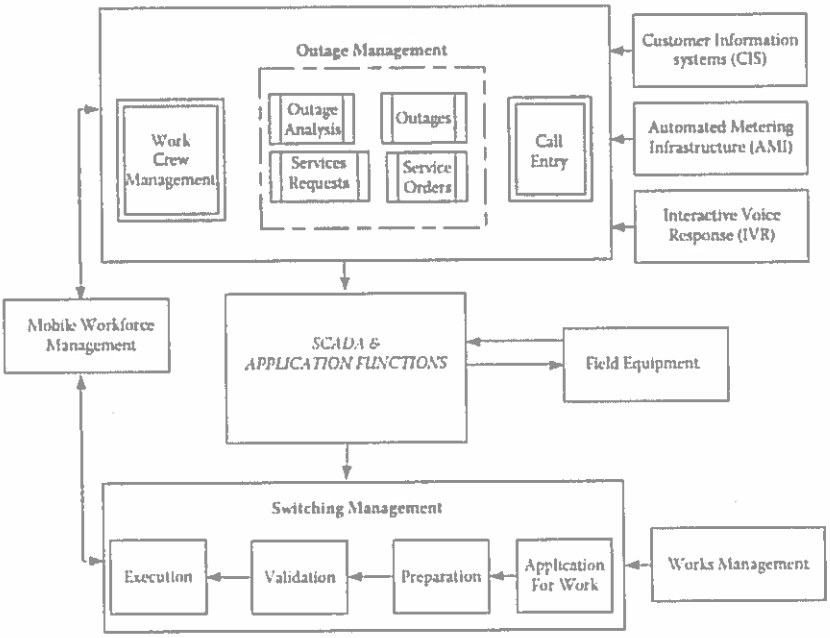


Figure 6.5 Outage management systems.

11. Describe the concept of distribution management system.

Distribution management systems include the real-time functionalities of distribution SCADA coupled with the relevant application functions with support from the corporate process systems such as customer information systems (CISs) and geographical information systems (GISs). DMSs are also integrated with outage management systems (OMSs) and asset management systems (AMSs). In the present scenario, advanced metering infrastructure (AMI) is an integral part of any distribution management planning and discussion, and AMI is integrated with DMS for common information sharing and activity. The subsystems integrated with DMS for providing quality power supply to the customers with maximum reliability with minimal cost to the utility are shown in Figure 6.4. The figure clearly shows the data integration of the subsystems.

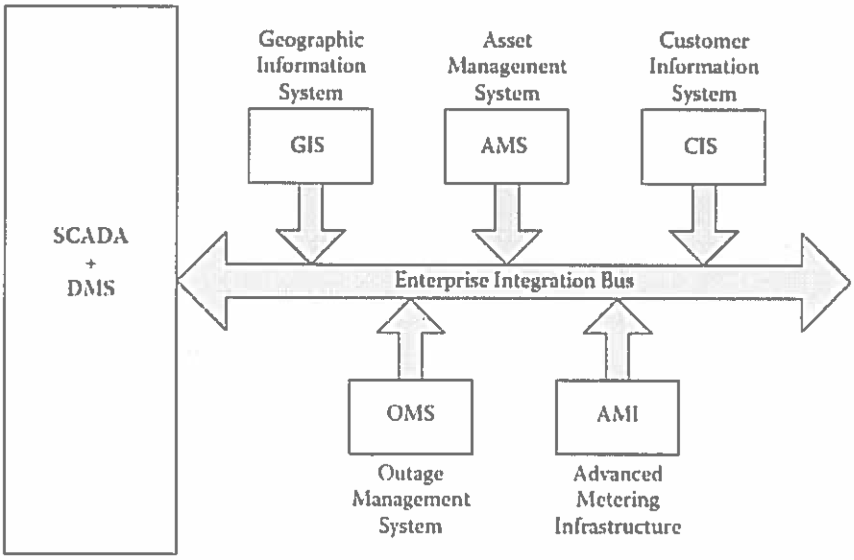
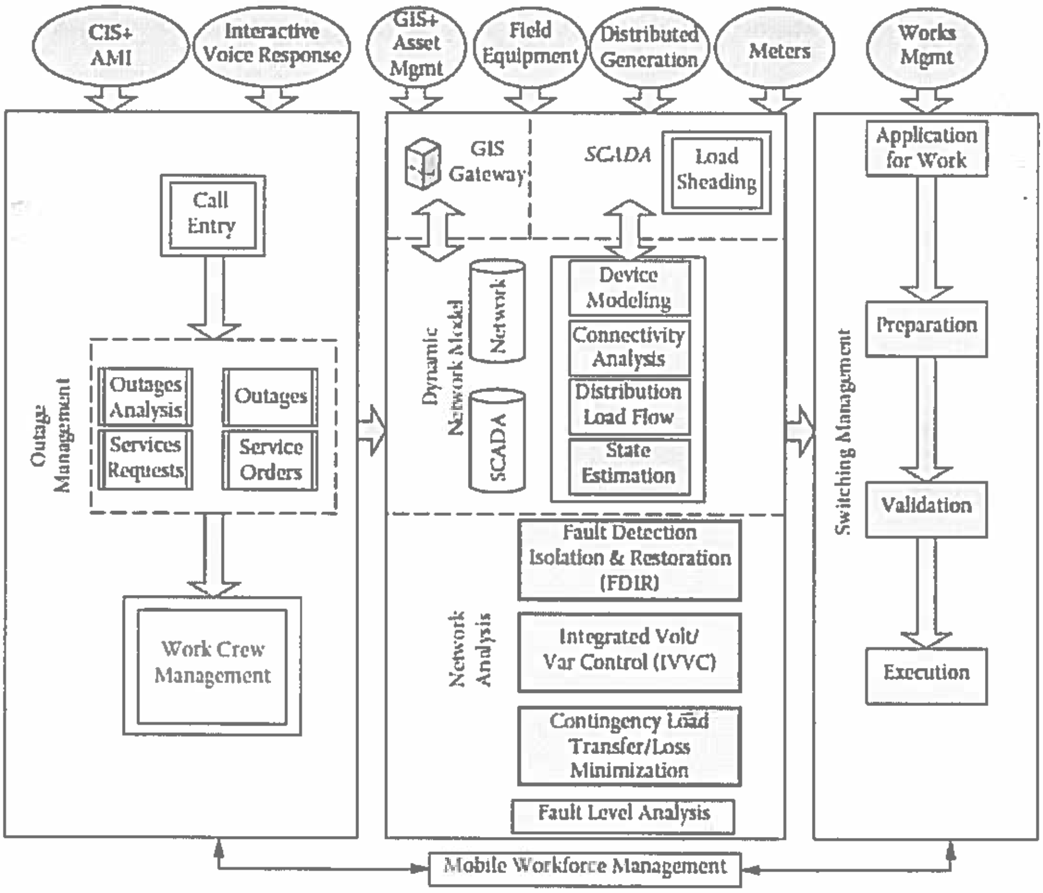


Figure 6.4 SCADA plus DMS integration with other subsystems in a distribution control center.



12. Compare the role of conventional metering and smart metering while involved in demand side management applications

13. Describe the concept of Advanced Metering infrastructure (AMI) in smart grid

The smart grid concept revolves around motivating and including customers in the grid management in various ways. The two-way communication between the utility and the customer, seen as the paradigm shift in the way customers are engaged by the utility, is achieved by the deployment of the AMI. The AMI systems measure, collect, and analyze energy usage, from advanced devices such as electricity meters, gas meters, and water meters through various communication media. However, for a power utility, the AMI network provides the communication link between the customer and the utility and provides measurements and system observability. Thus, AMI is not a single technology, but rather an integration of many technologies that provides an intelligent connection between consumers and system operators. AMI gives consumers the information they need to make intelligent decisions, the ability to execute those decisions, and a variety of choices leading to substantial benefits they do not currently enjoy. Through the integration of multiple technologies such as smart metering, home area networks (HANs), integrated communications, data management applications, and standardized software interfaces with existing utility operations and asset management processes, AMI provides an essential link between the grid, consumers and their loads, generation, and storage resources. Such a link is a fundamental requirement of a modern grid. Figure 7.15 depicts the vision of the modern grid and AMI is the first step toward grid modernization.

Components of AMI

AMI is composed of components that have been integrated to perform as a single platform to provide inputs to other automation systems such as distribution automation, outage management, and customer services. Figure 7.16 shows the AMI structure and the data flow and interface.

a) Smart meters.
b) Intelligent collectors (ICs)
c) AMI head end
d) Meter data management system (MDMS)
e) Communication infrastructure

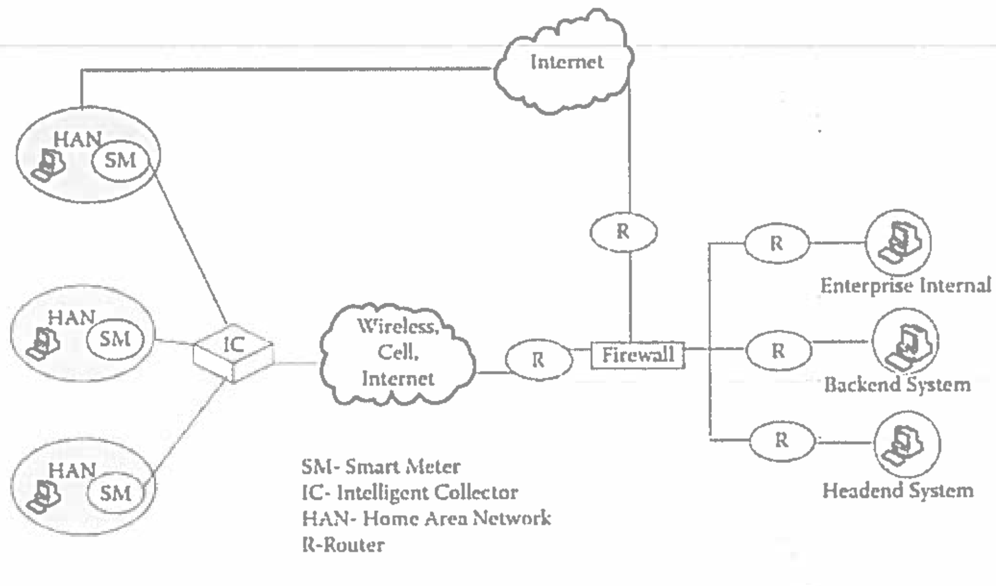


Figure 7.16 AMI structure and data flow.

14.Explain the important role of Local Area Network and PLC in the smart grid systems.

PLC is the first reliable communication medium available to utilities. It uses the power feeder lines as communication media. PLC transmits the radio frequency signals in the range of 30 to 500 kHz. The main components of PLC links are transmitter and receiver terminals, coaxial cable, impedance matching devices, and coupling capacitor for insulation and to inject high-frequency signal onto the distribution line. Line traps are also installed on the power conductor to block the signals entering the substation through an undesired path. PLC equipment is located within the substation, and thus the security is very high. This medium supports services such as voice, telemetry, SCADA, and relaying communication on 220/230 kV, 110/115 kV, or 66 kV interconnected power transmission network at an available data transmission rate up to 9600 baud. There are two types of PLC: analog and digital. Digital PLC requires more maintenance as compared to analog and it is not recommended for noisy power lines. But digital PLC can be increased from one to three channels within the same RF bandwidth. Digital PLC has the capacity for three to four channels (e.g., two voice and one high-speed data), whereas analog PLC has the capacity for two channels (e.g., one voice and one "speech plus" low-speed data). The main disadvantage is that it is not independent of the power line. The availability of fewer channels may be a disadvantage of PLC, and it is expensive on the per-channel basis.

LAN:

The substation LAN must meet industry standards to allow interoperability and the use of plug-and- play devices. Open-architecture principles should be followed, including the use of industry standard protocols (e.g., IEEE 802.x [Ethernet]). The LAN technology employed must be applicable to the substation environment and facilitate interfacing to process-level equipment (IEDs, PLCs) while providing immunity and isolation to substation noise.

15. (a) Illustrate the role of Broadband over power line (BPl) in the Smart grid operation.

Broadband over power line

BPLC technology provides a solution for real-time communication for automating electrical systems, improving service reliability. Intelligent electronic devices (IEDs) and systems that protect receive sensor data and can issue control commands, if they detect abnormal voltage, current or frequency, voltage raised or decreased levels in order to maintain the desired voltage quality. Some of the benefits obtained with the automation of power grids are listed below:

- Reduced O&M Expenditures. Using IEDs to monitor power factor in real-time will save on generation and reduce generation emissions.
- Reliability. Detection fault location, fault insulation and service restoration functionality.
- Flexibility in network topology. Controlling bidirectional energy flows that occur with the inclusion of new sources of power generation to the grid, for example photo voltaic cells, electric cars, wind turbines, among others.
- Efficiency. One of the improvements of intelligent network is to improve the efficiency of energy use, particularly with the use of a smart power management in order to obtain homogeneity in the load curve, for example turning off air conditioners during short-term spikes in electricity price. The intelligent use of technology to improve the efficiency of electrical networks and manage the balance between supply and demand reduces the need for emergency generators. In addition, demand management and the

creation of better storage allow the flexibility to manage the intermittency of renewable energy sources.

- Sustainability. The flexibility of Smart Grids allows the penetration of new renewable energy sources to electricity generation and thus has a variety of ability to meet demand.
- Market. Bidirectional communication between consumers and energy suppliers, allowing greater flexibility in their operating strategies.
- Reliability and Quality of Service.
- Monitoring and measurement of energy quality in real time.
- Increased Availability: The Micro-Grids can be switched based on network's condition or fee. Reconnect consumer choice.
- Energy Storage. Energy storage has implicit time dependence: a node may inject or extract energy depending on weather and operating conditions.
- Demand response. Allows generators and loads to interact automatically in real time, coordinating demand to flatten spikes. Eliminating the fraction of demand that occurs in these spikes, demand management eliminates the cost of adding reserve generators, cuts wear and extends the equipment's lifespan. Demand management allows users to reduce their energy bills down saying devices priority to use energy only when it is cheaper too

(b) Explain the important features of house area network (HAN) used in smart grid.

Home Area Network (HAN): A home area network connects all the components of the HEM system, including the sensors, measuring devices, smart appliances, and any displays into a network for implementation by transferring the monitoring and control data as required. There are different technologies used in building the HAN backbone, depending on the communication technology and protocol used. Efforts are ongoing to standardize the technologies used, and the three technologies include (1) Zigbee wireless standards that connect the widest range of home devices, to work together with the control facility; (2) using the power line wiring in the network with smart plugs that will have specific IP addresses and can be monitored and controlled by the HEMs; (3) using the Z wave open standard for wireless which will enable the compatible devices to communicate and build an effective HAN. However, integrating various technologies for a homogeneous HAN is still a challenge as interoperability is an issue and also the security and privacy of the customer information must be ensured.

NSRIT

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	**Program**	CSE (Honors)	**Academic Year**	2021 - 2022
Course Code	20CSH01	**Test Duration**	3 Hrs. Max. Marks 70	**Semester**	IV
Course	**Advanced Computer Architecture**				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define Amdahl's law.	20CSH01.1	L1
2	Outline the basic structure of memory hierarchy.	20CSH01.2	L2
3	Distinguish pipelining from parallelism.	20CSH01.3	L2
4	Classify the vector instruction types.	20CSH01.4	L2
5	What is operand forwarding?	20CSH01.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Discuss in detail about the SIMD and multi vector systems.	12M	20CSH01.1	L1
	OR			
7	Explain the difficulties faced by parallel processing programs.	12M	20CSH01.1	L1
8	Discuss in detail about the memory hierarchy technologies with necessary illustrations.	12M	20CSH01.2	L2
	OR			
9 (a)	Explain the virtual memory address translation and TLB with necessary diagram.	6M	20CSH01.2	L2
9 (b)	Examine the concept of paging and segmentation.	6M	20CSH01.2	L2
10	What is dynamic scheduling and compare how it is different from static pipeline scheduling.	12M	20CSH01.3	L2
	OR			
11	Explain in detail about the pipelining and super scalar techniques with necessary illustrations.	12M	20CSH01.3	L2
12	Discuss the steps involved in the address translation of virtual memory with necessary illustrations.	12M	20CSH01.4	L2
	OR			
13 (a)	Explain Multicore architecture of computers.	6M	20CSH01.4	L2
13 (b)	Explain the three generations of multi computers.	6M	20CSH01.4	L2
14	Discuss in detail about the instruction level parallelism.	12M	20CSH01.5	L2
	OR			
15	Illustrate the following i. Operand Forwarding ii. Branch Prediction	12M	20CSH01.5	L2

NSRIT

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

ADVANCED COMPUTER ARCHITECTURE

Part A (Short Answer Questions 5 x 2 = 10 Marks)

1.Define Amdahl's law

Amdahl's law, named after a computer architect named Gene Amdahl and his work in the 1960s, is a law showing how much latency can be taken out of a Performance task by introducing parallel computing. In parallel computing, Amdahl's law is mainly used to predict the theoretical maximum speed up for Program processing using multiple processors.

Formula

Amdahl's Law can be expressed in mathematically as follows –

$Speedup_{MAX} = 1/((1-p) + (p/s))$

$Speedup_{MAX}$ = maximum performance gain

s = performance gain factor of p after implement the enhancements.

p = the part which performance needs to be improved.

2.Outline the basic structure of memory hierarchy.

Level 0: CPU registers

Level 1: Cache memory

Level 2: Main memory or primary memory

Level 3: Magnetic disks or secondary memory

Level 4: Optical disks or magnetic types or tertiary Memory

3. Distinguish pipelining from parallelism.

Parallelism involves replicated hardware (exploiting space). Pipelining involves re-using hardware optimally based on data flows (exploiting time). Parallelism can do two calculations at the same time in separate compute units. Pipelining can do a calculation twice in incremental time over doing it once, using a single set of compute units.

In pipelining independent computations are executed in an interleaved manner, while parallel processing achieves the same using duplicate hardware. Parallel processing systems are also referred to as block processing systems. The block size indicates the number of inputs processed simultaneously.

4. Classify the vector instruction types.

A Vector operand contains an ordered set of n elements, where n is called the length of the vector. All elements in a vector are same type scalar quantities, which may be a floating point number, an integer, a logical value, or a character.

Four primitive types of vector instructions are:

f1: V --> V

f2: V --> S

f3: V x V --> V

f4: V x S --> V

5. What is operand forwarding?

Operand forwarding (or data forwarding) is an optimization in pipelined CPUs to limit performance deficits which occur due to pipeline stalls. A data hazard can lead to a pipeline stall when the current operation has to wait for the results of an earlier operation which has not yet finished. Data is forwarded when it is ready. · The previous clock cycle must complete before data being forwarded. · Use split phase access if data is ready-...

Part B (Long Answer Questions 5 x 12 = 60 Marks)

6. Discuss in detail about the SIMD and multi vector systems.[12M]

SIMD Computers

In SIMD computers, 'N' number of processors are connected to a control unit and all the processors have their individual memory units. All the processors are connected by an interconnection network.

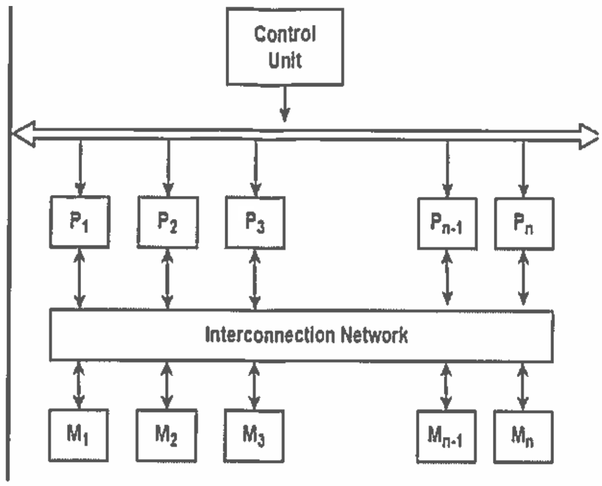
SIMD represents single-instruction multiple-data streams. The SIMD model of parallel computing includes two parts such as a front-end computer of the usual von Neumann style, and a processor array as displayed in the figure.

The processor array is a collection of identical synchronized processing elements adequate for simultaneously implementing the same operation on various data. Each processor in the array has a small amount of local memory where the distributed data resides while it is being processed in parallel.

The processor array is linked to the memory bus of the front end so that the front end can randomly create the local processor memories as if it were another memory.

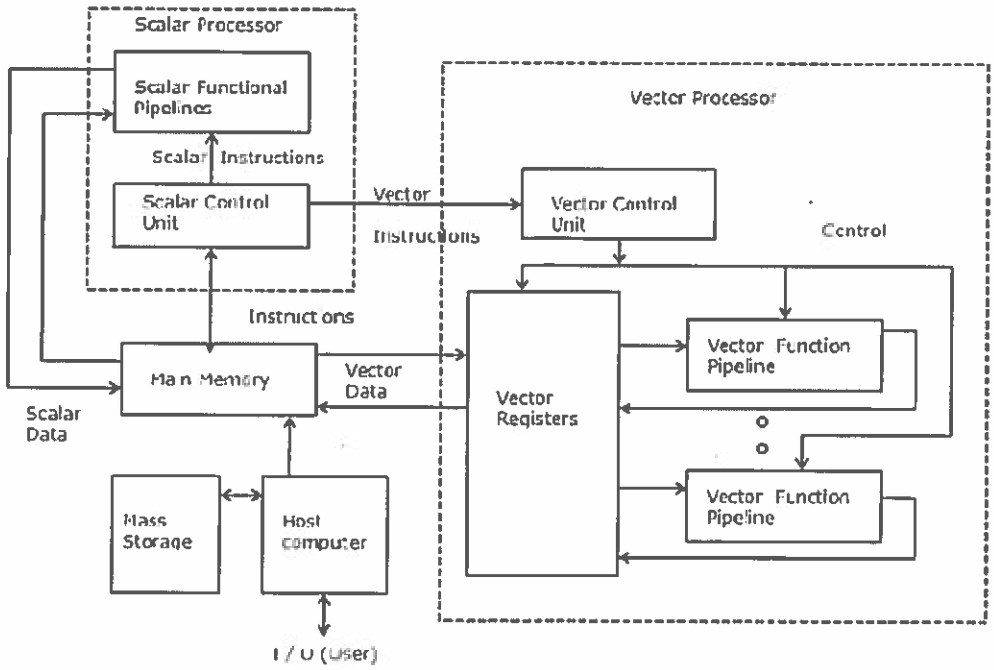
Two main configurations have been applied in SIMD machines. In the first scheme, each processor has its local memory. Processors can interact with each other through the interconnection network. If the interconnection network does not support direct connection between given groups of processors, then this group can exchange information via an intermediate processor.

In the second SIMD scheme, processors and memory modules communicate with each other via the interconnection network. Two processors can send information between each other via intermediate memory module(s) or possibly via intermediate processor(s). The BSP (Burroughs' Scientific Processor) used the second S.



Multi Vector Computer:

- In a vector computer, a vector processor is attached to the scalar processor as an optional feature.
- The host computer first loads program and data to the main memory.
- Then the scalar control unit decodes all the instructions.
- If the decoded instructions are scalar operations or program operations, the scalar processor executes those operations using scalar functional pipelines.
- On the other hand, if the decoded instructions are vector operations then the instructions will be sent to vector control unit.



OR

7. Explain the difficulties faced by parallel processing programs. [12M]

Amount of Parallelizable CPU-Bound Work

The number one requirement for parallelization is that the program must have enough work that can be performed in parallel. If only half of the work can be parallelized, Amdahl's Law dictates that we are not going to be able to speed up the program by more than a factor of two. Also, additional CPUs thrown at a task will help the most if the CPU was the performance bottleneck. If the program spends 90% of its time waiting for a server to execute SQL queries, then parallelizing the program likely will not achieve significant benefits

Task Granularity

Even if a program does a lot of parallelizable work, we must be careful to ensure that we will split up the work into appropriately-sized chunks which will execute in parallel. If we create too many chunks, the overheads of managing and scheduling the chunks will be large. If we create too few chunks, some cores on the machine will have nothing to do.In some parts of the Parallel Extensions API, such as Parallel. For and PLINQ, the code in our library is responsible for deciding on the proper granularity of tasks. In other parts of the API, such as tasks and futures, it is the responsibility of the user code.

Memory Allocations and Garbage Collection

Some programs spend a lot of time in memory allocations and garbage collections. For example, programs that manipulate strings tend to allocate a lot of memory, particularly if they are not designed carefully to prevent unnecessary allocations.

Unfortunately, allocating memory is an operation that may require synchronization. After all, we need to ensure that memory regions allocated by different threads will not overlap.

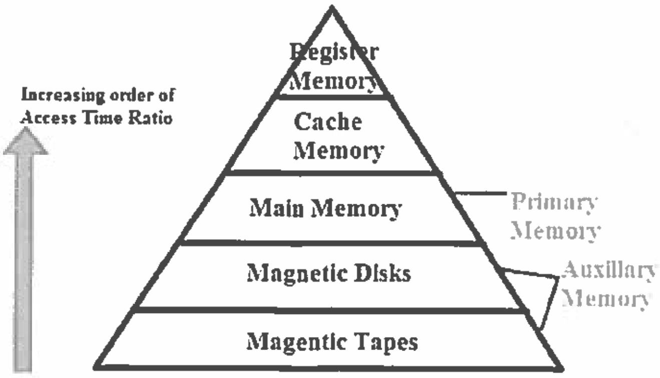
In order to explain this particular performance problem of parallel programs, let's quickly review a few details about how caches work on today's mainstream computers. When a CPU reads a value from the main memory, it copies the value to cache, so that subsequent accesses to that value are much faster. In fact, rather than just bringing in that particular value into cache, the CPU will bring in also nearby memory locations. It turns out that if a program read a particular memory location, chances are that it is going to read nearby values too. So, values are moved between main memory and cache in chunks called *cache lines*, typically of size 64 or 128 bytes.

One problem that arises on machines with multiple cores is that if one core invalidates a particular memory location, the version of that memory location cached by another core gets invalidated. Then, the core with an invalid cached copy must go all the way to the main memory on the next read of that memory location. So, if two cores keep writing and reading a particular memory location, they may end up continuously invalidating each other's caches, sometimes dramatically reducing the performance of the program.

8 .Discuss in detail about the memory hierarchy technologies with necessary illustrations. [12M]

The memory in a computer can be divided into five hierarchies based on the speed as well as use. The processor can move from one level to another based on its requirements. The five hierarchies in the memory are registers, cache, main memory, magnetic discs, and magnetic tapes. The first three hierarchies are volatile memories which mean when there is no power, and then automatically they lose their stored data. Whereas the last two hierarchies are not volatile which means they store the data permanently. A memory element is the set of storage devices which stores the binary data in the type of bits. In general, the storage of memory can be classified into two categories such as volatile as well as non- volatile.

Memory Hierarchy in Computer Architecture design in a computer system mainly includes different storage devices. Most of the computers were inbuilt with extra storage to run more powerfully beyond the main memory capacity. The following memory hierarchy diagram is a hierarchical pyramid for computer memory. The designing of the memory hierarchy is divided into two types such as primary (Internal) memory and secondary (External) memory.



The memory hierarchy in computers mainly includes the following.

Registers

Usually, the register is a static RAM or SRAM in the processor of the computer which is used for holding the data word which is typically 64 or 128 bits. The program counter register is the most important as well as found in all the processors. Most of the processors use a status word register as well as an accumulator. A status word register is used for decision making, and the accumulator is used to store the data like mathematical operation. Usually, computers like complex instruction set computers have so many registers for accepting main memory, and RISC- reduced instruction set computers have more registers.

Cache Memory

Cache memory can also be found in the processor, however rarely it may be another IC (integrated circuit) which is separated into levels. The cache holds the chunk of data which are frequently used from main memory. When the processor has a single core then it will have two (or) more cache levels rarely. Present multi-core processors will be having three, 2-levels for each one core, and one level is shared.

Main Memory

The main memory in the computer is nothing but, the memory unit in the CPU that communicates directly. It is the main storage unit of the computer. This memory is fast as well as large memory used for storing the data throughout the operations of the computer. This memory is made up of RAM as well as ROM.

Magnetic Disks

The magnetic disks in the computer are circular plates fabricated of plastic otherwise metal by magnetized material. Frequently, two faces of the disk are utilized as well as many disks may be stacked on one spindle by read or write heads obtainable on every plane. All the disks in computer turn jointly at high speed. The tracks in the computer are nothing but bits which are stored within the magnetized plane in spots next to concentric circles. These are usually separated into sections which are named as sectors.

Magnetic Tape

This tape is a normal magnetic recording which is designed with a slender magnetizable covering on an extended, plastic film of the thin strip. This is mainly used to back up huge data. Whenever the computer requires accessing a strip, first it will mount to access the data. Once the data is allowed, then it will be uncounted. The access time of memory will be slower within magnetic strip as well as it will take a few minutes for accessing a strip.

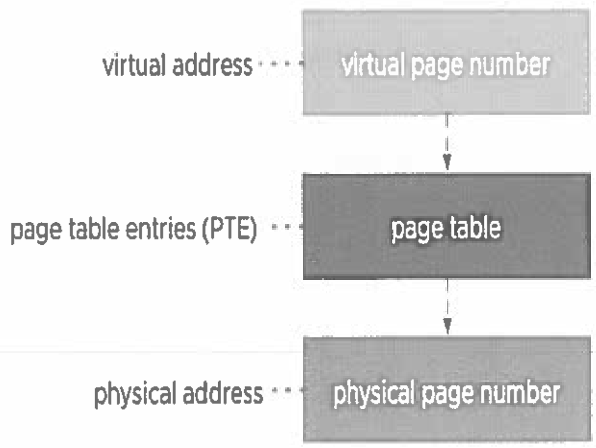
OR

9(a) Explain the virtual memory address translation and TLB with necessary diagram.[6M]

Virtual Memory Translations

The physical address space is your system RAM, the memory modules inside your ESXi hosts, also referred to as the global system memory. When talking about virtual memory, we are talking about the memory that is controlled by an operating system, or a hypervisor like vSphere ESXi. Whenever workloads access data in memory, the system needs to look up the physical memory address that matches the virtual address. This is what we refer to as memory translations or mappings.

To map virtual memory addresses to physical memory addresses, page tables are used. A page table consists of numerous page table entries (PTE).

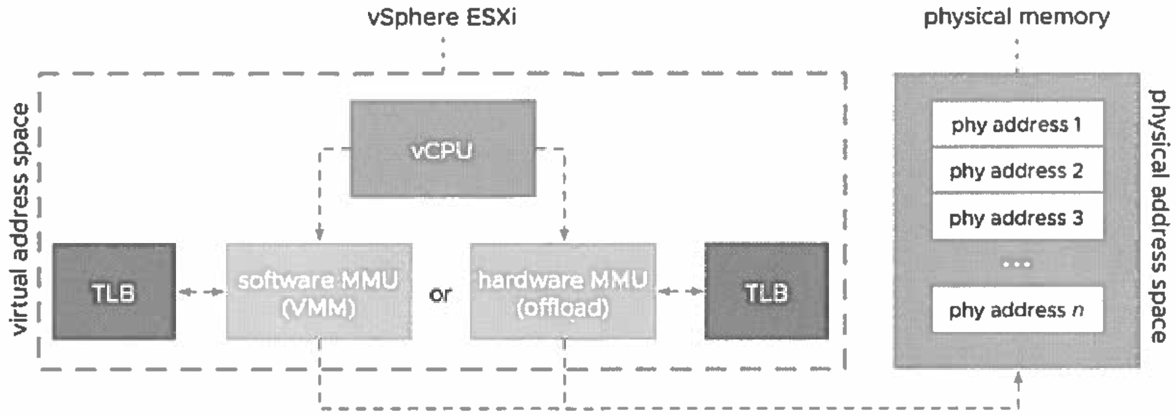


One memory page in a PTE contains data structures consisting of different sizes of 'words'. Each type of word contains multiple bytes of data (*WORD* (16 bits/2 bytes), *DWORD* (32 bits/4 bytes) and *QWORD* (64 bits/8 bytes)). Executing memory translations for every possible word, or virtual memory page, into physical memory address is not very efficient as this could potentially be billions of PTE's. We need PTE's to find the physical address space in the system's global memory, so there is no way around them.

To make memory translations more efficient, we use page tables to group chunks of memory addresses in one mapping. Looking at an example of a DWORD entry of 4 bytes; a page table covers 4 kilobytes instead of just the 4 bytes of data in a single page entry. For example, using a page table, we can translate virtual address space 0 to 4095 and say this is found in physical address space 4096 to 8191. Now we no longer need to map all the PTE's separately, and be far more efficient by using page tables.

MMU and TLB

The page tables are managed by a Memory Management Unit (MMU). All the physical memory references are passed through the MMU. The MMU is responsible for the translation between virtual memory addresses and physical memory addresses. With vSphere ESXi, a virtual machine's vCPU will call out to MMU functionality by the Virtual Machine Monitor (VMM) process, or a hardware MMU supported by a vendor specific CPU offloading instruction.



The Memory Management Unit (MMU) works with the Translation Lookaside Buffer (TLB) to map the virtual memory addresses to the physical memory layer. The page table always resides in physical memory, and having to look up the memory pages directly in physical memory, can be a costly exercise for the MMU as it introduces latency. That is where the TLB comes into play.

TLB in Detail

The TLB acts as a cache for the MMU that is used to reduce the time taken to access physical memory. The TLB is a part of the MMU. Depending on the make and model of a CPU, there's more than one TLB, or even multiple levels of TLB like with memory caches to avoid TLB misses and ensuring as low as possible memory latency.

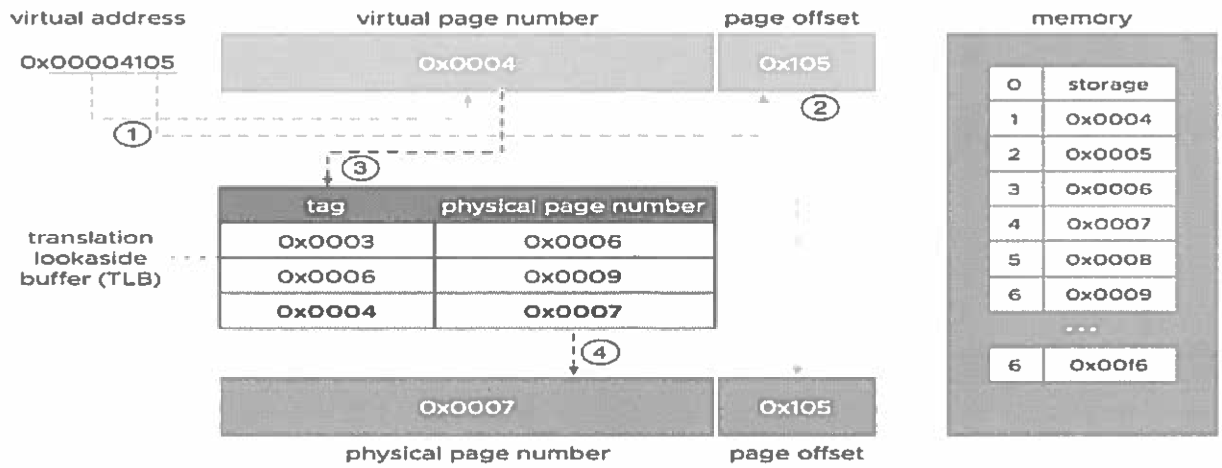
In essence, the TLB stores recent memory translations of virtual to physical. It is a cache for page tables. Because it is part of the MMU, the TLB lives inside the CPU package. This is why the TLB is faster than main memory, which is where the page tables exist. Typically access times for a TLB are ~10 ns where main memory access times are around 100 ns.

Now that we covered the basics on memory translation, let's take a look at some example scenarios for the TLB.

TLB hit

A virtual memory address comes in, and needs to be translated to the physical address. The first step is always to dissect the virtual address into a virtual page number, and the page offset. The offset consists of the last bits of the virtual address. The offset bits are not translated and passed through to the physical memory address. The offset contains bits that can represent all the memory addresses in a page table.

So, the offset is directly mapped to the physical memory layer, and the virtual page number matches a tag already in the TLB. The MMU now immediately knows what physical memory page to access without the need to look into the global memory.

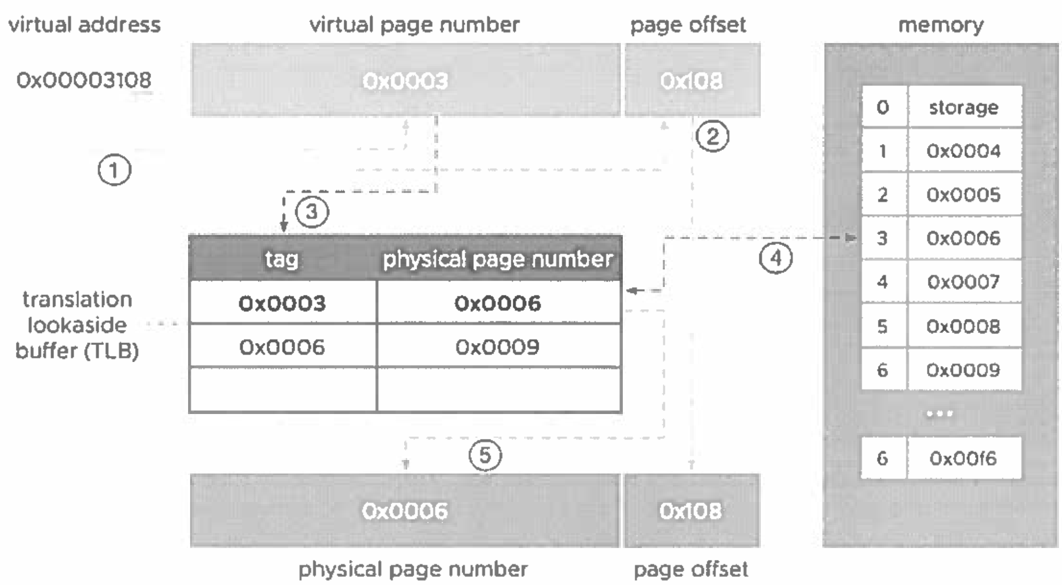


In the example provided in the above diagram, the virtual page number is found in the TLB, and immediately translated to the physical page number.

1. The virtual address is dissected in the virtual page number and the page offset.
2. The page offset is passed through as it is not translated.
3. The virtual page number is looked up in the TLB, looking for a tag with the corresponding number.
4. There is an entry in the TLB (hit), meaning we immediately can translate the virtual to the physical address.

TLB miss

What happens when a virtual page number is not found in the TLB, also referred to as a TLB miss? The TLB needs to consult the system's global memory to understand what physical page number is used. Reaching out to physical memory means higher latency compared to a TLB hit. If the TLB is full and a TLB miss occurs, the least recent TLB entry is flushed, and the new entry is placed instead of it. In the following example, the virtual page number is not found in the TLB, and the TLB needs to look into memory to get the page number.

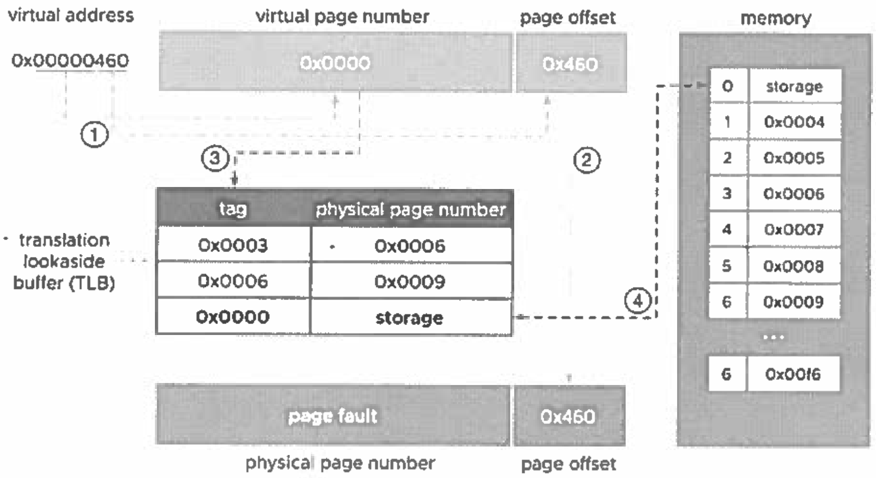


1. The virtual address is dissected in the virtual page number and the page offset.
2. The page offset is passed through as it is not translated.

3. The virtual page number is looked up in the TLB, looking for a tag with a corresponding number. In this example, the TLB does not yet have a valid entry.
4. TLB reaches out to memory to find page number 3 (because of the tag, derived from the virtual page number). Page number 3 is retrieved in memory with value 0x0006.
5. The memory translation is done and the entry is now cached in the TLB.

Retrieve from storage

A TLB miss is not ideal, but the worst-case scenario is data that is not residing in memory but on storage media (flash or disk). Where we are talking nanoseconds to retrieve data in caches or global memory, getting data from storage media will quickly run into milliseconds or seconds depending on the media used.



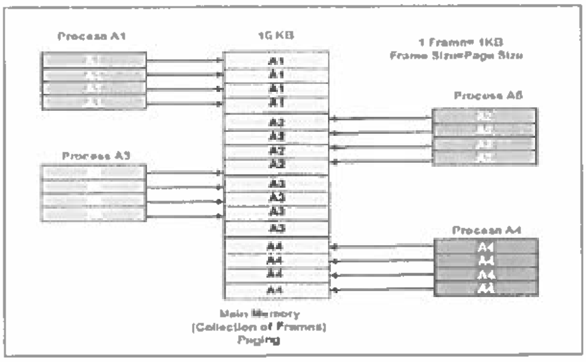
1. The virtual address is dissected in the virtual page number and the page offset.
2. The page offset is passed through as it is not translated.
3. The virtual page number is looked up in the TLB, looking for a tag with a corresponding number. In this example, the TLB does not yet have a valid entry.
4. TLB reaches out to memory to find page number 0 (because of the tag, derived from the virtual page number). Page number 0 is retrieved in memory but finds that the data does not resides in memory, but on storage. A page fault is triggered, because we cannot translate memory pages for data that is not in memory. We need to wait for the data from storage.

9(b) Examine the concept of paging and segmentation. [6M]

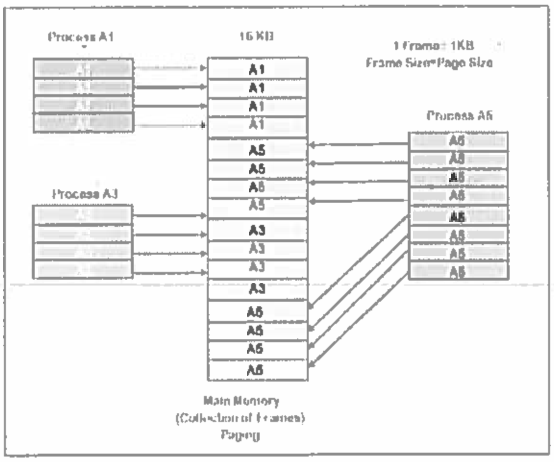
Paging is a storage technique used for memory management. In paging, the (OS) Operating System retrieves the processes from the secondary memory into the main memory, and the memory is in the form of the pages. In this technique, we split the main memory into the small blocks of physical memory which are called frames. The size of the frames is fixed. In paging, for the maximum usage of the main memory and to prevent external fragmentation, the frame size must be the same as the page size. Paging is a logical concept, and it helps us to access the data faster.

Example of Paging

Suppose we have main memory, and the size of the main memory is 16 KB, and the size of the frame is 1 KB. In this, the main memory is split into 16 frames, and each frame is of 1 KB. In the system, we have four distinct processes, and the processes are A1, A2, A3, and A4, and the size of each process is 4 KB. In this, we split or divide all the pages into the pages of size 1KB so that the OS can store 1page in 1 frame.



When the process started its execution, all the frames were vacant to store process pages in a contiguous manner. The below figure shows the frames, pages, and the mapping between the frames and the pages. We can see in the following example that after some time, the process A2 and the process A4 are moved into the waiting state. So, the eight frames will become vacant, and we need to load or put other pages in those vacant blocks. The process A5 is having a size of eight pages (8 KB), which are waiting in the ready queue.



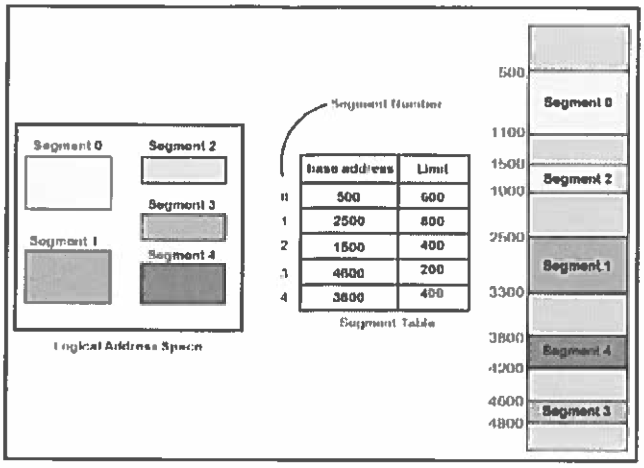
We can see in the following example that we have eight non-contiguous frames that are existing in the memory, with the help of paging; we can store the processes at different places. Due to this, we can load the pages of the A5 process instead of process A2 and Process A4.

Segmentation

Segmentation is a technique of memory management. It is just like the Paging technique except the fact that in segmentation, the segments are of variable length but, in Paging, the pages are of fixed size. In segmentation, the memory is split into variable-length parts. Each part is known as segments. The information which is related to the segment is stored in a table which is called a segment table.

There are two types of information stored in the segment table:
Limit
Base
Limit: - The limit is the length or size of the segment
Base: - The base is the base address of the segment.
A segment of the program comprises of the utility function, data structure, and the main function of the program for each process. The operating system preserves a segment map table for mapping. The table

consists of segment number, list of the memory blocks which are free along with its size, and its memory location in the virtual memory or the main memory.



Types of Segmentation:
There are two types of Segmentation:
1. Simple Memory Segmentation
2. Virtual Memory Segmentation

Simple Memory Segmentation: - In simple memory segmentation, each process is split into different segments, and at the run time, all the processes are loaded. Also, not all the processes need to be loaded into a contiguous way.

Virtual Memory Segmentation: - As simple memory segmentation, in virtual memory segmentation, each process is split into different segments, but not all of them are residents at any point of time.

10. What is dynamic scheduling and compare how it is different from static pipeline scheduling. [12M]

In a dynamically scheduled pipeline, all instructions pass through the issue stage in order; however they can be stalled or bypass each other in the second stage and thus enter execution out of order. Instructions will also finish out-of-order.

Dynamic Scheduling is a technique in which the hardware rearranges the instruction execution to reduce the stalls, while maintaining data flow and exception behavior. The advantages of dynamic scheduling are:
• It handles cases when dependences are unknown at compile time
– (e.g., because they may involve a memory reference)
• It simplifies the compiler
• It allows code compiled for one pipeline to run efficiently on a different pipeline
• Hardware speculation, a technique with significant performance advantages, builds on dynamic scheduling

In a dynamically scheduled pipeline, all instructions pass through the issue stage in order; however they can be stalled or bypass each other in the second stage and thus enter execution out of order. Instructions will also finish out-of-order.

The three steps in a dynamic scheduler are listed below
• Issue
• Get next instruction from FIFO queue
• If available RS, issue the instruction to the RS with operand values if available
• If a RS is not available, it becomes a structural hazard and the instruction stalls
• If an earlier instruction is not issued, then subsequent instructions cannot be issued
• If operand values are not available, the instructions will wait in the RSs looking at CDBs for operands
• Execute

• When operand becomes available on the CDB, store it in any reservation station waiting for it
• When all operands are ready, the instruction is executed by the respective functional unit
• Loads and store are maintained in program order through the effective address
• No instruction allowed initiating execution until all branches that precede it in program order have completed
• Write result
• Write result on CDB into reservation stations and store buffers
• Stores must wait until address and value are received
Assumptions (for now):
– 1 instruction issue / cycle
– Several pipelines with a common IF and ID
• Ideal CPI still 1, but real CPI won't be 1 but will be closer to 1 than before
• Same techniques will be used when we look at multiple issues

Differences:

- Static scheduling (optimized by compiler)
  – When there is a stall (hazard) no further issue of instructions
  – Of course, the stall has to be enforced by the hardware
- Dynamic scheduling (enforced by hardware)
  Instructions following the one that stalls can issue if they do not produce structural hazards.

- A linear pipeline processor is a series of processing stages and memory access. A nonlinear pipelining (also called dynamic pipeline) can be configured to perform various functions at different times. In a dynamic pipeline there is also feed forward or feedback connection. Non-linear pipeline also allows very long instruction words.
- The performance of static pipelines is severely degraded when the operations change often, since this requires the pipeline to be drained and refilled each time. A dynamic pipeline can perform more than one operation at a time.

OR

11 .Explain in detail about the pipelining and super scalar techniques with necessary illustrations. [12M]

Pipelining is the process of accumulating instruction from the processor through a pipeline. It allows storing and executing instructions in an orderly process. It is also known as pipeline processing. Pipelining is a technique where multiple instructions are overlapped during execution. Pipeline is divided into stages and these stages are connected with one another to form a pipe like structure. Instructions enter from one end and exit from another end. Pipelining increases the overall instruction throughput. In pipeline system, each segment consists of an input register followed by a combinational circuit. The register is used to hold data and combinational circuit performs operations on it. The output of combinational circuit is applied to the input register of the next segment.

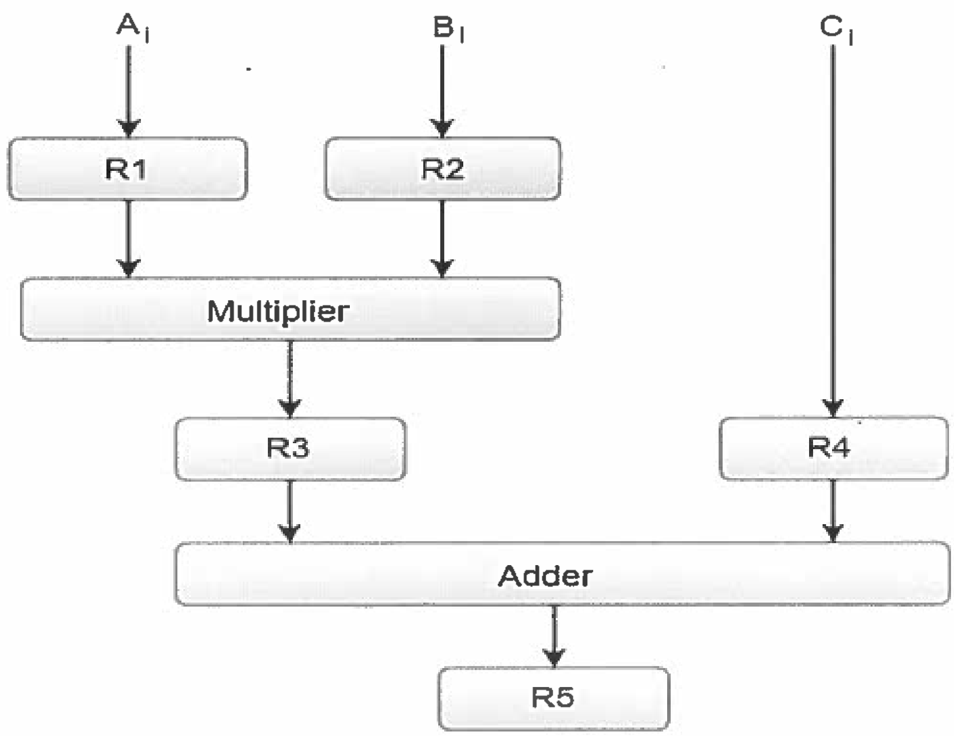
The term Pipelining refers to a technique of decomposing a sequential process intosub-operations, with each sub-operation being executed in a dedicated segment thatoperatesconcurrentlywithallothersegments.
The most important characteristic of a pipeline technique is that several computationscanbeinprogressindistinctsegmentsatthesametime.Theoverlappingofcomputationismade possible by associating a register with each segment in thepipeline.Theregistersprovideisolationbetweeneachsegmentsothateachcanoperateondistinctdatasimultaneously.
Thestructureofapipelineorganizationcanberepresentedsimplybyincludinganinputregisterforeachsegmentfollowedbyacombinationalcircuit.

Let us consider an example of combined multiplication and addition operation to get abetterunderstandingofthepipelineorganization.
The combined multiplication and addition operation is done with a stream of numberssuchas:
$A_i*B_i+C_i$fori=1,2,3, ,7

The operationtobe performedonthenumbersisdecomposedintosub-operationswitheachsub-operationtobeimplementedinasegmentwithinapipeline.
Thesub-operationsperformedineachsegmentofthepipelinearedefinedas:$R1 \leftarrow A_i, R2 \leftarrow B_i$ InputA_i,andB_i
$R3 \leftarrow R1*R2, R4 \leftarrow C_i$ Multiply,andinputC_i
$R5 \leftarrow R3+R4$ Add$C_i$toproduct

Thefollowingblockdiagramrepresentsthecombinedaswellasthesub-operationsperformedineachsegmentofthepipeline.

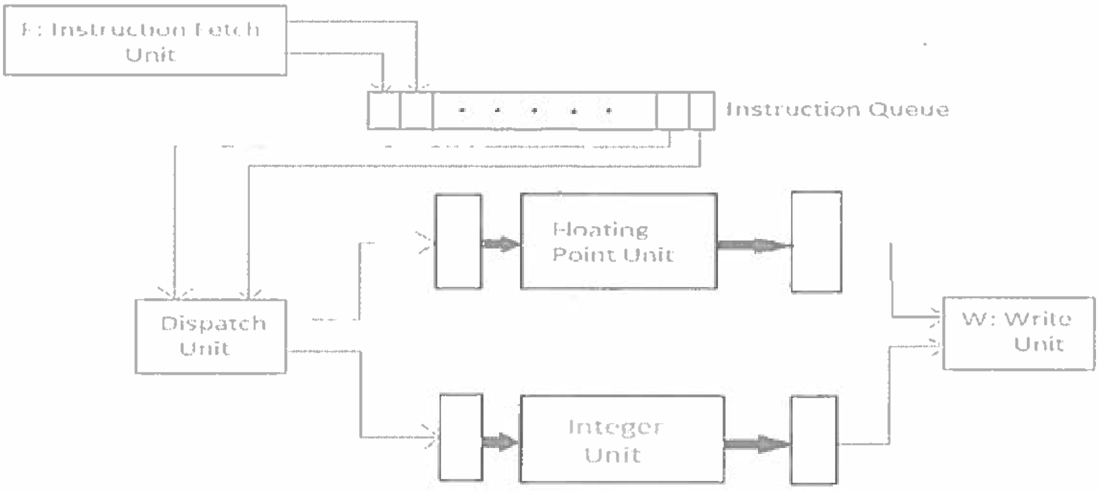
Pipeline Processing:



Registers R1, R2, R3, and R4 hold the data and the combinational circuits operate in a particular segment. The output generated by the combinational circuit in a given segment is applied as an input register of the next segment. For instance, from the block diagram, we can see that the register R3 is used as one of the input registers for the combinational adder circuit.

Superscalar Architecture

A more aggressive approach is to equip the processor with multiple processing units to handle several instructions in parallel in each processing stage. With this arrangement, several instructions start execution in the same clock cycle and the process is said to use multiple issue. Such processors are capable of achieving an instruction execution throughput of more than one instruction per cycle. They are known as 'Superscalar Processors'.



Processor with Two Execution Units

In the above diagram, there is a processor with two execution units; one forinteger and one for floating point operations. The instruction fetch unit is capableof reading the instructions at a time and storing them in the instruction queue. Ineach cycle, the dispatch unit retrieves and decodes up to two instructions fromthe front of the queue. If there is one integer, one floating point instruction and nohazards,boththeinstructionsaredispatchedinthesameclockcycle.

AdvantagesofSuperscalarArchitecture:
The compiler can avoid many hazards through judicious selection andorderingofinstructions.
The compiler should strive to interleave floating point and integerinstructions.Thiswouldenablethedispatchunittokeepboththeintegerandfloating-pointunitsbusymostofthetime.
Ingeneral,highperformanceisachievedifthecompilerisabletoarrange program instructions to take maximum advantage of theavailablehardwareunits.

DisadvantagesofSuperscalarArchitecture:
InaSuperscalarProcessor,thedetrimentaleffectonperformanceofvarioushazardsbecomesevenmorepronounced.

12. Discuss the steps involved in the address translation of virtual memory with necessary illustrations. [12M]

Virtual Memory is a storage allocation scheme in which secondary memory can be addressed as though it were part of the main memory. The addresses a program may use to reference memory are distinguished from the addresses the memory system uses to identify physical storage sites, and program-generated addresses are translated automatically to the corresponding machine addresses.

The size of virtual storage is limited by the addressing scheme of the computer system and the amount of secondary memory is available not by the actual number of the main storage locations.

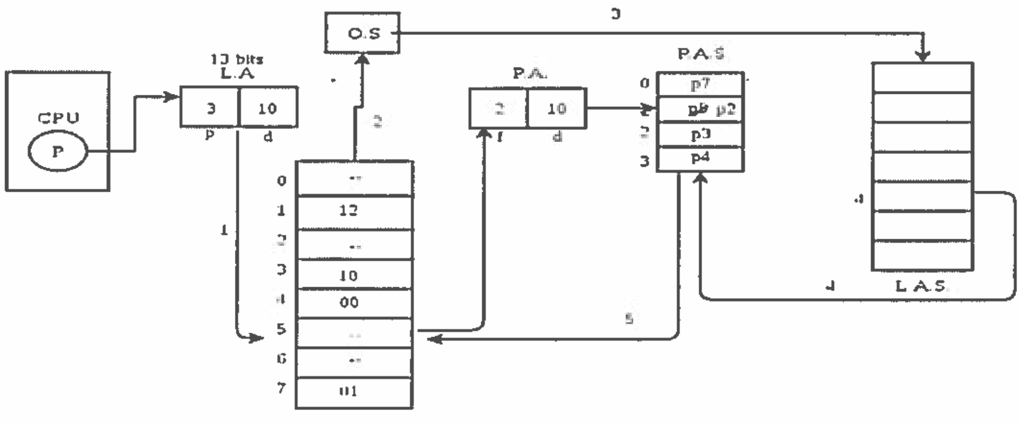
It is a technique that is implemented using both hardware and software. It maps memory addresses used by a program, called virtual addresses, into physical addresses in computer memory.

1. All memory references within a process are logical addresses that are dynamically translated into physical addresses at run time. This means that a process can be swapped in and out of the main memory such that it occupies different places in the main memory at different times during the course of execution.
2. A process may be broken into a number of pieces and these pieces need not be continuously located in the main memory during execution. The combination of dynamic run-time address translation and use of page or segment table permits this.

If these characteristics are present then, it is not necessary that all the pages or segments are present in the main memory during execution. This means that the required pages need to be loaded into memory whenever required. Virtual memory is implemented using Demand Paging or Demand Segmentation.

Demand Paging:

The process of loading the page into memory on demand (whenever page fault occurs) is known as demand paging.



1. If the CPU tries to refer to a page that is currently not available in the main memory, it generates an interrupt indicating a memory access fault.
2. The OS puts the interrupted process in a blocking state. For the execution to precede the OS must bring the required page into the memory.
3. The OS will search for the required page in the logical address space.
4. The required page will be brought from logical address space to physical address space. The page replacement algorithms are used for the decision-making of replacing the page in physical address space.
5. The page table will be updated accordingly.
6. The signal will be sent to the CPU to continue the program execution and it will place the process back into the ready state.

Hence whenever a page fault occurs these steps are followed by the operating system and the required page is brought into memory.

OR

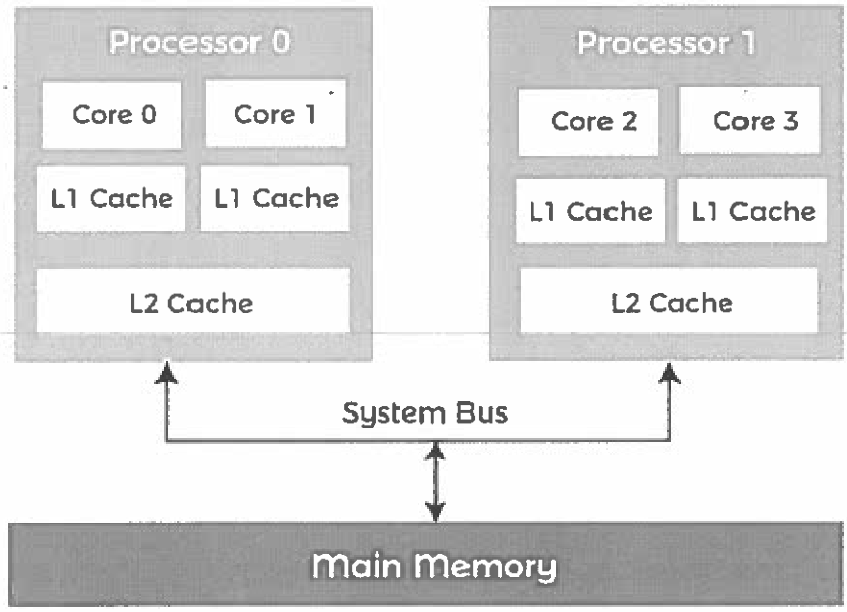
13(a).Explain Multicore architecture of computers. [6M]

A multi-core processor is an integrated circuit with two or more processors connected to it for faster simultaneous processing of several tasks, reduced power consumption, and for greater performance. Generally, it is made up of two or more processors that read and execute program instructions.

In other words, on a single chip, a multi-core processor comprises numerous processing units, or "Cores," each of which has the potential to do distinct tasks. For instance, if you are performing many tasks at once, such as watching a movie and using WhatsApp, one core will handle activities like watching a movie while the other handles other responsibilities like WhatsApp.

Architecture of Multicore Processor

A multi-core processor's design enables the communication between all available cores, and they divide and assign all processing duties appropriately. The processed data from each core is transmitted back to the computer's main board (Motherboard) via a single common gateway once all of the processing operations have been finished. This method beats a single-core CPU in terms of total performance.



13(b).Explain the three generations of multi computers. [6M]

Three Generations of Multicomputer

Design Choices in the Past

While selecting a processor technology, a multicomputer designer chooses low-cost medium grain processors as building blocks. Majority of parallel computers are built with standard off-the-shelf microprocessors. Distributed memory was chosen for multi-computers rather than using shared memory, which would limit the scalability. Each processor has its own local memory unit.

For interconnection scheme, multicomputers have message passing, point-to-point direct networks rather than address switching networks. For control strategy, designer of multi-computers chooses the asynchronous MIMD, MPMD, and SMPD operations. Caltech's Cosmic Cube (Seitz, 1983) is the first of the first-generation multi-computers.

Present and Future Development

The next generation computers evolved from medium to fine grain multicomputers using a globally shared virtual memory. Second generation multi-computers are still in use at present. But using better processor like i386, i860, etc. second-generation computers have developed a lot.

Third generation computers are the next generation computers where VLSI implemented nodes will be used. Each node may have a 14-MIPS processor, 20-Mbytes/s routing channels and 16 Kbytes of RAM integrated on a single chip.

The Intel Paragon System

Previously, homogeneous nodes were used to make hypercube multicomputers, as all the functions were given to the host. So, this limited the I/O bandwidth. Thus to solve large-scale problems efficiently or with high throughput, these computers could not be used.The Intel Paragon System was designed to overcome this difficulty. It turned the multicomputer into an application server with multiuser access in a network environment.

14.Discuss in detail about the instruction level parallelism.[12M]

Instruction Level Parallelism (ILP) is used to refer to the architecture in which multiple operations can be performed parallel in a particular process, with its own set of resources – address space, registers, identifiers, state, program counters. It refers to the compiler design techniques and processors designed to execute operations, like memory load and store, integer addition, float multiplication, in parallel to improve the performance of the processors. Examples of architectures that exploit ILP are VLIWs, Superscalar Architecture.
ILP processors have the same execution hardware as RISC processorsthe machines without ILP have complex hardware which is hard to implement. A typical ILP allows multiple-cycle operations to be pipelined.
Example:
Suppose, 4 operations can be carried out in single clock cycle. So there will be 4 functional units, each attached to one of the operations, branch unit, and common register file in the ILP execution hardware. The sub-operations that can be performed by the functional units are Integer ALU, Integer Multiplication, Floating Point Operations, Load, and Store. Let the respective latencies be 1, 2, 3, 2, and 1.
Let the sequence of instructions be –

1. y1 = x1*1010
2. y2 = x2*1100
3. z1 = y1+0010
4. z2 = y2+0101
5. t1 = t1+1
6. p = q*1000
7. clr = clr+0010
8. r = r+0001

Sequential record of execution vs. Instruction-level Parallel record of execution –

CYCLE	OPERATION
1	y1 = x1*1010
2	nop
3	nop
4	y2 = x2*1100
5	nop
6	nop
7	z1 = y1+0010
8	z2 = y2+0101
9	t1 = t1+1
10	p = q*1000
11	clr = clr+0010
12	r = r+0001

Fig. a

CYCLE	INT ALU	INT ALU	FLOAT ALU	FLOAT ALU
1	t1 = t1+1	clr = clr+0010	y1 = x1*1010	y2 = x2*1100
2	r = r+0001		p = q*1000	
3	nop			
4	z1 = y1+0010	z2 = y2+0101		

Fig. b

The 'nop's or the 'no operations' in the above diagram are used to show idle time of processor. Since latency of floating-point operations is 3, hence multiplications take 3 cycles and processor has to remain idle for that time period. However, in Fig. b processor can utilize those nop's to execute other operations while previous ones are still being executed.

While in sequential execution, each cycle has only one operation being executed, in processor with ILP, cycle 1 has 4 operations, cycle 2 has 2 operations. In cycle 3 there is 'nop' as the next two operations are dependent on first two multiplication operations. The sequential processor takes 12 cycles to execute 8 operations whereas processor with ILP takes only 4 cycles.

Architecture:

Instruction Level Parallelism is achieved when multiple operations are performed in single cycle that is done by either executing them simultaneously or by utilizing gaps between two successive operations that is created due to the latencies.

Now, the decision of when to execute an operation depends largely on the compiler rather than hardware. However, extent of compiler's control depends on type of ILP architecture where information regarding parallelism given by compiler to hardware via program varies. The classification of ILP architectures can be done in the following ways –

1. **Sequential Architecture:**
   Here, program is not expected to explicitly convey any information regarding parallelism to hardware, like superscalar architecture.
2. **Dependence Architectures:**
   Here, program explicitly mentions information regarding dependencies between operations like dataflow architecture.
3. **Independence Architecture:**
   Here, program gives information regarding which operations are independent of each other so that they can be executed instead of the 'nop's.

In order to apply ILP, compiler and hardware must determine data dependencies, independent operations, and scheduling of these independent operations, assignment of functional unit, and register to store data.

OR

15.Illustrate the following[12M]

1. **Operand Forwarding**
2. **Branch Prediction**

Operand forwarding (or **data forwarding**) is an optimization in pipelinedCPUsto limit performance deficits which occur due to pipeline stallsA data hazardcan lead to a pipeline stall when the current operation has to wait for the results of an earlier operation which has not yet finished.

Example

ADD A B C#A=B+C
SUB D C A#D=C-A

If these two assembly pseudocode instructions run in a pipeline, after fetching and decoding the second instruction, the pipeline stalls, waiting until the result of the addition is written and read.

Without operand forwarding

1	2	3	4	5	6	7	8
Fetch ADD	Decode ADD	Read Operands ADD	Execute ADD	Write result			
	Fetch SUB	Decode SUB	*stall*	*stall*	Read Operands SUB	Execute SUB	Write result

With operand forwarding

1	2	3	4	5	6	7
Fetch ADD	Decode ADD	Read Operands ADD	Execute ADD	Write result		
	Fetch SUB	Decode SUB	*stall*	Read Operands SUB: use result from previous operation	Execute SUB	Write result

In some cases all stalls from such read-after-write data hazards can be completely eliminated by operand forwarding:

1	2	3	4	5
Fetch ADD	Decode ADD	Read Operands ADD	Execute ADD	Write result
	Fetch SUB	Decode SUB	Read Operands SUB: use result from previous	Execute SUB

2. Branch Prediction

Conditional Branches present in the programs significantly affect the performance of the system. So we need to come up with efficient branch prediction mechanism so as to get the branch target address with high accuracy and thus minimizing the stalls associated with control hazards.

In case if failure in correctly predicting target address, penalty will occur in terms of flushing the pipeline and bringing back the processors to a state that was there earlier when it was executing branch instruction.

Types of Branch Prediction Technique –

Branch prediction technique can be of two types:

Static Branch Prediction Technique

Dynamic Branch Prediction Technique

These are explained as following below.

1. Static Branch Prediction Technique:

In case of Static branch prediction technique underlying hardware assumes that either the branch is not taken always or the branch is taken always.

2. Dynamic Branch Prediction Technique:
In Dynamic branch prediction technique prediction by underlying hardware is not fixed, rather it changes dynamically.This technique has high accuracy than static technique.

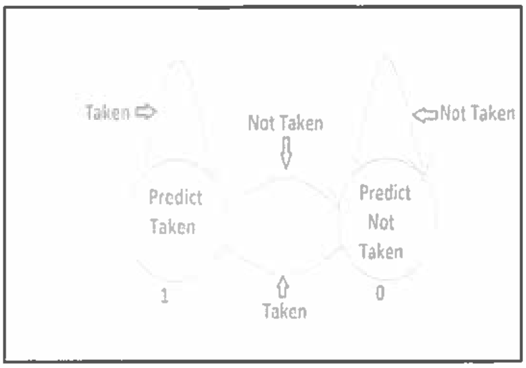
Some dynamic branch prediction techniques are:

1. 1-bit branch prediction technique
2. 2-bit branch prediction technique

3. Correlating branch prediction technique

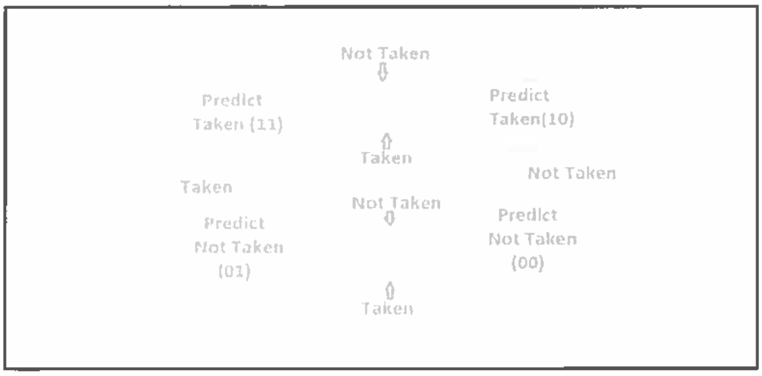
These are explained as following below.

- **1-bitBranch Prediction Technique –**
  In this technique hardware changes its assumption just after one false assumption.For example if hardware assumes branch to be taken but actually branch is not taken, then in next step hardware assumes branch to be not taken and vice-versa.
  1-bit branch prediction machine is shown in the fig below:



- **2-bit Branch Predictor –**
  In this technique the underlying hardware does not changes its assumption just after one incorrect assumption, rather it changes its assumption after two consecutive wrong assumption and vice-versa.

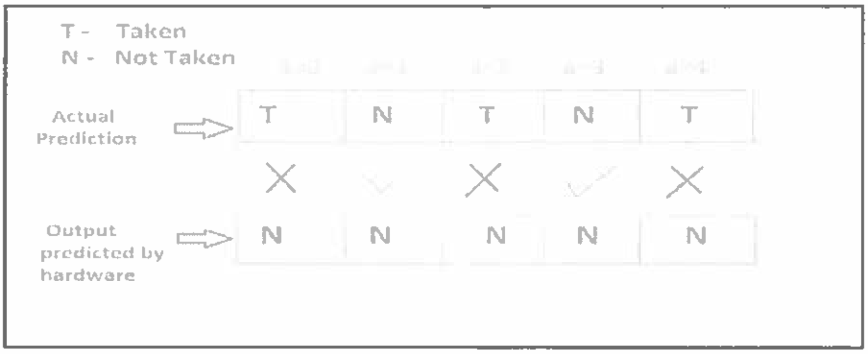
2-bit branch prediction machine is shown in the figure:



Explanation –

1. Let's say when a=0 everything is reset (00) and so hardware assume branch not to be taken and branch is taken. So current state is (01)
2. When a=1, hardware assumes branch not to be taken and branch is not taken. So current state is (00)
3. When a=2, hardware assumes branch not to be taken and branch is taken.So current State is (01)
4. When a=3, hardware assumes branch not to be taken and branch is not taken. So current State is (10)

5. When a=4, hardware assumes branch not to be taken and branch is taken.So current State is(00)



Correlating Branch Prediction–

We cannot get significant accuracy from 2-bit branch predictor also due to interference with other branches.So correlating branch prediction comes into picture which is also known as two-level branch predictor in which prediction accuracy is improved as it takes into consideration the recent behavior of other branches also.

Information Source –

- It uses k least significant bits of Branch Target Address which is fetched before.
- It also uses Local History Table (LHH) which is table of shift registers where shift register refers to the last outcome of m branches having same k least significant bits.
- It also uses Local Prediction Table to predict the outcome depending on the state in which it is present.

NSRIT

Semester End Regular Examination, June , 2022

Degree	B. Tech. (U. G.)	**Program**	ECE (Honors)			**Academic Year**	2021 - 2022
Course Code	20ECH01	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	IV
Course	**Low Power VLSI Design**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Give the need for Low Power design In VLSI systems?	20ECH01.1	L1
2	What is Constant voltage scaling?	20ECH01.2	L1
3	Differentiate static power and Dynamic Power in VLSI circuits?	20ECH01.3	L1
4	Draw the carry equation of adder using CMOS logic?	20ECH01.4	L3
5	List the different multiplier architectures?	20ECH01.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	What is switching power dissipation? Explain it with a CMOS Inverter.	8M	20ECH01.1	L2
6 (b)	Define short circuit Power dissipation.	4M	20ECH01.1	L2
	OR			
7	Explain the leakage and glitching power dissipation in a CMOS inverter.	12M	20ECH01.1	L2
8	Explain the MT CMOS technique.	12M	20ECH01.2	L2
	OR			
9	Explain the role of parallel and pipeline processing in Architectural low power design.	12M	20ECH01.2	L2
10	Discuss the various power reduction techniques used in Gate level design.	12M	20ECH01.3	L2
	OR			
11 (a)	Explain the types of Parasitic capacitance in detail.	10M	20ECH01.3	L2
11 (b)	Give the formulae for capacitive power dissipation.	2M	20ECH01.3	L2
12	Compare Ripple carry Adder and Carry look ahead adder for a 4 bit input.	12M	20ECH01.4	L2
	OR			
13	Draw the architecture of 16 bit Carry select Adder and explain the reasons for its low power consumption when compared to other Adders.	12M	20ECH01.4	L2
14	Explain the working of Braun Multiplier with its structure.	12M	20ECH01.5	L2
	OR			
15	Explain about the Booth Multiplier and draw its VLSI Structure.	12M	20ECH01.5	L2

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(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

1	Circuit diagram of small signal high frequency CE model of a transistor and list its elements.	2M
2	Expression for current gain for Darlington pair.	2M
3	Identification of three advantages of negative feedback amplifier.	2M
4	List of four types of oscillators	2M
5	Identify of factors that influences on the selectivity of a single tuned amplifier.	2M
6 (a)	Circuit diagram Derivation of expression for the CE short circuit current gain A_I	3M 6M
6 (b)	Circuit diagram for Hybrid - π model for a transistor in the CB configuration	3M
7 (a)	Statement for Miller's theorem Explanation of Miller's theorem.	3M 5M
7 (b)	Circuit diagram for Hybrid - π model for a transistor in the CE configuration	4M
8 (a)	Three circuit diagrams Explanation of three types of coupling methods used in multistage amplifiers	3M 5M
8 (b)	Circuit diagram of cascade (Two stage RC coupled) amplifier with Without biasing circuit Advantages.	3M 1M
9	Circuit diagram and explanation of Darlington emitter follower i. current gain ii. input impedance iii. voltage gain iv. output impedance Comparison with emitter follower.	2M 2M 2M 2M 2M 2M
10	Circuit for voltage shunt feedback amplifier Justification of the type of feedback Derivation for A_V and β Expression for input and output resistance with feedback	2M 1M 6M 3M
11	Circuit for Voltage series feedback amplifier derivation for A_f and β Advantages	3M 7M 2M
12	Circuit diagram for a FET based RC Phase shift oscillator Derivation for frequency of oscillation condition for sustained oscillations	2M 7M 3M
13	Circuit diagrams of Hartley oscillator using bipolar junction transistor	3M 3M

	Operation of Hartley oscillator Derivation for frequency of oscillations	6M
14 (a)	Circuit diagram Operation of class B push pull amplifier Explanation of crossover distortion minimization	3M 4M 2M
14 (b)	Identification of effects of Harmonic distortions in power amplifiers.	3M
15 (a)	Circuit diagram of cascade (staggered) tuned amplifier Explanation	3M 5M
15 (b)	Four features of single tuned amplifier	4M

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

Degree	B. Tech. (U. G.)	Program	ECE (Honors)	Academic Year	2021 - 2022
Course Code	20ECH01	Test Duration	3 Hrs. Max. Marks 70	Semester	IV
Course	Low Power VLSI Design				

ANSWER KEY

Part A (Short Answer Questions 5 x 2 = 10 Marks)

1 Give the need for Low Power design In VLSI systems?

Answer: The need for low-power design is also becoming a major issue in high-performance digital systems, such as microprocessors, digital signal processors (DSPs) and other applications. Increasing chip density and higher operating speed lead to the design of very complex chips with high clock frequencies

2 What is Constant voltage scaling?

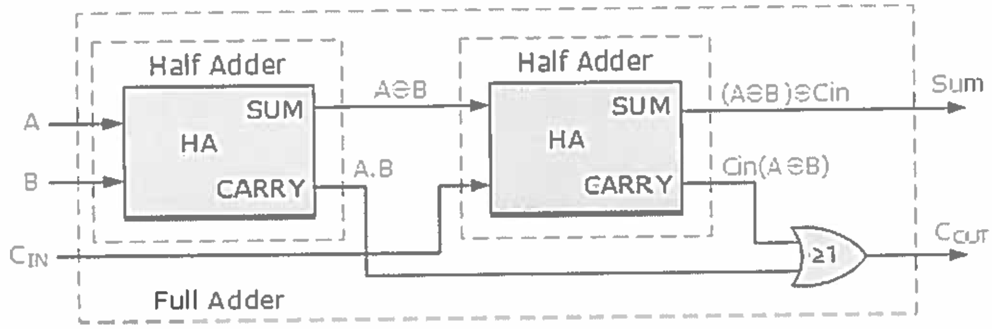
Answer: In constant voltage scaling, VDD is kept constant, and the process is scaled. For constant field scaling, the device dimensions are scaled by the parameter λ. The most important point in this scaling is the supply voltage is scaled but the electric field remains constant hence the same constant field scaling is given.

3 Differentiate static power and Dynamic Power in VLSI circuits?

Answer: Power dissipation in CMOS circuits arises from two different mechanisms: static power, which is primarily leakage power and is caused by the transistor not completely turning off, and dynamic power, which is largely the result of switching capacitive loads between two different voltage state

4 Draw the carry equation of adder using CMOS logic?

Answer: Boolean expression for a full adder is as follows. For the CARRY-OUT (Cout) bit: CARRY-OUT = A AND B OR Cin(A XOR B) = A.B + Cin(A)



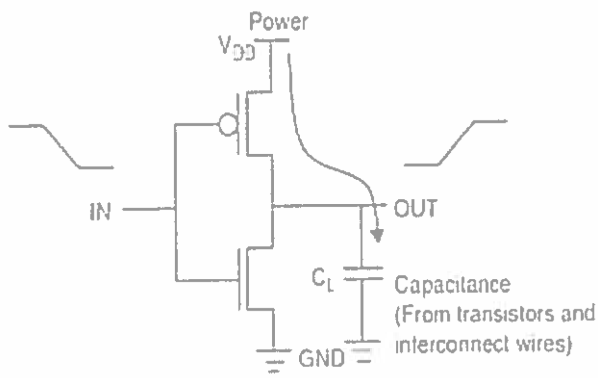
5 **List the different multiplier architectures?**

Answer: the three multipliers architecture are array multiplier, a column bypass multiplier, and a array multiplier using Reversal Logic schemes.

Part B (Long Answer Questions 5 x 12 = 60 Marks)

6 (a) what is switching power dissipation? Explain it with a CMOS Inverter.

Answer: The first and primary source of dynamic power consumption is the Switching power dissipation occurs due the power required to charging and discharging of the output capacitance on a gate. Figure illustrates switching power for charging a capacitor



The energy per transition is given by

$$Energy/Transition = \frac{1}{2} \times C_L \times V_{dd}^2$$

Where CL is the load capacitance and Vdd is the supply voltage

Switching power is therefore expressed as:

$$P_{switch} = \frac{Energy}{Transition} \times f = C_L \times V_{dd}^2 \times P_{trans} \times f_{clock}$$

Where f is the frequency of transitions, Ptrans is the probability of an output transition

and f clock is the frequency of the system clock

In addition to the switching power dissipation for charging and discharging the load

Capacitance, switching power dissipation also occurs for charging and discharging of the

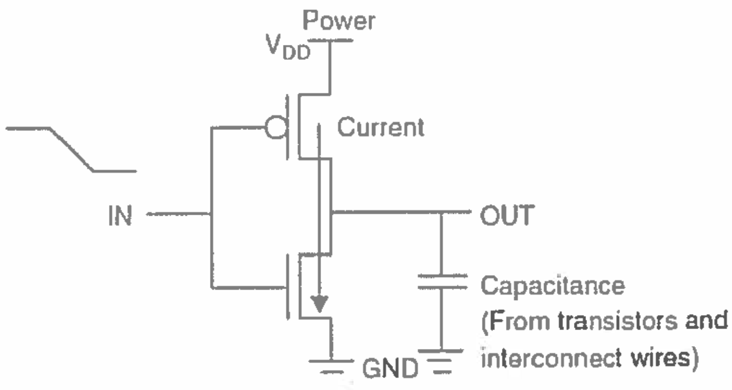
Internal node capacitance. Thus, total switching power dissipation is given by

$$P_{totalswitch} = C_L \times V_{dd}^2 \times P_{trans} \times f_{clock} + \sum \alpha_i \times C_i \times V_{dd} \times (V_{dd} - V_{th}) \times f_{clock}$$

Where αi and Ci are the transition probability and capacitance, respectively, for an internal node i

6 (b) Define short circuit Power dissipation.

Answer: short-circuit currents. Short-circuit currents occur when both the negative metal–oxide–semiconductor (NMOS) and positive metal–oxide–semiconductor (PMOS) transistors are ON. Let Vtn be the threshold voltage of the NMOS transistor and Vtp is the threshold voltage of the PMOS transistor. Then, in the period when the voltage value is between Vtn and Vdd–Vtp, while the input is switching either from 1 to 0 or vice versa, both the PMOS and the NMOS transistors remain ON, and the short-circuit current follows from Vdd to ground (GND)



The expression for short-circuit power is given by $\boldsymbol{Pshortcircuit = tsc \times Vdd \times Ipeak \times fclock = \mu\varepsilon oxW\ 12LD \times (Vdd - Vth)\ 3 \times tsc \times fclock}$ (1.4) ✓ Where tsc is the rise/fall time duration of the short-circuit current ✓ Ipeak is the total internal switching current (short-circuit current plus the current to charge the internal capacitance) ✓ μ is the mobility of the charge carrier ✓ εox is the permittivity of the silicon dioxide ✓ W is the width ✓ L is the length ✓ D is the thickness of the silicon dioxide From the above equation it is evident that the short-circuit power dissipation depends on the supply voltage, rise/fall time of the input and the clock frequency apart from the physical parameters. So the short-circuit power can be kept low if the ramp (rise/fall) time of the input signal is short for each transition. Then the overall dynamic power is determined by the switching power.

7. Explain the leakage and glitching power dissipation in a CMOS inverter.

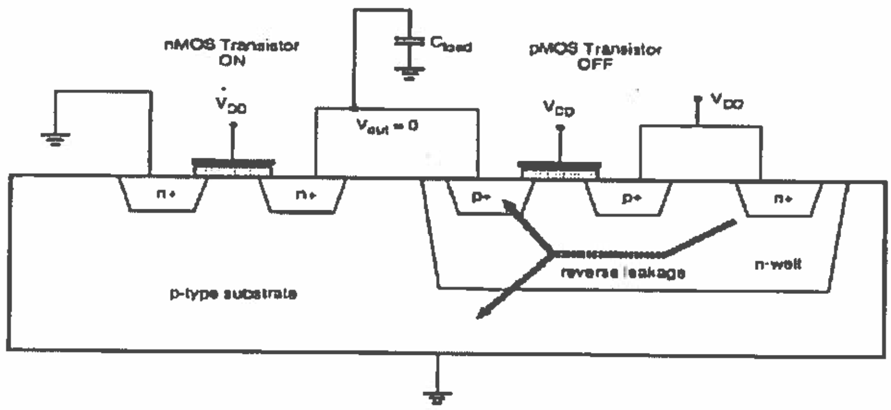
Answer: Leakage Power Dissipation The nMOS and pMOS transistors used in a CMOS logic gate generally have nonzero reverse leakage and sub threshold currents. In a CMOS VLSI chip containing a very large number of transistors, these currents can contribute to the overall power dissipation even when the transistors are not undergoing any switching event. The magnitude of the leakage currents is determined mainly by the processing parameters.

Two main leakage current components found in a MOSFET are

(1) Reverse diode leakage current

(2) Sub threshold leakage current Reverse diode leakage current:

The reverse diode leakage occurs when the pn-junction between the drain and the bulk of the transistor is reversely biased. The reverse-biased drain junction then conducts a reverse saturation current which is eventually drawn from the power supply. Consider a CMOS inverter with a high input voltage, where the nMOS transistor is turned on and the output node voltage is discharged to zero. Although the pMOS transistor is turned off, there will be a reverse potential difference of VDD between its drain and the n-well, causing a diode leakage through the drain junction. The n-well region of the pMOS transistor is also reverse-biased with VDD, with respect to the p-type substrate. Therefore, another significant leakage current component exists due to the n-well junction

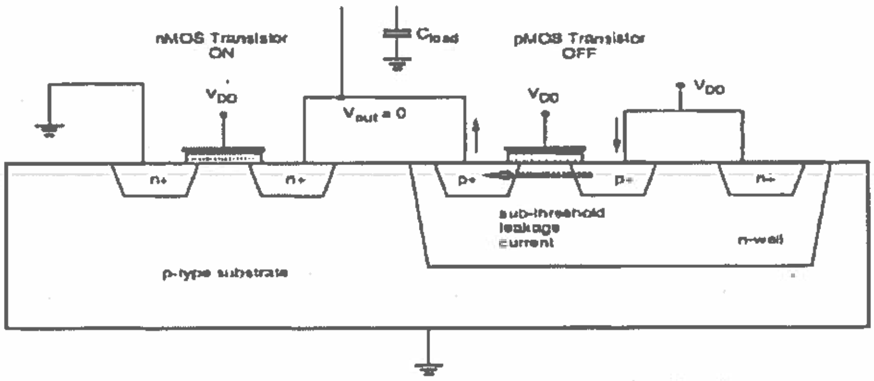


A similar situation can be observed when the input voltage is equal to zero, and the output voltage is charged up to VDD through the pMOS transistor. Then, the reverse potential difference between the nMOS drain region and the p-type substrate causes a reverse leakage current which is also drawn from the power supply (through the pMOS transistor). The magnitude of the reverse leakage current of a pn-junction is given by the following expression

$$I_{reverse} = A \cdot J_S \left(e^{\frac{qV_{bias}}{kT}} - 1 \right)$$

Where Vbias is the magnitude of the reverse bias voltage across the junction, JS is the reverse saturation current density and the A is the junction area. The typical magnitude of the reverse saturation current density is 1 - 5 pA/mm2 , and it increases quite significantly with temperature. Note that the reverse leakage occurs even during the stand-by operation when no switching takes place. Hence, the power dissipation due to this mechanism can be significant in a large chip containing several million transistors

Sub threshold leakage current: Another component of leakage currents which occur in CMOS circuits is the subthreshold current, which is due to carrier diffusion between the source and the drain region of the transistor in weak inversion. An MOS transistor in the subthreshold operating region behaves similar to a bipolar device and the sub threshold current exhibits an exponential dependence on the gate voltage. The amount of the sub threshold current may become significant when the gateto source voltage is smaller than, but very close to the threshold voltage of the device. In this case, the power dissipation due to sub threshold leakage can become comparable in magnitude to the switching power dissipation of the circuit. The sub threshold leakage current is shown in Fig. below



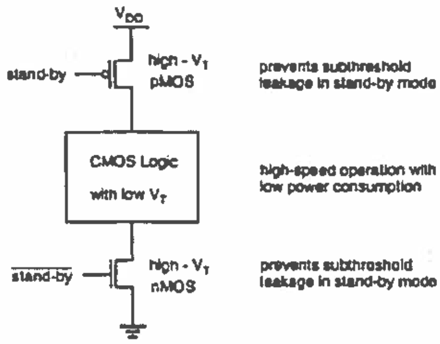
Subthreshold leakage current path in a CMOS inverter with high input voltage.

Note the sub threshold leakage current also occurs when there is no switching activity in the circuit, and this component must be carefully considered for estimating the total power dissipation in the stand-by operation mode. The sub threshold current expression is given below, in order to illustrate the exponential dependence of the current on terminal voltages.

$$I_D(subthreshold) \cong \frac{qD_n W x_c n_0}{L} \cdot e^{\frac{q\phi_r}{kT}} \cdot e^{\frac{q}{kT}(A \cdot V_{GS} + BV_{DS})}$$

8 Explain the MT CMOS technique.

Answer : Another technique which can be applied for reducing leakage currents in low voltage circuits in the stand-by mode is based on using two different types of transistors (both n-MOS and p-MOS) with two different threshold voltages in the circuit. Here, low-VT transistors are typically used to design the logic gates where switching speed is essential, whereas high- VT transistors are used to effectively isolate the logic gates in stand-by and to prevent leakage dissipation. The generic circuit structure of the MTCMOS logic gate is shown

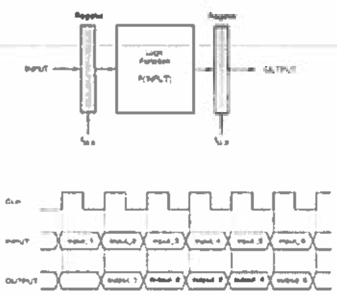


In the active mode, the high-VT transistors are turned on and the logic gates consisting of low-VT transistors can operate with low switching power dissipation and small propagation delay. When the circuit is driven into stand-by mode, on the other hand, the high-VT transistors are turned off and the conduction paths for any sub-threshold leakage currents that may originate from the internal low-VT circuitry are effectively cut off. Figure shows a simple D-latch circuit designed with the MTCMOS technique. The critical signal propagation path from the input to the output consists exclusively of low- VT transistors, while a cross-coupled inverter pair consisting of highVT transistors is used for preserving the data in the stand-by mode

The MTCMOS technique is conceptually easier to apply and to use compared to the VTCMOS technique, which usually requires a sophisticated substrate bias control mechanism. It does not require a twin-well or triple-well CMOS process; the only significant process-related overhead of MTCMOS circuits is the fabrication of MOS transistors with different threshold voltages on the same chip. One of the disadvantages of the MTCMOS circuit technique is the presence of series-connected stand-by transistors, which increase the overall circuit area and also add extra parasitic capacitance. While the VTCMOS and MTCMOS circuit techniques can be very effective in designing low-power/low-voltage logic gates, they may not be used as a universal solution to low-power CMOS logic design. In certain types of applications where variable threshold voltages and multiple threshold voltages are infeasible due to technological limitations, system-level architectural measures such as pipelining and hardware replication techniques offer feasible alternatives for maintaining the system performance (throughput) despite voltage scaling.

8. Explain the role of parallel and pipeline processing in Architectural low power design.

Answer: First, consider the single functional block shown in Fig. which implements a logic function F(INPUT) of the input vector, INPUT. Both the input and the output vectors are sampled through register arrays, driven by a clock signal CLK. Assume that the critical path in this logic block (at a power supply voltage of VDD) allows a maximum sampling frequency off CLK; in other words, the maximum input-to-output propagation delay pmax of this logic block is equal to or less than TCLK = lfCLK. Figure shows a simplified timing diagram of the circuit. A new input vector is latched into the input register array at each clock cycle, and the output data becomes valid with a latency of one cycle.



Single-stage implementation of a logic function and its simplified timing diagram.

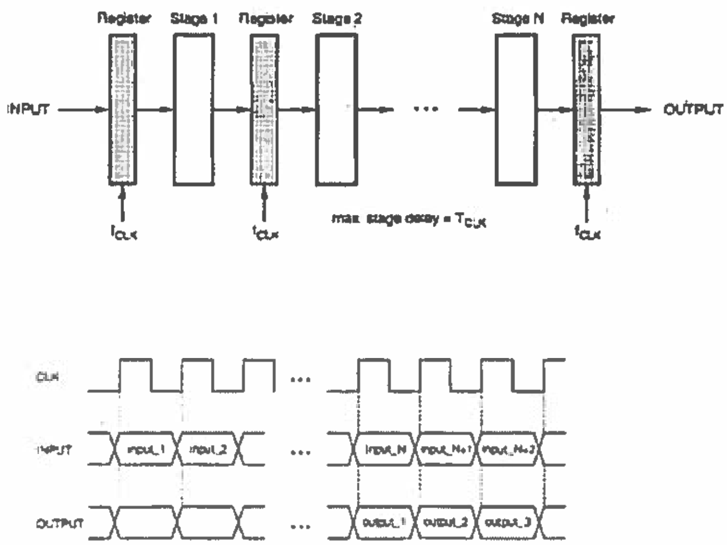
Let C total be the total capacitance switched every clock cycle. Here, C total, consists of (i)the capacitance switched in the input register array, (ii) the capacitance switched to implement the logic function, and (iii) the capacitance switched in the output register array. Then, the dynamic power consumption of this structure can be found as

$$P_{reference} = C_{total} \cdot V_{DD}^{2} \cdot f_{CLK}$$

The logic function F(INPUT) has been partitioned into N successive stages, and a total of (N- 1) register arrays have been introduced, in addition to the original input and output registers, to create the pipeline. All registers are clocked at the original sample rate, *fCLK*. If all stages of the partitioned function have approximately equal delays of

$$\tau_P(pipeline\_stage) = \frac{\tau_{P,max}(input - to - output)}{N} = T_{CLK}$$

Then the logic blocks between two successive registers can operate N-times slower while *maintaining the same functional throughput* as before. This implies that the power supply



N-stage pipeline structure realizing the same logic function as shown in Fig. The maximum pipeline stage delay is equal to the clock period, and the latency is N clock cycles.

The dynamic power consumption of the N-stage pipelined structure with a lower supply voltage and with the same functional throughput as the single-stage structure can be approximated by

$$P_{pipeline} = \left[C_{total} + (N-1)C_{reg}\right] \cdot V_{DD,new}^{2} \cdot f_{CLK}$$

10. Discuss the various power reduction techniques used in Gate level design.

- Answer: The most popular gate-level analysis is based on the so called event-driven logic simulation Events are zero-one logic switching of nets in a circuit at a particular simulation time point. As one switching event occurs at the input of a logic gate, it may trigger other events at the output of the gate after a specified time delay.

- Computer simulation of such events provides a very accurate pre-fabrication logic analysis and verification of digital chips. Most gate-level simulation also supports other logic states such as, "unknown," "don't care" and "high-impedance," to help the designer to simulate the circuit in a more realistic manner.

- Some simulators offer an extensive set of logic states for added accuracy in timing analysis. Verilog and VHDL are two popular languages used to describe gate-level design. Recently, the cycle-based simulators are being introduced into the design community. Such simulators assume that circuits are driven by synchronous master clock signals. Instead of scheduling events at arbitrary time points

- Certain nets of the circuit are only allowed a handful of events at a given clock cycle. This reduces the number of events to be simulated and results in more efficient analysis Many gate-level simulators are so mature that special purpose computer hardware has been used to speed up the simulation algorithms

- The idea is similar to the graphic coprocessor in a computer system. Instead of using a general purpose CPU to execute the simulation program, special purpose hardware optimized for logic simulation is used This hardware acceleration technology generally results in several factors of speedup compared to using a general purpose computing system

- Another technology that offers several orders of magnitude speedup in gate-level analysis is called "**hardware emulation** "Instead of simulating switching events using software programs, the logic network is partitioned into smaller manageable sub blocks.

- The Boolean function of each sub-block is extracted and implemented with a hardware table mapping mechanism such as RAM or FPGA. A reconfigurable interconnection network, carrying the logic signals, binds the sub-blocks together. Circuits up to a million gates can be emulated with this technology but this is also the most expensive type of logic simulator to operate and maintain because of the sophisticated high-speed hardware required. The simulation speed is only one to two orders of magnitude slower than the actual VLSI chips to be fabricated. For example, a 200MHz CPU can be emulated with a 2MHz clock rate, permitting moderate real-time simulation.

11 (a) Explain the types of Parasitic capacitance in detail.

Answer: Capacitance is the most important physical attribute that affects the power dissipation of CMOS circuits. Capacitance also has a direct impact on delays and signal slopes of logic gates. Changes in gate delays may affect the switching characteristics of the circuit and influence power dissipation. Short-circuit current is affected by the input signal slopes and output capacitance loading (see Section 1.3). Thus, capacitance has a direct and indirect impact on power analysis. The accurate estimation of capacitance is important for power analysis and optimization. Two types of parasitic capacitance exist in CMOS circuits: 1. device parasitic capacitance; 2. wiring capacitance

The parasitic capacitance of MOS devices can be associated with their terminals. The gate capacitance is heavily dependent on the oxide thickness of the gate that is process dependent. The design dependent factors are the width, length and the shape of the gate. Typically, the shape of a transistor gate is rectangular and the width and length of the gate determine its capacitance. For a gate that "bends," e.g., L-shaped, a correction factor can be used to find its equivalent rectangular width and length. The source and drain capacitance is also estimated from a similar method. The primary capacitance contribution of source and drain terminals is the area and shape of the diffusion regions. In general, a larger transistor has more capacitance in all of its terminals.

11 (b) Give the formulae for capacitive power dissipation.

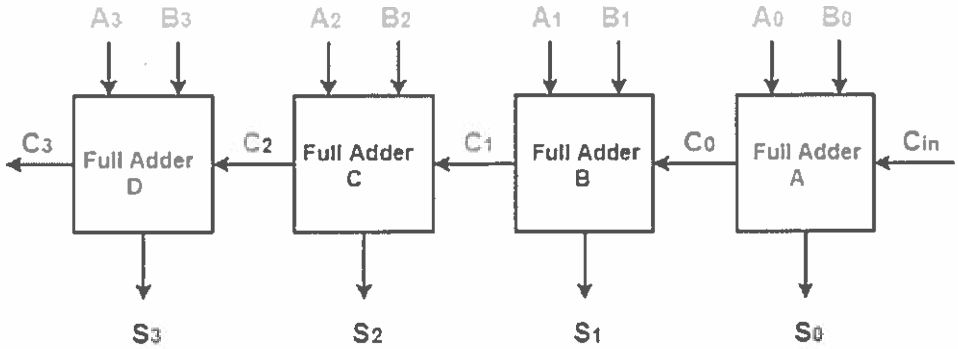
Answer: n the cell-based design environment, the design and layout of the library cells are made available before the chip design. The capacitance of each pin of a cell is therefore fixed by its circuit and layout. The pin capacitance of a cell can be accurately measured and stored in the cell library. One way to measure the pin capacitance is to use SPICE circuit simulation with the help of the capacitor I-V equation $i = Cdv/dt$

We vary the pin voltage ~ V of the cell in time ~T and observe the current i to obtain the capacitance C. This measurement can be performed during the characterization of the cell. The second source of parasitic capacitance is wiring capacitance. Wiring capacitance depends on the layer, area and shape of the wire. Typically, the width of routing wires is set to the minimum and the wiring capacitance is estimated from the lengths of the wires. In practice, the process dependent factors of wiring capacitance are expressed by a capacitance-per-unit-length. parameter that depends on the thickness of the wire, its distance from the substrate and its width. Once the length of a wire is known, wiring capacitance can be computed. Since wiring capacitance depends on the placement and routing of the gate-level netlist, accurate estimation cannot be obtained before the physical design phase

12 Compare Ripple carry Adder and Carry look ahead adder for a 4 bit input.

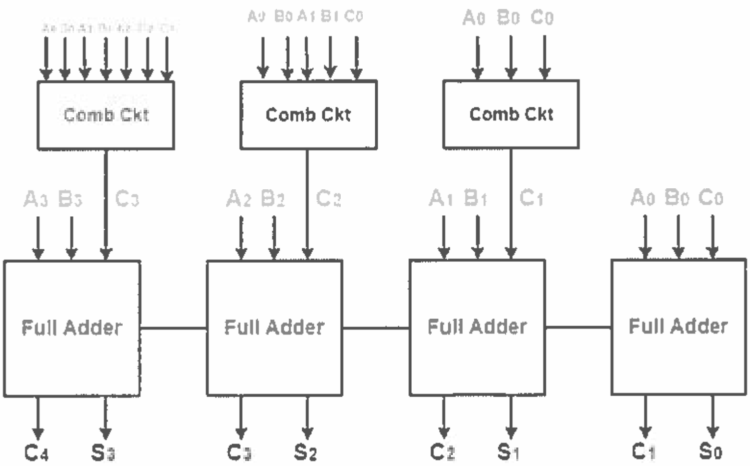
Answer: In Ripple Carry Adder,

- Each full adder has to wait for its carry-in from its previous stage full adder.
- Thus, n^{th} full adder has to wait until all (n-1) full adders have completed their operations.
- This causes a delay and makes ripple carry adder extremely slow.
- The situation becomes worst when the value of n becomes very large.
  - To overcome this disadvantage, Carry Look Ahead Adder comes into play.



4-bit Ripple Carry Adder

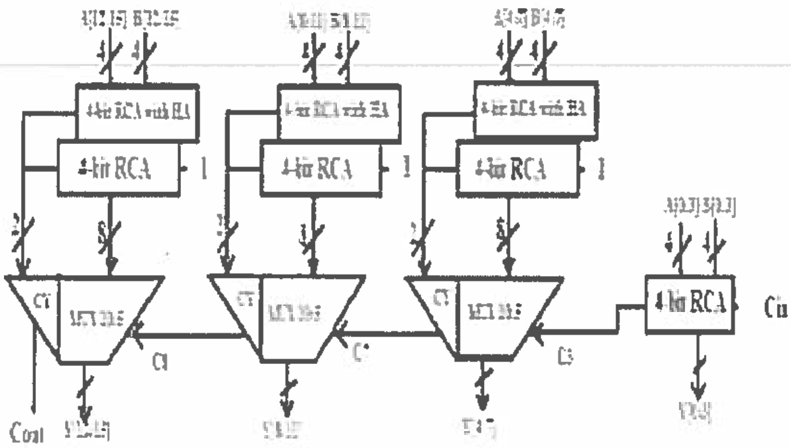
- **Carry look ahead adder**
- Carry Look Ahead Adder is an improved version of the ripple carry adder.
- It generates the carry-in of each full adder simultaneously without causing any delay.
- The time complexity of carry look ahead adder = Θ (logn).



13. Draw the architecture of 16 bit Carry select Adder and explain the reasons for its low power consumption when compared to other Adders

Answer: A carry select adder is a particular way to implement an adder, which is a logic element that computes the (n+1)-bit sum of two n-bit numbers. The carry select adder is simple but rather fast

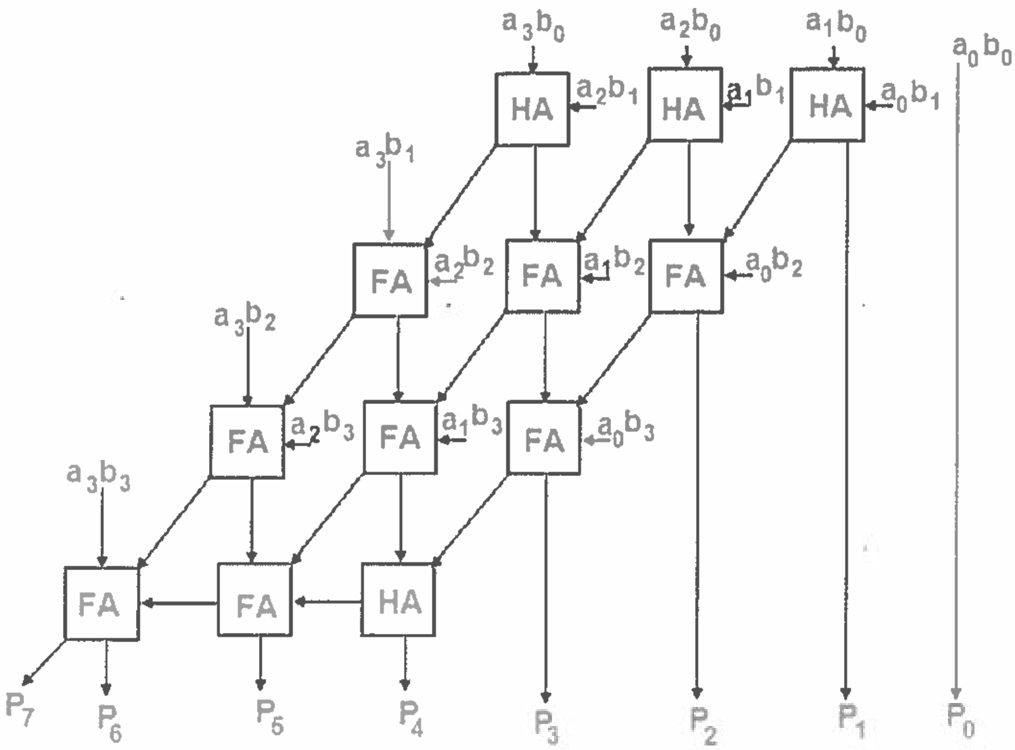
A carry-select adder performs two additions in parallel, one assuming a Cin= 0, the other a Cin= 1. The speed of the Carry Select Adder is improved by predicting the carry input and performs the addition. Carry select adder presume the carry as „0" and „1" and calculate carry and sum. Ultimate result is determined by selecting accurate carry by Multiplexer. CSL approach reduces "addition time" to "addition selection time" for the higher stages. The internal logic schematic of a carry select adder constructed using the conventional 4-bit ripple carry adder (RCA) shown in fig.. The RCA uses multiple full adders and half adder to perform addition operation. Although Half adder is used only at very first stage (LSB) of the Binary adder, where Cin=0. Each full adder inputs a carryin, which is the carry-out of the preceding adder. The CSLA divides the bits to be added into blocks and forms two sums for each block in parallel, one with assumed carry in (Cin) of 0 and the other with Cin of 1. As shown in Fig. the carry-out from one stage of 4-bit RCA is used as the select signal for the multiplexer. This selects the corresponding sum bit from the next block of data. This speeds-up the computation process of the Binary adder. Thus, the Carry Select Adder achieves higher speed of operation at the cost of increased number of devices used in the circuit. This in turn increases the area and cost of the circuits.



14 Explain the working of Braun Multiplier with its structure.

Answer: Braun multiplier is a **type of parallel array multiplier**. The architecture of Braun multiplier mainly consists of some Carry Save Adders, array of AND gates and one Ripple Carry Adder.

In Braun multiplier, the partial products are first computed in parallel, then collected through a cascade of different types of adders. It consists of an array of AND gates and adders arranged in an iterative structure which doesn't require any logic registers.



4x4 Row bypassing multiplier The Row bypassing multiplier reduces the switching activity by bypassing the row in which the multiplicand bit is zero. That means in the multiplier if a bit is zero then that row of adders will get disabled.

15 Explain about the Booth Multiplier and draw its VLSI Structure.

Answer: the Booth multiplier algorithm is used for **multiplication of both signed as well as unsigned binary values in 2's complement form**. This algorithm is introduced by Andrew Donald Booth in the 1950s. A multiplier shows great efficiency in area, power consumption and scalability

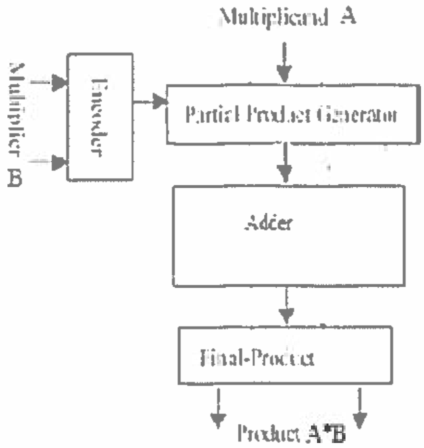
It is a powerful algorithm for signed-number multiplication, which treats both positive and negative numbers uniformly For the standard add-shift operation; each multiplier bit generates one multiple of the multiplicand to be added to the partial product. If the multiplier is very large, then a large number of multiplicands have to be added. In this case the delay of multiplier is determined mainly by the number of additions to be performed. If there is a way to reduce the number of the additions, the performance will get better. Booth algorithm is a method that will reduce the number of multiplicand multiples. For a given range of numbers to be represented, a higher representation radix leads to fewer digits. Since a k-bit binary number can be interpreted as K/2-digit radix-4 number, a K/3-digit radix-8 number, and so on, it can deal with more than one bit of the multiplier in each cycle by using high radix multiplication. This is shown for Radix-4 in the example below

Multiplicand	A =	● ● ● ●	
Multiplier	B =	(●●)(●●)	
Partial product bits		● ● ● ●	$(B_1B_0)_2 A4^0$
		● ● ● ●	$(B_3B_2)_2 A4^1$
Product	P =	● ● ● ● ● ● ● ●	

Radix-4 multiplication in dot notation.

as shown in the figure above, if multiplication is done in radix 4, in each step, the partial product term (Bi+1Bi)2 A needs to be formed and added to the cumulative partial product. Whereas in radix-2 multiplication, each row of dots in the partial products matrix represents 0 or a shifted version of A must be included and added

The 'multiplier' is successfully shifted and gates the appropriate bit of the 'multiplicand'. They are then added using the different technique of adder to form the product bit for the particular form Fig 1 shows the structure of modified booth multiplier. Functionality of each block described



Multiplicand A
Multiplier B
Encoder
Partial Product Generator
Adder
Final-Product
Product A*B

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Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	**Program**	CSE (AI & ML &DS) - Honors		**Academic Year**	2021 - 2022
Course Code	20DSH01	**Test Duration**	3 Hrs.	Max. Marks 70	**Semester**	IV
Course	**Text Analytics**					

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What is natural language processing?	20DSH01.1	L1
2	List the applications of multi class classification system.	20DSH01.2	L1
3	Recall latent Dirichlet allocation.	20DSH01.3	L1
4	Compare Manhattan and Euclidean distance.	20DSH01.4	L2
5	Distinguish between supervised and unsupervised learning.	20DSH01.5	L2

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Interpret the language syntax and structure in natural language.	6M	20DSH01.1	L2
6 (b)	Illustrate language semantics with suitable example.	6M	20DSH01.1	L2
	OR			
7 (a)	Explain first order logic with an example.	6M	20DSH01.1	L2
7 (b)	Summarize the steps involved in text categorization and text analytics.	6M	20DSH01.1	L2
8 (a)	Analyze support vector machine for automated text classification.	6M	20DSH01.2	L3
8 (b)	Explain the text classification with building process of automated text classification.	6M	20DSH01.2	L3
	OR			
9 (a)	Interpret TF-IDF mode in text classification.	6M	20DSH01.2	L3
9 (b)	Compare the performance and features of dependency based parsing and constituency based parsing.	6M	20DSH01.2	L3
10 (a)	What is text normalization? List the features to be extracted for text normalization.	6M	20DSH01.3	L2
10 (b)	Illustrate weighted tag based phrase extraction with an example.	6M	20DSH01.3	L2
	OR			
11 (a)	Explain the process of automated document summarization.	6M	20DSH01.3	L2
11 (b)	Explain unsupervised learning techniques in text summarization.	6M	20DSH01.3	L2
	be			
12 (a)	Elaborate the process of feature extraction in text summarization.	6M	20DSH01.4	L3
12 (b)	Compare the performance of the Manhattan and Euclidean distance similarity measures.	6M	20DSH01.4	L3
	OR			
13 (a)	Analyze K-means clustering algorithm for document clustering using suitable dataset.	6M	20DSH01.4	L3
13 (b)	Explain about Ward's agglomerative hierarchical clustering with suitable dataset.	6M	20DSH01.4	L3
14 (a)	Analyze lexical semantic relations with necessary illustrations.	6M	20DSH01.5	L3
14 (b)	Distinguish between first order and propositional logics for semantic analysis.	6M	20DSH01.5	L3
	OR			
15 (a)	Examine named entity recognition for semantic analysis.	6M	20DSH01.5	L3
15 (b)	Compare the performance of unsupervised sentiment analysis models.	6M	20DSH01.5	L3

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(AUTONOMOUS)
SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

COURSE: TEXT ANALYTICS
COURSE CODE: 20DSH01

2MARKS

1. **What is natural language processing?**

NLP stands for Natural Language Processing, which is a part of Computer Science, Human language, and Artificial Intelligence. It is the technology that is used by machines to understand, analyse, manipulate, and interpret human's languages. It helps developers to organize knowledge for performing tasks such as translation, automatic summarization, Named Entity Recognition (NER), speech recognition, relationship extraction, and topic segmentation

2. **List the applications of multi class classification system.**

speech recognition, handwriting recognition, biometric identification, document classification.

3. **Recall latent Dirichlet allocation.**

In natural language processing, **Latent Dirichlet Allocation (LDA)** is a generative statistical model that explains a set of observations through unobserved groups, and each group explains why some parts of the data are similar. LDA is an example of a topic model. In this, observations (e.g., words) are collected into documents, and each word's presence is attributable to one of the document's topics. Each document will contain a small number of topics.

4. **Compare Manhattan and Euclidean distance.**

Euclidean Distance: Euclidean distance is calculated as the square root of the sum of the squared differences between a new point (x) and an existing point (y).

- The Euclidean distance or Euclidean metric is the "ordinary" (i.e.straight-line) distance between two points in Euclidean space.
- The Euclidean distance between points p and q is the length of the line segment connecting them.

Manhattan Distance: This is the distance between real vectors using the sum of their absolute difference.

- It is the sum of the lengths of the projections of the line segment between the points onto the coordinate axes.

5. **Distinguish between supervised and unsupervised learning.**

Supervised learning is a machine learning method in which models are trained using labeled data. In supervised learning, models need to find the mapping function to map the input variable (X) with the output variable (Y).

$$\boxed{Y = f(X)}$$

Supervised learning needs supervision to train the model, which is similar to as a student learns things in the presence of a teacher.

Unsupervised learning is another machine learning method in which patterns inferred from the unlabeled input data. The goal of unsupervised learning is to find the structure and patterns from the input data. Unsupervised learning does not need any supervision. Instead, it finds patterns from the data by its own.

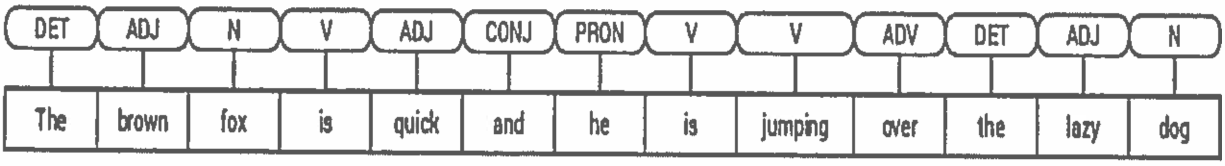
PART B

6 a) Interpret the language syntax and structure in natural language.

Knowledge about the structure and syntax of language is helpful in many areas like text processing, annotation, and parsing for further operations such as text classification or summarization. Typical parsing techniques for understanding text syntax are mentioned below.

- **Parts of Speech (POS) Tagging**
- **Shallow Parsing or Chunking**
- **Constituency Parsing**
- **Dependency Parsing**

We will be looking at all of these techniques in subsequent sections. Considering our previous example sentence ***"The brown fox is quick and he is jumping over the lazy dog"***, if we were to annotate it using basic POS tags, it would look like the following figure.



POS tagging for a sentence

Thus, a sentence typically follows a hierarchical structure consisting the following components,

sentence → clauses → phrases → words

Tagging Parts of Speech

Parts of speech (POS) are specific lexical categories to which words are assigned, based on their syntactic context and role. Usually, words can fall into one of the following major categories.

- ***N(oun):*** This usually denotes words that depict some object or entity, which may be living or nonliving. Some examples would be fox , dog , book , and so on. The POS tag symbol for nouns is **N**.
- ***V(erb)***: Verbs are words that are used to describe certain actions, states, or occurrences. There are a wide variety of further subcategories, such as auxiliary, reflexive, and transitive verbs (and many more). Some typical examples of verbs would be running , jumping , read , and write . The POS tag symbol for verbs is **V**.
- ***Adj(ective)***: Adjectives are words used to describe or qualify other words, typically nouns and noun phrases. The phrase beautiful flower has the noun (N) flower which is described or qualified using the adjective (ADJ) beautiful . The POS tag symbol for adjectives is **ADJ** .
- ***Adv(erb):*** Adverbs usually act as modifiers for other words including nouns, adjectives, verbs, or other adverbs. The phrase very beautiful flower has the adverb (ADV) very , which modifies the adjective (ADJ) beautiful , indicating the degree to which the flower is beautiful. The POS tag symbol for adverbs is **ADV**.

Besides these four major categories of parts of speech , there are other categories that occur frequently in the English language. These include pronouns, prepositions, interjections, conjunctions, determiners, and many others. Furthermore, each POS tag like the *noun* (**N**) can be further subdivided into categories like *singular nouns* (**NN**), *singular proper nouns*(**NNP**), and *plural nouns* (**NNS**).

The process of classifying and labeling POS tags for words called *parts of speech tagging* or *POS tagging* . POS tags are used to annotate words and depict their POS, which is really helpful to perform specific analysis, such as narrowing down upon nouns and seeing which ones are the most prominent, word sense disambiguation, and grammar analysis. We will be leveraging both **nltk** and **spacy** which usually use the ***Penn Treebank notation*** for POS tagging.

	Word	POS tag	Tag type
0	US	NNP	PROPN
1	unveils	VBZ	VERB
2	world	NN	NOUN
3	's	POS	PART
4	most	RBS	ADV
5	powerful	JJ	ADJ
6	supercomputer	NN	NOUN
7	,	,	PUNCT
8	beats	VBZ	VERB
9	China	NNP	PROPN

SpaCy POS tagging

	Word	POS tag
0	US	NNP
1	unveils	VBZ
2	world's	VBZ
3	most	RBS
4	powerful	JJ
5	supercomputer,	JJ
6	beats	NNS
7	China	NNP

NLTK POS tagging

POS tagging a news headline

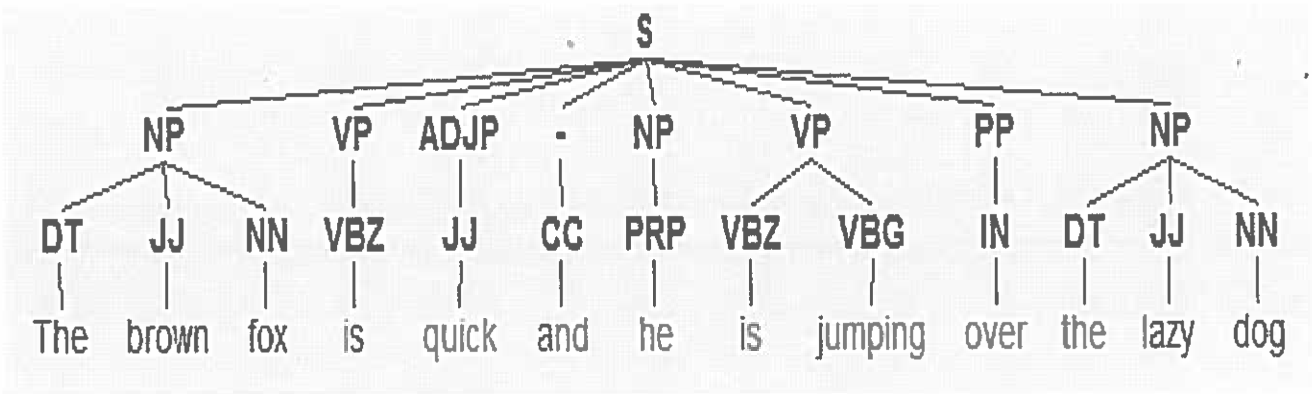
We can see that each of these libraries treat tokens in their own way and assign specific tags for them. Based on what we see, **spacy** seems to be doing slightly better than **nltk**.

Shallow Parsing or Chunking

Based on the hierarchy we depicted earlier, groups of words make up phrases. There are five major categories of phrases:

- **Noun phrase (NP):** These are phrases where a noun acts as the head word. Noun phrases act as a subject or object to a verb.
- **Verb phrase (VP):** These phrases are lexical units that have a verb acting as the head word. Usually, there are two forms of verb phrases. One form has the verb components as well as other entities such as nouns, adjectives, or adverbs as parts of the object.
- **Adjective phrase (ADJP):** These are phrases with an adjective as the head word. Their main role is to describe or qualify nouns and pronouns in a sentence, and they will be either placed before or after the noun or pronoun.
- **Adverb phrase (ADVP):** These phrases act like adverbs since the adverb acts as the head word in the phrase. Adverb phrases are used as modifiers for nouns, verbs, or adverbs themselves by providing further details that describe or qualify them.
- **Prepositional phrase (PP):** These phrases usually contain a preposition as the head word and other lexical components like nouns, pronouns, and so on. These act like an adjective or adverb describing other words or phrases.

Shallow parsing, also known as light parsing or chunking, is a popular natural language processing technique of analyzing the structure of a sentence to break it down into its smallest constituents (which are tokens such as words) and group them together into higher-level phrases. This includes POS tags as well as phrases from a sentence.



An example of shallow parsing depicting higher level phrase annotations

6b) Illustrate language semantics with suitable example.

The simplest definition of semantics is the study of meaning.
Linguistics has its own subfield of linguistic semantics, which deals with the study of meaning in language, the relationships between words, phrases, and symbols, and their indication, meaning, and representation of the knowledge they signify.
In simple words, semantics is more concerned with the facial expressions, signs, symbols, body language, and knowledge that are transferred when passing messages from one entity to another.

7a) Explain first order logic with an example.

- In propositional logic, we can only represent the facts, which are either true or false. PL is not sufficient to represent the complex sentences or natural language statements. The propositional logic has very limited expressive power. Consider the following sentence, which we cannot represent using PL logic.
  - "Some humans are intelligent", or
  - "Sachin likes cricket."
- To represent the above statements, PL logic is not sufficient, so we required some more powerful logic, such as first-order logic.
- FOL is sufficiently expressive to represent the natural language statements in a concise way.
- First-order logic is also known as Predicate logic or First-order predicate logic. First-order logic is a powerful language that develops information about the objects in a more easy way and can also express the relationship between those objects.
- First-order logic (like natural language) does not only assume that the world contains facts like propositional logic but also assumes the following things in the world:
- Objects: A, B, people, numbers, colors, wars, theories, squares, pits, wumpus, ...
- Relations: It can be unary relation such as: red, round, is adjacent, or n-any relation such as: the sister of, brother of, has color, comes between
- Function: Father of, best friend, third inning of, end of, ...
- As a natural language, first-order logic also has two main parts:
- Syntax
- Semantics

Table 1-3. Representation of Natural Language Statements Using First Order Logic

Sl No.	FOL Representation	Natural Language Statement
1	¬ eats(John, fish)	John does not eat fish
2	is_hot(pie) ∧ is_delicious(pie)	The pie is hot and delicious
3	is_hot(pie) ∨ is_delicious(pie)	The pie is either hot or delicious
4	study(John, exam) → pass(John, exam)	If John studies for the exam, he will pass the exam
5	∀x student(x) → pass(x, exam)	All students passed the exam
6	∃x student(x) ∧ fail(x, exam)	There is at least one student who failed the exam
7	(∃x student(x) ∧ fail(x, exam) ∧ (∀y fail(y, exam) → x=y))	There was exactly one student who failed the exam
8	∀x (spider(x) ∧ black_widow(x)) → poisonous(x)	All black widow spiders are poisonous

7b) Summarize the steps involved in text categorization and text analytics.

Text categorization process includes five main steps:

Document Preprocessing

In this step, html tags, rare words and stop words are removed, and some stemming is needed; this can be done easily in English, but it is more difficult in Arabic, Chinese, Japanese and some other languages. Word's root extraction methods may help in this step in order to normalize the document's words. There are several root extraction methods, including morphological analysis ofthe words and using N-gram technique [40].

Document Representation

Before classification, documents must be transformed into a format that is recognized by a computer, vector space model (VSM) is the most commonly used method. This model takes the document as a multi-dimension vector, and the feature selected from the dataset as a dimension ofthis vector.

Dimension Reduction

There are tens of thousands of words in a document, so as features it is infeasible to do the classification for all of them; also, the computer cannot process such amount of data. That is why it is important to select the most meaningful and representative features for classification, the most commonly selection methods used includes Chi square statistics [4][38], information gain, mutual information, document frequency, latent semantic analysis.

Model Training

This is the most important part of text categorization. It includes choosing some documents from corpus to comprise the training set, performs the learning on the training set, and then generates the model.

1.1.4. Testing and Evaluation

This step uses the model generated from the model training step, and performs the classification on the testing set, then chooses appropriate index to do evaluations.

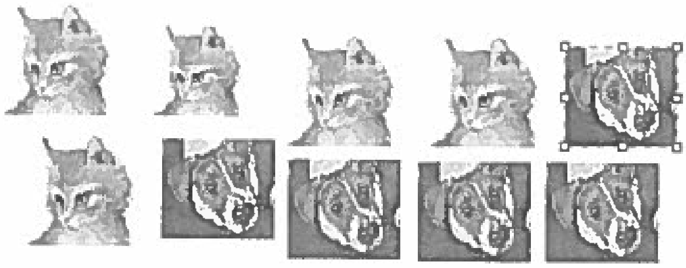
Text analytics, also known as text mining, is the methodology and process followed to derive quality and actionable information and insights from textual data.
This involves using NLP, information retrieval, and machine learning techniques to parse unstructured text data into more structured forms and deriving patterns and insights from this data that would be helpful for the end user. Text analytics comprises a collection of machine learning, linguistic, and statistical techniques that are used to model and extract information from text primarily for analysis needs, including business intelligence, exploratory, descriptive, and predictive analysis.

There are 7 basic steps involved in preparing an unstructured text document for deeper analysis:

1. Language Identification
2. Tokenization
3. Sentence Breaking
4. Part of Speech Tagging
5. Chunking
6. Syntax Parsing
7. Sentence Chaining

8a) Analyze support vector machine for automated text classification.

Support Vector Machine is a supervised classification algorithm where we draw a line between two different categories to differentiate between them. SVM is also known as the support vector network.Consider an example where we have cats and dogs together.



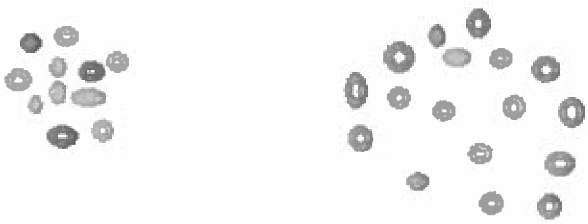
We want our model to differentiate between cats and dogs.

There are many cases where the differentiation is not so simple as shown above. In that case, the hyperplane dimension needs to be changed from 1 dimension to the Nth dimension. This is called Kernel. To be more simple, its the functional relationship between the two observations. It will add more dimensions to the data so we can easily differentiate among them.

We can have three types of kernels.

1. Linear Kernels
2. Polynomial Kernels
3. Radial Basis Function Kernel

In practical life, it's very difficult to get a straight hyperplane. Consider the image below where the points are mixed together. You cannot separate the points using a straight 2d hyperplane.



8b) Explain the text classification with building process of automated text classification.

Text classification also known as *text tagging* or *text categorization* is the process of categorizing text into organized groups. By using Natural Language Processing (NLP), text classifiers can automatically analyze text and then assign a set of pre-defined tags or categories based on its content.

Automated Text Classification:

- Consider several humans doing the task of going through each document and classifying it. They would then be a part of the text classification system we are talking about. However, that would not scale very well once there were millions of text documents to be classified quickly.

- To make the process more efficient and faster, we can consider automating the task of text classification, which brings us to automated text classification. To automate text classification, we can make use of several ML techniques and concepts.

- There are mainly two types of ML techniques that are relevant to solving this problem:

• Supervised machine learning • Unsupervised machine learning

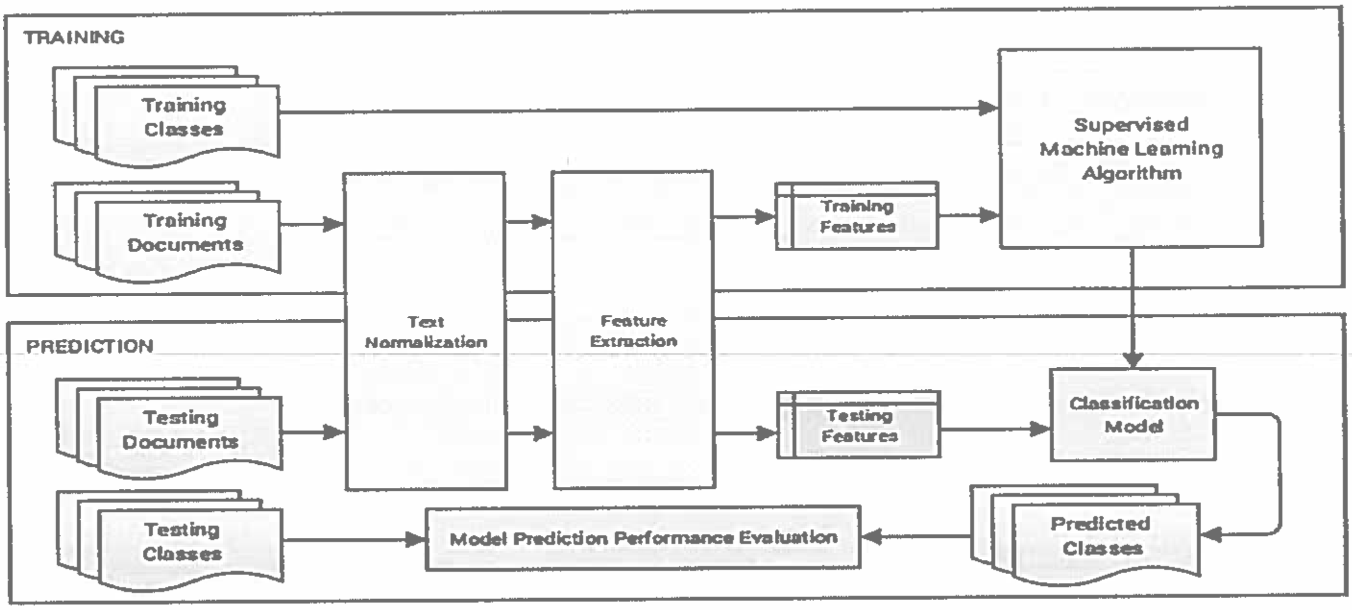


Figure: Blueprint for building an automated text classification system

9a) Interpret TF-IDF mode in text classification.

TF-IDF stands for **"Term Frequency — Inverse Document Frequency"**. This is a technique to quantify words in a set of documents. We generally compute a score for each word to signify its importance in the document and corpus. This method is a widely used technique in Information Retrieval and Text Mining.

The process to find meaning of documents using TF-IDF is very similar to Bag of words,

1. Clean data / Preprocessing — Clean data (standardise data) , Normalize data(all lower case) , lemmatize data (all words to root words).

2. Tokenize words with frequency

3. Find TF for words

4. Find IDF for words

5. Vectorize vocab

9b) Compare the performance and features of dependency based parsing and constituency based parsing.

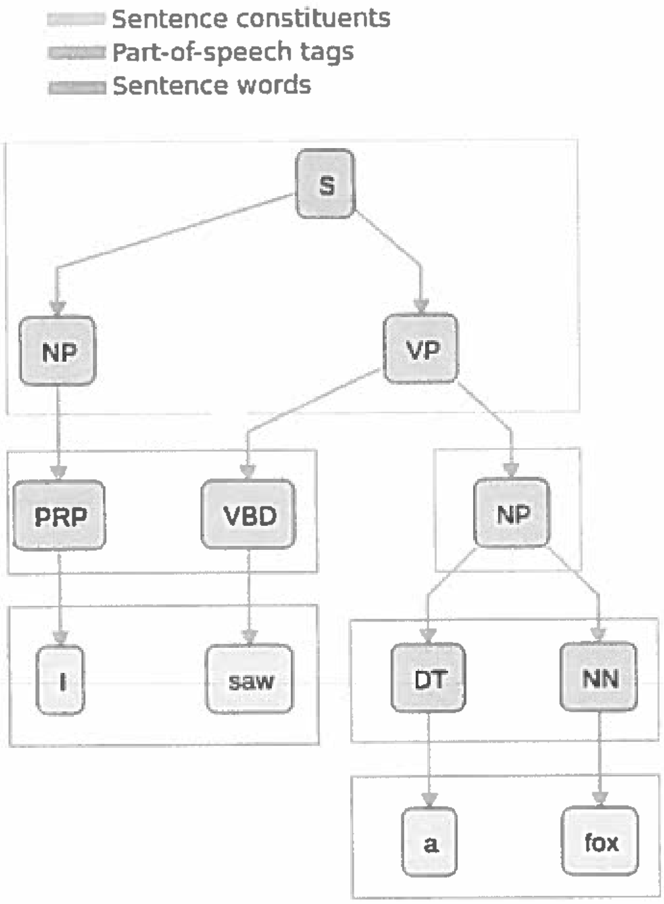
The constituency parse tree is based on the formalism of context-free grammars. In this type of tree, the sentence is divided into constituents, that is, sub-phrases that belong to a specific category in the grammar.

In English, for example, the phrases "a dog", "a computer on the table" and "the nice sunset" are all noun phrases, while "eat a pizza" and "go to the beach" are verb phrases.

The grammar provides a specification of how to build valid sentences, using a set of rules. As an example, the rule means that we can form a verb phrase (VP) using a verb (V) and then a noun phrase (NP).

While we can use these rules to generate valid sentences, we can also apply them the other way around, in order to extract the syntactical structure of a given sentence according to the grammar.

Let's dive straight into an example of a constituency parse tree for the simple sentence, "I saw a fox":



A constituency parse tree always contains the words of the sentence as its terminal nodes. Usually, each word has a parent node containing its part-of-speech tag (noun, adjective, verb, etc...), although this may be omitted in other graphical representations.

All the other non-terminal nodes represent the constituents of the sentence and are usually one of verb phrase, noun phrase, or prepositional phrase (PP).

In this example, at the first level below the root, our sentence has been split into a noun phrase, made up of the single word "I", and a verb phrase, "saw a fox". This means that the grammar contains a rule like , meaning that a sentence can be created with the concatenation of a noun phrase and a verb phrase.

Similarly, the verb phrase is divided into a verb and another noun phrase. As we can imagine, this also maps to another rule in the grammar.

To sum things up, constituency parsing creates trees containing a syntactical representation of a sentence, according to a context-free grammar. This representation is highly hierarchical and divides the sentences into its single phrasal constituents.

Dependency Parsing

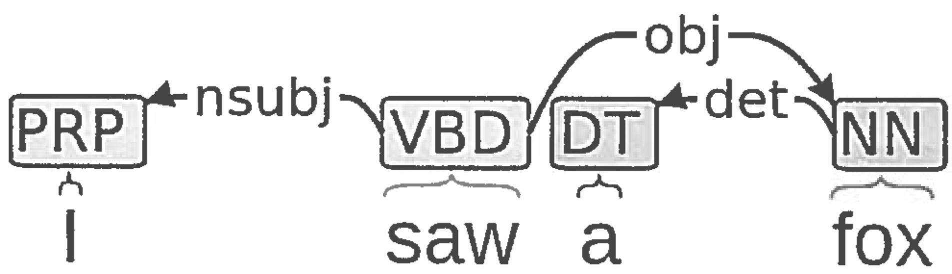
As opposed to constituency parsing, **dependency parsing doesn't make use of phrasal constituents or sub-phrases**. Instead, the syntax of the sentence is expressed in terms of dependencies between words — that is, directed, typed edges between words in a graph.

More formally, a dependency parse tree is a graph where the set of vertices contains the words in the sentence, and each edge in connects two words. The graph must satisfy three conditions:

1. There has to be a single root node with no incoming edges.
2. For each node in , there must be a path from the root to .
3. Each node except the root must have exactly 1 incoming edge.

Additionally, each edge in has a type, which defines the grammatical relation that occurs between the two words.

Let's see what the previous example looks like if we perform dependency parsing:



As we can see, the result is completely different. With this approach, the root of the tree is the verb of the sentence, and edges between words describe their relationships.

For example, the word "saw" has an outgoing edge of type *nsubj* to the word "I", meaning that "I" is the nominal subject of the verb "saw". In this case, we say that "I" *depends on* "saw".

10 a) What is text normalization? List the features to be extracted for text normalization.

Text normalization is a pre-processing step aimed at improving the quality of the text and making it suitable for machines to process. Four main steps in text normalization are case normalization, tokenization and stop word removal, Parts-of-Speech (POS) tagging, and stemming.

Some of the most popular methods of feature extraction are :

- Bag-of-Words
- TF-IDF

Bag of Words: Bag-of-Words is one of the most fundamental methods to transform tokens into a set of features. The BoW model is used in document classification, where each word is used as a feature for training the classifier. For example, in a task of review based sentiment analysis, the presence of words like *'fabulous'*, *'excellent'* indicates a positive review, while words like *'annoying'*, *'poor'* point to a negative review . There are 3 steps while creating a BoW model :

1. The first step is **text-preprocessing** which involves:
    1. converting the entire text into lower case characters.
    2. removing all punctuations and unnecessary symbols.
2. The second step is to **create a vocabulary** of all unique words from the corpus. Let's suppose, we have a hotel review text. Let's consider 3 of these reviews, which are as follows :

1. *good movie*
2. *not a good movie*
3. *did not like*

1. Now, we consider all the unique words from the above set of reviews to create a vocabulary, which is going to be as follows :

{good, movie, not, a, did, like}

1.In the third step, we **create a matrix of features** by assigning a separate column for each word, while each row corresponds to a review. This process is known as **Text Vectorization**. Each entry in the matrix signifies the presence(or absence) of the word in the review. We put **1** if the word is present in the review, and **0** if it is not present.
or the above example, the matrix of features will be as follows :

good	movie	not	a	did	like
1	1	0	0	0	0
1	1	1	1	0	0
0	0	1	0	1	1

10b) Illustrate weighted tag based phrase extraction with an example.

This method borrows concepts from a couple of papers, namely K. Barker and N. Cornachhia's "Using Noun Phrase Heads to Extract Document Keyphrases" and "KEA: Practical Automatic KeyphraseExtraction" by Ian Witten et al., which you can refer to for further details on their experimentations and approaches. We follow a two-step process in our algorithm here:

1. Extract all noun phrases chunks using shallow parsing
2. Compute TF-IDF weights for each chunk and return the topweighted phrases

For the first step, we will use a simple pattern based on parts of speech (POS) tagsto extract noun phrase chunks. You will be familiar with this from Chapter 3 where we explored chunking and shallow parsing. Before discussing our algorithm, let us define the corpus on which we will be testing our implementation. We use a sample description of elephants taken from Wikipedia as shown in the following code:

```
toy_text = """
Elephants are large mammals of the family Elephantidae
and the order Proboscidea. Two species are traditionally recognised,the African elephant and
the Asian elephant. Elephants are scatteredthroughout sub-Saharan Africa, South Asia, and
Southeast Asia. Male African elephants are the largest extant terrestrial animals. All elephants
have a long trunk used for many purposes,
particularly breathing, lifting water and grasping objects. Theirincisors grow into tusks, which
can serve as weapons and as toolsfor moving objects and digging. Elephants' large ear
flaps help to control their body temperature. Their pillar-like legs can carry their great weight.
African elephants have larger ears
and concave backs while Asian elephants have smaller earsand convex or level
backs.
"""
```

Now that we have our corpus ready, we will use the pattern, " NP: {<DT>? <JJ>* <NN.*>+}" for extracting all possible noun phrases from our corpus of documents/ sentences. You can always experiment with more sophisticated patterns later, incorporating verb, adjective, or even adverb phrases. However, I will keep things simpleand concise here to focus on the core logic. Once we have our pattern, we will define a function to parse and extract these phrases using the following snippet.

11a) Explain the process of automated document summarization.

Automated document summarization is the process of using a computer program or algorithm based on statistical and ML techniques to summarize a document or corpus of documents such that we obtain a short summary that captures all the essential concepts and themes of the original document or corpus. A wide variety of techniques for building automated document summarizers exist, including various extraction- and abstraction-based techniques. The key concept behind all these algorithms is to find a representative subset of the original datasetsuch that the core essence of the dataset from the semantic and conceptual standpoints is contained in this subset. Document summarization usually involves trying to extract and construct
an executive summary from a single document. But the same algorithms can be extended to multiple documents, though usually the idea is not to combine several diverse documents together, which would defeat the purpose of the algorithm. Thesame concept is not only applied in text analytics but also to image and video summarization.

There are mainly two broad approaches towards document summarization usingautomated techniques:

- *Extraction-based techniques*: These methods use mathematical and statistical concepts like SVD to extract some key subset of content from the original document such that this subset of content contains the core information and

acts as the focal pointof the entire document. This content could be words, phrases,
or sentences. The end result from this approach is a short executive summary of a couple of lines are taken or extracted from the original document. No new content is generated in thistechnique—hence the name *extraction-based.*

- *Abstraction-based techniques*: These methods are more complex and sophisticated and leverage language semantics to create representations. They also make use of NLG techniques where themachine uses knowledge bases and semantic representations to generate text on its own and creates summaries just like a human would write them.

12a) Elaborate the process of feature extraction in text summarization.
Text summarization and information extraction deal with trying to extract key importantconcepts and themes from a huge corpus of text, essentially reducing it in the process. Before we dive deeper into the concepts and techniques, we should first understand the need for text summarization. The concept of information overload is one of the prime reasons behind the demand for text summarization.

Information overload, then, is the presence of excess data or information, which consumers find difficult to process in making well-informed decisions. The overload occurs when the amount of information as input to the system starts exceeding the processing capability of the system. We as humans have limited cognitive processing capabilities and are also wired in such a way that we cannot spend a long time readinga single piece of information or data because the mind tends to wander every now andthen. Thus when we get loaded with information, it leads to a reduction in making qualitative decisions.

Techniques:

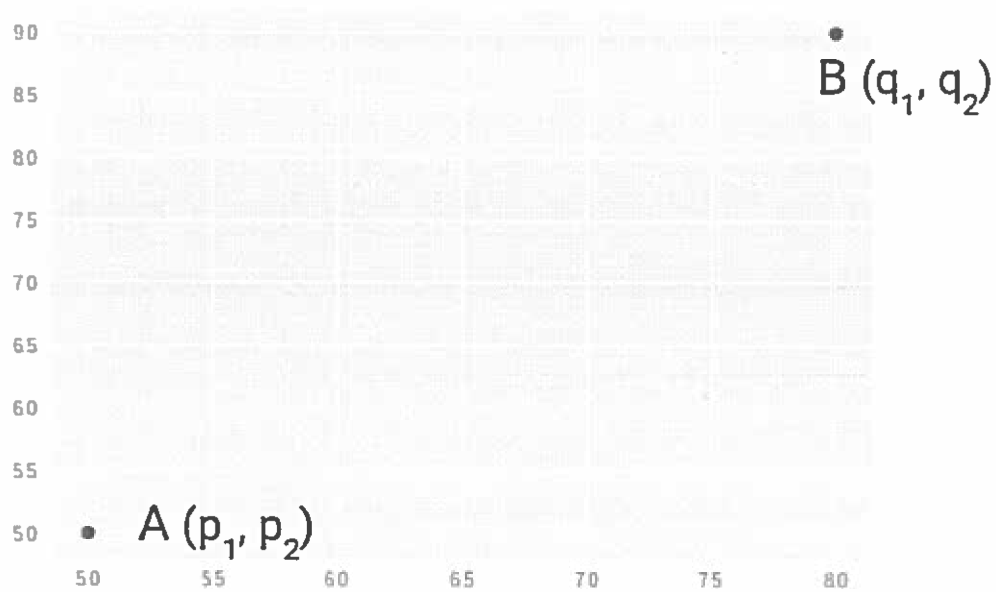
- *Keyphrase extraction* is perhaps the simplest out of the three techniques. It involves extracting keywords or phrases from a textdocument or corpus that capture its main concepts or themes. This can be said to be a simplistic form of topic modeling. You might have seen keywords or phrases described in a research paper or even some product in an online store that describes the entity in a few words or phrases, capturing its main idea orconcept.
- *Topic modeling* usually involves using statistical and mathematical modeling techniques to extract main topics, themes, or concepts from a corpus of documents. Note here theemphasis on *corpus* of documents because the more diverse setof documents you have, the more topics or concepts you can generate—unlike with a single document where you will not gettoo many topics or concepts if it talks about a singular concept. Topic models are also often known as *probabilistic statistical models*, which use specific statistical techniques including singular valued decomposition and latent dirichlet allocation to discover connected latent semantic structures in text data that yield topics and concepts. They are used extensively in textanalytics and even bioinformatics.
- *Automated document summarization* is the process of using a computer program or algorithm based on statistical and ML techniques to summarize

a document or corpus of documents such that we obtain a short summary that captures all the essential concepts and themes of the original document or corpus. A wide variety of techniques for building automated document summarizers exist, including various extraction- and abstraction-based techniques. The key concept behind all these algorithms is to find a representative subset of the original datasetsuch that the core essence of the dataset from the semantic and conceptual standpoints is contained in this subset. Document summarization usually involves trying to extract and construct an executive summary from a single document. But the same algorithms can be extended to multiple documents, though usually the idea is not to combine several diverse documents together, which would defeat the purpose of the algorithm. The same concept is not only applied in text analytics but also to image and video summarization.

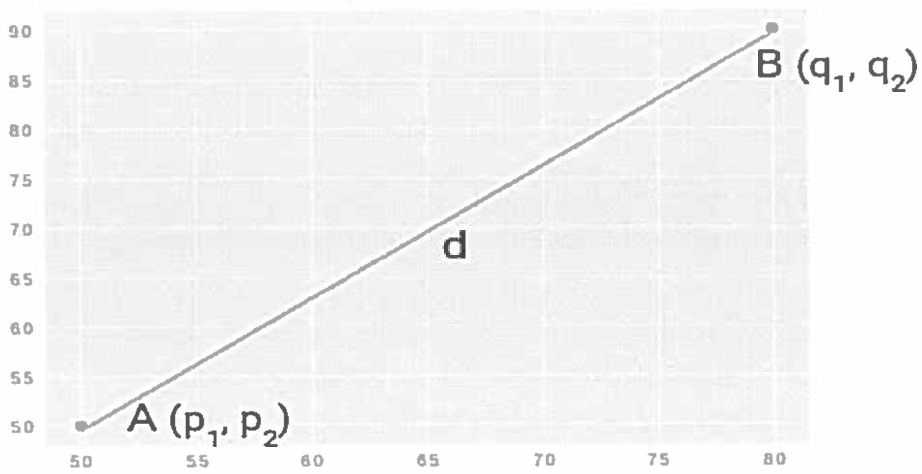
12b) Compare the performance of the Manhattan and Euclidean distance similarity measures.

Euclidean Distance represents the shortest distance between two points.

Most machine learning algorithms including K-Means use this distance metric to measure the similarity between observations. Let's say we have two points as shown below:



So, the Euclidean Distance between these two points A and B will be:



Here's the formula for Euclidean Distance:

$$d = ((p_1 - q_1)^2 + (p_2 - q_2)^2)^{1/2}$$

We use this formula when we are dealing with 2 dimensions. We can generalize this for an n-dimensional space as:

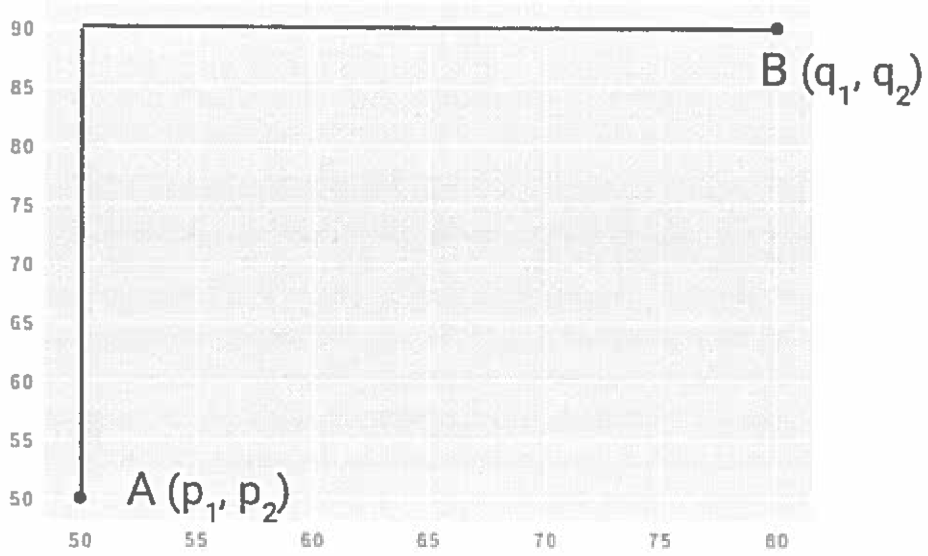
$$D_e = \left(\sum_{i=1}^{n} (p_i - q_i)^2 \right)^{1/2}$$

Where,

- n = number of dimensions
- pi, qi = data points

Manhattan Distance is the sum of absolute differences between points across all the dimensions.

We can represent Manhattan Distance as:



Since the above representation is 2 dimensional, to calculate Manhattan Distance, we will take the sum of absolute distances in both the x and y directions. So, the Manhattan distance in a 2-dimensional space is given as:

$$d = |p_1 - q_1| + |p_2 - q_2|$$

And the generalized formula for an n-dimensional space is given as:

$$D_m = \sum_{i=1}^{n} |p_i - q_i|$$

Where,

- n = number of dimensions
- pi, qi = data points

```
Manhattan Distance b/w (1, 2, 3) and (4, 5, 6) is:  9
```

Note that **Manhattan Distance is also known as city block distance.** SciPy has a function called *cityblock* that returns the Manhattan Distance between two points.

13a) Analyze K-means clustering algorithm for document clustering using suitable dataset.

The *k-means clustering algorithm* is a centroid-based clustering model that tries to cluster data into groups or clusters of equal variance. The criteria or measure that this algorithm tries to minimize is *inertia*, also known as *within-cluster sum-of-squares*. Perhaps the one main disadvantage of this algorithm is that the number of clusters *k* need to be specified in advance, as is the case with all other centroid-based clustering models. This algorithm is perhaps the most popular clustering algorithm out there and is frequently used due to its ease of use as well as the fact that it is scalable with large amounts of data.

We can now formally define the k-means clustering algorithm along with its mathematical notations. Consider that we have a dataset X with N data points or samples and we want to group them into K clusters where K is a user-specified parameter. The k-means clustering algorithm will segregate the N data points into K disjoint separate clusters C_k, and each of these clusters can be described by the means of the cluster samples. These means become the cluster centroids μ_k such that these centroids are not bound by the condition that they have to be actual data points from the N samples in X. The algorithm chooses these centroids and builds the clusters in such a way that the inertia or within-cluster sums of squares are minimized. Mathematically, this can be represented as

$$\min \sum_{i=1}^{K} \sum_{x_n \in C_i} \| x_n - \mu_i \|^2$$

with regard to clusters C_i and centroids μ_i such that $i \in \{1, 2, \ldots, k\}$. This optimization is an NP *hard problem* for all you algorithm enthusiasts out there. Lloyd's algorithm is a solution to this problem, which is an iterative procedure consisting of the following steps.

1. Choose initial k centroids μ_k by taking k random samples from the dataset X.
2. Update clusters by assigning each data point or sample to its nearest centroid point. Mathematically, we can represent this as $C_k = \{x_n : \| x_n - \mu_k \| \le \text{all } \| x_n - \mu_l \| \}$ where C_k denotes the clusters.

13b) Explain about Ward's agglomerative hierarchical clustering with suitable dataset.

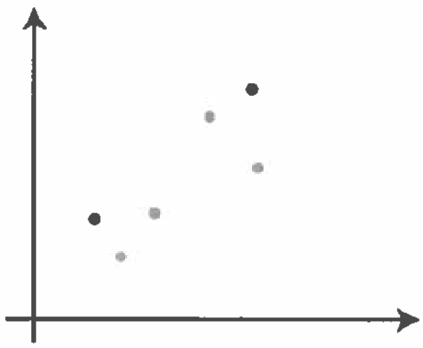
he agglomerative hierarchical clustering algorithm is a popular example of HCA. To group the datasets into clusters, it follows the **bottom-up approach**. It means, this algorithm considers each dataset as a single cluster at the beginning, and then start combining the closest pair of clusters together. It does this until all the clusters are merged into a single cluster that contains all the datasets.

This hierarchy of clusters is represented in the form of the dendrogram.

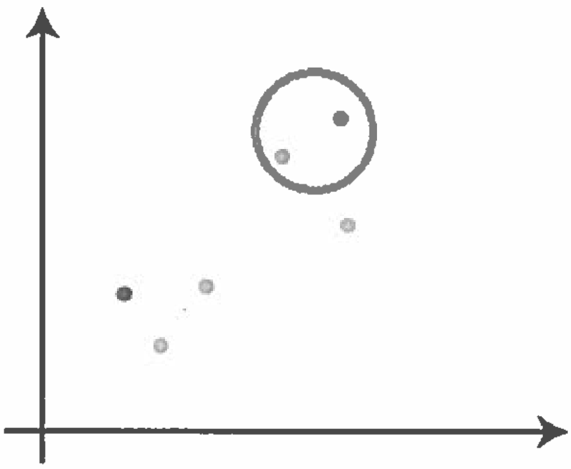
How the Agglomerative Hierarchical clustering Work?

The working of the AHC algorithm can be explained using the below steps:

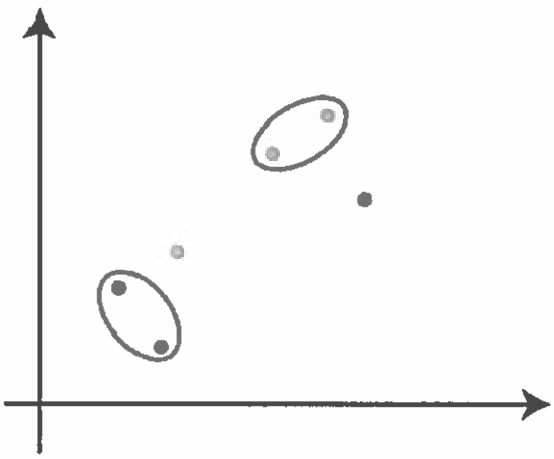
- **Step-1:** Create each data point as a single cluster. Let's say there are N data points, so the number of clusters will also be N.



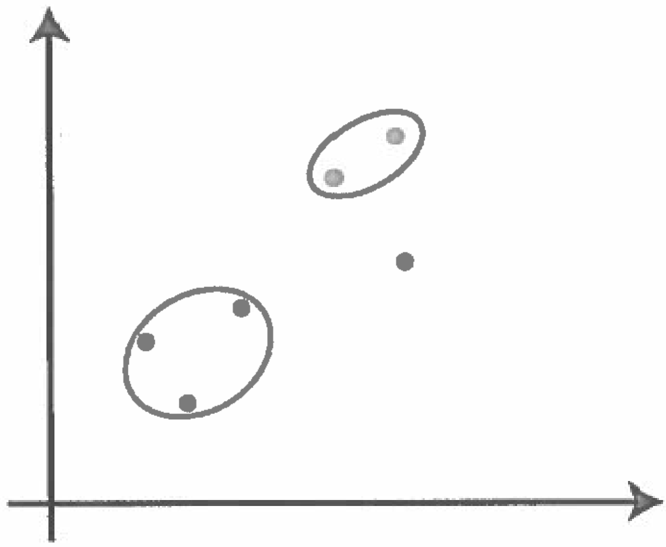
- **Step-2:** Take two closest data points or clusters and merge them to form one cluster. So, there will now be N-1 clusters.

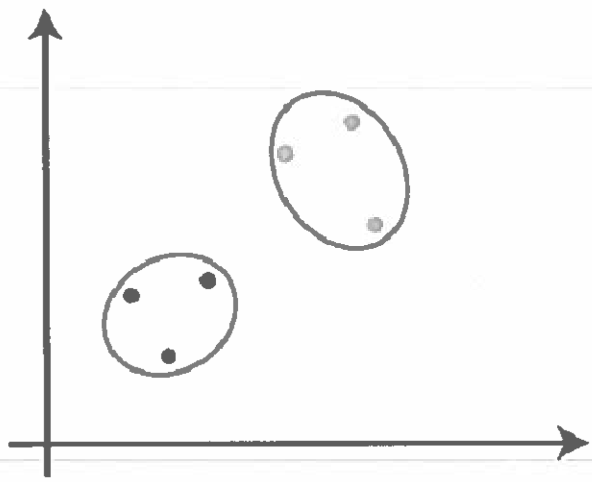


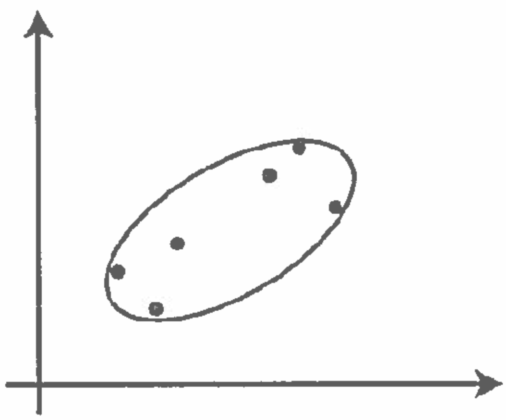
- **Step-3**: Again, take the two closest clusters and merge them together to form one cluster. There will be N-2 clusters.



- **Step-4:** Repeat Step 3 until only one cluster left. So, we will get the following clusters. Consider the below images:







- **Step-5:** Once all the clusters are combined into one big cluster, develop the dendrogram to divide the clusters as per the problem.

14a) Analyze lexical semantic relations with necessary illustrations.

Lexical semantics

It is usually concerned with identifying semantic relations between lexical units in a language and how they are correlated to the syntax and structure of the language.

Each lexical unit has its own syntax, form, and meaning. They also derive meaning from their surrounding lexical units in phrases, clauses, and sentences.

A lexicon is a complete vocabulary of these lexical units.

A lemma is also known as the canonical or citation form for a set of words.
The lemma is usually the base form of a set of words, known as a lexeme in this context.
A simple example would the lexeme {eating, ate, eats}, which contains the wordforms, and their lemma is the word eat .

Homograph: one of two or more words spelled alike but different in meaning or derivation or pronunciation.

Ex:

Bear - To endure ; Bear - Animal.

Lean - Thin ; Lean - Rest against.

Lead - Metal ; Lead - Start off in front.

Fair - Appearance ; Fair - Reasonable.

Homonyms: are defined as words that share the same spelling or pronunciation but have different meanings.
Ex:
The bat hangs upside down from the tree and That baseball bat is really sturdy.

Homophone is a word that is pronounced the same as another word but differs in meaning.
A *homophone* may also differ in spelling. The two words may be spelled the same, a for example *rose* (flower) and *rose* (past tense of "rise")

Heteronyms are words that have the same written form or spelling but different pronunciations and meanings. Examples of heteronyms are the words lead (metal, command) and tear (rip off something, moisture from eyes).

Heterographs are words that have the same pronunciation but different meanings and spellings. Some examples include the words to, too , and two , which sound similar but have different spellings and meanings.

Polysemes are words that have the same written form or spelling and different but very relatable meanings. While this is very similar to homonymy, the difference is subjective and depends on the context, since these words are relatable to each other. A good example is the word bank which can mean (1) a financial institution, (2) the bank of the river, (3) the building that belongs to the financial institution, or (4) a verb meaning to rely upon. These examples use the same word bank and are homonyms. But only (1), (3), and (4) are polysemes representing a common theme (the financial organization representing trust and security).

Capitonyms are words that have the same written form or spelling but have different meanings when capitalized. They may or may not have different pronunciations. Some examples include the words march (March indicates the month, and march depicts the action of walking) and may (May indicates the month, and may is a modal verb).

Synonyms are words that have different pronunciations and spellings but have the same meanings in some or all contexts. Ex: That milkshake is really (big/large/huge).

Antonyms are pairs of words that define a binary opposite relationship.
Ex: fat, skinny;
divide, unite.

Hyponyms and Hypernyms:
Hyponyms are words that are usually a subclass of another word.
Hypernyms are the words that act as the superclass to hyponyms and have a more generic sense compared to the hyponyms. An example would be the word fruit, which is a hypernym, and the words mango, orange, and pear would be possible hyponyms. The relationships depicted between these words are often termed hyponymy and hypernymy.

14b) Distinguish between first order and propositional logics for semantic analysis.

	Propositional Logic	Predicate Logic
1	Propositional logic is the logic that deals with a collection of declarative statements which have a truth value, true or false.	Predicate logic is an expression consisting of variables with a specified domain. It consists of objects, relations and functions between the objects.
2	It is the basic and most widely used logic. Also known as Boolean logic.	It is an extension of propositional logic covering predicates and quantification.
3	A proposition has a specific truth value, either true or false.	A predicate's truth value depends on the variables' value.
4	Scope analysis is not done in propositional logic.	Predicate logic helps analyze the scope of the subject over the predicate. There are three quantifiers : Universal Quantifier (∀) depicts for all, Existential Quantifier (∃) depicting there exists some and Uniqueness Quantifier (∃!) depicting exactly one.
5	Propositions are combined with Logical Operators or Logical Connectives like Negation(¬), Disjunction(∨), Conjunction(∧), Exclusive OR(⊕), Implication(⇒), Bi-Conditional or Double Implication(⇔).	Predicate Logic adds by introducing quantifiers to the existing proposition.
6	It is a more generalized representation.	It is a more specialized representation.
7	It cannot deal with sets of entities.	It can deal with set of entities with the help of quantifiers.

15a) Examine named entity recognition for semantic analysis.

Named Entity Recognition is a task of finding the named entities that could possibly belong to categories like persons, organizations, dates, percentages, etc., and categorize the identified entity to one of these categories.

Named Entity Recognition Working:

When we read a text, we naturally recognize named entities like people, values, locations, and so on. For example, in the sentence *"Mark Zuckerberg is one of the founders of Facebook, a company from the United States"* we can identify three types of entities:

- *"Person": Mark Zuckerberg*
- *"Company": Facebook*
- *"Location": United States*

For computers, however, we need to help them recognize entities first so that they can categorize them.

This is done through machine learning and Natural Language Processing (NLP).

NLP studies the structure and rules of language and creates intelligent systems capable of deriving meaning from text and speech, while machine learning helps machines learn and improve over time.

To learn what an entity is, an NER model needs to be able to detect a word, or string of words that form an entity (e.g. New York City), and know which entity category it belongs to.

So first, we need to create entity categories, like *Name, Location, Event, Organization*, etc., and feed an NER model relevant training data. Then, by tagging some word and phrase samples with their corresponding entities, you'll eventually teach your NER model how to detect entities itself.

Named entity recognition (NER) helps you easily identify the key elements in a text, like names of people, places, brands, monetary values, and more. Extracting the main entities in a text helps sort unstructured data and detect important information, which is crucial if you have to deal with large datasets.

Here are some interesting use cases of named entity recognition:

Categorize Tickets in Customer Support

If you're dealing with a rising number of customer support tickets, you can use named entity recognition techniques to handle customer requests faster.

Automate repetitive customer service tasks, like categorizing customers' issues and queries, and save you valuable time that will help improve your resolution rates and boost customer satisfaction.

You can also use entity extraction to pull relevant pieces of data, like product names or serial numbers, making it easier to route tickets to the most suitable agent or team for handling that issue.

Gain Insights from Customer Feedback

Online reviews are a great source of customer feedback: they can provide rich insights about what clients like and dislike about your products, and the aspects of your business that need improving.

NER systems can be used to organize all this customer feedback and pinpoint recurring problems. For example, you could use NER to detect locations that are mentioned most often in negative customer feedback, which might lead you to focus on a particular office branch.

Content Recommendation

Many modern applications (like Netflix and YouTube) rely on recommendation systems to create optimal customer experiences. A lot of these systems rely on named entity recognition, which is able to make suggestions based on user search history.

For example, if you watch a lot of comedies on Netflix, you'll get more recommendations that have been classified as the entity *Comedy*.

Process Resumes

Recruiters spend many hours of their day going through resumes, looking for the right candidate. Each resume contains the same type of information, but they're often organized and formatted differently: a classic example of unstructured data.

By using an entity extractor, recruitment teams can instantly extract the most relevant information about candidates, from personal information (like name, address, phone number, date of birth and email), to data related to their training and experience (such as certifications, degree, company names, skills, etc).

NSRIT

Semester End Regular Examination, June, 2022

Degree	B. Tech. (U. G.)	**Program**	Common to All (Minor)			**Academic Year**	2021 - 2022
Course Code	20AIM01	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	IV
Course	**Fundamentals of Neural Networks**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Compare biological neuron with artificial neuron.	20AIM01.1	L2
2	Distinguish between supervised and unsupervised learning.	20AIM01.2	L2
3	What is perceptron?	20AIM01.3	L1
4	Identify the role of neuron in multilayer neural network.	20AIM01.4	L3
5	List out any four applications of BPN.	20AIM01.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Write the various benefits of neural networks. Explain the benefits of neural networks.	12M	20AIM01.1	L2
	OR			
7	Describe McCulloch-Pitts neuron model.	12M	20AIM01.1	L2
8	Explain how weights are adjusted in the different types of learning law. (Both supervised and unsupervised learning).	12M	20AIM01.2	L2
	OR			
9 (a)	Write the differences between conventional computer program and ANN.	6M	20AIM01.2	L2
9 (b)	List the advantages and disadvantages of Artificial Neural Networks.	6M	20AIM01.2	L2
10	Describe preceptron learning rule and delta learning rule.	12M	20AIM01.3	L2
	OR			
11	Elaborate the various learning processes used in the neural networks.	12M	20AIM01.3	L2
12	What is Multi-layer feed forward networks? What is the importance of hidden and output layers in it?	12M	20AIM01.4	L1
	OR			
13 (a)	Explain the steps involved in the back propagation algorithm.	6M	20AIM01.4	L1
13 (b)	What are the pattern recognition tasks that can be performed by back propagation network?	6M	20AIM01.4	L1
14	Explain Hebbian learning with necessary illustrations.	12M	20AIM01.5	L2
	OR			
15	Explain the architecture and function of Bidirectional Associative memory (BAM).	12M	20AIM01.5	L2

Semester End Exam June 2022 (1)

Academic Year 2021-22

B.Tech Minors AI & ML

Fundamentals of Neural Networks 20AIM01

Part A

(1) Structure of biological neuron and artificial neuron are more or less same. Functionally also, artificial neuron acts as a summing unit.

(2) Supervised learning needs information about class labels and it is a guided learning. Unsupervised learning does not need any guidance for learning

(3) Perceptron is a single layer ANN. That can be used for linear classification.

(4) Neurons in whichever layer they are function in the same way. The output of neurons of one layer are given as inputs to neurons of subsequent layers.

(5) BPN - image classification
handwritten character recognition
Pattern recognition and grouping

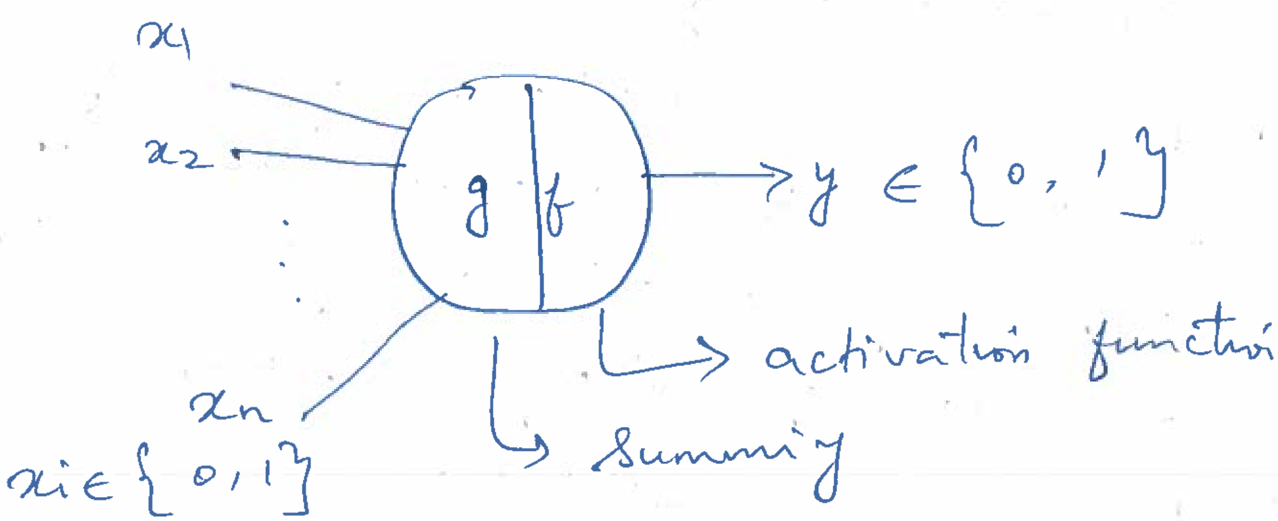
Part B

(6) Benefits of neural networks

- Storage of information (2)
- fault tolerant systems (2)
- ability to work with insufficient knowledge. (2)
- distributed memory (2)
- parallel processing (2)
- any application where large data is available (2)

- Explanation on each of these.

(7) McCulloh Pitt Neuron model



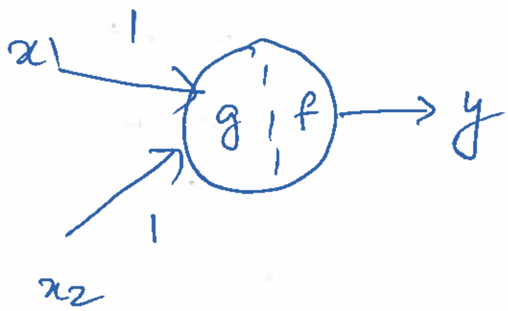
$$g(x_1, x_2, \ldots x_n) = \sum_{i=1}^{n} x_i \qquad (6) \qquad 3$$

$$y = f(g(x)) = \begin{cases} 1 & \text{if } g(x) \geq \theta \\ 0 & \text{otherwise} \end{cases}$$

- usage of this model for realising logical gates OR and AND (6)

OR

x_1	x_2	y
0	0	0
0	1	1
1	0	1
1	1	1



x_1	x_2	$g(x)$	y
0	0	0	0
0	1	1	1
1	0	1	1
1	1	2	1

$\Rightarrow \theta = 1$

(4)

(8) Superwised leaning (6)

definition

network example

Function / training / weight adjustment

Unsupervised learning (6)

definition

network example

weight adjustment

(9) a. Conventional program V's ANN

Conventional programs function with the logic and set of Rules and calculations

ANN can work with trained and unseen data, images, pictures and concepts

It also works for insufficient data

b. Advantages

a) ability to learn by themselves and produce output

b) loss of data doesnot affect its working

c) works even for missing data

Disadvantages

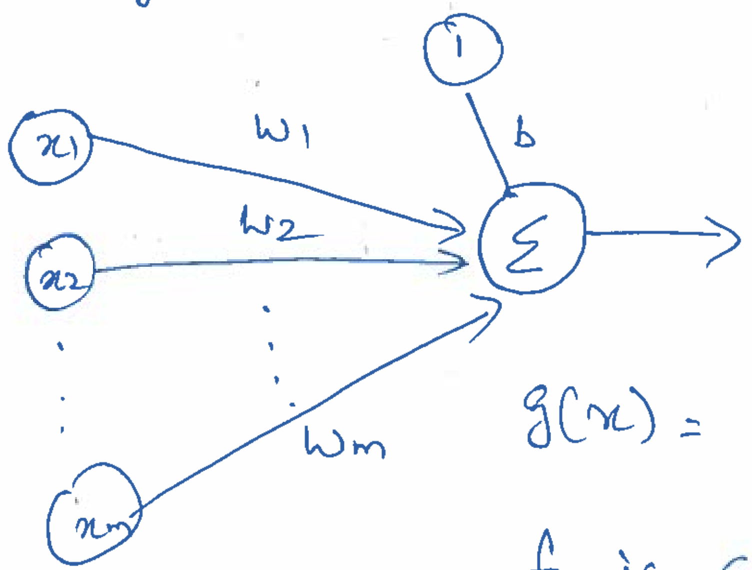
a) needs training

b) struetline of network

c) high processing time

10 Perceptron learning rule

- Single layer network



f $y = f(g(x))$.

$$g(x) = \sum_{i=0}^{m} x_i w_i$$

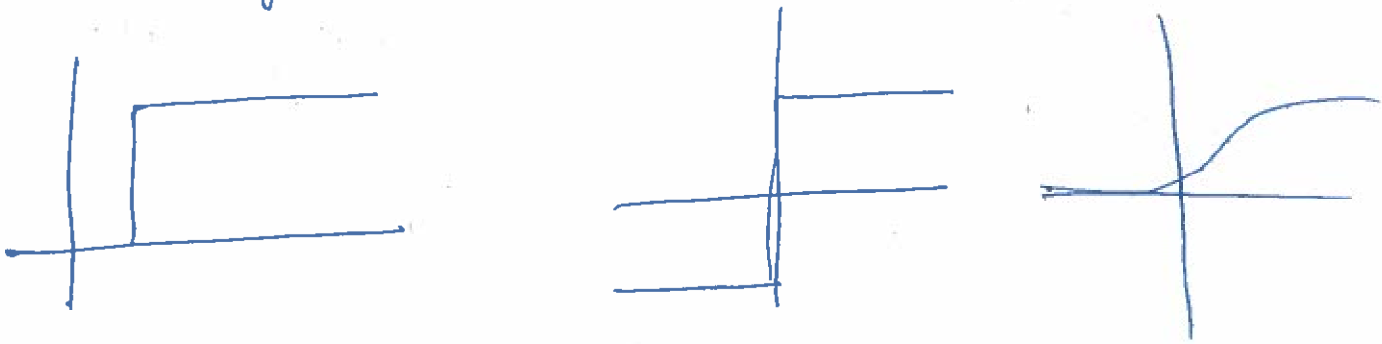
f is called activation function

It can be

step function

sign "

sigmoid function



8

11 Learning processes

Supervised learning (3)

unsupervised " (3)

semi supervised " (3)

~~ae~~ reinforced learning (3)

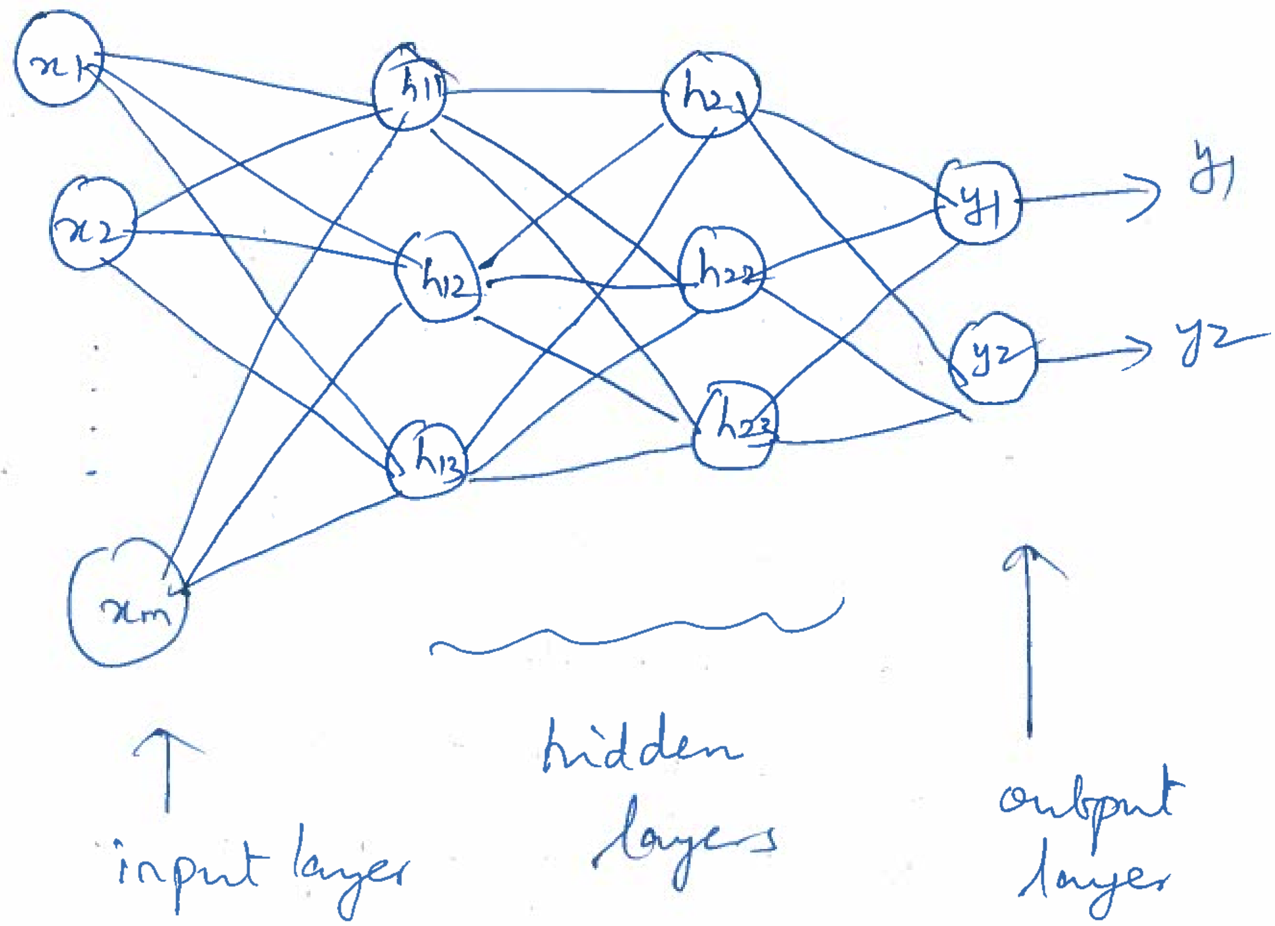
- definition
network
example

} for all the above listed learnings.

(12) Multilayer feed forward networks

- layers in addition to input and output layer (6)
- no of neurons in each layer may be different
- function of all neurons are same
- activation used may be varying

(6)



13. (a) Backpropagation algorithm (8)

- input layer working (6)
- summation
- activation
- output
- processing by layers
- error calculation
- back propagation of errors to minimize the error

(b) Pattern recognition with BPN (6)

- forward pass of pattern 1
- find the actual output
- compare with expected output and calculate error
- back propagate error to minimize it by weight adjustments

- do these for all patterns

14 Hebbian learning

- first learning rule (4)
- single layer network

one i/p layer and one o/p layer
(can have many units) (only one unit)

Learning algorithm. (4)

a) set weights to 0
and bias to 0

b) for each input vector, (s, t)
do the following

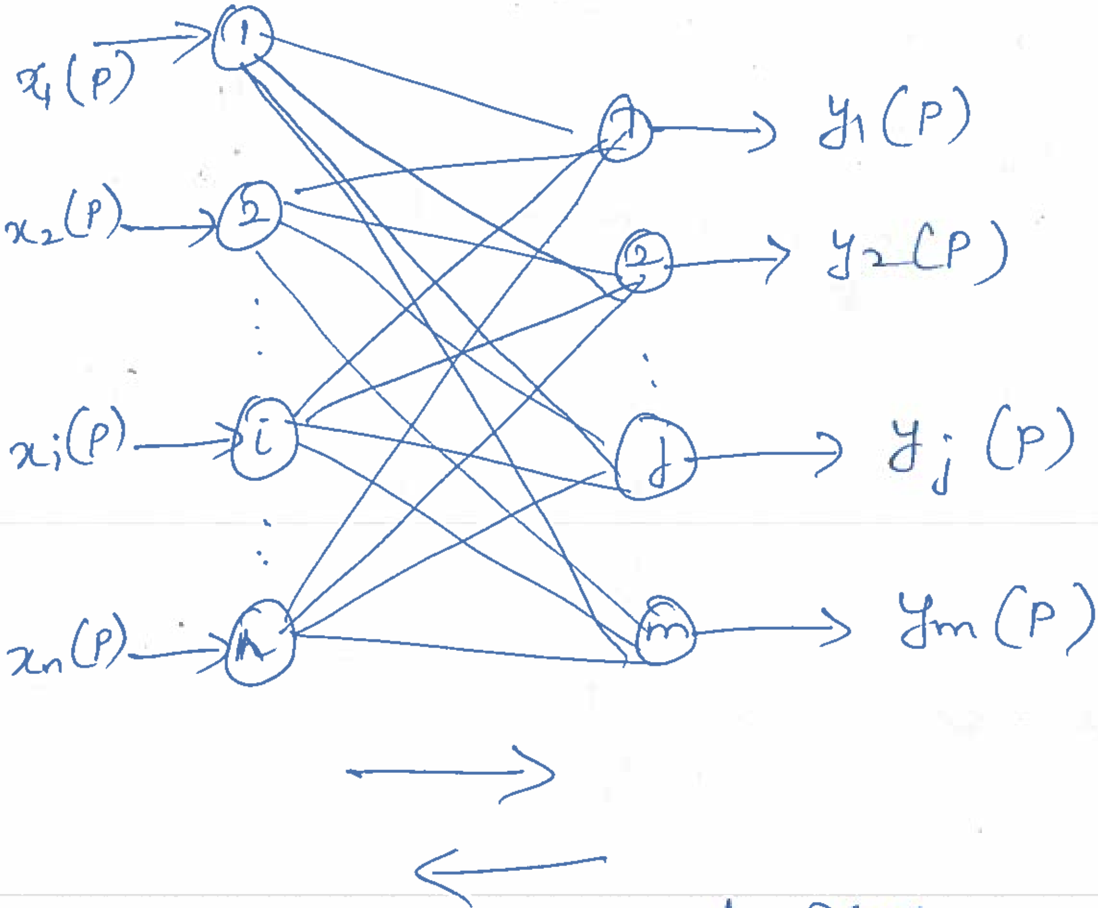
i) set $x_i = ~~x~~ s_i$
$y = t$

ii) Update weight
$$w_i(new) = w_i(old) + x_i y$$
$$b(new) = b(old) + y$$

AND Gate example (4)

(15) BAM - bidirectional associative memory (10)

(6)



(3)

Learning rules — forward pass

— backward "

Limitations of BAM. (3)