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2 MARK AND 16 MARK QUESTION BANK

Numerical Methods
UNIT- I
SOLUTION OF EQUATIONS AND EIGEN VALUE PROBLEMS.

1. What are the two types of errors involved in numerical computation?
Solution: i)Round off error and ii) Truncation error

2. Define Round off error
While dealing with decimal numbers, it is very inconvenient to work with all decimal places. So we take approximates to
facilitate calculation work. These approximations lead to error in final result known as Round off error

3. Define Truncation error
The error caused by using approximate formula in computations is known as Truncation error.

$$Example: \quad e = 1 + \frac{1}{1!} + \frac{1}{2!} + \frac{1}{3!} + ...$$

If we write $e = 1 + \frac{1}{1!} + \frac{1}{2!} + \frac{1}{3!}$ (approxiamtely) We get truncation error

4. In what form is the coefficient matrix transformed into when $AX = B$ is solved by Gauss elimination method.
Solution: Upper triangular matrix

5. When Gauss elimination method fails?
Solution: This method fails if the element in the top of the first column is zero. We can rectify this by interchanging
the rows of the matrix

6. In what form is the coefficient matrix transformed into when $AX = B$ is solved by Gauss Jordan method.
Solution: Diagonal matrix

WriteasufficientconditionforGaussSeidalmethod to converge.(or)
Stateasufficientcondition forGaussJacobimethod to converge.
Solution: Theprocess ofiterationbyGaussSeidalmethod willconvergeif in eachequation of thesystem theabsolute
value ofthelargestcoefficientisgreater than the sum oftheabsolute valuesof theremainingcoefficients

38. **Givetwo indirectmethod to solveasystem oflinear equations?**
Solution: i)GaussJacobimethod(ii) Gauss Seidalmethod

39. Explain theterm pivoting.
Solution: Intheeliminationprocess if anyoneofthe
pivotelementsa_{11} ,a_{22},........a_{nn}vanishesarebecomeverysmall comparedto otherelementsinthatcolumn ,then weattempttorearrangetheremainingrows soas to obtain anon
vanishingpivotorto avoid
themultiplicationbyalargenumber.thisstrategyiscalledpivoting.Thepivotingisoftwo
types 1) Partialpivoting 2) Complete pivoting.

14 What are elementary transforms?
Solution: Elementary transforms:

← Interchange the ith and jthb row.
← Multiply all the elements in the ith row by a number K.
← Adding the elements in the ith row to the corresponding elements in the jth row multiply by a constant K.

15 Explain briefly Gauss Jordan iteration to solve simultaneous equations.
Solution: Conisder the system of equations $AX=B$ If A is diagonal matrix the given system reduces to

$$\begin{bmatrix} a_{11} & 0 & \dots & \dots & 0 \\ 0 & a_{22} & \dots & \dots & 0 \\ \dots & \dots & \dots & \dots & \dots \\ 0 & 0 & \dots & \dots & a_{nn} \end{bmatrix} \begin{bmatrix} x_1 \\ x_2 \\ \dots \\ x_n \end{bmatrix} = \begin{bmatrix} b_1 \\ b_2 \\ \dots \\ b_n \end{bmatrix}$$

This system reduces to the following n equations

$a_{11}x_1 = b_1, a_{22}x_2 = b_2, \dots a_{nn}x_n = b_n$

Hence we get the solution directly as $x_1 = \frac{b_1}{a_{11}}, x_2 = \frac{b_2}{a_{22}}, \dots x_n = \frac{b_n}{a_{nn}}$ This method of obtaining the solution of equations by reducing the matrix A to a diagonal matrix is known as Gauss Jordan elimination method.

12. State TrueorFalse: "GaussSeidaliterationconverges onlyif thecoefficientmatrixisdiagonallydominant" **Solution:** True

13. Solve the following system of equations by Gauss Jordan method.

5x + 4y = 15, 3x + 7y =12 **Solution:** The given system is equivalent to $\begin{bmatrix} 5 & 4 \\ 3 & 7 \end{bmatrix} \begin{bmatrix} x \\ y \end{bmatrix} = \begin{bmatrix} 15 \\ 12 \end{bmatrix}$

i.e.,AX=B . Here $[A, B] = \begin{bmatrix} 5 & 4 & 15 \\ 3 & 7 & 12 \end{bmatrix}$ Now we will make the matrix A as a diagonal

matrix. $\begin{bmatrix} 5 & 4 & 15 \\ 0 & 23 & 15 \end{bmatrix} R_2 \leftrightarrow 5R_2 - 3R_1 \begin{bmatrix} 115 & 0 & 285 \\ 0 & 23 & 15 \end{bmatrix} R_1 \leftrightarrow 23R_1 - 4R_2$ $\therefore 115x = 285$

$\Rightarrow$ x=2.4783 , 23y=15 $\Rightarrow$ y=0.6522

42.State a sufficient condition for Gauss Jacobi method to converge.
Solution: Let the given equation be

$a_1x + b_1y + c_1z = d_1$, $a_2x + b_2y + c_2z = d_2$, $a_3x + b_3y + c_3z = d_3$

The sufficient condition is $|a_1| \geq |b_1| + |c_1|$, $|b_2| \geq |a_2| + |c_2|$, $|c_3| \geq |a_3| + |b_3|$

15. WhyGauss Seidalmethod is better than Jacobi'siterative method?
Solution : Sincethecurrentvaluesofthe unknowns ateach stageofiteration areusedinproceeding to thenextstage ofiteration, theconvergencein GaussSeidalmethod willbe morerapid than in GaussJacobimethod.

43. State True or FalseGaussSeidal iteration converges only if the coefficient matrix is diagonally dominant.Solution: True

44. Gauss elimination and Gauss Jordan are direct methods while …. and …… are iterative methods.
Solution: i) Gauss Seidal method ii) Gauss Jacobi method

18. Write the first iteration values of x,y,z when the equations $27x + 6y - z = 85,\ 6x + 15y + 2z = 72,\ x + y + 5z = 110$ are solved by Gauss-Seidal method.

Solution: $x = \frac{1}{27}[85 - 6y + z] = \frac{85}{27}$ [Putting $y = z = 0$]; $y = \frac{1}{15}[72 - 6x - 2z] = \frac{1}{15}\left[72 - 6\left(\frac{85}{27}\right)\right]$ [Putting z=0]

45. $= \frac{1}{5}[110 - x - y] = \frac{1}{5}\left[110 - \frac{85}{27} - \frac{1}{15}\left(72 - 6\left(\frac{85}{27}\right)\right)\right]$

19. Saytrueorfalse:The convergencein theGauss –Seidalmethod isthriceas fastas in jacobi’smethod.

Solution : False. The rate ofconvergenceofGauss Seidalmethod isroughlytwicethatoftheJacobi’smethod.

20. Compare Gauss elimination and Gauss Seidal methods.

Solution : Gauss elimination method has the advantage that it is finite and works in theory for any non singular set of equations.GaussSeidal iteration method converges only for special system of equations. For some systems, elimination is the only course available.In general, the round off error is smaller in the iteration methods, Iteration is a self correcting method. Any errors made at any step in the computation are corrected in the subsequent iterations.

. **Distinguish between direct and iterative methods of solving simultaneous equations.**

Solution: There are numerical methods of solving simultaneous equations. They are particularly suited for computer operations. These numerical methods are of two types, direct or iterative. Direct methods involve certain amount of fixed computation and they are exact Solutions. Iterative or indirect methods are those in which the Solution is got by successive approximations. But the method of iteration is not applicable to all systems of equations.

. **“In an iterative method, the amount of computation depends on the degree of accuracy required”.**

Say whether this is true or false?Solution: The statement is True

23. Compare Gauss Jacobi and Gauss Seidal method for solving linear systems of the form $AX = B$.

Solution : Gauss elimination is direct method

Gauss Seidal is iterative method.

24. What do you mean by diagonally dominant?

Solution: A matrix is diagonally dominant if the numerical value of the leading diagonal element in each row is greater than or equal to the sum of the numerical values of the other values of the other element in that row.

25. Define Eigen value and Eigen vector.

Solution: Let $A = \left[a_{ij} \right]$ be a square matrix of order n.

If there exists a non-zero(non-null) column vector X and a scalar λ such that AX= λX then λ is called an eigen value of the matrix A and X is called eigen vector corresponding to the eigen value λ.

49. State TrueorFalse: "GaussSeidaliterationconverges onlyif thecoefficientmatrixisdiagonallydominant"Solution : True.

50. Is theiterationmethod aself correctingmethod always?
Solution : Ingeneraliterationisaself correctingmethod sincethe round offerrorissmaller.

ii) **Statethe principle usedin Gauss Jordan method?**
Solution : Coefficientmatrixistransformed into diagonalmatrix.

iii) Thenumericalmethodsof solvinglinearequationsareoftwo types:oneisdirectandtheother is…………
Solution : iterative.

50. Compare Gauss Seidal andGausseliminationmethod? Solution :

S.No.	GaussJacobimethod	GaussSeidalmethod.
1.	Convergencemethod isslow	TherateofconvergenceofGauss Seidalmethodisroughlytwicethatof GaussJacobi.
2.	Directmethod	Indirectmethod
3.	Conditionforconvergenceis the coefficientmatrixdiagonallydominant	Conditionforconvergenceis the coefficientmatrixdiagonallydominant

31. For solving alinear system,compareGauss eliminationmethod andGaussJordanmethod.
Solution :

S.No.	Gauss eliminationmethod	Gauss Jordan method.
1.	Coefficientmatrixis transformed into uppertriangularmatrix	Coefficientmatrixistransformed into diagonalmatrix
2.	Directmethod	Directmethod
3.	Weobtain theSolutionbyback substitutionmethod	No need for substitution method

32. Distinguishbetween directanditerative method ofsolvingsimultaneousequation.
SolutionSolution:Solution: :

S.No.	Directmethod	Iterative method.

1.	WegetexactSolution	Approximate solution
2.	Simpletakelesstime	Timeconsuminglaborious
3.	This method determine all the roots at the same time	This method determine only one root at a time.

1. **Explain Gauss Seidal method to solve a system of simultaneous equations Solution:** Let the rearranged form of a given set of equations be

$x = \frac{1}{a_1}(d_1 - b_1 y - c_1 z)$ (1), $y = \frac{1}{b_2}(d_2 - a_2 x - c_2 z)$ (2), $z = \frac{1}{c_3}(d_3 - a_3 x - b_3 y)$ ....(3) We start with the

initial values $y^{(0)}, z^{(0)}$ for y and z and get $x^{(1)}$ from (1) i.e., $x^{(1)} = \frac{1}{a_1}\left(d_1 - b_1 y^{(0)} - c_1 z^{(0)}\right)$

While using (2), we use

$z^{(0)}$ for z and $x^{(1)}$ for $x^{(0)}$ and get $y^{(1)} = \frac{1}{b_2}\left(d_2 - a_2 x^{(1)} - c_2 z^{(0)}\right)$ Now having known

$x^{(1)}$ and $y^{(1)}$ we use

$x^{(1)}$ for x and y$^{(1)}$ for y in (3) and get $z^{(1)} = \frac{1}{c_3}\left(d_3 - a_3 x^{(1)} - b_3 y^{(1)}\right)$ The process may becontinued until

convergence is assured to all the solutions.

34.What are the merits of Newton's method of iteration?

Solution: It is applicable to the solution of equations involving algebraical functions as well as transcendental functions. It

is successfully used to improve the result obtained by the other methods.

35. What is the condition for the convergence of the iteration method for solving x = $\phi(x)$

Solution: $|\phi'(x)| < 1$ is the range

7. **Using solve Gauss elimination method x+ y=2 , 2x+3y=5. Solution:** x=1 ,y=1

37.State the iterative formula for method of false position to solve f(x) =0

Solution: If f(x) =0 has a root in the interval (a,b) then the approx $x = \frac{af(b) - bf(a)}{f(b) - f(a)}$

38.Solve by**Gauss elimination method 2x+ y=4 , x+2y=5.**

Solution: [A,B] = $\left[\begin{array}{cc|c} 1 & 2 & 5 \\ 2 & 1 & 4 \end{array}\right] \left[\begin{array}{cc|c} 1 & 2 & 5 \\ 0 & -3 & -6 \end{array}\right]$ **=>x+2y=5 ; -3y=-6**=>x=1 ,y=2

39.Find the dominant eigen value of A= $\begin{bmatrix} 1 & 2 \\ 3 & 4 \end{bmatrix}$ **by power method**

Solution: Let aninitial arbitrary vector be $X_1 = \begin{bmatrix} 0 \\ 1 \end{bmatrix}$

A $X_1 = \begin{bmatrix} 1 & 2 \\ 3 & 4 \end{bmatrix}\begin{bmatrix} 0 \\ 1 \end{bmatrix} = 4\begin{bmatrix} 0.5 \\ 1 \end{bmatrix} = 4X_2$; **A** $X_2 = \begin{bmatrix} 1 & 2 \\ 3 & 4 \end{bmatrix}\begin{bmatrix} 0.5 \\ 1 \end{bmatrix} = 7.5\begin{bmatrix} 1 \\ \frac{1}{3} \\ 1 \end{bmatrix} = 7.5X_3$; A $X_3 = \begin{bmatrix} 1 & 2 \\ 3 & 4 \end{bmatrix}\begin{bmatrix} 1 \\ \frac{1}{3} \\ 1 \end{bmatrix} = 5\begin{bmatrix} 7 \\ \overline{15} \\ 1 \end{bmatrix} = 5X_4$

$AX_4=\begin{bmatrix}1 & 2\\3 & 4\end{bmatrix}\begin{bmatrix}7\\15\end{bmatrix} = 5.4\begin{bmatrix}0.4567\\1\end{bmatrix} =5.4X_5$ 8. $AX_5=\begin{bmatrix}1 & 2\\3 & 4\end{bmatrix}\begin{bmatrix}0.4567\\1\end{bmatrix} = 5.3704\, X_6$; $AX_6=\begin{bmatrix}1 & 2\\3 & 4\end{bmatrix}\begin{bmatrix}0.4575\\1\end{bmatrix}$ $=5.3724X_7$; $AX_7=\begin{bmatrix}1 & 2\\3 & 4\end{bmatrix}\begin{bmatrix}\\1\end{bmatrix} =5.3723X_8$; $AX_8=\begin{bmatrix}1 & 2\\3 & 4\end{bmatrix}\begin{bmatrix}0.4574\\1\end{bmatrix}=5.3723\begin{bmatrix}0.4574\\1\end{bmatrix}$;

Hence eigen vector $X_1 = \begin{bmatrix}0.4574\\1\end{bmatrix}$

40.State the condition for the convergence of Gauss Seidel iterative method for solving system of equations.

Solution: $|a_{11}|>|a_{12}|+|a_{13}|,\ |a_{22}|>|a_{21}|+|a_{23}|,\ |a_{33}|>|a_{31}|+|a_{32}|$

41.Why Gauss Seidel iteration is a method of successive corrections?

Solution: Because we replace approximations by corresponding new ones as soon the later have been computed.

42.By Newton's method find an iterative formula to find $\sqrt{N}$ (where N is a positive number)

Solution: Let x $=\sqrt{N}$ $\therefore$ **x^2- N =0. Let f(x)=x^2- N** $\Rightarrow f'(x)=2x$

$$\alpha_{i+1}=\alpha_i-\frac{\alpha_i^2-N}{2\alpha_i}=\frac{1}{2}\left[\alpha_i+\frac{N}{\alpha_i}\right]$$

43. What type of eigen value can be obtained using power method

Solution: Power method is used to determine numerically largest eigen value and the corresponding eigen vector of a matrix A

44.State the Newton's formula and order of convergence of that method

Solution: $x_{i+1}=x_i-\dfrac{f(x_i)}{f'(x_i)}$ Here the convergence is quadratic and is of order 2

9. Write the convergence condition and order ofconvergence of Newton – Raphson method

Solution: Convergence condition is $|f(x)\, f''(x)|<|f'(x)|^2$ and order is 2

46.Write the Descartes rule of signs

Sol: 1)An equation $f(x) = 0$ cannot have more number of positive roots than there arechanges of sign in the terms of the polynomial $f(x)$.

2)An equation $f(x) = 0$ cannot have more number of positive roots than there arechanges of sign in the terms of the polynomial $f(x)$.

47.When would we not use N-R method .

Sol: If x_1 is the exact root and x_0 is its approximate value of the equation $f(x) = 0$.we know that $x_1 = x_0 - \dfrac{f(x_0)}{f'(x_0)}$ If $f'(x_0)$is small,the error 16. $\dfrac{f(x_0)}{f'(x_0)}$ will be large and the computation of the root by this,method will be a slow process or may even be impossible.Hence the method should not be used in cases where the graph of the function when itcrosses the x axis is nearly horizontal.

48.If g(x) is continuous in [a , b] then under what condition the iterative method x = g(x) hasa unique solution in [a , b].

Sol: Let x = r be a root of x = g(x) .Let I = [a , b] be the given interval combining thepoint x = r.if $|g'(x)|<1$ for all x in I, the sequence of approximation $x_0, x_1,\dots,x_n$ willconverge to the root r,provided that the initial approximation x_0 is chosen in r.

.49.In the case of fixed point iteration method ,the convergence is ………..

Sol: Linear.

18. If the eigen values of A are 1,3,4 then the dominant eigen value of A is ……

Sol:4

19. If the eigen values of A are 1,3,-4 then the dominant eigen value of A is ……

Sol:-4

20. If the eigen values of A are 1,3,-3 then the dominant eigen value of A is ……

Sol: No Dominant eigen value.

21. The power method will work satisfactory only if A has a ……..

Sol: Dominanteigen value.

- **Write down the procedure to find the numerically smallest eigen value of a matrix by power method**

Solution: By power method the largest eigen value of A^{-1} can be found then the smallest eigen value of A is the reciprocal of largest eigen value of A^{-1}

55.Determine the largest eigen value and the corresponding eigen vector of the matrix $\begin{pmatrix} 1 & 1 \\ 1 & 1 \end{pmatrix}$ correct to 2 decimal places using power method

Solution: $AX_1 = \begin{pmatrix} 1 & 1 \\ 1 & 1 \end{pmatrix}\begin{pmatrix} 1 \\ 1 \end{pmatrix} = \begin{pmatrix} 2 \\ 2 \end{pmatrix} = 2\begin{pmatrix} 1 \\ 1 \end{pmatrix} = 2X_2$; $AX_2 = \begin{pmatrix} 1 & 1 \\ 1 & 1 \end{pmatrix}\begin{pmatrix} 1 \\ 1 \end{pmatrix} = \begin{pmatrix} 2 \\ 2 \end{pmatrix} = 2\begin{pmatrix} 1 \\ 1 \end{pmatrix} = 2X_3$

Therefore the largest eigen value =2 and the corresponding eigen vector = $\begin{pmatrix} 1 \\ 1 \end{pmatrix}$

56.Can we find a real root of the equation cos x -2x +3=0 in $\left[0, \frac{\pi}{2}\right]$ by the method of iteration

Solution: Given: cos x -2x +3=0 => 2x= cos x + 3

$x = \frac{1}{2}[\cos x + 3] = f(x) \Rightarrow f'(x) = \frac{1}{2}[-\sin x]$

$|f'(x)| = \left|\frac{1}{2}\right|$ sinx is not less than 1 in $\left[0, \frac{\pi}{2}\right]$

Hence iterative method can not be applied

- **What is Newton's algorithm to solve the equations $x^2 = 12$**

Solution: let f(x) =x^2 -12 , $f'(x) = 2x$

By N.R rule, if x_r is the r^{th} iterate. $x_{r+1} = x_r - \frac{f(x_r)}{f'(x_r)} = x_r - \frac{x_r^2 - 12}{2(x_r)} = \frac{2x_r^2 - x_r^2 - 12}{2(x_r)} = \frac{x_r^2 + 12}{2(x_r)}$

23. To what kind of a matrix can the Jacobi's method be applied to obtain the eigen values of a matrix

Solution: Rotational matrix

59. Can we find a real root of the equation 2x-cos x =5 in $\left[0, \frac{\pi}{2}\right]$

Solution: Given: 2x-cos x =5 => 2x= cos x + 5

$x = \frac{1}{2}[\cos x + 5] = f(x) \Rightarrow f'(x) = \frac{1}{2}[-\sin x]$

$|f'(x)| = \left|\frac{1}{2}\right|$ sinx is not less than 1 in $\left[0, \frac{\pi}{2}\right]$

Hence iterative method can not be applied

UNIT –II
INTERPOLATON AND APPROXIMATION

1. Derive Newton's forward difference formula by using operator method.Solution:

$$P_n(x) = P_n(x_0 + uh) = E^u P_n(x_0) = E^u y_0 = (1+\Delta)^u y_0$$

$$= 1 + \frac{u}{1!}\Delta y_0 + \frac{u(u-1)}{2!}\Delta^2 y_0 + \frac{u(u-1)(u-2)}{3!}\Delta^3 y_0 + \ldots \quad \text{where } u = \frac{x - x_0}{h}$$

2. Derive Newton's forward difference formula by using operator method.Solution:

$$P_n(x) = P_n(x_n + vh) = E^v P_n(x_n) = (1-\nabla)^{-v} y_n \quad \text{since } E = (1-\nabla)^{-1}$$

$$= 1 + \frac{v}{1!}\nabla y_n + \frac{v(v+1)}{2!}\nabla^2 y_n + \frac{v(v+1)(v+2)}{3!}\nabla^3 y_n + \ldots \quad \text{where v} = \frac{x - x_n}{n}$$

3. State Gregory Newton's forward difference interpolation formula.Solution:

$$P_n(x) = P_n(x_0 + uh) = 1 + \frac{u}{1!}\Delta y_0 + \frac{u(u-1)}{2!}\Delta^2 y_0 + \frac{u(u-1)(u-2)}{3!}\Delta^3 y_0 + \ldots \quad \text{where } u = \frac{x - x_0}{h}$$

4. State Gregory Newton's backward difference interpolation formulaSolution:

$$P_n(x) = P_n(x_n + vh) = 1 + \frac{v}{1!}\nabla y_n + \frac{v(v+1)}{2!}\nabla^2 y_n + \frac{v(v+1)(v+2)}{3!}\nabla^3 y_n + \ldots \quad \text{where v} = \frac{x - x_n}{n}$$

5. When Newton's backward interpolation formula is used?Solution:
The formula is used mainly to interpolate the values of y near the end of a set of tabular values and also for extrapolating the values of y a short distance ahead (to the right) of y_0 .

27. Newton's forward difference interpolation formula is used only for …….. intervals.
Solution: Equidistant intervals (or) equal intervals.

28. Say True or False.Newton's interpolation formula are not suited to estimate the value of a function near the middleof a table.Solution: True

29. Say True or False.Newton's forward and backward interpolation formulae are applicable for interpolation nearthe beginning and end respectively of tabulated values.**Solution**: The statement is True

30. When Newton's forward interpolation formula is used?
Solution: The formula is used mainly to interpolate the values of y near the beginning of the table value and also for
extrapolating the values of y a short distance ahead (to the left) of y_0

. **If interpolation is required near the middle of the table, we use ……… formula.**
Solution: Stirling's formula

35. Using Lagrange's interpolation find the polynomial through (0,0), (1,1) and (2,2)

Solution : Lagrange's interpolation formula is

$$y=\frac{(x-x_1)(x-x_2)}{(x_0-x_1)(x_0-x_2)}y_0+\frac{(x-x_0)(x-x_2)}{(x_1-x_0)(x_1-x_2)}y_1+\frac{(x-x_0)(x-x_1)}{(x_2-x_0)(x_2-x_1)}y_2$$

$$=\frac{(x-1)(x-2)}{(0-1)(0-2)}(0)+\frac{(x-0)(x-2)}{(1-0)(1-2)}(1)+\frac{(x-0)(x-1)}{(2-0)(2-1)}(2)$$

$$=-x^2+2x+x^2-x=x$$

12. Stirling's formula is the average of two …….. formula

Solution: Gauss

13. Using Newton's forward difference formula, find $\sin(0.1604)$ **from the following table**

x	**0.160**	**0.161**	**0.162**
sinx	**0.1593182066**	**0.1603053541**	**0.1612923412**

Solution:

x	y	Δy	$\Delta^2 y$
0.160	0.1593182066		
		0.0009871475	
0.161	0.1603053541		-0.000001604
		0.0009869871	
0.162	0.1612923412		

There are only 3 data given. Hence the polynomial of degree 2.

$$y(x)=y_0+\frac{u}{1!}\Delta y_0+\frac{u(u-1)}{2!}\Delta^2 y_0 \quad \text{where } u=\frac{x-x_0}{h}$$

$x_0=0.160$, $h=0.161-0.160=0.001$, $x=0.1604$, $u=\frac{0.1604-0.160}{0.001}=0.4$

37.

$$y(x)=0.1593182066+(0.4)(0.0009871475)+\frac{(0.4)(0.4-1)}{2}(-0.000001604)$$

$$=0.1597130849$$

$\sin(0.1604)==0.1597130849$

14. What is the error in Newton's forward interpolation formula.

Solution: $f(x)-\phi_n(x)=\frac{r(r-1)(r-2)\ldots(r-n)}{(n+1)!}\Delta^{n+1}f(z)$

15. What is the error in Newton's backward interpolation formula.

Solution: $f(x)-\phi_n(x)=\frac{r(r+1)(r+2)\ldots(r+n)}{(n+1)!}h^{n+1}f^{n+1}(z)$

16. Obtain the interpolation quadratic polynomial for the given data by using Newton's forward difference formula.

x:	**0**	**2**	**4**	**6**
y:	**-3**	**5**	**21**	**45**

Solution:

x	y	Δy	$\Delta^2 y$	$\Delta^3 y$

0	-3			

		8		
2	5		8	
		16		0
4	21		8	
		24		
6	45			

$$n = \frac{x - x_0}{h} = \frac{x-0}{2} = \frac{x}{2}$$

$$y(x) = y_0 + n\Delta y_0 + \frac{n(n-1)}{2}\Delta^2 y_0 = -3 + \left(\frac{x}{2}\right)(8) + \frac{\left(\frac{x}{2}\right)\left(\frac{x}{2}-1\right)}{2}(8)$$

$$= -3 + 4x + \frac{\frac{x^2}{4} - \frac{x}{2}}{2}(8) = -3 + 4x + x^2 - 2x = x^2 + 2x - 3$$

38. What is the assumption we make when Lagrange's formula is used?

Solution: Lagrange's interpolation formula can be used whether the values of x, the independent variable are equally spaced or not whether the difference of y become smaller or not

18. Given $f(0) = -2, f(1) = 2$ and $f(2) = 8$. Find the root of the Newton's interpolating polynomial equation $f(x) = 0$

Solution:

x	y	Δy	$\Delta^2 y$
0	-2		
		4	
1	2		2
		6	
2	8		

There are only three datas given. Hence the polynomial of degree 2.

$$y(x) = P_2(x) = y_0 + \frac{u}{1!}\Delta y_0 + \frac{u(u-1)}{2!}\Delta^2 y_0 \quad \text{where } u = \frac{x - x_0}{h}$$

$x_0 = 0, h = 1 - 0 = 1, u = x$

Here $y(x) = -2 + \frac{x}{1!}(4) + \frac{x(x-1)}{2!}(2) = -2 + 4x + x^2 - x$

$= x^2 + 3x - 2$

The roots of the equation $f(x) = 0$

$i.e., x^2$ 44. $3x - 2 = 0$ $i.e., x = \frac{-3 \pm \sqrt{9+8}}{2} = \frac{-3 \pm \sqrt{17}}{2}$

1 State Lagrange's interpolation formula

Solution: Let $y = f(x)$ be a function which takes the values $y_0, y_1, y_2 \dots y_n$ corresponding to

51. $= x_0, x_1, \dots x_n$

Then Lagrange's interpolation formula is

$$y = f(x) = \frac{(x-x_1)(x-x_2)\dots(x-x_n)}{(x_0-x_1)(x_0-x_2)\dots(x_0-x_n)}y_0 + \frac{(x-x_0)(x-x_2)\dots(x-x_n)}{(x_1-x_0)(x_1-x_2)\dots(x_1-x_n)}y_1 + \dots\dots + \frac{(x-x_0)(x-x_1)\dots(x-x_{n-1})}{(x_n-x_0)(x_n-x_1)\dots(x_n-x_{n-1})}y_n$$

1. What is the Lagrange's formula to find y, if three sets of values $(x_0, y_0), (x_1, y_1)$ and (x_2, y_2) are given.

Solution: $y = \frac{(x-x_1)(x-x_2)}{(x_0-x_1)(x_0-x_2)}y_0 + \frac{(x-x_0)(x-x_2)}{(x_1-x_0)(x_1-x_2)}y_1 + \frac{(x-x_0)(x-x_1)}{(x_2-x_0)(x_2-x_1)}y_2$

2. What advantage has Lagrange's formula over Newton's?

Solution: The forward and backward interpolation formulae of Newton can be used only when the values of independent variable x are equally spaced can also be used when the differences of the dependent variable y become smaller ultimately. But Lagrange's interpolation formula can be used whether the values of x, the independent variable are equally spaced or not whether the difference of y become smaller or not.

22. Find the second degree polynomial fitting the following data.

x	1	2	4
y	4	5	13

Solution: Here $x_0 = 1, x_1 = 2, x_2 = 4$; $y_0 = 4, y_1 = 5, y_3 = 13$

By Lagrange's formula we get, $y = \frac{(x-x_1)(x-x_2)}{(x_0-x_1)(x_0-x_2)} y_0 + \frac{(x-x_0)(x-x_2)}{(x_1-x_0)(x_1-x_2)} y_1 + \frac{(x-x_0)(x-x_1)}{(x_2-x_0)(x_2-x_1)} y_2$

$$y = \frac{(x-2)(x-4)}{(-1)(-3)}(4) + \frac{(x-1)(x-4)}{(1)(-2)}(5) + \frac{(x-1)(x-2)}{(3)(2)}(13) = \frac{1}{6}\left(6x^2 - 12x + 30\right)$$

23. What is inverse interpolation?

Solution: Suppose we are given a table of values of and y. Direct interpolation is the process of finding the values of y corresponding to a value of x, not present in the table. Inverse interpolation is the process of finding the values of x corresponding to a value of y, not present in the table.

24. Construct a linear interpolating polynomial given the points (x_0, y_0) and (x_1, y_1)

Solution: $y = \frac{(x-x_1)}{(x_0-x_1)} y_0 + \frac{(x-x_0)}{(x_1-x_0)} y_1$

25. Find the polynomial which takes the following values

x	0	1	2
y	1	2	1

Solution:

x	y	Δy	$\Delta^2 y$
0	1		
		1	
1	2		-2
		-1	
2	1		

Newton' forward interpolation formula is

$$y(x) = y_0 + n\Delta y_0 + \frac{n(n-1)}{2}\Delta^2 y_0 = 1 + x(1) + \frac{x(x-1)}{2!}(-2) = 1 + 2x - x^2$$

4. What is the disadvantage in practice in applying Lagrange's interpolation formula?

Solution: Though Lagrange's formula is simple and easy, its application is not speedy. It requires close attention to sign and there are always a chance of committing some errors due to number of positive and negative sign in the numerator and the denominator.

5. Give the inverse of Lagrange's interpolation formula.

Solution:

$$x = \frac{(y-y_1)(y-y_2)\dots(y-y_n)}{(y_0-y_1)(y_0-y_2)\dots(y_0-y_n)}x_0 + \frac{(y-y_0)(y-y_2)\dots(y-y_n)}{(y_1-y_0)(y_1-y_2)\dots(y_1-y_n)}x_1 + \dots + \frac{(y-y_0)(y-y_1)\dots(y-y_{n-1})}{(y_n-y_0)(y_n-y_1)\dots(y_n-y_{n-1})}x_n$$

6. Use Lagrange's formula, to find the quadratic polynomial that takes these values.

x:	**0**	**1**	**3**
y:	**0**	**1**	**0**

Solution: Given $x_0 = 0, x_1 = 1, x_2 = 3$ and $y_0 = 0, y_1 = 1, y_2 = 0$

By Lagrange's interpolation formula

$$y = \frac{(x-x_1)(x-x_2)}{(x_0-x_1)(x_0-x_2)}y_0 + \frac{(x-x_0)(x-x_2)}{(x_1-x_0)(x_1-x_2)}y_1 + \frac{(x-x_0)(x-x_1)}{(x_2-x_0)(x_2-x_1)}y_2$$

8. $$\frac{(x-1)(x-3)}{(0-1)(0-3)}(0) + \frac{(x-0)(x-3)}{(1-0)(1-3)}(1) + \frac{(x-0)(x-1)}{(3-0)(3-1)}(0) = \frac{x(x-3)}{-2} = \frac{-1}{2}(x^2-3x)$$

29. Explain briefly about interpolation

Solution: The process of finding the value of the function inside the given range is called interpolation.

Interpolating function:

Let a set of tabular values of a function $y = f(x)$ where the explicit nature of the function is not known, then $f(x)$ is replaced by a simpler function $\phi(x)$ such that $f(x)$ and $\phi(x)$ agree with the set of tabulated points. Any other value may be calculated from $\phi(x)$. This function $\phi(x)$ is known as an interpolating function.

∴ **Can you use Lagrange's interpolation formula when the intervals are equal?**
Solution: Yes.

∴ **What is a cubic spline?**

Solution: A cubic polynomial which has continuous slope and curvature is called a cubic spline.

10. Find the polynomial for the following data by Newton's backward difference formula.

x	**0**	**1**	**2**	**3**
f(x)	**-3**	**2**	**9**	**18**

Solution:

x	$f(x)$	∇f	$\nabla^2 f$	$\nabla^3 f$
0	-3			
		5		
1	2		2	
		7		0
2	9		2	
		9		
3	18			

Newton's backward difference formula is

1. $$y(x) = f_n + p\nabla f_n + \frac{p(p+1)}{2!}\nabla^2 f_n + \dots \text{ where } p = \frac{x-x_n}{h} = x-3$$

$$f(x)=18+(x-3)(9)+\frac{(x-3)(x-3+1)}{2!}(2)=18+9x-27+\frac{(x-3)(x-2)}{2!}(2)=x^2+4x-3$$

33. What is a natural cubic spline?

Solution : A cubic spline fitted to the given data such that the end cubics approach linearity at their extremities is
called a natural cubic spline.

3. **Define acubicspline S(x) whichiscommonlyusedfor interpolation**.
   Solution : Wedefineacubic,S(x) asfollows:

i)S(x) isapolynomialof degreeone for X <X0 andX>Xn

ii)S(x) isatmostacubicpolynomialineachinterval(xi-1,xi) ,i=1,2,3,…….,n

iii)S(x),S'(x) andS''(x) are continuousat eachpoint (xi,yi), i=0,1,2,…….nand

iv)S(xi)=yi i =0,1,2,3………n

35. Write the endconditionsonMi(x)innaturalcubicsplines

Solution: M0(x)=0,Mn(x)=0

36. Writethe relationbetweenthe secondderivativesMi(x)incubicsplineswithequalmeshspacing.

Solution:Mi-1+4Mi +Mi+1=6/h^2 [yi-1– 2yi +yi+1],i =1,2,3,…..n-1

4. Write the difference between **Lagrange's formula and Newton's forward difference formula**
   Solution:

S.No.	**Lagrange's formula**	**Newton's forward difference formula**
1.	Interval of differencing need not be uniform	Interval of differencing should be uniform
2.	Used to find f(x) at any place	Used to find f(x) at the beginning
3.	$y=f(x)=\frac{(x-x_1)(x-x_2)\dots(x-x_n)}{(x_0-x_1)(x_0-x_2)\dots(x_0-x_n)}y_0+\frac{(x-x_0)(x-x_2)\dots(x-x_n)}{(x_1-x_0)(x_1-x_2)\dots(x_1-x_n)}y_1+\dots\dots+\frac{(x-x_0)(x-x_1)\dots(x-x_{n-1})}{(x_n-x_0)(x_n-x_1)\dots(x_n-x_{n-1})}y_n$	$y(x)=y_0+\frac{u}{1!}\Delta y_0+\frac{u(u-1)}{2!}\Delta^2 y_0$ where $u=\frac{x-x_0}{h}$

38.A third degree polynomial passes through (0,-1), (1,1) ,(2,1) and (3,-2) find its value at x=4?

Solution: -9

39. Define the terms interpolation and extrapolation

Solution: Interpolation is the technique of estimating the value of a function for any intermediate value at the independent variable while the process of computing the value of a function outside the given range is called extrapolation

40.Write the Newton's divided difference interpolation formula for unequal intervals.

Solution: f(x)= f(x_0) +(x-x_0)f(x_0,x_1)+(x-x_0)(x-x_1)f(x_0,x_1,x_2) +…+ (x-x_0)(x-x_1)…(x-x_{n-1}) f(x_0,x_1,x_2,…,x_n)

41. Write the formula for $\frac{dy}{dx}$ at x= x_n using backward difference operator

Solution: $$\left(\frac{dy}{dx}\right)_{x=x_n}=\frac{1}{h}\left[\nabla y_n+\frac{1}{2}\nabla^2 y_n+\frac{1}{3}\nabla^3 y_n+\dots\right]$$

42.Obtain the divided difference table for the following data:

x	**2**	**3**	**5**
Y	**0**	**14**	**102**

Solution:

x	y	Δy	$\Delta^2 y$
2	0		
		14	
3	14		10
		44	
5	102		

43. Obtain the divided difference table for the following data:

x	**-1**	**0**	**2**	**3**
f(x)	**-8**	**3**	**1**	**12**

Solution:

6.	$f(x)$	$\Delta f(x)$	10. $^2 f(x)$	$\Delta^3 f(x)$
9. 1	-8			
		11		
0	3			
		-1	-4	2
2	1			
		11	4	
3	12			

44.What is the nature of n^{th} divided difference of a polynomial of n^{th} degree?

Solution: The n^{th} divided difference of a polynomial of n^{th} degree are **constants**

45.Find the second divided differences with arguments a, b, c if $f(x) = \frac{1}{x}$

Solution : If $f(x) = \frac{1}{x} \Rightarrow f(a) = \frac{1}{a}$

$$f(a,b) = \Delta\left[\frac{1}{a}\right] = \frac{\frac{1}{b} - \frac{1}{a}}{b-a} = \frac{-1}{ab} \text{ ; } f(a,b,c) = \Delta^2\left[\frac{1}{a}\right] = \frac{f(b,c) - f(a,b)}{c-a} = \frac{1}{abc}$$

1. **Form the divided difference table for the data (0,1) ,(1,4),(3,40), (4,85) Solution** :

x	$f(x)$	$\Delta f(x)$	$\Delta^2 f(x)$	$\Delta^3 f(x)$
0	1			
		3		
1	4		5	
		8		0.44
3	40		6.75	

4	85	45		

47.Obtain the divided difference table for the following data:

x	5	15	22
Y	7	36	160

Solution:

2.	y	Δy	1.
5	7		
		2.9	
z	36		$\frac{14.8}{17}$
		17.7	
y	160		

$2y$

If y(x) = y, i= 0,1,…,n write down the for the formula for the cubic spline polynomial y(x) valid in $x_{i-1}< x< x_i$

Solution : y(x)= $\frac{1}{6}\left[(x-x_i)^3 M_{i-1} + (x-x_{i-1})^3 M_i\right] + (x-x_i)\left[y_{i-1} - \frac{1}{6}M_{i-1}\right] + (x-x_{i-1})\left[y_i - \frac{1}{6}M_i\right]$

49.When to use Newton's forward interpolation and when to use Newton's backward interpolation?

Solution :(i) The formula is used to interpolate the values of y near the beginning of the table value and also for extrapolating
the values of y short distance ahead (to the left of)y_0

(ii)The formula is used to interpolate the values of y near the end of a set of tabular values and also for extrapolating
the values of y short distance ahead (to the right of)y_0

50. Simplify $\Delta^2 y_2 - \Delta y_1 - E^2 y_0$

Solution :

$\Delta^2 y_2 - \Delta y_1 - \frac{y_0}{E} = \Delta(\Delta y_2) - (y_2 - y_1) - y_2 = \Delta(y_3 - y_2) - 2y_2 + y_1 = y_4 - 2y_3 - y_2 + y_1$

Which metods are used for finding the polynomial if the intervals are unequal?

Solution : Divided difference

52.Show that the divided differences are symmetrical in their arguments.

Solution : $f(x_0, x_1) = \frac{f(x_1) - f(x_0)}{x_1 - x_0} = \frac{f(x_0) - f(x_1)}{x_0 - x_1} = f(x_1, x_0)$

53.Evaluate $\Delta^{10}(1-x)(1-2x)(1-3x)...(1-10x)$ **by taking h=1**

Solution : $\Delta^{10}(1-x)(1-2x)(1-3x)...(1-10x)$

$= \Delta^{10}(10!\,x^{10}$ + terms involving lesser degree$) = 10!10! = 0 = (10!)^2$

54.Show that the divided difference operator Δ is linear

Solution : Let α *and* β are 2 constants f(x) and g(x) are 2 functions then

$$\Delta[\alpha f(x)+\beta g(x)] = \frac{[\alpha f(x_1)+\beta g(x_1)] - [\alpha f(x_0)+\beta g(x_0)]}{x_1 - x_0}$$

$$\alpha \frac{f(x_1)-f(x_0)}{x_1-x_0} + \beta \frac{g(x_1)-g(x_0)}{x_1-x_0} = \alpha \Delta f(x) + \beta \Delta g(x)$$

55.State any 2 properties of divided difference

Solution : (i) The divided differences are symmetrical in all their arguments. i.e., the value of any difference is independent of the order of the arguments

Thedivided differences of the sum or difference of 2 functions is equal to the sum or difference of the corresponding separate divided difference

UNIT- III
NUMERICAL DIFFERENTIATION AND INTEGRATION

1. **State Newton's formula to find $f'(x)$ using the forward differences.**

Solution : $\left(\frac{dy}{dx}\right)_{x=0} = f'(x_0) = \frac{1}{h}\left[\Delta y_0 - \frac{\Delta^2 y_0}{2} + \frac{\Delta^3 y_0}{3} \ldots\right]$

2. **Using forward differences, write the formula for $f'(a)$**

Solution: $f'(a) = \frac{1}{h}\left[\Delta f(a) - \frac{\Delta^2 f(a)}{2} + \frac{\Delta^3 f(a)}{3} - \ldots\right]$

3. **Numerical differentiation can be used only when the difference of some order…….**

Solution: Numerical differentiation can be used only if it is clear from the tabulated values that the differences of some order are constant.

4. **Why is Trapezoidal rule so called?**

Solution: The Trapezoidal rule is so called, because it approximates the integral by the sum of n trapezoids.

Find $\frac{dy}{dx}$ at $x=1$ from the following table.

x:	1	2	3	4
y:	1	8	27	64

Solution:

The forward difference table is as follows:

x	y	Δy	$\Delta^2 y$	$\Delta^3 y$
1	1			
		7		
2	8		12	
		19		6
3	27		18	
		37		
4	64			

$$\left(\frac{dy}{dx}\right)_{x=0} = f'(x_0) = \frac{1}{h}\left[\Delta y_0 - \frac{\Delta^2 y_0}{2} + \frac{\Delta^3 y_0}{3} \ldots\right] \quad \text{……..(1)}$$

Here $h = 1, x_0 = 1, \Delta y_0 = 7, \Delta^2 y_0 = 12, \Delta^3 y_0 = 6$

Substituting the values in (1) we get,

$$\left(\frac{dy}{dx}\right)_{x=1} = \frac{1}{1}\left(7 - \frac{12}{2} + \frac{6}{3}\right) = 3$$

Using Newton's backward difference formula, write the formulae for the first and second order derivatives at the end value $x = x_n$ **upto the fourth order difference term.**

Solution:

$$\left(\frac{dy}{dx}\right)_{x=x_n} = \frac{1}{h}\left[\nabla y_n + \frac{1}{2}\nabla^2 y_n + \frac{1}{3}\nabla^3 y_n + \ldots\right]$$

$$\left(\frac{d^2y}{dx^2}\right)_{x=x_n} = \frac{1}{h^2}\left[\nabla^2 y_n + \nabla^3 y_n + \frac{11}{12}\nabla^4 y_n + \ldots\right]$$

7. **If** $f(x) = a^x\ (a \neq 0)$ **is given for** x **=0,0.5,1 show by numerical differentiation that** $f'(0) = 4\sqrt{a} - a - 3$

Solution:For $x = 0$, 0.5, 1 the values of $f(x) = a^x$ are respectively a^0, $a^{0.5}$, a^1 i.e., $1, \sqrt{a}, a$

The formal difference table is as follows:

Y		Δ	Δ^2
01			
		$\sqrt{a}$ -1	
0.5	$\sqrt{a}$		a-$2\sqrt{a}$+1
1	a	a-$\sqrt{a}$	

$$\left(\frac{dy}{dx}\right)_{x=0} = f'(x_0) = \frac{1}{h}\left[\Delta y_0 - \frac{\Delta^2 y_0}{2} + \frac{\Delta^3 y_0}{3} \ldots\right] \ldots..(1)$$

Here $h = 5$, $x_0 = 0$, $\Delta y_0 = \sqrt{a} - 1$, $\Delta^2 y_0 = a - 2\sqrt{a} + 1$, $\Delta^3 y_0 = 0$

Substituting the values in (1) we get,

$$f'(0) = \frac{1}{5}\left[\sqrt{a} - 1 - \frac{1}{2}\left(a - 2\sqrt{a} + 1\right)\right] = 2\left[\sqrt{a} - 1 - \frac{a}{2} + \sqrt{a} - \frac{1}{2}\right] = 4\sqrt{a} - a - 3$$

By differentiating the Newton's backward difference formula, find the first derivative of the function (x) .

Solution: We know that the Newton's backward difference interpolation formula is

$$y(x) = y(x_n + nh) = y_n + \frac{n}{1!}\nabla y_n + \frac{n(n+1)}{2!}\nabla^2 y_n + \frac{n(n+1)(n+2)}{3!}\nabla^3 y_n + \ldots \quad \text{where } n = \frac{x - x_n}{h}$$

Differentiating w.r.to x we get

$$y'(x) = \left(\frac{dy}{dx}\right) = \frac{1}{h}\left[\nabla y_n + \left(\frac{2n+1}{2}\right)\nabla^2 y_n + \left(\frac{3n^2 + 6n + 2}{3}\right)\nabla^3 y_n + \ldots\right]$$

9. **Using Trapezoidal rule evaluate** $\int_0^{\pi} \sin x dx$ **by dividing the range into 6 equal parts.**

Solution:

x	0	$\pi/6$	$2\pi/6$	$3\pi/6$	$4\pi/6$	$5\pi/6$	π
y	0	0.5	0.8660	1	0.8660	0.5	0

$$\int_0^{\pi} \sin x dx = \frac{h}{2}\left[\left(y_0 + y_6\right) + 2\left(y_1 + y_2 + y_3 + y_4 + y_5\right)\right]$$

$$= \frac{\pi}{36}\left[\left(0 + 0\right) + 2\left(0.5 + 0.8660 + 1 + 0.8660 + 0.5\right)\right] = 0.65136$$

10. What is the Geometrical interpretation of trapezoidal rule?

Solution: We are finding the area of the curve enclosed by $y=f(x)$, the X axis, the ordinates x=a and x=b
by using the area of trapezium.

11. Find the error in the derivative of $f(x)=\cos x$ by computing directly and using the approximation $f(x)=\dfrac{f(x+h)-f(x-h)}{2h}$ at x=0.8 choosing h=0.01.

Solution: Given $f x)=\cos x \quad \cos(0.8) \quad =0.999902$. Here x=0.8 and h=0.01

$$(0.8)=\frac{(0.8+0.01)-f(20.8-0.01)}{(0.01)}=-0.000211782$$

12. When can numerical differential be used.

Solution: When the function is given in the form of table of values instead of giving analytical expression we use numerical differentiation.

13. Write NewtonCotes quadratic formula.

Solution: $$\int_{x_0}^{x_0+nh} y(x)\,dx=h\left[ny_0+\frac{n^2}{2}\Delta y_0+\frac{1}{2}\left(\frac{n^3}{3}-\frac{n^2}{2}\right)\Delta^2 y_0+\frac{1}{6}\left(\frac{n^4}{4}-n^3+n^2\right)\Delta^3 y_0+\ldots\right]$$

14. Evaluate $\int_1^4 f(x)dx$ from the table by Simpson's three eighth rule.

x	1	2	3	4
f(x)	1	8	27	64

Solution :

$$\int_1^4 f(x)dx=\frac{3h}{8}\left[f_0+3f_1+3f_2+f_3\right]=\frac{3}{8}\left[+3(8)+3(27)+64\right]=63.75$$

15. State Romberg's method of integration formula to find the value of $I=\int_a^b f(x)dx$ using h and h/2.

Solution: Let $I=\int_a^b f(x)dx$. Let I_h and $I_{h/2}$ be the values of I with the sub intervals h and $\frac{h}{2}$.
Then I can be
improved by the formula $I(h,h/2)=\frac{1}{3}[4I_{h/2}-I_h]$.

Compare Trapezoidal rule and Simpson's 1/3 rule for evaluating numerical integration.

Solution: In Trapezoidal rule, y is a linear function of x and it is least accurate result. In Simpson's 1/3 rule, y is a polynomial of degree two, and it is more accurate result.

17. How the accuracy can be increased in trapezoidal rule of evaluating a given definite integral?

Solution: If the number of points of the base segment b-a (the range of integration) is increased, a better
approximation to the area given by the definite integral will be obtained.

Using Trapezoidal rule, find $f \int_0^6 f(x)dx$ from the following set of values of x and (x)

x:	0	1	2	3	4	5	6
$f(x)$:	1.56	3.64	4.62	5.12	7.08	9.22	10.44

Solution: Here $h = 1, y_1 = 1.56, y_2 = 3.64$ etc.

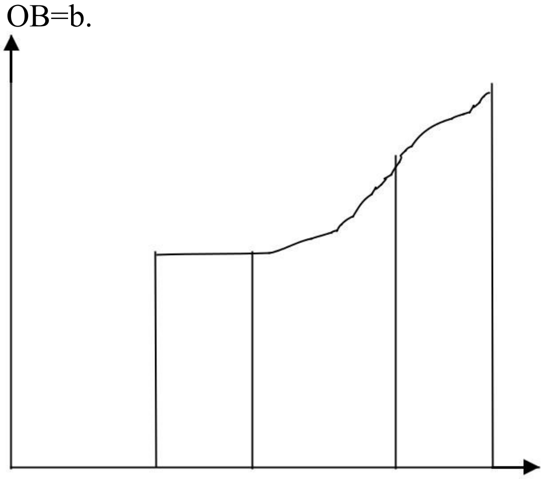
A=Sum of the first and last ordinates = 1.56 + 10.44 =12

B = Sum of the remaining ordinates = 3.64 + 4.62 + 5.12 + 7.08 + 9.22 = 29.68

$$\int_{1/2}^{1} \frac{1}{x}dx = \frac{h}{2}(A + 2B) = \frac{1}{2}(12 + 2 \times 29.68) = 35.68$$

19. State the Trapezoidal rule to evaluate $\int_a^b f(x)dx$

Solution: Let DC be the curve $y = f(x)$ and DA , CB be the terminal ordinates. Let OA=a and OB=b.



Then AB=OB-OA=b-aDivide AB into n equal parts $AA_1, AA_2, \ldots A_{n-1}B$ so that each part = $\frac{b-a}{n} = h$ (say). Draw the ordinates through $AA_1 AA_2, \ldots \tilde{An}\, 1B$ and let that can be $y_1, y_2, \ldots y_n, y_{n+1}$ respectively.Then $\int_{x_0}^{x} f(x)dx = \frac{h}{2}(A\ 2B)$ called approximately...... (1)Where

$A = y_1 + y_{n+1}$ = sum of the first and last ordinates and $B = y_2 + y_3 + \ldots + y_n$ =sum of the remaining ordinates (1) is known as Trapezoidal rule.

20. Evaluate $\int_{\frac{1}{2}}^{1} \frac{1}{x} dx$ **by Trapezoidal rule, dividing the range into 4 equal parts.**

Solution:

Here $h = \frac{1 - 1/2}{4} = \frac{1}{8}$; $y = \frac{1}{x}$

Here $x_0 = \frac{1}{2} = \frac{4}{8}, x_n = 1, x_1 = \frac{5}{8}, x_2 = \frac{6}{8}, x_3 = \frac{7}{8}, x_4 = \frac{8}{8}$; $y_1 = \frac{1}{4}, y_2 = \frac{1}{5}, y_3 = \frac{8}{6}, y_4 = \frac{8}{7}, y_5 = \frac{8}{8}$

A=Sum of the first and last ordinates = $\frac{8}{4} + \frac{8}{8} = 3$

B = Sum of the remaining ordinates = $\frac{8}{5}+\frac{8}{6}+\frac{8}{7}=\frac{856}{210}$

$$\int_{1/2}^{1}\frac{1}{x}dx=\frac{h}{2}(A+2B)=\frac{1}{16}\left(3+\frac{856 X 2}{210}\right)=\frac{1171}{1680}=0.6971$$

When does Simpson's rule give exact result?

Solution: Simpson's rule will give exact result , if the entire curve $y=f(x)$ is itself a parabola.

What is the general Newton-cotes quadrature formula? How is the trapezoidal rule its special case?

Solution: The general Newton-cotes quadrature formula is

$$\int_{a}^{a+nh} f(x)dx = nh\left[1+\frac{n\Delta}{2}+\frac{n(2n-3)\Delta^2}{12}+\frac{n(n-2)^2\Delta^3}{24}+...\right]f(a)$$ This is also known as the general Gauss-Legendre integration formula. Putting n=1 and omitting the second and hyper differences in the above, we get

$$\int_{a}^{a+h} f(x) = \frac{h}{2}\left[f(a)+f(a+h)\right] = \frac{h}{2}(y_1+y_2)$$ which is the Trapezoidal rule.

23. What is the order of error in Trapezoidal formula.

Solution: Error in Trapezoidal formula is of order h^2

What is the order of error in Simpson's formula.

Solution: Error in Simpson's formula is of order h^4

25. State Trapezoidal rule to evaluate $\int_{x_0}^{x_n} f(x)dx$

Solution: $\int_{x_0}^{x_1} f(x)dx = \frac{h}{2}\left[(y_0+y_n)+2(y_1+y_2+...y_{n-1})\right]$

For what type of functions, Simpson's rule and direct integration will give the same result?

Solution: Simpon's rule will give exact result, if the entire curve $y=f(x)$ is itself a parabola.

Error in Simpson's rule is of order ……Solution: h^4

Six set of values of x and y are given (being equally spaced). Write the formula to get $\int_{x_1}^{x_6} ydx$ **.**

Solution: $\int_{x_1}^{x_6} ydx = \frac{h}{2}\left[(y_1+y_6)+(y_2+y_3+y_4+y_5)\right]$

Which one is more reliable, Simpson's rule or Trapezoidal rule

Solution: Simpson's rule

What are the errors in Trapezoidal and Simpson's rules of numerical integration?

Solution: Error in Trapezoidal rule $|E| < \frac{(b-a)h^2}{12}M$ in the interval $(a, b), h = \frac{b-a}{n}$.

Error in Simpson's rule $|E| < \frac{(b-a)}{180}h^4 M$.

31. In order to evaluate $\int_{x_0}^{x_n} y\,dx$ by Simpson's $\frac{1}{3}$ rule as well as by Simpson's $\frac{3}{8}$ rule, what is the restriction in the number of intervals?

Solution: Let n = interval For using Simpson's $\frac{1}{3}$ rule, the number of ordinates is odd (or) the intervals number is even.For Simpson's $\frac{3}{8}$ rule, n is a multiple of 3.

32. State True or False:
Whenever Trapezoidal rule is applicable Simpson's rule can be applied.
Solution: False.

33. Using Trapezoidal rule evaluate $\int_0^{\pi} \sin x\,dx$ by dividing the range into 6 equal parts.

Solution: Here $y(x) = \sin x,\ h = \frac{\pi}{6}$

x	0	$\frac{\pi}{6}$	$\frac{2\pi}{6}$	$\frac{3\pi}{6}$	$\frac{4\pi}{6}$	$\frac{5\pi}{6}$	π
Y	0	0.5	0.866	1	0.866	0.5	0

$$\int_0^{\pi} \sin x\,dx = \frac{h}{2}\left[(y_0 + y_n) + 2(y_1 + y_2 + \dots + y_{n-1})\right] = \frac{\pi/6}{2}\left[(0+0) + 2(0.5 + 0.866 + 1 + 0.866 + 0.5)\right]$$

$$\frac{\pi}{12}[7.464] = 0.622\pi$$

What approximation is used in deriving Simpson's rule of integration?
Solution: Simpson's one third rule approximates the area of two adjacent strips by the area under a quadratic parabola.

35. Write the Trapezoidal rule to evaluate $\int_1^6 f(x)\,dx$ with $h = 0.5$.

Solution: Here $f(x) = h = 0.5$

x	1	1.5	2	2.5	3	3.5	4	4.5	5	5.5	6
$y=f(x)$	f(1)	f(2)	f(3)	f(4)	f(5)	f(6)	f(7)	f(8)	f(9)	f(10)	f(11)

By Formula $\int_1^6 f(x)\,dx = \frac{h}{2}\left[(\text{Sum of first and last ordinates}) + 2(\text{remaining ordintes})\right]$

$$\frac{0.5}{2}\left[(f(1)+f(11)) + 2(f(2)+f(3)+f(4)+f(5)+f(6)+f(7)+f(8)+f(9)+f(10))\right]$$

The velocity of a particle which starts from rest is given by the following table:

t (sec)	0	2	4	6	8	10	12	14	16	18	20
V(ft/sec)	0	16	29	40	46	51	32	18	8	3	0

Estimate using Trapezoidal rule the total distance traveled in 20 sec.

Solution: We know $v = \frac{ds}{dt}$ where s is the distance traveled. $\therefore ds = vdt \Rightarrow s = \int vdt$

Therefore total distance traveled in 20 seconds is $s = \int_0^{20} vdt$

Using Trapezoidal rule

$$\int_0^{20} vdt = \frac{h}{2}\left[(y_0 + y_{10}) + 2(y_1 + y_2 + \ldots + y_9)\right] = \frac{2}{2}\left[0 + 2(16 + 29 + 40 + 46 + 51 + 32 + 18 + 8 + 3)\right] = 486 \text{ feet}$$

37. Using Simpson's rule find $\int_0^4 e^x dx$ **, given** $e^0 = 1, e^1 = 2.72, e^2 = 7.39, e^3 = 20.09$ and $e^4 = 54.6$

Solution: Let $= e^x$ and h = 1 by Simpson's rule

$$\int_0^4 e^x dx = \frac{h}{3}\left[(y_0 + y_4) + 2y_2 + 4(y_1 + y_3)\right] = \frac{1}{3}\left[(1 + 54.6) + 2(7.39) + 4(2.72 + 20.09)\right] = 53.8733$$

From the following table find the area bounded by the curve and the x axis from $= 2$ to $x = 7$

x	2	3	4	5	6	7
(x)	8	27	64	125	216	343

Solution: Here h = 1 and only 6 ordinates are given. Therefore we use Trapezoidal rule

$$Area = \int_2^7 ydx = \frac{h}{2}\left[(y_0 + y_5) + 2(y_1 + y_2 + y_3 + y_4)\right] = \frac{1}{2}\left[(8 + 343) + 2(27 + 64 + 125 + 216)\right] = 607.5 \text{ sq.units}$$

39. What are the errors involved in Simpson's $\frac{1}{3}, \frac{3}{8}$ **rules for the evaluation of a integral of the form** $\int_a^b f(x)dx$ **?**

Solution:

Rule	Error	Order
Simpson's $\frac{1}{3}$ rule	$\lvert E \rvert < \frac{(b-a)h^4}{180} M$	h^4
Simpson's $\frac{3}{8}$ rule	$\frac{-3}{8} h^5 y^{iv}(\xi)$	h^5

Why Simpson's one third rule is called a closed formula?

Solution: Since the end point ordinates y_0 and y_n are included in the Simpson's is $\frac{1}{3}$ rule, it called closed formula.

What approximation is used in deriving Simpson's rule of integration.

Solution: Simpson's $\frac{1}{3}$ rule approximates the area of two adjacent strips by the area under a quadratic parabola.

If $I_1 = 0.7083$ **and** $I_2 = 0.6970$ **find** I **using Romberg's method.**

Solution: Romberg's formula $I = I_2 + \left[\frac{I_2 - I_1}{3}\right] = 0.6970 + \frac{0.6970 - 0.7083}{3} = 0.6932$

43. State Trapezoidal rule for evaluating $\int_a^b \int_c^d f(x, y)\,dxdy$.

Solution: I = $\frac{hk}{4}$ [(Sum of values of *f* at the four corners)+2(Sum of values of *f* at the remaining nodes of the boundary)+4(Sum of values of *f* at the interior nodes)]

State Simpson's rule for evaluating $\int_a^b \int_c^d f(x, y)\,dxdy$.

Solution: I = $\frac{hk}{9}$ [(Sum of values of *f* at the four corners)+2 (Sum of values of *f* at the odd positions on the boundary except the corners)+4(Sum of values of *f* at the even positions on the boundary)+{4(Sum of values of *f* at the odd positions)+8(Sum of values of *f* at the even positions) on the odd row of the matrix except boundary rows}+{8(Sum of values of *f* at the odd positions)+ 16 (Sum of values of *f* atthe even positions) on the even rows of the matrix}]

45.Evaluate $\int_0^1 \frac{dx}{1+x}$ **with h=0.5 using Trapezoidal rule**

Solution: $\int_0^1 \frac{dx}{1+x}$

x:	1	2	3	4
y:	1	8	27	64

Trapezoidal rule

$$= \frac{h}{2}\left[(y_n + y_{n+1}) + 2(y_2 + y_n)\right] = \frac{5}{2}[(1+0.5)+2(0.667)] = 0.708$$

45. Obtain the divided difference table for the following and hence find f(9)

x:	-	0	2	5	10
y:	-2	-1	7	124	999

Solution:

x	y				
-1	-2				
		1			
0	-1		7		
		4		1	0
2	7		7		
		39		1	
5	124		17		
		175			
10	999				

Here, $f(x_0, x_1) = 1$

$f(x_0, x_1, x_2) = 1$

$f(x_0, x_1, x_2, x_3) = 1$

$f(x_0)+(x-x_0)+f(x_0,x_1)+\ldots+(x-x_0)(x-x_1)(x-x_3)f(x_0, x_1, x_2, x_3)$

$-2+(9+1)(1)+(9+1)(9-0)(1)+(9+1)(9-0)(9-2)(1)= 728$

What is the order of error in Trapezoidal rule ?

Solution: The error in Trapezoidal rule is of order h^2

47 Compute $\int_{-2}^{2} e^{-x^2/2}dx$ **using Gaussian 2 point formula**

Solution: By Gaussian 2 point formula , n=2,

$$I_g = \frac{b-a}{2}[w_1 g(z_1) + w_2 g(z_2)]$$

$$x = \frac{b-a}{2}z + \frac{b+a}{2} = 2z \qquad \therefore g(z)=e^{-2z/2}e^{-z}$$

$$For\ a\ 2\ pt\ formula\ w_1 = w_2 = 1 \Rightarrow z_1 = \frac{-1}{\sqrt{3}};\ z_2 = \frac{1}{z}$$

$$\therefore I_g = 2\left[\exp\left(\frac{1}{\sqrt{3}}\right) + \exp\left(\frac{-1}{\sqrt{3}}\right)\right] = 4.6853922$$

47. State 2 point Gaussian quadrature formula to evaluate $\int_{-1}^{1} f(x)dx$

Solution: 2 pointGaussian quadrature formula is $\int_{-1}^{1} f(x)\,dx = f\left[\frac{-1}{\sqrt{3}}\right] + f\left[\frac{1}{\sqrt{3}}\right]$

This formula is exact for polynomials upto degree 3

Write down the Simpon's 3/8 rule of integration given (n+1) data

Solution: $\int_{x_0}^{x} f(x)dx = \frac{h}{2}\left[(y_0 + y_n) + 2(y_1 + y_2 + \ldots y_{n-1})\right]$

49. Using 2 point Gaussian quadrature formula evaluate $\int_{-1}^{1} \frac{1}{1+x^2} dx$

Solution: $\int_{-1}^{1} f(x)dx = \frac{5}{9}\left[f\left(-\sqrt{\frac{3}{5}}\right) + f\left(\sqrt{\frac{3}{5}}\right)\right] + \frac{8}{9} f(0) = 1.5833$

In numerical integration what should be the number of intervals to apply Simpson's one-third and Simpson's three-eight rule

Solution: For Simpson's one-third rule , the number of intervals must be even. For Simpson's three-eight rule the number of intervals is a multiple of 3

UNIT- IV
INITIAL VALUE PROBLEMS FOR ORDINARY DIFFERENTIAL EQUATIONS

State the disadvantage or demerits of Taylor's series method.

Solution: In the differential equation $\frac{dy}{dx} = f(x, y)$, the function $f(x, y)$ may have a complicated algebraical structure. Then the evaluation of higher order derivatives may become tedious. This is the demerit of this method.

2. Write the fourth order Taylor's algorithm

Solution: $y_{n+1} = y_n + hy'_n + \frac{h^2}{2!} y''_n + \frac{h^3}{3!} y'''_n + \frac{h^4}{4!} y^{iv}_n$

Write the merits and demerits of Taylor's method of Solution.

Solution: The method gives a straight forward adaptation of classic calculus to develop the Solutionas infinite series. It is a powerful single step method if we are able to find the successive derivatives easily. If (x, y) involves some complicated algebraical structure then the evaluation of higher orderderivatives may become tedious and the method fails. This is the major drawback of this method. However the method will be very useful for finding the starting values for powerful methods like Runge-kutta method, Milne's method etc.

Which is better Taylor's method or R.K method?

Solution: R.K method do not require prior calculation of higher derivatives of $y(x)$, as the Taylor'smethod does.

Since the differential equations using the application are often complicated, the calculation of derivatives may be

difficult. Also the R.K formula involve the computation of (x, y) at various positions, instead of derivatives
and this function occurs in the given equation.

Taylor series method will be very useful to give some ……… for powerful numerical methods such as Runge-kutta method, Milne's method etc.
Solution: Initial starting values.

Name the method which is Taylor's series method of first order.
Solution: Runge-kutta method, Milne's method etc.

State Taylor's series algorithm for the first order differential equation.

Solution: To find the numerical **Solution** of $\frac{dy}{dx} = f(x, y)$ with the condition $y(x_0) = y_0$, we expand
$y(x)$ at a general point x_n in a Taylor's series, getting

$$y_{n+1} = y_n + hy_n' + \frac{h^2}{2!}y_n'' + \frac{h^3}{3!}y_n''' + \frac{h^4}{4!}y_n^{iv} + \ldots .$$

Here y_n^r denotes the r^{th} derivatives w.r.to x at the point (x_n, y_n).

8. Solve the differential equation $\frac{dy}{dx} = x + y + xy$, $y(0) = 1$ by Taylor's series method to get the value of y at $x = h$.

Solution: Given $y' = x + y + xy$, $x_0 = 0, y_0 = 1$.

We know that the Taylor's series formula for y_1 is $y_1 = y_0 + hy_0' + \frac{h^2}{2!}y_0'' + \frac{h^3}{3!}y_0''' + \ldots$

$y' = x + y + xy$ $\quad y_0' = x_0 + y_0 + x_0 y_0 = 0+1+0=1$

$y'' = 1 + y' + xy' + y$ $\quad y_0'' = 1 + y_0' + x_0 y_0' + y_0 1 + 1 + 0 + 1 = 3$

$y''' = y'' + xy'' + y' + y'$ $\quad y_0''' = y_0'' + x_0 y_0'' + 2y_0' = 1 + 0 + 2 = 3$

$\quad = y'' + xy'' + 2y'$

$$y_1 = 1 + \frac{h}{1!}(1) + \frac{h^2}{2!}(3) + \frac{h^3}{3!}(3) + \ldots$$

9. What is the Truncation error in Taylor's series method.

Solution: Error $= \frac{(x - x_0)^n}{n!} y^n(\xi) = 0(h^n)$

10. What is meant by initial value problem and give an example for it.
Solution: Problems in which all the initial conditions are specified at the initial point only are called
initial value problems.
Example : $y' = f(x, y)$ with $y(x_0) = y_0$.

11. Write the Euler algorithm to the differential equation $\frac{dy}{dx} = f(x, y)$.

Solution: $y_{n+1} = y_n + hf(x, y)$ when n = 0,1,2,…

This is the Euler algorithm and it can also be written as $y(x + h) = y(x) + hf(x, y)$

12. State True or False:
In Euler's method, if h is small, the method is too slow and if it is large, it gives inaccurate values.
Solution: The statement is True.

13. Using Euler's method find $y(0.2)$ **from** $\frac{dy}{dx} = x + y,\ y(0) = 1$ with $h = 0.2$

Solution: By Euler's algorithm, $y_1 = y_0 + hf(x_0, y_0) = 1 + (0.2)(x_0 + y_0) = 1.2$

14. State the modified Euler's algorithm to solve $y' = f(x, y)$ with $y(x_0) = y_0$ at $x = x_0 + h$.

Solution: $y_{n+1} = y_n + hf\left(x_n + \frac{h}{2}, y_n + \frac{h}{2} f(x_n, y_n)\right)$, $y_1 = y_0 + hf\left(x_0 + \frac{h}{2}, y_0 + \frac{h}{2} f(x_0, y_0)\right)$

Using Modified Euler's method, find $y(0.1)$ if $\frac{dy}{dx} = x^2 + y^2,\ y(0) = 1$.

Solution: Given $f(x, y) = x^2 + y^2$, $x_0 = 0, y_0 = 1, h = 0.1, x_1 = 0.1$

By Modified Euler method $y_{n+1} = y_n + hf\left(x_n + \frac{h}{2}, y_n + \frac{h}{2} f(x_n, y_n)\right)$

$y_1 = y_0 + hf\left(x_0 + \frac{h}{2}, y_0 + \frac{h}{2} f(x_0, y_0)\right)$

$f(x_0, y_0) = x_0^2 + y_0^2 = 0 + 1 = 1$

$y_1 = 1 + (0.1) f\left[0 + \frac{0.1}{2}, 1 + \frac{0.1}{2}(1)\right] = 1 + (0.1) f(0.05, 1.05) = 1 + (0.1)\left[(0.05)^2 + (1.05)^2\right] = 1.1105$

State True or False: The modified Euler method is based on the average of points.
Solution: The statement is True.

What is the error of Euler's method.

Solution: Error at $(x = x_1) = \frac{h^2}{2!} y''(x_1, y_1)$ Error $= O(h^2) =$ Error is of order h^2.

What are the limitations of Euler's method?
Solution: 1. The attainable accuracy is limited by length of step h.
The method is slow and has limited accuracy.

What is the Error in Modified Euler's method.

Solution: Error $= -\frac{h^3}{12}$ X constant Error $= o(h^3)$, order of h^3

20. Write the Rungekutta algorithm of second order for solving $y' = f(x, y), y(x_0) = y_0$.
Solution: Let h denote the interval between equidistant values of x. If the initial values are (x_0, y_0), the first increment
in y is computed from the formulas. $k_1 = hf(x_0, y_0)$, $k_2 = hf\left(x_0 + \frac{h}{2}, y_0 + \frac{k_1}{2}\right)$ and $\Delta y = k_2$.

Then $x_1 = x_0 + h, y_1 = y_0 + \Delta y$. The increment of y in the second interval is computed in a similar manner using the
same three formulas, using the values x_1, y_1 in the place of x_0, y_0 respectively.

State the third order R.K method algorithm to find the numerical Solution of the first order differential equation.

Solution: To solve the differential equation $y' = f(x, y)$ by the third order RK method, we use the following algorithm

$k_1 = hf(x, y)$, $k_2 = hf\left(x + \frac{h}{2}, y + \frac{k_1}{2}\right)$

$k_3 = hf(x + h, y + 2k_2 + k_1)$ and $\Delta y = \frac{1}{6}(k_1 + 4k_2 + k_3)$

22. Write the RK formula of fourth order to solve $\frac{dy}{dx} = f(x, y)$ with $y(x_0) = y_0$

Solution:Let h denote the interval between equidistant values of x. If the initial values are (x_0, y_0), the first increment in y is computed from the formulas

$k_1 = hf(x_0, y_0)$, $k_2 = hf\left(x_0 + \frac{h}{2}, y_0 + \frac{k_1}{2}\right)$

$k_3 = hf\left(x_0 + \frac{h}{2}, y_0 + \frac{k_2}{2}\right)$, $k_4 = hf(x_0 + h, y_0 + k_3)$ and $\Delta y = \frac{1}{6}(k_1 + 2k_2 + 2k_3 + k_4)$

Then $x_1 = x_0 + h$, $y_1 = y_0 + \Delta y$ The increment of y in the second interval is computed in a similar manner using the same three formulas, using the values x_1, y_1 in the place of x_0, y_0 respectively.

23. State the special advantage of Rungekutta method over Taylor series method.

Solution: Rungekutta method do not require prior calculation of higher derivatives of $y(x)$, as the Taylor method does.

Since the differential equations using in applications are often complicated, the calculation of derivatives may be difficult.Also, the Rungekutta formulas involve the computation of $f(x, y)$ at various positions, instead of derivatives and this function occurs in the given equation.

State True or False:Modified Euler's method is the Rungekutta method of second order. Solution:True.

Is Euler's formula, a particular case of second order Rungekutta method.

Solution: Yes, Euler's formula is a particular case of second order Rungekutta method

The fourth order Rungekutta methods are used widely in ……… to differential equations.

Solution: Getting numerical **Solution**s

27. In the deviation of fourth order Rungekutta formula, why it is called fourth order.

Solution: It is called fourth order formula since the parameters are determined such that y_{n+1} obtained by RK method agrees upto h^4 term in Taylor's method.

Write the formula to solve second order differential equation using Rungekutta method of fourth order.

Solution: The Solution of $y'' = f(x, y, y')$ given $y(x_0) = y_0$, $y'(x_0) = y_0'$

Now set $y' = z$ and $y'' = z'$. Hence the equation reduces to $f(x, y, y') = f(x, y, z)$.

$k_1 = hf(x_0, y_0, z_0)$ $k_2 = hf\left(x_0 + \frac{h}{2}, y_0 + \frac{k_1}{2}, z_0 + \frac{l_1}{2}\right)$ $k_3 = hf\left(x_0 + \frac{h}{2}, y_0 + \frac{k_2}{2}, z_0 + \frac{l_2}{2}\right)$ $k_4 = hf(x_0 + h, y_0 + k_3, z_0 + l_3)$	$l_1 = hf(x_0, y_0, z_0)$ $l_2 = hf\left(x_0 + \frac{h}{2}, y_0 + \frac{k_1}{2}, z_0 + \frac{l_1}{2}\right)$ $l_3 = hf\left(x_0 + \frac{h}{2}, y_0 + \frac{k_2}{2}, z_0 + \frac{l_2}{2}\right)$ $l_4 = hf(x_0 + h, y_0 + k_3, z_0 + l_3)$
$\Delta y = \frac{1}{6}(k_1 + 2k_2 + 2k_3 + k_4)$	$\Delta z = \frac{1}{6}[l_1 + 2l_2 + 2l_3 + l_4]$
$y_1 = y_0 + \Delta y$	$z_1 = z_0 + \Delta z$

29. What are the values of k_1 and l_1 to solve $y'' + xy' + y' = 0, y(0) = 1, y'(0) = 0$ by Rungekutta method of fourth order.

Solution: Given $y'' + xy' + y = 0, y(0) = 1, y'(0) = 0$. The equation becomes $y'' = z' = -xz - y$, setting $y' = z$, i.e., $x_0 = 0, y_0 = 1, z_0 = y'_0 = 0$

$\frac{dy}{dx} = z = f_1(x, y, z)$ and $\frac{dz}{dx} = f_2(x, y, z)$

By algorithm

$k_1 = hf_1(x_0, y_0, z_0) = hf_1(0, 1, 0) = h(0) = 0$

$l_1 = hf_2(x_0, y_0, z_0) = hf_2(0, 1, 0) = h(-0, -1) = -h$

30. Error table:

R.K method	Error
First order	$O(h^2)$
Second order	$O(h^3)$
Fourth order	$O(h^5)$

R.K method of first order is nothing but ……….Solution: Euler's method

R.K method of Second order is nothing but ……….Solution: Modified Euler's method

How many prior values are required to predict the next value in Milne's method Solution: Four prior values.

Write Milne's predictor corrector formula

Solution: Milne's predictor formula is $y_{n+1} = y_{n-3} + \frac{4h}{3}(2y'_{n-2} - y'_{n-1} + 2y'_n) + \frac{14h^5}{45} y^5(\xi_1)$ Where ξ_1 lies between x_{n-3} and x_{n+1}. Milne's corrector formula is $y_{n+1} = y_{n-1} + \frac{h}{3}(y'_{n-1} + 4y'_n + y'_{n+1}) - \frac{h^5}{90} y^5(\xi_2)$ Where ξ_2 lies between x_{n-1} and x_{n+1}.

Say True or False:Milne's method is a self starting method.Solution: The statement is false.

Say True or False: Predictor corrector methods are single step methods. Solution: The statement is false

Pick out the correct answer:The error term in Milne's predictor formula is

a) $\frac{14h}{45}\Delta^4 y_0'$ b) $\frac{14h}{45}\Delta^4 y_0$ c) $-\frac{19h}{45}\Delta^4 y_0'$ d) $\frac{h}{90}\Delta^4 y_0'$ **Solution**: The error term is (a) $\frac{14h}{45}\Delta^4 y_0'$.

How many prior values are required to predict the next value in Adam's method.Solution: Four prior values.

What is the error term in Milne's corrector formula.

Solution: The error term is $-\frac{h}{90}\Delta^4 y_0'$.

40. Predictor corrector methods are ………. Starting methods.Solution: not self

41. Error table:

Milne's method	Error
Predictor	$\underline{h^5} y_n(\xi) = o(h^5)$
Corrector	$\frac{-h}{90} y_n(\xi) = o(h^5)$

Write Adam's Bashforth formula.

Solution: Adam's predictor corrector formula $y_{k+1,p} = y_k + \frac{h}{24}(55y_k' - 59y_{k-1}' + 37y_{k-2}' - 9y_{k-2}')$

and $y_{k+1,c} = y_k + \frac{h}{24}(9y_{k+1}' + 19y_k' - 5y_{k-1}' + y_{k-2}')$

Say True or False:Adam's Bashforth method is a self starting method.Solution: The statement is false.

What is a predictor corrector method of solving a differential equation?

Solution: Predictor corrector methods are methods which require the values of y at $x_n, x_{n-1}, x_{n-2,\ldots}$ for computing the value of y at x_{n+1} . We first use a formula to find the value of y at x_{n+1} and this is known as a predictor formula. The value of y so got is improved or corrected by another formula known as corrector formula.

What is the condition to apply Adam's Bashforth method.

Solution: To use Adam's Bashforth method atleast four values of y, prior to the desired value are required.

46. Error table:

Adam's method	Error
Predictor	$\frac{251}{720}h^5 f^{(iv)}(\xi)$
Corrector	$\frac{-19}{720}h^5 f^{(iv)}(\xi)$

47. What do you mean by saying that a method is self starting/ Not self starting?

Self starting method	Not self starting method

1. To find (x_{n+1}, y_{n+1}) we use only the information at $x_n, y_n)$	We need past values, based on that only we get the next value.
2. Example Taylor’s series method	Milne’s method, Adam’s method

48. Explain the terms initial and boundary value problems.
Solution: Initial value problem : In solving a differential equation analytically, we usually find a general Solution
containing arbitrary constants and then evaluate the arbitrary constants so that the expression agrees with the initial
conditions, the problem is called initial value problem.Boundary value problem: When the differential equation is to be
solved satisfying the conditions specified at the end points of an interval, the problem is called boundary value problem

compare the Milne’s predictor corrector and Adam’s Bashforth predictor corrector methods
for solving ordinary differential equations.
Solution:
i)To apply Milne’s and Adam’s Bashforth methods, we require four starting values of y which are calculated by means
of Taylor’s series method or Euler’s method or RK method.
The Adam’s method is a method that does not have the same instability problem as the Milne’smethod, but is efficient.

to $_5 ho$ iii)Milne’smethodissimpleandhasagoodlocalerror). However, it is subject
instability in certain cases the
errors do not tend to zero as h is made smaller. Because of this instability, another method, a modification of Adams
method, is more widely used than Milne’s.

Mention the multi step methods available for solving ordinary differential equation.
Solution: Milne’s predictor-corrector, Adams Bashforth predictor-corrector formulae are multi step methods

51. Find y(1.1) by Taylor’s series given $\frac{dy}{dx} = x + y, \, y(1) = 0$

Solution: $\frac{dy}{dx} = x + y, \, y(1) = 0$ here h=0.1 , x_0=1 , y_0=0

$$y(1.1) = y_1 = y_0 + \frac{h}{1!} y_0' + \frac{h^2}{2!} y_0'' + \ldots \Rightarrow y_0' = x_0 + y_0 = 1; \, y_0'' = 1 + y_0' = 1 + 1 = 2; \, y(1.1) = 11$$

52.Give a comparison of Adam’s Bashforth method with Runge-Kutta methods
Solution:
AB’s method requires 4 starting values of y.Hence it is referred as a multi step method, while R.K methods are single step method
R.K method is self starting as it does not require more than the initial value of y, where as the AB method are not self startings as values of y at prior pts are used.
R.K method permits on easy change in the step size while it is not so in AB method

53.State Adam’s predictor and corrector formula for solving initial value problem

Solution: $y_{n+1} = y_{n-3} + \frac{4h}{3}(2y'_{n-2} - y'_{n-1} + 2y'_{n})$ and $y_{n+1} = y_{n-1} + \frac{h}{3}(y'_{n-1} + 4y'_{n} + y'_{n+1})$

54. Explain one step method and multi step methods

Solution: **One step method:** In each step , we use the data of just one proceeding step.

Multi step method: In each step , we use the data from more than one of the proceeding step.

UNIT- V
BOUNDARY VALUE PROBLEMS IN ORDINARY AND PARTIAL DIFFERENTIAL EQUATIONS

1. State the condition for the equation $Au_{xx} + Bu_{xy} + Cu_{yy} + Du_x + Eu_y + Fu = G$ where A,B,C,D,E,F,G are function of x and y to be (i) elliptic (ii) parabolic (iii) hyperbolic.

Solution: The given equation is said to be i)elliptic at a point (x, y) in the plane $B^2 - 4AC < 0$
ii)parabolic if $B^2 - 4AC = 0$ iii)hyperbolic if $B^2 - 4AC > 0$

2. State the condition for the equation $Au_{xx} + Bu_{xy} + Cu_{yy} = f(u_x, u_y, x, y)$ to be (i) elliptic (ii) parabolic (iii) hyperbolic when A, B, C are the functions of x and y.

Solution: The equation is elliptic if $(2B^2) - 4AC < 0$, Parabolic if $B^2 - 4AC = 0$ and hyperbolic if $B^2 - 4AC > 0$

3. Fill up the blanks.The equation $y u_{xx} + u_{yy} = 0$ is hyperbolic in the region …….

Solution:

Here $A = y$, $B = 0$, $C = 1$

$B^2 - 4AC = 0 - 4y = -4y$ $B^2 - 4AC > 0$ i.e., -4y>0 or y<0It is

The equation is hyperbolic in the region (x,y)
where hyperbolic in the region y<0.

4. What is the classification of $f_x - f_{yy} = 0$?

Solution: Here A=0, B=0, C=-1.
$B^2 - 4AC = 0 - 4 \times 0 \times -1 = 0$ So the equation is parabolic.

5. Give an example of parabolic equation.

Solution: The one dimensional heat equation $\frac{\partial u}{\partial t} = \alpha^2 \frac{\partial^2 u}{\partial x^2}$ is parabolic.

6. Write the general and simplest forms of the difference equation corresponding to hyperbolic equation $u_{tt} = a^2 u_{xx}$.

Solution: The general form of the difference equation to solve the equation $u_{tt} = a^2 u_{xx}$ is

$u_{i,j+1} = 2(1 - \lambda^2 a^2)u_{i,j} + \lambda^2 a^2 (u_{i+1,j} + u_{i-1,j}) - u_{i,j-1}$ ……….(1)

If $\lambda^2 a^2 = 1$, the coefficient of $u_{i,j}$ in (1) is = 0.The recurrence equation (1) takes the simplest form

$u_{i,j+1} = u_{i+1,j} + u_{i-1,j} - u_{i,j-1}$

State Schmidt's explicit formula for solving heat flow equation

Solution: $u_{i,j+1} = \lambda u_{i+1,j} + (1-2\lambda)\, u_{i,j} \ \lambda u_{i-1,j}$ If $\lambda = \frac{1}{2}, u_{i,j+1} = \frac{1}{2}\left[u_{i+1,j} + u_{i-1,j}\right]$

Write an explicit formula to solve numerically the heat equation (parabolic equation) $u_{xx} - au_t = 0$

Solution: $u_{i,j+1} = \lambda u_{i+1,j} + (1-2\lambda)\, u_{i,j} + \lambda u_{i-1,j}$ where $= k\frac{h^2}{a}$ (h is the space for the variable x and k is the space in the time direction)

Bender Schmidt recurrence scheme is useful to solve …… equation

Solution: One dimensional heat

10. What is the value of k to solve $\frac{\partial u}{\partial t} = \frac{1}{2} u_{xx}$ by Bender Schmidt method with h=1 if h and k are the increment of x and t respectively?

Solution: Given $u_{xx} = 2\frac{\partial u}{\partial t}$. Here $\alpha^2 = 2, h = 1$

$\lambda = \frac{k\alpha^2}{h^2} = \frac{k(2)}{1} = 2k \Rightarrow \lambda = 2k = \frac{1}{2} \Rightarrow k = \frac{1}{4}$

11. What is the classification of one dimensional heat flow equation.

Solution: One dimensional heat flow equation is $\frac{\partial^2 u}{\partial x^2} = \frac{1}{\alpha^2}\frac{\partial u}{\partial t}$

Here A=1, B=0, C=0 $B^2 - 4AC = 0$ Hence the one dimensional heat flow equation is parabolic.

12. Write the Crank Nicholson formula to solve $u_t = u_{xx}$

Solution: $\frac{1}{2}\lambda u_{i+1,j+1} + \frac{1}{2}\lambda u_{i-1,j+1} - (\lambda+1)u_{i,j+1} = -\frac{1}{2}\lambda u_{i+1,j} - \frac{1}{2}\lambda u_{i-1,j} + (\lambda-1)u_{i,j}$

Or $\lambda\left(u_{i+1,j+1} + \lambda u_{i-1,j+1}\right) - 2(\lambda+1)u_{i,j+1} = 2(\lambda-1)u_{i,j} - \lambda\left(u_{i+1,j} + u_{i-1,j}\right)$

13. Write down the implicit formula to solve one dimensional heat flow equation $u_{xx} = \frac{1}{c^2}u$

Solution:

$\lambda\left(u_{i+1,j+1} + \lambda u_{i-1,j+1}\right) - 2(\lambda+1)u_{i,j+1} = 2(\lambda-1)u_{i,j} - \lambda\left(u_{i+1,j} + u_{i-1,j}\right)$

14. Fill up the blanks:

In the parabolic equation $u_t = \alpha^2 u_{xx}$ if $\lambda = \frac{k\alpha^2}{h^2}$ where $k = \Delta t$ and $h = \Delta x$, then

(i) **explicit method is stable only if $\lambda = ...$**

(ii) **implicit method is convergent when $\lambda = ...$**

Solution :

(i) explicit method is stable only if $\lambda < \frac{1}{2}$

(ii) implicit method is convergent when $\lambda = \frac{1}{2}$

15. Why Crank Nicholson scheme called an implicit scheme?
Solution: The schematic representation of Crank Nicholson method is shown below.

The solution value at any point $(i, j+1)$ on the $(j+1)^{th}$ level is dependent on the solution values at the neighboring points on the same level and on three values on the j^{th} level. Hence implicit method. it is an

What type of equations can be solved by using Crank Nicholson's difference formula?

Solution : Crank Nicholson's formula is used to solve parabolic equations of the form $u_{xx} = au_t$

17. Write the Crank Nicholson difference scheme to solve $u_{xx} = au_t$ with $u(0,t) = T_0, u(l,t) = T_1$ and the initial condition as $u(x,0) = f(x)$

Solution : The scheme is $\frac{1}{2}\lambda u_{i+1,j+1} + \frac{1}{2}\lambda u_{i-1,j+1} - (\lambda+1)u_{i,j+1} = -\frac{1}{2}\lambda u_{i+1,j} - \frac{1}{2}\lambda u_{i-1,j} + (\lambda-1)u_{i,j}$

For what purpose Bender Schmidt recurrence relation is used?

Solution : To solve one dimensional heat equation.

Write a note on the stability and convergence of the solution of the difference equation corresponding to the hyperbolic equation $u_{tt} = a^2 u_{xx}$

Solution : For $= \frac{1}{a}$, the solution of the difference equation is stable and coincides with the solution of the difference equation.

For $\lambda > \frac{1}{a}$, the solution is unstable. For $\lambda < \frac{1}{a}$, the solution is stable but not convergent.

20. State the explicit scheme formula for the solution of the wave equation.

Solution : The formula to solve numerically the wave equation $a^2 u_{xx} - u_{tt} = 0$ is

$$u_{i,j+1} = 2\left(1-\lambda^2 a^2\right)u_{i,j} + \lambda^2 a^2\left(u_{i+1,j} + u_{i-1,j}\right) - u_{i,j-1}$$

The schematic representation is shown below:

The solution value at any point $(i, j+1)$ on the $(j+1)^{th}$ level is expressed in terms of solution values on the previous j and (j-1) levels. Hence this is an explicit difference formula.

21. For what value of λ the explicit method of solving the hyperbolic equation $\frac{\partial^2 u}{\partial x^2} = \frac{1}{c^2}\frac{\partial^2 u}{\partial t^2}$ is stable, where $\lambda = \frac{C\Delta t}{\Delta x}$.

Solution: The explicit method is stable only of $= \frac{1}{c}$

22. If u satisfies Laplace equation and u=100 on the boundary of a square what will be the value of u at an interior grid point.

Solution:

Since u satisfies Laplace equation and u=100 on the boundary of a square

$$u_{i,j} = \frac{1}{4}(100 + 100 + 100 + 100) = 100$$

The number of conditions required for solving the Laplace equation is …..

Solution: Four

Wrtie the diagonal five point formula to solve the Laplace equation $u_{xx} + u_{yy} = 0$

Solution:

$$u_{i,j} = \frac{1}{4}\left[u_{i-1,j-1} + u_{i-1,j-1} + u_{i+1,j-1} + u_{i+1,j+1}\right]$$

25. Write down the standard five point formula to solve Laplace equation $\frac{\partial^2 u}{\partial x^2} + \frac{\partial^2 u}{\partial y^2} = 0$

Solution: The standard five point formula is $u_{i,j} = \frac{1}{4}\left[u_{i-1,j} + u_{i+1,j} + u_{i,j-1} + u_{i,j+1}\right]$

26. Write the difference scheme for solving the Laplace equation.
Solution: The five point difference formula for $\nabla^2\phi = 0$ is $u_{i,j} = \frac{1}{4}\left[u_{i-1,j} + u_{i+1,j} + u_{i,j-1} + u_{i,j+1}\right]$

What is the purpose of Liebmann's process?
Solution:
The purpose of Liebmann's process is to find the solution of the Laplace equation $u_{xx} + u_{yy} = 0$

28. For the following mesh in solving $\nabla^2 u = 0$ find the one set of rough values of u at the interior mesh points.

Solution: By symmetry $u_2 = u_3$. Assume $u_2 = 3$ (u_2 is at $\frac{1}{3}$ distance from u=2)

Therefore the rough values are

$u_1 = \frac{1}{4}[1 + 1 + 2u_2] = 2,\ u_2 = 3,\ u_3 = \frac{1}{4}[5 + 5\ 2u_2] = 4,\ u_4 = \frac{1}{4}[u_1 + u_4 + 2 + 4] = 3$

29. Write the Laplace equation $u_{xx} + u_{yy} = 0$ in difference quotients.
Solution: $\frac{u_{i-1,j} - 2u_{i,j} + u_{i+1,j}}{h^2} + \frac{u_{i,j-1} - 2u_{i,j} + u_{i,j+1}}{k^2} = 0$

30. Define a difference quotient.
Solution: A difference quotient is the quotient obtained by dividing the difference between two
values of a function by the difference between two corresponding values of the independent variable.

31. State Liebmann's iteration process formulae.
Solution: $u_{i.j}^{n+1} = \frac{1}{4}\left[u_{i-1,j}^{(n+1)} + u_{i+1,j}^{(n)} + u_{i,j-1}^{(n)} + u_{i,j+1}^{(n+1)}\right]$

32. Write the finite difference form of the equation $\nabla^2 u = f(x, y)$

Solution: $u_{i-1,j} + u_{i+1,j} + u_{i,j-1} + u_{i,j+1} - 4u_{i,j} = h^2 f(ih, jh)$

33. State the five point formula to solve the poisson equation $u_{xx} + u_{yy} = 100$.

Solution: $u_{i-1,j} + u_{i+1,j} + u_{i,j+1} + u_{i,j-1} - 4u_{i,j} = h^2 f(ih, jh) = 100$

34. Write the difference scheme for $\nabla^2 u = f(x, y)$
Solution: Consider a square mesh with the interval of differencing as h. Taking $x = ih,\ y = jh$ the
difference equation reduces to

$$\frac{u_{i-1,j} - 2u_{i,j} + u_{i+1,j}}{h^2} + \frac{u_{i,j-1} - 2u_{i,j} + u_{i,j+1}}{h^2} = f(ih, jh)$$

$i.e., u_{i-1,j} + u_{i+1,j} + u_{i,j+1} - 4u_{i,j} = h^2 f(ih, jh)$

State the general form of Poisson's equation in partial derivatives.
Solution:

$$\frac{\partial^2 u}{\partial x^2}+\frac{\partial^2 u}{\partial y^2}=f(x,y)$$

36. Obtain the finite difference scheme for differential equation $2\frac{d^2y}{dx^2}+y=5$

Solution: Given $2\frac{d^2y}{dx^2}+y=5 \Rightarrow 2\frac{y_{i+1}-2y_i+y_{i-1}}{h^2}+y_i=5 \Rightarrow 2y_{i+1}-(4-h^2)y_i-2y_{i-1}=5h^2$

Discuss diagonal five point formula and standard five point formula.

Solution: These two formulas are used to solve elliptic equations

SFPF is $u_{i,j}=\frac{1}{4}\left[u_{i-1,j}+u_{i+1,j}+u_{i,j-1}+u_{i,j+1}\right]$

DFPF is $u_{i,j}=\frac{1}{4}\left[u_{i-1,j-1}+u_{i-1,j-1}+u_{i+1,j-1}+u_{i+1,j+1}\right]$

Mention any two single step methods for solving an ordinary differential equation subject to initial condition.

Solution: 1. Bender Schmidt 2. Crank Nicholson

39. What is the condition of stability for the Schmidt method?

Solution: λ satisfies the condition $\lambda \le \frac{1}{2}$

40. What is the order of the Crank Nicholson method for solving the heat conduction equation?

Solution: $o(k^2+h^2)$

41. What is the condition for stability for the Crank Nicholson method?

Solution: The Crank Nicholson method is stable for the values of the mesh ratio parameter λ. This
method is also called an unconditionally stable method.

42. What type of system of equations do we get when we apply the Crank Nicholson method to solve the one dimensional heat conduction equation.

Solution: We obtain a linear tridiagonal system of algebraic equations.

43. Solve $y''-xy=0$ **given** $y(0)=-1, y(1)=2$ **by finite difference method taking n=2.**

Solution: If n=2, then h=1/2 since the range is (0,1)

The nodal points are $x_0=0, x_1=0.5, x_2=1$

The differential equation reduces to $\frac{y_{i+1}+y_{i-1}-2y_i}{h^2}-x_iy_i=0$ i.e., $y_{i+1}-(2+h^2x_i)y_i+y_{i-1}=0$

Where i=1, h=1/2, x_i 0.5, y_0 $-1, y_2=2 \therefore y_2-\left(2+\frac{1}{8}\right)y_1+y_0=0 \Rightarrow y_1=\frac{8}{17}=0.4706$

What is the error for solving Laplace and Poisson's equation by finite difference method.

Solution: The error in replacing $\frac{\partial^2 u}{\partial x^2}$ by the difference expression is of order $o(h^2)$. Since h=k,
the error in replacing $\frac{\partial^2 u}{\partial y^2}$ by the difference expression is of order $o(h^2)$.

45. If $b^2-4ac<0$ and $a>0$**, then the PDE** $af_{xx}+bf_{xy}+cf_{yy}+\phi(x,y,f_x,f_y)=0$ **is**

a) Elliptic equation **b) Hyperbolic equation**

c) Parabolic equation d) None
Solution : (a) Elliptic equation

46. To solve $\frac{\partial^2 u}{\partial x^2}+\frac{\partial^2 u}{\partial y^2}=0$ **, we use**

a) Crank Nicholson method b) Bender Schmidt scheme c) Liebmann's method
Solution: c) Liebmann's method

47. Obtain the finite difference scheme for the difference equation $2\frac{\partial^2 y}{\partial x^2}+y=5$

Solution: The given differential equation can be written as $2y''(x)+y(x)=5$ …..(1)
Using the central difference approximation, we have $2y''=\frac{y_{k-1}-2y_k+y_{k+1}}{h^2}$ …..(2)

Substituting (2) in (1), we get $\frac{y_{k-1}-2y_k+y_{k+1}}{h^2}+y_k=5$ i.e., $y_{k-1}-2y_k+y_{k+1}+h^2y_k=5h^2$

48. Classify the equation $x^2f_{xx}+(1-y^2)f_{yy}=0, -\infty<x<\infty, -1<y<1$

Solution: From the equation $A=x^2, B=0, C=1-y^2$
Therefore $B^2-4AC=-4x^2(1-y^2)$

Case (i): If x=0 and y takes any value, then $B^2-4AC=0$ Therefore the equation is Parabolic.
Case (ii): If $-\infty<x<\infty, y=-1$ or $y=1$, then $B^2-4AC=0$ Therefore the equation is Parabolic

Case (iii): If $-\infty<x<\infty, -1<y<1$, then $B^2-4AC=-ve$ Therefore the equation is Elliptic

49. What is the purpose of Liebmann's iterative formula?
Solution: To correct (improved) values of the interior mesh points.

50. The PDE $\frac{\partial^2 u}{\partial x^2}+\frac{\partial^2 u}{\partial y^2}=f(x,y)$ **is called**

Heat equation b) Wave equation c) Laplace equation d) Poisson's equation
Solution : d) Poisson's equation

Derive the forward finite difference formula for u_x
Solution: Keeping y fixed, using Taylor's series

$u(x,y_0)=u(x_0,y_0)+(x-x_0)u(x_0,y_0)+\frac{(x-x_0)^2}{2!}u_{xx}(x_0,y_0),\quad x_0<\varepsilon<x$

Put $x_0+h=x$, and re-writing

$\frac{u(x_0+h,y_0)-u(x_0,y_0)}{h}=u_x(x_0,y_0)$ with $-\frac{h}{2}u_{xx}(x_0,y_0)$ Truncation error $x_0<E<x_0+h$

52. Give the initial and boundary conditions for the wave equation and an explicit formula for its solution
Solution: Wave equation: $a^2u_{xx}-u_{tt}=0$
Boundary condition : u(0,t) =0; u(l,t)=0
Initial condition: u(x,0)=f(x); $u_t(x,0)=0$
Explicit formula : $U_{i,j+1}=U_{i-1,j}+U_{i+1,j}-U_{i,j-1}$

53.Name atleast 2 numerical methods that are used to solve one dimensional diffusion equation
Solution: (i) Bender –Schmidt method (ii) Crank- Nicholson method

54.State finite difference approximation for $\frac{d^2y}{dx^2}$ **and state the order of truncation error**

Solution: $\frac{d^2y}{dx^2} = \frac{2y_{i+1} - 2y_i + y_{i-1}}{h^2}$, **Error =O(h^2)**

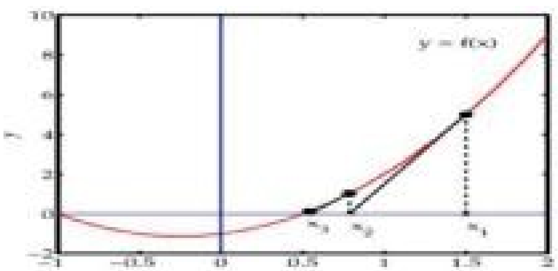
55.For what points of x and y the equation $xf_{xx} + yf_{yy} = 0, x > 0, y > 0$ **is elliptic**

Solution:A=x ,B=0, C=4

B^2-4AC=-4xy (- ve) => It is elliptic

May / June 2015

Interpret Newton Rapshon method geometrically.



Which of the iterative methods for solving linear system of equations converge faster? Why?

The convergence of Gauss seidel method is faster that the convergence of Gauss – Jacobi iterative method.

Because the current value of the unknowns at each stage of iterations are used in proceeding to the next

stage of iteration. The convergence in gauss – Seidel method will be more rapid than in Gauss – Jacobi method.

Given =3,= , = , = , = .**Find** Δ .

y	Δ	Δ	Δ	Δ
3	9	60	-10	-259
12	69	50	-269	
81	119	-219		
200	-100			
100				

Distinguish between Newton divided difference interpolation and Lagrange's interpolation.

Lagrange’s Method	Newton’s method
We can apply both equally and unequally spaced arguments.	The arguments are equally spaced.
Can be used to interpolate any where in the range	Newton’s forward formula is suitable to interpolate near the beginning. Newton’s backward formula is suitable to interpolate near and of the value.

5. **Find ′( ) from the following table.**

x	0	1	2	3	4	5
y	4	8	15	7	6	2

Solution:

Here h = 1

x	y	Δ	Δ	Δ	Δ	Δ
0 1 2 3 4 5	4 8 15 7 6 2	4 7 -8 -1 -4	3 -15 7 -3	-18 22 -10	40 -32	-72

$$'(0) = \frac{1}{h}\{\Delta_0 - \frac{1}{2}\Delta^2_0 + \frac{1}{3}\Delta^3_0 - \frac{1}{4}\Delta^5_0\}$$

$$= \frac{1}{1}(4 - \frac{3}{2} - \frac{18}{3} - \frac{40}{4} - \frac{72}{5})$$

$$= -27.9$$

6. Use two point Gaussian quadrature formula evaluate $I = 4\int_{-1}^{1} \sin(\ + 4)$.

Solution:

The Gaussian two point quadrature formula is

$$\int_{-1}^{1}(\ ) = (\frac{-1}{\sqrt{3}}) + (\frac{1}{\sqrt{3}})$$

Given $(\ ) = \sin(\frac{-1\ +}{4})$

$(\frac{-1}{\sqrt{3}}) = -0.6727$ and $(\frac{1}{\sqrt{3}}) = 0.4054$

$$= \frac{1}{4}\int_{-1}^{1}\sin(\frac{+}{4}) = -0.6727 + 0.4054 = -0.2673$$

Find by Taylor’s series method, the value of at = . from = + , ( ) = . Solution:

Given

x	0	0.1
y	1	?

Let $_0 = 0$ and $_0 = 1$ here h= 0.1

′( ) = ²+	′ = ² + 0 0 0 = 1
″ = 2 ′ + 1	″ = 2 0 0 0 + 1 = 3
‴ = 2(″ + ²)	‴ = 2(″ + ²) = 8

By Taylor's series

( ) = (0.1) = 1 + (0.1)(1) + $\frac{(0.1)^2}{2}$ (3) + $\frac{(0.1)^3}{6}$ (8)

(0.1) = 1.1162.

Distinguish between single step methods and multi step methods.

Solution:

Single step method	Multi step method
A series for y interms of powers of x,from which the value of y can be obtained by direct substitution.	The values of y are computed by short steps for equal intervals h of the independent variable. These values are iterated till we get desired accuracy.
Example: 1. Taylor's series 2. Picards method	Example: 1. Euler method 2. R- K method

9. **Classify the following equation:** + + − + = .

Solution:

Here A = 1 , B = 4 , C= 4

² − 4 = 16 − 4(1)(4) = 0

Given PDE is of parabolic type.

Express = **in terms of difference approximation.** Solution:

Given ² − = 0

² ²(+1, − 2 , + −1,) − , +1 + 2 , − , −1 = 0

M/J 2016

What is the criterion for the convergence of Newton Raphson method?

Soln:

The sequence 1, 2, 3, … … converges to the exact value if | ′( )| < 1

ie., if | ( ) ″( )| < | ′( )|².

2. Find the inverse of A=(1 3 / 2 7) by Gauss Jordan method.

Soln:

[,] = [1 3 | 1 0 / 2 7 | 0 1]

=[1 3 | 1 0 / 0 1 | −2 1] → − 2

$= [^1 \quad 0| \qquad {}^7 \; {}^{-3}] \rightarrow -3$

$^{-1} = [-^7{}_2 \qquad -{}_1{}^3]$

State Newton' forward difference formula for equal intervals.

Soln:

$$P_n(x) = P_n(x_0 + uh) = E^u P_n(x_0) = E^u y_0 = (1+\Delta)^u y_0$$

$$= 1 + \frac{u}{1!}\Delta y_0 + \frac{u(u-1)}{2!}\Delta^2 y_0 + \frac{u(u-1)(u-2)}{3!}\Delta^3 y_0 + \dots \quad \text{where } u = \frac{x - x_0}{h}$$

Find the divided difference of ( ) = ³ − ² + 3 + 8 for the arguments 0,1,4,5. Soln:

x	0	1	4	5
f(x)	8	11	68	123

x	F(x)	Δ( )	Δ²( )	Δ³( )
0	8			
1	11	3		
4	68	19	4	1
5	123	55	9	

5. Write down the expression for — and ² — at = by Newton's backward difference formula.

Soln:

$$\left(\frac{dy}{dx}\right)_{x=x_n} = \frac{1}{h}\left[\nabla y_n + \frac{1}{2}\nabla^2 y_n + \frac{1}{3}\nabla^3 y_n + \dots\right]$$

$$\left(\frac{d^2 y}{dx^2}\right)_{x=x_n} = \frac{1}{h^2}\left[\nabla^2 y_n + \nabla^3 y_n + \frac{11}{12}\nabla^4 y_n + \dots\right]$$

6. Taking h=0.5, evaluate $\int_1^2 \frac{}{1+\;^2}$ using Trapezoidal rule.

Soln:

$( ) = \frac{1}{1+\;^2}, \quad h = 0.5$

x:	1	1.5	2
Y:	0.5	0.3077	0.2

By Trapezoidal Rule,

$$\int_1^2 \frac{1}{1+\;^2} = h/2[(\;_0 + \;_2) + 2(\;_1)]$$

$$= \frac{}{2}[0.5 + 0.2 + 2(0.3077)]$$

$$0.3289$$

Find (0.1) if = 1 + , (0) = 1 using Taylor's series method.

Soln:

Let h=0.1; $_1 = 0.1$

By Taylor's series is,

$( ) = \quad + (- \quad)' + \frac{(- \;_0)^2}{2!} + \dots$

′( ) = 1 +	′ = 1 + = 2
″( ) = ′	″ = ′ = 2
‴( ) = ″	‴ = 2

$$= 1 + (0.1)(2) + \frac{(0.1)^2}{2!}(2) + \cdots$$

$_1 = 1.2103$

8. State the fourth order Runge-Kutta algoritham.

Soln:

Let h denote the interval between equidistant values of x. If the initial values are (x_0, y_0), the first increment in y is computed from the formulas

$k_1 = hf(x_0, y_0)$, $k_2 = hf\left(x_0 + \frac{h}{2}, y_0 + \frac{k_1}{2}\right)$

$k_3 = hf\left(x_0 + \frac{h}{2}, y_0 + \frac{k_2}{2}\right)$, $k_4 = hf(x_0 + h, y_0 + k_3)$ and $\Delta y = \frac{1}{6}(k_1 + 2k_2 + 2k_3 + k_4)$

Then $x_1 = x_0 + h$, $y_1 = y_0 + \Delta y$ The increment of y in the second interval is computed in a similar manner using the same three formulas, using the values x_1, y_1 in the place of x_0, y_0 respectively

Obtain the finite difference scheme for the differential equation 2 ″ + = 5. Soln:

The given differential equation can be written as $2y''(x) + y(x) = 5$ …..(1)

Using the central difference approximation, we have $2y'' = \frac{y_{k-1} - 2y_k + y_{k+1}}{h^2}$ …..(2)

Substituting (2) in (1), we get $\frac{y_{k-1} - 2y_k + y_{k+1}}{h^2} + y_k = 5$ i.e., $y_{k-1} - 2y_k + y_{k+1} + h^2 y_k = 5h^2$

Write Liebmann's iteration process.

Soln:

$$u_{i,j}^{n+1} = \frac{1}{4}\left[u_{i-1,j}^{(n+1)} + u_{i+1,j}^{(n)} + u_{i,j-1}^{(n)} + u_{i,j+1}^{(n+1)}\right]$$

PART – B
UNIT: I
SOLUTION OF EQUATIONS AND EIGEN VALUE PROBLEMS
PART-B
PROBLEMS RELATED TO REGULA FALSI METHOD

1. Find the positive root of $x^3 = 2x + 5$ by the method of false position, correct to 4 decimals.
2. Find an approximate root of $x \log_{10} x = 1.2$ by false position method, correct to four decimal places.
3. Find the root of $xe^x = 3$ by Regula Falsi method.
4. Find the root of $x - \cos x = 0$ by Regula Falsi method.

PROBLEMS RELATED TO NEWTON-RAPHSON METHOD

Apply Newton's method to find the positive root of $x = \cos x$.

2. Using Newton's method, find the real root of $3x - \cos x - 1 = 0$.
3. Using Newton's method, find a positive root of $f(x) = x^3 - 5x + 3 = 0$.
4. Find the root between 1 and 2 of $2x^3 - 3x - 6 = 0$ UsingNewton's method correct to 5 decimal.
5. Find the root between 2 and 3 of $x^3 - 5x - 7 = 0$ UsingNewton's method correct to 3 decimal.
6. Apply Newton's method to find the positive root of $x^4 - x - 10 = 0$.
7. Using Newton's method, find a positive root of $xe^x = 1$, correct to four decimal places.

PROBLEMS RELATED TO FIXED POINT ITERATION METHOD

1. Find a real root of $x^3 + x^2 - 100 = 0$ by using iteration method.

PROBLEMS RELATED TO GAUSS ELIMINATION METHOD

1. Apply Gauss elimination method to solve the equation $3x + y - z = 3$; $2x - 8y + z = -5$; $x - 2y + 9z = 8$
2. Apply Gauss elimination method to solve the equation $-x + y + 10z = 35.61; 10x + y - z = 11.19; x + 10y + z = 20.08$

PROBLEMS RELATED TO GAUSS JORDAN METHOD

1. Solve the system of equations $4x + 2y + z = 14$; $x + 5y + z = 10$; $x + y + 8z = 20$ By Gauss Jordan method.

Solve the system of equations $10x + y + z = 12$; ; $2x + 10y + = 13$; $x + + 5z = 7$ By Gauss Jordan method.

Solve the system of equations $x + y + z = 9$; $2x - 3y + 4z = 13$; $3x + 4y \; 5z = 40$ by Gauss Jordan method.

PROBLEMS RELATED TO GAUSS JACOBI METHOD

1. Solve by Jacobi iteration method correct to two decimal places $10x + y - z = 11.09$; $x + 10y + z = 28.08$; $-x + y + 10z = 35.61$

PROBLEMS RELATED TO GAUSS- SEIDEL METHOD

Solve by Gauss- Seidel method, $8x + y + z = 8$; $4y + z \; 4y$; $x + 3 + 5z = 5$.

Solve by Gauss- Seidel method, $x + y + 54z \; \frac{2x}{110}$; $6 \; -z \; 85 \; 27x +$; $6x + 15y \; 2z = 72$

Correct to three decimal places

3. Solve the system of equations $10x\ 5y - 2z = 3$; $+ 6y + 10z = -3$; $4x - 10y + 3z = -3$

by Gauss- Seidel method

4. Solve by Gauss- Seidel method, $28x + 4y - z = 32$; $x + 3y + 10z = 24$; $2x + 17y + 4z = 35$.

PROBLEMS RELATED TOINVERSE OF THE MATRIX BY GAUSS-JORDAN METHOD

1. Find the inverse of $A=\begin{pmatrix} 1 & -2 & 3 \\ 0 & -1 & 4 \\ -2 & 2 & 0 \end{pmatrix}$ by Gauss-Jordan method.

2. Find the inverse of the matrix using elimination process $\begin{bmatrix} 3 & -1 & 1 \\ -15 & 6 & -5 \\ 5 & -2 & 2 \end{bmatrix}$

3. Using Gauss-Jordan method, find the inverse of the matrix $A=\begin{pmatrix} 1 & -1 & 1 \\ 1 & -2 & 4 \\ 1 & 2 & 2 \end{pmatrix}$.

4. Using Gauss-Jordan method, find the inverse of the matrix $A=\begin{pmatrix} 4 & 1 & 2 \\ 2 & 3 & -1 \\ 1 & -2 & 2 \end{pmatrix}$.

5. Find the Eigen values and Eigen vectors of $A=\begin{pmatrix} 2 & 3 \\ 3 & 2 \end{pmatrix}$ by Jacobi method.

PROBLEMS RELATED TO POWER METHOD

1. Using Power method find the largest Eigen value and its corresponding Eigen vector $\begin{pmatrix} 1 & 3 & -2 \\ 4 & 4 & -1 \\ 6 & 3 & 5 \end{pmatrix}$.

2. Find the largest Eigen value and the corresponding Eigen vector for the matrix $A=\begin{pmatrix} 3 & 2 & 1 \\ 2 & 4 & 2 \\ 1 & 2 & 3 \end{pmatrix}$

By Power method.

3. Using power method, find the largest Eigen value and corresponding Eigen vector of $A=\begin{pmatrix} 25 & 1 & 2 \\ 1 & 3 & 0 \\ 2 & 0 & -4 \end{pmatrix}$

4. Find the numerically largest Eigen value of $A=\begin{pmatrix} 1 & -3 & \\ 4 & 4 & 1 \\ 6 & 3 & \end{pmatrix}$ by power method.

5. Find the numerically largest Eigen value and the corresponding Eigen vector using power method given matrix $A=\begin{pmatrix} 1 & 6 & 1 \\ 1 & 2 & 0 \\ 0 & 0 & 3 \end{pmatrix}$.

6. Find the dominant Eigen value and the corresponding Eigen vector using power method given matrix $A=\begin{pmatrix} 5 & 4 & 3 \\ 10 & 8 & 6 \\ 20 & -4 & 22 \end{pmatrix}$ starting vector is $\begin{pmatrix} 1 \\ 1 \\ 1 \end{pmatrix}$

7. Obtain by power method the numerically largest Eigen value of the matrix $\begin{pmatrix} 15 & -4 & -3 \\ -10 & 12 & -6 \\ -20 & 4 & -2 \end{pmatrix}$

UNIT: II
INTERPOLATION AND APPROXIMATION
PART-B

PROBLEMS BASED ON LAGRANGE'S & INVERSE LAGRANGE'S INTERPOLATION METHOD

Using Lagrangian's interpolation formula finds the values of y at $x = 10$ from the following data:

x	5	6	9	11
y	12	13	14	16

2. Using Lagrangian's interpolation formula, find x corresponding to y=85 given

x	2	5	8	14
y	94.8	87.9	81.3	68.7

3. Find a Lagrangian's interpolating polynomial y= $f(x)$, and find f(5).

x	1	3	4	6
y	-3	0	30	132

4. Fit a polynomial in x for $f(x)$, given the following data:

x	0	2	3	4	7	9
y	4	26	58	112	466	922

5. Using Lagrange's formula fit a polynomial to the data:

x	0	1	3	4
y	-12	0	6	12

PROBLEMS BASED ON DIVIDED DIFFERENCE

1. If $f(x) = \frac{1}{x}$, find the divided differences $f(a, b)$, $f(a,b,c)$ and $f(a,b,c,d)$.

PROBLEMS BASED ON NEWTON'S DIVIDED DIFFERENCE FORMULA

1. Using Newton's divided difference formula, find the value of $f(8)$ from the following data:

x	4	5	7	10	11	13
$f(x)$	48	100	294	900	1210	2028

2. Find $f(1)$, $f(5)$ and $f(9)$ using Newton's divided difference formula from the following data:

x	0	2	3	4	7	8
y	4	26	58	112	466	668

3. Using Newton's divided difference formula, find the polynomial of the given data

x	-1	0	1	3
y	2	1	0	-1

4. Given the values

x	5	7	11	13	17
y	150	392	1452	2366	5202

Evaluate f(9) using Newton' divided difference formula.

If f(1)=0,f(2)=-12 , f(4)=0, f(5)=600, and f(7)=7308, find a polynomial that satisfies this data using Newton's
divided difference interpolation formula. Hence find f (6).

PROBLEMS BASED ON NEWTON'S FORWARD INTERPOLATION FORMULA 1. The following are data from the steam table:

$Temp\ ^\circ C$	140	150	160	170	180
$Pressure\ (kgf\ cm^2)$	3.685	4.854	6.302	8.076	10.225

Using Newton's formula, find the pressure of the steam for a temperature of 142^o .

2. From the following table, find the value of $\tan(0.12)$

x	0.10	0.15	0.20	0.25	0.30
$y = \tan x$	0.1003	0.1511	0.2027	0.2553	0.3093

3. Find $y(2)$ from the following table:

x	1	3	5	7	9
y	2	10	26	50	82

4. Find a polynomial of degree two for the data by Newton's forward difference method:

x	0	1	2	3	4	5	6	7
y	1	2	4	7	11	16	22	29

5. Find y(12) using Newton's forward difference formula given:

x	10	20	30	40	50
y	46	66	81	93	101

6. From the following table of half-yearly premium for policies maturing at different ages, estimate the premium for policies maturing at age 46.

Age(X)	45	50	55	60	65
Premium (y)	114.84	96.16	83.32	74.48	68.48

7. From the given table, the values of y are consecutive terms of a series of which 23.6 is the 6^{th} term. Find the first and tenth terms of the series.

x	3	4	5	6	7	8	9
y	4.8	8.4	14.5	23.6	36.2	52.8	73.9

PROBLEMS BASED ON NEWTON'S BACKWARD INTERPOLATION FORMULA

1. Find the value of y at $x = 28$ from the following data:

x	20	23	26	29
y	0.3420	0.3907	0.4384	0.4848

Find $f(2.25)$ using Newton's backward difference formula from the following data:

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y	0.3679	0.2865	0.2231	0.1738	0.1353

3. Find (0.9) from the following table by using Newton's method.

x	0.2	0.4	0.6	0.8	1.0
$f(x)$	0.9798622	0.9177710	0.8080348	0.6386093	0.3483735

4.Find the sixth term of the sequence 8,12,19,29,42.

PROBLEMS BASED ON NEWTON'S BACKWARD & FORWARD INTERPOLATION FORMULA

From the given table, the values of y are consecutive terms of a series of which 23.6 is the 6^{th} term. Find the first and tenth terms of the series.

x	3	4	5	6	7	8	9
y	4.8	8.4	14.5	23.6	36.2	52.8	73.9

2. The following are data from the steam table:

$Temp\,^{\circ}C$	140	150	160	170	180
Pressure $\left(kg\ f\ cm^2\right)$	3.685	4.854	6.302	8.076	10.225

Using Newton's formula, find the pressure of the steam for a temperature of 142^o and 175^o

PROBLEMS BASED ON CUPIC SPLINE APPROXIMATION

1. Fit a cubic spline curve for the points $(2,11)$, $(3,49)$ and $(4,123)$. Hence find $y(2.5)$ and $y(3.5)$. Assume that $y''(2) = 0$ and $y''(4) = 0$.

2. Using cubic spline, find $y(0.5)$ and $y(1.5)$ from the following data, assuming that $y''(0) = 0$ and $y''(2) = 0$.

x	0	1	2
y	-5	-4	3

3. Fit the cubic spline for the data:

x	1	2	3	4
$f(x)$	1	2	5	11

Assume that $y''(1) = 0$ and $y''(4) = 0$.

4. Find the cubic spline interpolation:

x	1	2	3	4	5
$f(x)$	1	0	1	0	1

5. Find the cubic spline approximation for the function given below, assuming that $y''(1) = 0$ and $y''(3) = 0$.

x	1	2	3
y	-8	-1	18

6. Find the cubic spline approximation for the function y=f(x) from the data,given that $y_0'' = y_3'' = 0$

x	-1	0	1	2
y	-1	1	3	35

7.Given the following table , find f(2.5)using cubic spline functions:

i	0	1	2	3
x_i	1	2	3	4
$f(x_i)$	0.5	0.3333	0.25	0.2

8. Fit a natural cubic spline for the following data:

x	0	1	2	3
Y	1	4	0	-2

UNIT-III

NUMERICAL DIFFERENTIATION AND INTEGRATION

PART-B

PROBLEMS BASED ON NEWTON'S FORWARD DIFFERENCE FORMULA (EQUAL INTERVALS)

1. Find the first and second derivatives of the function tabulated below at $x = 0.4$

$x:$	0.4	0.5	0.6	0.7	0.8
$y:$	1.5836	1.7974	2.0442	2.3275	2.6511

2. For the following values of x and y , find the first derivative at $x = 1.05$

$x:$	1	1.05	1.1	1.15	1.2	1.25	1.3
$y:$	1	1.025	1.049	1.072	1.095	1.118	1.140

3. Find $\frac{dy}{dx}$ and $\frac{d^2y}{dx^2}$ for $x = 1.2$ from the following data

$x:$	1.2	1.4	1.6	1.8	2.0	2.2
$y:$	3.3201	4.0552	4.9530	6.0496	7.3891	9.0250

PROBLEMS BASED ON NEWTON'S BACKWARD DIFFERENCE FORMULA(EQUAL INTERVALS)

4. Find y' and y'' at x=1.25 for the data given:

$x:$	1.00.	1.05	1.10	1.15	1.20	1.25	1.30
$y:$	1.00000	1.02470	1.04881	1.07238	1.09544	1.11803	1.14017

Find $\frac{dy}{dx}$ and $\frac{d^2y}{dx^2}$ at $x = 10$ for the following data:

$x:$	2	4	6	8	10
$y:$	6	54	134	246	390

6. Find the first and second derivatives of y w.r.to.x at x=4 and x=10

$x:$	3	5	7	9	11
Y	31	43	57	41	27

PROBLEMS BASED ON LAGRANGE'S METHOD (UNEQUAL INTERVALS)

7. Using the given data find $f'(5)$

$x:$	0	2	3	4	7	9
f(x):	4	26	58	112	466	922

(OR) Find $f'(6)$ and the maximum value of y=f(x) given the data:

$x:$	0	2	3	4	7	9
f(x):	4	26	58	112	466	922

PROBLEMS BASED ON TRAPEZOIDAL , SIMPSON'S RULE

8. Evaluate $\int_0^1 \frac{dx}{1+x^2}$ using Trapezoidal rule with 10 subintervals. Hence approximate the value of π.

9. Find the value of $\log_e 2$ from $\int_0^1 \frac{x^2 dx}{1+x^3}$ using Simpson's $\frac{1}{3}$ rule with h=0.25.

10. Evaluate $\int_0^{\pi/2} \sin x dx$ using

(i) Simpson's $\frac{1}{3}$ rule and

(ii) Simpson's $\frac{3}{8}$ rule, by dividing the range into six equal subintervals.

11. Evaluate $\int_4^{5.2} \log_e x dx$ using Simpson's $\frac{3}{8}$ rule.

12. Evaluate $\int_0^6 \frac{dx}{1+x^2}$ by Simpson's $\frac{3}{8}$ rule.

13. Compute the value of $\int_{0.2}^{1.4} (\sin x - \log x + e^x) dx$ taking h=0.2 and using Simpson's rules.

14. Evaluate $\int_0^1 \frac{dx}{1+x^2}$ using Trapezoidal and Simpson's $\frac{1}{3}$ rule with 8 subintervals.

PROBLEMS BASED ON ROMBERG'S METHOD

15. Evaluate $\int_0^1 \frac{dx}{1+x^2}$ using Romberg's method. Hence deduce an approximate value of π.

(OR)Use Romberg's method to compute $\int_0^1 \frac{1}{1+x^2}dx$ correct to 4 decimal places by taking h=0.5, 0.25 and 0.125.

PROBLEMS BASED ON GAUSSIAN TWO AND THREE POINT FORMULA

16. By Gaussian three point formula, evaluate $\int_2^3 \frac{dt}{1+t}$.

17. Evaluate $I=\int_0^1 \frac{dx}{1+x}$ by two and three point Gaussian formulae.

18. Evaluate $I=\int_3^7 \frac{dx}{1+x^2}$ by two and three point Gaussian formulae.

PROBLEMS BASED ON TRAPEZOIDAL , SIMPSON'S RULE FOR DOUBLE INTEGRAL

19. Evaluate $\int_0^1\int_0^1 \frac{dxdy}{xy+1}$ by using Trapezoidal rule taking (i) $h=0.5$ and $k=0.25$.(ii)h=k=0.5.

20. Evaluate $\int_1^{1.2}\int_1^{1.4} \frac{dxdy}{x+y}$ by Trapezoidal rule taking $h=0.1$ and $k=0.1$.

21. Evaluate $\int_1^2\int_1^2 \frac{dxdy}{x+y}$ by Trapezoidal rule taking 4 subintervals.

Evaluate $\int_2^{2.4}\int_4^{4.4} xydxdy$ using Simpson's $1/3$ rule, dividing the range of x and y into 4 equal parts.

Evaluate $\int_1^2\int_3^4 \frac{dxdy}{(x+y)^2}$ by Simpson's rule taking $h=k=0.5$.

24. Evaluate $\int_0^{\pi/2}\int_0^{\pi/2} \cos(x+y)dxdy$ using Simpson's rule by taking $h=k=\pi/4$.

UNIT: IV

INITIAL VALUE PROBLEMS FOR ORDINARY DIFFERENTIAL EQUATIONS

PART-B

PROBLEMS BASED ON TAYLOR SERIES METHOD

using Taylor series method, find y at $x=0.1, 0.2$ given that $\frac{dy}{dx}=x^2-y$ and $y(0)=1$.

By Taylor series method find y(0.1) given that $y''=y+xy'$, $y(0)=1$, $y'(0)=0$.

using Taylor series method, find y at x=0.1 given $\frac{dy}{dx}=e^x-y^2$, $y(0)=1$.

PROBLEMS BASED ON EULER AND MODIFIED EULER METHOD

1. Compute $y(4.1)$ and $y(4.2)$ by using Euler method given that $5x\frac{dy}{dx}+y^2-2=0$, $y(4)=1$.
2. using modified Euler method, find $y(0.1)$and $y(0.2)$ given $\frac{dy}{dx}=x^2+y^2$, $y(0)=1$.
3. Compute y at $x=0.25$by modified Euler method, given $\frac{dy}{dx}=2xy$, $y(0)=1$.

4. Solve $\frac{dy}{dx} = \log_{10}(x+y)$, $y(0) = 2$ by Euler's modified method and find the values of y(0.2),y(0.4) and y(0.6), taking h=0.2.

PROBLEMS BASED ON RUNGE- KUTTA METHOD OF FOURTH ORDER

1. Given $y'' + xy' + y = 0$, $y'(0) = 0$, $y(0) = 1$, find the value of y(0.1) by Runge-Kutta method of fourth order.

2. Solve the equation $\frac{dy}{dx} = \frac{1}{x+y}$, $y(0)=1$ for $y(0.1)$ and $y(0.2)$ using Runge-Kutta method of fourth order.

3. Apply Runge-Kutta fourth order method to find an approximate value of y when $x = 0.2$ where it is given that $\frac{dy}{dx} = x + y$ and $y(0) = 1$.

4. Compute y(0.2) given $\frac{dy}{dx} = \frac{y^2 - x^2}{y^2 + x^2}$, $y(0) = 1$, by Runge-Kutta method of fourth order, taking h= 0.2.

5. Compute $y(0.3)$ given $\frac{dy}{dx} + y + xy^2 = 0$, $y(0) = 1$ by taking $h = 0.1$, using Runge-Kutta fourth order method. (Correct to 4 decimals).

6. Find $y(0.1)$, $y(0.2)$ and $y(0.3)$ from $\frac{dy}{dx} = xy + y^2$, $y(0)=1$ by using Runge-Kutta fourth order method, Correct to 4 decimals.

PROBLEMS BASED ON MILNE'S PREDICTOR – CORRECTOR METHOD

1. Using Milne's predictor-corrector method, find $y(4.4)$ given $5xy' + y^2 - 2 = 0$ given $y(4) = 1$, $y(4.1) = 1.0049$, $y(4.2) = 1.0097$, $y(4.3)$ 1.0143 .

2. Given $y' = 1 - y$, $y(0) = 0$ and $y(0.1) = 0.1$. Obtain $y(0.2)$ by improved Euler method $y(0.3)$ and by Runge-Kutta fourth order method. Hence find $y(0.4)$ by Milne's method.

3. Find $y(2)$ if $y(x)$ satisfies the differential equation $2\frac{dy}{dx} - x = y$, given $y(0) = 2$, $y(0.5) = 2.636$, $y(1) = 3.595$ and $y(1.5) = 4.968$ using Milne's method.

4. Solve $\frac{dy}{dx}$ $xy + y^2$, $y(0) = 1$, using Milne's Predictor – Corrector formulae and find y(0.4). Use Taylor series method to find y(0.1), y(0.2) ,y(0.3).

5. Solve $\frac{dy}{dx} = x(x^2 + y^2)e^{-x}$, $y(0) = 1$, using Milne's Predictor – Corrector formulae and find y(0.4). Use Taylor series method to find y(0.1), y(0.2) ,y(0.3).

6.Solve $\frac{dy}{dx} = y - x^2$, $y(0) = 1$ Find y(0.1) and y(0.2) by R.K method for order 4.Find y(0.3) by Euler's method. Find y(0.4) by Milne's Predictor Corrector method.

PROBLEMS BASED ON ADAM'S PREDICTOR – CORRECTOR METHOD

1. Using Taylor series method, find y at $x = 1.1, 1.2$ and 1.3 given that $\frac{dy}{dx} = x^2(1+y), y(1) = 1$.

Also compute $y(1.4)$ by Adam's method.

2. Using Adam's Bash forth modified method, find y at $x = 1.4$ given that $\frac{dy}{dx}$ x $(1+\ )$ and $y(1) = 1$, $y(1.1) = 1.233$, $y(1.2) = 1.548$ and $y(1.3) = 1.979$.

3. Given $\frac{dy}{dx} = y - x^2$, $y(0) = 1$, $y(0.2)$ 1.1218 , $y(0.4)$ 1.4682 , $y(0.6)$ 1.7379 ,estimate $y(0.8)$

by Adam's method.

Evaluate $y(0.4)$ using Adam's method, given that $\frac{dy}{dx}$ x^2 $- y$ given $y(0) = 1$, $y(0.1) = 0.9052$, $y(0.2) = 0.8213$ and $y(0.3) = 0.7492$.

5. Find $y(0.1), y(0.2), y(0.3)$ *from* $\frac{dy}{dx}$ x^2 $- y, y(0) = 1$ by using Taylor's method and hence

obtain
y(0.4) using Adam's method.

UNIT: V
BOUNDARY VALUE PROBLEMS IN ORDINARY AND PARTIAL DIFFERENTIAL EQUATIONS

PART-B

1. Solve, by finite difference method, $\frac{d^2y}{dx^2} = y$ where $y(0) = 0$ and $y(1) = 1$, taking $h = \frac{1}{4}$.

2. Solve $\frac{d^2y}{dx^2} = xy$ given $y(0) = -1, y(1) = 2$ by finite difference method with $h = \frac{1}{2}$.

3. Solve $y'' - xy = 0$ given $y(0) = -1$; $y(1) = 2$ by finite difference method taking $n = 2$.

4. Derive Bender-Schmidt recurrence formula to solve one dimensional heat equation.

5. Solve $\frac{\partial^2 u}{\partial x^2} = 2\frac{\partial u}{\partial t}$ given $u(0,t) = 0$, $u(4,t) = 0$ and $u(x,0) = 3x(4-x)$. Assuming $h = 1$, find the values of u up to $t = 5$ by Bender-Schmidt method.

6. Using Bender-Schmidt recurrence method, solve numerically the parabolic equation $2u_{xx} = u_t$, subject to boundary and initial conditions

(i) $u(0,t) = 0$, $t > 0$

(ii) $u(12,t) = 0$, $t > 0$ and

(iii) $u(x,0) = 3x(12-x)$, $0 \le x \le 12$.

Assuming $h = 2$, find the values of u up to $t = 5$ properly choosing the step size k in the time direction.

7. Obtain the simplest explicit scheme to solve $\frac{\partial u}{\partial t} = \alpha^2 \frac{\partial^2 u}{\partial x^2}$. Find the values of u up to 3 seconds ,taking the step size for x as $h = 1$, given that $\frac{\partial^2 u}{\partial x^2} = \frac{\partial u}{\partial t}$, $u(0,t) = u(5,t) = 0$ and $u(x,0) = x^2(25 - x^2)$.

8. Solve $25u_{xx} = u_{tt}$ for u at the pivotal points given $u(0,t) = u(5,t) = 0$, $u_t(x,0) = 0$ and

$\begin{cases} 2x & for\ 0 \le x \le 2 \end{cases}$

$$u(x,0)=\begin{cases} 10-2x \ \ for\ 3 \le x \le 5 \end{cases}$$

up to $t=1$ seconds taking $h=1$ and $k=\frac{1}{5}$.

9. Solve $u_{xx}+u_{yy}=0$, $0\le x\le 4$, $0\le y\le 4$ Given that $u(0,y)=0$, $u(4,y)=8+2y$, $u(x,0)=\frac{x^2}{2}$ and $u(x,4)=2$ taking $h=k=1$. Obtain the result correct to one decimal.

10. Solve the Laplace equation $u_{xx}+u_{yy}=0$ inside the square region bounded by the lines $x=0$, $x=4$, $y=0$ and $y=4$ given that $u=x^2y^2$ on the boundary.

11. Solve $u_{xx}+u_{yy}=0$ in the square region bounded by $x=0$, $x=4$, $y=0$, $y=4$ and with boundary conditions $u(0,y)=0$, $u(4,y)=8+2y$, $u(x,0)=\frac{x^2}{2}$ and $u(x,4)=x^2$ taking $h=k=1$. by Liebmann's method .Obtain the values of u at the interior mesh points by always correcting the computed values to two Places of decimals.

12. Solve the Poisson's equation $u_{xx}+u_{yy}=-4(x^2+y^2)$ over the square mesh with sides $x=0$, $y=0$, $x=3$ and $y=3$ with $u=0$ on the boundary and mesh length 1 unit.

13. Solve $\nabla^2 u=-10(x^2+y^2+10)$ over the square mesh with sides $x=0$, $y=0$, $x=3$ and $y=3$ with $u=0$ on the boundary and mesh length 1 unit.

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EE6401- ELECTRICAL MACHINES I
UNIT-1

1. What is magnetic circuit?

The closed path followed by magnetic flux is called magnetic circuit

2. Define magnetic flux?

The magnetic lines of force produced by a magnet is called magnetic flux it is denoted as Φ and its unit is Weber

3. Define magnetic flux density?

It is the flux per unit area at right angles to the flux it is denoted by B and unit is Weber/m^2

4. Define magneto motive force?

MMF is the cause for producing flux in a magnetic circuit. the amount of flux setup in the core decent upon current(I)and number of turns(N).the product of NI is called MMF and it determine the amount of flux setup in the magnetic circuit

MMF=NI ampere turns (AT)

5. Define reluctance?

The opposition that the magnetic circuit offers to flux is called reluctance. It is defind as the ratio of MMF to flux. It is denoted by S and its unit is AT/m

6. What is retentivity?

The property of magnetic material by which it can retain the magnetism even after the removal of inducing source is called retentivity

7. Define permeance?

It is the reciprocal of reluctance and is a measure of the cause the ease with which flux can pass through the material its unit is wb/AT

8. Define magnetic flux intensity?

It is defined as the mmf per unit length of the magnetic flux path. it is denoted as H and its unit is AT/m
H=NI/L

9. Define permeability?

Permeability of a material means its conductivity for magnetic flux. Greater the permeability of material, the greaters its conductivity for magnetic flux and vice versa

Define relative permeability?

It is equal to the ratio of flux density produced in that material to the flux density produced in air by the same magnetizing force $\mu_r=\mu/\mu_0$

What is mean by leakage flux?

The flux does not follow desired path in a magnetic circuit is called leakage flux

What is leakage coefficient?

Leakage coefficient=total flux/useful flux

State faradays law of electromagnetic induction

Whenever a flux linking in the coil changes emf always induced in the conductor the

magnitude of induced emf is proportional to rate of change flux linkage $e = Nd\Phi/dt$

State Lenz law?

The law states that induced emf always opposite to applied voltage source

Define self inductance?

The property of a coil that opposes any change in the amount of current flowing through it is called self inductance

Define mutual inductance?

The property of a coil to produce emf in a coil due to change in the value of current or flux in it is called mutual inductance

Define coefficient coupling?

It is defined as the fraction of magnetic flux produced by the current in one coil that links the other coil

Give the expression for hysteresis loss and eddy current loss?

Hysteresis loss$=k_h\, b_{max}^{1.62}fv$ watts

Eddy current loss$=k_e b_{max}^2 f^2 t^2 v$ watts/unit volume

What is dynamically induced emf?

An induced emf is produced by the movement of the conductor in a magnetic field. this emf is called dynamically induced emf. The dynamically induced emf $e = Blv\sin\Theta$

What is fringing effect?

It is seen that the useful flux passing across the air gap tends to buldge outwords, there by increasing the effective area of the air gap and reducing the flux density in the gap is called fringing effect

State two types of IM?

Squirrel cage IM
Slip ring IM

State ohms law for magnetic circuits?

Ohms law for magnetic circuits mmf=flux X reluctance

What is statically induced emf?

Conductor is stationary and the magnetic field is moving or changing the induced emf is called stationary induced emf

How eddy current losses are minimized?

By laminating the core'

State types of electrical machines?

1.DC machines
2.AC machines
3.Special
machines

What is mean by stacking factor?

Magnetic cores are made up of thin, lightly insulated laminations to reduce the eddy current loss. As a result, the net cross sectional area of the core occupied by the magnetic material is less than its gross cross section; their ratio being is called the stacking factor. The stacking value is normally less than one .its value vary from 0.5 to 0.95 .the stacking factor value is also reaches to one as the lamination thickness increases

What are the magnetic losses?

Eddy current loss
Hysterisis loss

Types of induced emf?

Dynamically induced emf
Statically induced emf

UNIT-2

1. Define a transformer?

A transformer is a static device which changes the alternating voltage from one level to

another.

2 What is the turns ratio and transformer ratio of transformer?

Turns ratio = N_2/ N_1
Transformer = $E_2/E_1 = I_1/ I_2 =K$

3. Mention the difference between core and shell type transformers?

In core type, the windings surround the core considerably and in shell type the core surrounds the windings i.e winding is placed inside the core

What is the purpose of laminating the core in a transformer?

In order to minimise eddy current loss.

Give the emf equation of a transformer and define each term?

Emf induced in primary coil $E_1= 4.44f\Phi_m N_1$ volt
Emf induced in secondary Coil $E_2 =4.44 f\Phi_m N_2$.
f-----------freq of AC input
Φ---------maximum value of flux in the core
N_1, N_2----Number of primary & secondary turns.

6. Does transformer draw any current when secondary is open? Why?

Yes, it (primary) will draw the current from the main supply in order to magnetize the

core and to supply for iron and copper losses on no load. There will not be any current

in the secondary since secondary is open.

7. Define voltage regulation of a transformer?

When a transformer is loaded with a constant primary voltage, the secondary voltage

decreases for lagging PF load, and increases for leading PF load because of its internal resistance and leakage reactance. The change in secondary terminal voltage from no load to full load expressed as a percentage of no load or full load voltage is termed as
regulation. %regulation $=E_2-V_2/E_2 *100$
V2>E2 for leading p.f load
V2<E2 for lagging p.f load

8. Define all day efficiency of a transformer?

It is computed on the basis of energy consumed during a certain period, usually a day of 24 hrs. All day efficiency=output in kWh/input in kWh for 24 hrs.

9. Why transformers are rated in kVA?

Copper loss of a transformer depends on current & iron loss on voltage. Hence total losses depend on Volt-Ampere and not on PF. That is why the rating of transformers is in kVA and not in kW.

What determines the thickness of the lamination or stampings?

1.Frequenc
y 2.Iron
loss

What are the typical uses of auto transformer?

1.To give small boost to a distribution cable to correct for the voltage drop. 2. as induction motor starter.

What are the applications of step-up & step-down transformer?

Step-up transformers are used in generating stations. Normally the generated voltage will be either 11kV. This voltage (11kV) is stepped up to 110kV or 220kV or 400kV and transmitted through transmission lines (simply called as sending end voltage).

Step-down transformers are used in receiving stations. The voltage are stepped down to 11kV or 22kV are stepped down to 3phase 400V by means of a distribution transformer and made available at consumer premises. The transformers used at generating stations are called power transformers.

How transformers are classified according to their construction?

1. Core type 2.shell type. In core type, the winding (primary and secondary) surround the core and in shell type, the core surround the winding.

Explain on the material used for core construction?

The core is constructed by sheet steel laminations assembled to provide a continuous magnetic path with minimum of air gap included. The steel used is of high silicon content sometimes heat treated to produce a high permeability and a low hysteresis loss at the usual operating flux densities. The eddy current loss is minimized by
laminating the core, the laminations being used from each other by light coat of core-plate vanish or by oxide layer on the surface. The thickness of lamination varies from 0.35mm for a frequency of 50Hz and 0.5mm for a frequency of 25Hz.

How does change in frequency affect the operation of a given transformer? With a change in frequency, iron and copper loss, regulation, efficiency & heating varies so the operation of transformer is highly affected.

What is the angle by which no-load current will lag the ideal applied voltage? In an ideal transformer, there are no copper & core loss i.e. loss free core. The no load current is only magnetizing current therefore the no load current lags behind by angle 90^0. However the winding possess resistance and leakage reactance and therefore the no load current lags the applied voltage slightly less than 90^0.

List the arrangement of stepped core arrangement in a transformer?

To reduce the space effectively
To obtain reduced length of mean turn of the winding
To reduce I^2R loss.

Why are breathers used in transformers?

Breathers are used to entrap the atmospheric moisture and thereby not allowing it to pass on to the transformer oil. Also to permit the oil inside the tank to expand and contract as its temperature increases and decreases.

What is the function of transformer oil in a transformer?

It provides good insulation
Cooling.

Can the voltage regulation goes –ive? If so under what condition?

Yes, if the load has leading PF.

Distinguish power transformers & distribution transformers?

Power transformers have very high rating in the order of MVA. They are used in generating and receiving stations. Sophisticated controls are required. Voltage ranges will be very high. Distribution transformers are used in receiving side. Voltage levels will be medium. Power ranging will be small in order of kVA. Complicated controls are not needed.

Name the factors on which hysteresis loss depends?

. Frequency 2. Volume of the core 3. Maximum flux density

Why the open circuit test on a transformer is conducted at rated voltage?

The open circuit on a transformer is conducted at a rated voltage because core loss depends upon the voltage. This open circuit test gives only core loss or iron loss of the transformer.

What is the purpose of providing Taps in transformer and where these are provided?

In order to attain the required voltage, taps are provided, normally at high voltages side(low current).

What are the necessary tests to determine the equivalent circuit of the transformer?

Open circuit test
Short circuit test

Define efficiency of the transformer?

Transformer efficiency η= (output power/input power) x 100

Mention the difference between core and shell type transformers?

In core type, the windings surrounded the core considerably and in shell type the core surround the windings i.e winding is placed inside the core

Full load copper loss in a transformer is 1600W. What will be the loss at half load?

If n is the ratio of actual load to full load then copper loss = n^2 (F.L copper loss) P_c = (0.5)2 – 1600=400W.

Define all day efficiency of a transformer?

It is computed on the basis of energy consumed during a certain period, usually a day of
24 hrs. All day efficiency=output in kWh/input in kWh tor 24 hrs.

List the advantage of stepped core arrangement in a transformer?

1.To reduce the space effectively
2.To obtain reduce length of mean turn of the winding 3.To reduce I^2R loss.

Why are breathers used in transformers?

Breathers are used to entrap the atmospheric moisture and thereby not allowing it to pass on to the transformer oil. Also to permit the oil inside the tank to expand and contract as its temperature increases and decreases

UNIT-3

1. State the principle of electromechanical energy conversion?

The mechanical energy is converted in to electrical energy which takes place through either by magnetic field or electric field

2. Distinguish between statically induced emf and dynamically induced emf?

When emf induced in a conductor is stationary in a magnetic field then we call it statically induced emf.

If emf is induced in a conductor due to relative motion between conductor and the field then it call it as dynamically induced emf.

3. What does speed voltage mean?

It is that voltage generated in that coil, when there exists a relative motion between coil and magnetic field

4. Give example for single and multiple excited systems?

Single excited system-reluctance motor, single phase transformer, relay coil

Multiply excited system-alternator, electro mechanical transducer

Why do all practical energy conversion devices make use of the magnetic field as a coupling medium rather than electric field?

When compared to electric field energy can be easily stored and retrieved form a magnetic system with reduced losses comparatively. Hence most all practical energy conversion devices make use of magnetic medium as coupling

State necessary condition for production of steady torque by the interaction of stator and rotor field in electric machines?

1.The stator and rotor fields should not have any relative velocity or speed between each other

2.Airgap between stator and rotor should be minimum 3.Reluctance of iron path should be negligible

4.Mutual flux linkages should exist between stator and rotor windings

7. Write the application of single and doubly fed magnetic systems?

Singly excited systems are employed for motion through a limited distance or rotation through a prescribed angle

Whereas multiply excited systems are used where continues energy conversion take place and in ease of transducer where one coil when energized the care of setting up of flux and the other coil when energized produces a proportional signal either electrical or mechanical

Explain the following with respect to rotating electrical machines

Pole pitch
Charding angle

Pole pitch is that centre to centre distance between any two consecutive poles in a rotating machine, measured in slots per poles

Chording angle is that angle by which the coil span is short of full pitched in electrical degrees

Why energy stored in a magnetic material always occur in air gap

In iron core or steel core the saturation and aging effects form hindrance to storage

Built in air gap as reluctance as well permeability is constant, the energy storage takes place linearly without any complexity

Hence energy is stored in air gap in a magnetic medium

What is the significance of co energy?

When electrical energy is fed to coil not the whole energy is stored as magnetic energy

.the co energy gives a measure of other energy conversion which takes place in coil then magnetic energy storage
Field energy
Coenergy

Write the equation which relates rotor speed in electrical and mechanical radians per second?

$\dot{\omega}_e = \dot{\omega}_m (p/2)$
$\dot{\omega}_e$=speed in electrical radians per sec $\dot{\omega}_m$=speed in mechanical radians per sec p=no of poles

Relate co energy density and magnetic flux density?

Co energy density=$w_f=\int_0 \lambda$ (I, x)
di $w_f=1/2BH$

Short advantages of short pitched coil?

Hormonics are reduced in induced voltage
Saving of copper
End connections are shorter

What is the significance of winding factor?

Winding factor gives the net reduction in emf induced due to short pitched coil wound in distributed type
Winding factor
$k_w=k_pk_d$ k_p= pitch factor
k_d= distribution factor $k_p= \cos(\alpha/2)$
$k_d= \sin(m\gamma/2)/m\sin(\gamma/2)$

What is the necessity to determine the energy density in the design of rotating machines?

Energy density $w_f=B^2/2\mu$

Derive the relation between co energy and the phase angle between the rotor and stator fluxes of the rotating machines?

F_1, f_2 are the rotor and stator flux peak values respectively $F_r^2=f_1^2+f_2^2+2f_1f_2\cos\alpha$
Co energy= { $f_1^2+f_2^2+2f_1f_2\cos\alpha$}

Write the energy balance equation for motor?

Mechanical energy o/p-=electrical energy i/p-increase in field energy $F_f\, dx=id\lambda-dW_f$

Write the expression for the mechanical energy output when the armature moves from one position to other with constant coil current?

Let us assume armature moves from position x_a to x_b for a constant coil current The mechanical energy is

UNIT-4

1. What is prime mover?

The basic source of mechanical power which drives the armature of the generator is called prime mover.

2. Give the materials used in machine manufacturing?

There are three main materials used in m/c manufacturing they are steel to conduct magnetic flux copper to conduct electric current insulation.

3. What are factors on which hysteresis loss?

It depends on magnetic flux density, frequency & volume of the material.

4. What is core loss? What is its significance in electric machines?

When a magnetic material undergoes cyclic magnetization, two kinds of power losses occur on it. Hysteresis and eddy current losses are called as core loss. It is important in determining heating, temperature rise, rating & efficiency of transformers, machines & other A.C run magnetic devices.

5. What is eddy current loss?

When a magnetic core carries a time varying flux, voltages are induced in all possible path enclosing flux. Resulting is the production of circulating flux in core. These

circulating current do no useful work are known as eddy current and have power loss known as eddy current loss.

5. How hysteresis and eddy current losses are minimized?

Hysteresis loss can be minimized by selecting materials for core such as silicon steel & steel alloys with low hysteresis co-efficient and electrical resistivity. Eddy current losses are minimized by laminating the core.

7. How will you find the direction of emf using Fleming's right hand rule?

The thumb, forefinger & middle finger of right hand are held so that these fingers are mutually perpendicular to each other, then forefinger-field thumb-motion middle current.

How will you find the direction of force produced using Fleming's left hand rule?

The thumb, forefinger & middle finger of left hand are held so that these fingers are mutually perpendicular to each other, then forefinger-field thumb-motion middle-current.

9. Write down the emf equation for d.c.generator?

$E= (\Phi NZ/60)\ (P/A)$

V. p--->no of poles

Z--->Total noof conductor

Φ--->flux per pole, N--->speed in rpm.

Why the armature core in d.c machines is constructed with laminated steel sheets instead of solid steel sheets?

Lamination highly reduces the eddy current loss and steel sheets provide low reluctance path to magnetic field.

Why commutator is employed in d.c.machines?

Conduct electricity between rotating armature and fixed brushes, convert alternating emf into unidirectional emf (mechanical rectifier).

Distinguish between shunt and series field coil construction?

Shunt field coils are wound with wires of small section and have more no of turns. Series field coils are wound with wires of larger cross section and have less no of turns.

How does D.C. motor differ from D.C. generator in construction?

Generators are normally placed in closed room and accessed by skilled operators only. Therefore on ventilation point of view they may be constructed with large opening in the frame. Motors have to be installed right in the place of use which may have dust, dampness, inflammable gases, chemicals….etc. to protect the motors against these elements, the motor frames are made either partially closed or totally closed or flame proof.

How will you change the direction of rotation of d.c.motor?

Either the field direction or direction of current through armature conductor is reversed.

What is back emf in D.C. motor?

As the motor armature rotates, the system of conductor come across alternate north and South Pole magnetic fields causing an emf induced in the conductors. The direction of the emf induced in the conductor is in opposite to current. As this emf always opposes the flow of current in motor operation it is called as back emf.

What is the function of no-voltage release coil in D.C. motor starter?

As long as the supply voltage is on healthy condition the current through the NVR coil produce enough magnetic force of attraction and retain the starter handle in ON position against spring force. When the supply voltage fails or becomes lower than a prescribed value then electromagnet may not have enough force to retain so handle will come back to OFF position due to spring force automatically.

Enumerate the factors on which speed of a d.c.motor depends?

$N = (V - I_aR_a)/\Phi$ so speed depends on air gap flux, resistance of armature, voltage applied to armature.

Under What circumstances does a dc shunt generator fails to generate?

Absence of residual flux, initial flux setup by field may be opposite in direction to residual flux, shunt field circuit resistance may be higher than its critical field resistance; load circuit resistance may be less than its critical load resistance.

Define critical field resistance of dc shunt generator?

Critical field resistance is defined as the resistance of the field circuit which will cause the shunt generator just to build up its emf at a specified field.

Why is the emf not zero when the field current is reduced to zero in dc generator?

Even after the field current is reduced to zero, the machine is left out with some flux as residue so emf is available due to residual flux.

On what occasion dc generator may not have residual flux?

The generator may be put for its operation after its construction, in previous operation; the generator would have been fully demagnetized.

What are the conditions to be fulfilled by for a dc shunt generator to build back emf?

The generator should have residual flux, the field winding should be connected in such a manner that the flux setup by field in same direction as residual flux, the field resistance should be less than critical field resistance, load circuit resistance should be above critical resistance.

23 Define armature reaction in dc machines?

The interaction between the main flux and armature flux cause disturbance called as armature reaction.

What are two unwanted effects of armature reactions?

Cross magnetizing effect & demagnetizing effect.

What is the function of carbon brush used in dc generators?

The function of the carbon brush is to collect current from commutator and supply to external load circuit and to load.

What are the 2 types of 3phase induction motor?

Squirrel cage and slip ring induction motor.

Write two extra features of slip ring induction motor?

Rotor has 3phase winding, Extra resistance can be added in rotor circuit for improving PF with the help of three slip rings.

Why an induction motor is called as rotating transformer?

The rotor receives same electrical power in exactly the same way as the secondary of a two winding transformer receiving its power from primary. That is why induction motor is called as rotating transformer.

Why an induction motor never runs at its synchronous speed?

If it runs at sy.speed then there would be no relative speed between the two, hence no rotor emf, so no rotor current, then no rotor torque to maintain rotation.

What are slip rings?

The slip rings are made of copper alloys and are fixed around the shaft insulating it. Through these slip rings and brushes rotor winding can be connected to external circuit.

What is the advantage of cage motor?

Since the rotor has low resistance, the copper loss is low and efficiency is very high. On account of simple construction of rotor it is mechanically robust, initial cost is less; maintenance cost is less, simple starting arrangement.

What are different methods of speed control in D.C shunt motor?

1.Armature control 2.Flux or field control 3.Applied voltage control

When is a four point DC starter required in DC motors?

A four point DC starter is required for dc motor under field control

If speed is decreased in a dc motor, what happens to the back emf decreases and armature current?

If speed is decreased in a dc motor, the back emf decreases and armature current increases

How does a series motor develop high starting torque?

A dc series motor is always started with some load. Therefore the motor armature current increases. Due to this, series motor develops high starting torque.

What is the necessity of starter in dc motors?

When a dc motor is directly switched on, at the time of starting, the motor back emf is zero. Due to this, the armature current is very high. Due to the very high current, the motor gets damaged. To reduce the starting current of the motor a starter is used.

Mention the types of braking of dc motor?

1.Regenerative braking
2.Dynamic braking
3.Plugging

What are the losses in dc motor?

1.Copper losses 2.Iron losses 3.Mechanical losses

Name any 2 non-loading method of testing dc machines?

1.Swinburne'stest
2.Hopkinson test

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UNIT-5

1. What is prime mover?

The basic source of mechanical power which drives the armature of the generator is called prime mover.

What are the essential parts of a d.c generator?

Magnetic frame or yoke 2. Poles 3. Armature 4. Commutator, pole shoes, armature windings, interpoles 5. Brushes, bearings and shaft.

Give the materials used in machine manufacturing?

There are three main materials used in m/c manufacturing they are steel to conduct magnetic flux copper to conduct electric current insulation.

3. What are factors on which hysteresis loss?

It depends on magnetic flux density, frequency & volume of the material.

4. What is core loss? What is its significance in electric machines?

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6. How hysteresis and eddy current losses are minimized?

Hysteresis loss can be minimized by selecting materials for core such as silicon steel &

steel alloys with low hysteresis co-efficient and electrical resistivity. Eddy current losses are minimized by laminating the core.

How will you find the direction of emf using Fleming's right hand rule?

The thumb, forefinger & middle finger of right hand are held so that these fingers are mutually perpendicular to each other, then forefinger gives the direction of the lines of flux, thumb gives the direction of the relative motion of conductor and middle finger gives the direction of the emf induced.

8. How will you find the direction of force produced using Fleming's left hand rule?

The thumb, forefinger & middle finger of left hand are held so that these fingers are mutually perpendicular to each other, then forefinger gives the direction of magnetic field, middle finger gives the direction of the current and thumb gives the direction of the force experienced by the conductor.

9. What is the purpose of yoke in d.c machine?

1.It acts as a protecting cover for the whole machine and provides mechanical support for the poles.
2.It carries magnetic flux produced by the poles

What are the types of armature winding?

1.Lap winding, A=P, 2.Wave winding, A=2.

How are armatures windings are classified based on placement of coil inside the armature slots?

Single and double layer winding.

Write down the emf equation for d.c.generator?

E= (ΦNZ/60)(P/A)V. p---------no of poles

Z---------Total no of conductor Φ---------flux per pole N---------speed in rpm.

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current through
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handle in ON position against spring force. When the supply voltage fails or becomes lower than a prescribed value then electromagnet may not have enough force to retain so handle will come back to OFF position due to spring force automatically.

Enumerate the factors on which speed of a d.c.motor depends?

N= (V-IaRa)/Φ so speed depends on voltage applied to armature, flux per pole, resistance of armature.

Under what circumstances does a dc shunt generator fails to generate?

Absence of residual flux, initial flux setup by field may be opposite in direction to residual flux, shunt field circuit resistance may be higher than its critical field resistance, load circuit resistance may be less than its critical load resistance.

Define critical field resistance of dc shunt generator?

Critical field resistance is defined as the resistance of the field circuit which will cause the shunt generator just to build up its emf at a specified field.

Why is the emf not zero when the field current is reduced to zero in dc generator?

Even after the field current is reduced to zero, the machine is left out with some flux as residue so emf is available due to residual flux.

24. On what occasion dc generator may not have residual flux?

The generator may be put for its operation after its construction, in previous operation, the generator would have been fully demagnetized.

25. What are the conditions to be fulfilled by for a dc shunt generator to build back emf?

The generator should have residual flux, the field winding should be connected in such a manner that the flux setup by field in same direction as residual flux, the field resistance should be less than critical field resistance, load circuit resistance should be above critical resistance.

26. Define armature reaction in dc machines?

The interaction between the main flux and armature flux cause disturbance called as armature reaction.

27. What are two unwanted effects of armature reactions?

Cross magnetizing effect & demagnetizing effect.

28. What is the function of carbon brush used in dc generators?

The function of the carbon brush is to collect current from commutator and supply to external load circuit and to load.

What is the principle of generator?

When the armature conductor cuts the magnetic flux emf is induced in the conductor.

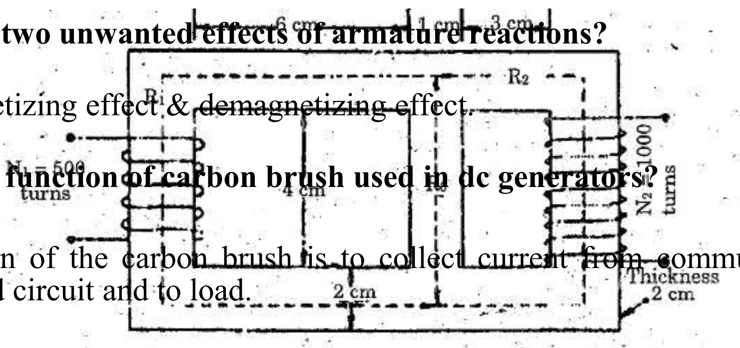
What is the principle of motor?

When a current carrying conductor is placed in a magnetic field it experiences a force tending to move it.

UNIT-I

PART: B

For the magnetic circuit as shown below, Calculate the self and mutual inductance between the two coils. Assume core permeability =1600 (16)



Explain the methods of energy conversion via Electric Field, with examples of Electrical Machines. (16)

(i)Specify the causes for Hysteresis and Eddy current losses in Electrical machines. Also give the methods in construction to minimize the above losses.(8)

(ii)List the properties of magnetic material suitable for fabrication Permanent Magnet and Electromegnet.(8)

19. (i)Describe the AC operation of magnetic circuits.(8)

Describe the principle of a typical magnetic circuit with air gap and explain. Also show that the core reluctance may be neglected in practice.(8)

The magnetic circuit has dimensions: A_c=4*4 cm2 I_g = 0.06 cm, I_c =40 cm and N= 600 turns. Assume the value of μ_r = 6000 for iron. Measure the exciting current for B_c = 1.2 T and the corresponding flux and flux linkages.(16)

A single phase 50 Hz , 100KVA transformer for 12000/240 V ratio has a maximum flux density of 1.2 Wb/m2 and an effective core section of 300 cm2 the magnetizing current is 0.2A.Identify the inductance of each wire on open circuit(16)

(i)Derive the expression for self and mutual inductance of the coil.(8)

Two coils A and B are wound on same iron core. There are 600 turns on A and 3600 turns on B. The current of 4 A through coil. A produces a flux of 500810-6 Wb in the core. If this current is reversed in0.02 sec. Identify the average emf induced in coils A and B. (8)

(i) Explain the losses in magnetic materials(8)

(ii)The field winding of dc electromagnets is wound with 800 turns and has a resistance of 40Ω when exciting voltage is 230V, magnetic flux around the coil is 0.004

Calculate self inductance and energy in magnetic field.(8)

(i) Give the expression for energy density in the magnetic field.(4)

Describe in detail “ Eddy-current loss”.(4)

The total core loss of a apecimen of silicon steel is found to be 1500W at 50 Hz. Keeping the flux density constant the loss becomes 3000 W when the frequency is raised to75 Hz. Calculate separately the hysteresis and eddy current loss at each of their frequencies. (8)

21. Compare the similarities and dissimilarities between electric and magnetic circuits.(16)

UNIT-II

PART: B

1.(i) Explain the principle of operation of a transformer. Derive its emf equation.(8)

A single phase transformer has 180 turns respectively in its secondary and primary windings. The respective resistances are 0.233 and 0.067. Calculate the equivalent resistance of a)the primary in terms of the secondary winding b)the secondary in terms of the primary winding c)the total resistance of the transformer in terms of the primary (8)

2.Explain the construction and working of core type and shell type transformers with neat sketches.(16)

3.evelop the equivalent circuit of a single phase transformer referred to primary and secondary.(16)

4.(i) Describe the phasor diagram of transformer when it is operating under load and explain.(8)

(ii)The parameters of approximate equivalent circuit of a 4 KVA, 200/400 V, 50 Hz single phase transformer are $R'_p = 0.15\ \Omega$; $X'_p = 0.37\ \Omega$; $R_o = 600\ \Omega$; $X_m = 300\ \Omega$ when a rated voltage of 200 V i_a applied to the primary, a current of 10A at lagging power factor of 0.8 flows in the secondary winding . Identify

(i)The current in the primary, I_p

(ii)The terminal voltage at the secondary side.(8)

5.(i) What is meant by Inrush current in Transformer? Describe the nature of inrush currents and its problem during transformer charging.(8)

A 500 KVA Transformer has a core loss of 2200 watts and a full load copper loss of 7500 watts. If the power factor of the load is 0.90 lagging, Evaluate the full load efficiency and the KVA load at which maximum efficiency occurs.(8)

6.(i) Summarize the generalised conditions for parallel operation of Transformer. Also explain the effect of load sharing due to impedance variation between transformers during parallel operation. (8)

(ii)A 100 KVA , 3300 V/240 V , 50 HZ single phase transformer has 990 turns on the primary. Identify the number of turns on secondary and the approximate value of primary and secondary full load currents.(8)

7.The voltage per turn of a single phase transformer is 1.1 volt, when the primary winding is connected to a 220 volt , 50 Hz AC supply the secondary voltage is found to be 550 volt. Identify the primary and secondary turns and core area if maximum flux density is 1.1 Tesla. (16)

8.Describe the principle of operation of a transformer. Draw the vector diagram to represent a load at UPF ,lagging and leading power factor.(16)

9.Obtain the equivalent circuit of a 200/400V 50 Hz single phase transformer from the following test data.

O.C.test: 200V, 0.7 W, 70W – on L .V

Side S.C. test: 15V, 10A, 85 W – on H.V

side

Calculate the secondary voltage when delivering 5 kW at 0.8 p.f. lagging. The primary voltage being 200V. (16

10.(i) Derive an expression for maximum efficiency of a transformer.(8)

A 500KVA transformer has 95% efficiency at full load and also at 60% of full load both at UPF.

a)Separate out the transformer losses.

b)Measure the transformer efficiency at 75% full load, UPF.(8)

UNIT-III

PART: B

1.Discuss the multiple excited magnetic field system in electromechanical energy conversion systems. Also obtain the expression for field energy in the system. (16)

2.Formulate the torque equation of a round rotor machine. Also clearly state the assumptions made. (16)

3.Describe in detail the production of mechanical force for an attracted armature relay excited by an electric source(16)

4.Explain briefly the production of rotating magnetic field. What are the speed and direction of rotation of the field ? Is the speed uniform? (16)

5.(i)Describe the concept of rotating MMF waves in AC machine. (8)

(ii)Obtain an expression for the mechanical force of field origin a typical Attracted armature relay.(8)

6. Derive an expression for the magnetic force developed in multiply excited magnetic systems.(16)

7.Derive an expression for co-energy density of an electromechanical energy conversion device. (16)

8.(i) Develop the torque in doubly excited magnetic system and show that is equal to the rate of increase of field energy with respect to displacement at constant current.(8) 9.The λ- I characteristics of singly excited electromagnet is given by i= 121 λ2x2 for $0<i<4$ A and $0<x<$ 10Cm. If the air gap is 5Cm and a current of 3A is flowing in the coil, Identify (a) Field Energy (b) Co- energy (c) Mechanical Force on the moving part.(8)

10. Describe the flow of energy in electromechanical devices. (8)

9.(ii)Describe about the 'field energy' and 'coenergy' in magnetic system.(4)

10.(iii)The magnetic flux density on the surface of an iron face is 1.6 T which is a typical saturation level value for ferromagnetic material. Identify the force density on the iron face.(4)

UNIT-IV

PART:B

(i) Draw and Explain the Load Characteristics of Differentially and Cumulatively compound DC generator. (8)

(ii) A 4 pole DC shunt generator with lap connected armature supplies 5 kilowatt at 230 Volts. The armature and field copper losses are 360 Watts and 200 Watts respectively. Calculate the armature current and generated EMF?

In a 400 volts, DC compound generator, the resistance of the armature, series and shunt windings are 0,10 ohm, 0.05 ohm and 100 ohms respectively. The machine supplies power to Nos. resistive heaters, each rated 500 watts, 400 volts. Identify the induced emf and armature currents when the generator is connected in (1) Short Shunt (2) Long Shunt. Allow brush contact drop of 2 volts per brush.

(i) Explain armature reaction and commutation in detail. (8)

Draw and explain the OCC Characteristics and External Characteristics of DC generator.(8)

4. Discuss the performance characteristics of different types of DC generators and explain them.

5. With neat sketch explain the Construction and principle of operation of DC Generator

6. A 6-pole DC generator has 150 slots. Each slots has 8 conductors and each conductor has resistance of 0.01Ω.The armature terminal current is 15 A. Calculate the current per conductor and the drop in armature for Lap and Wave winding connections.

7. (i)Show the condition for maximum efficiency of the DC generator.

(ii)Explain the following: (i) Self and separately excited DC generators (4) (ii) Commutation.(4)

8. A 400V DC shunt generator has a full load current of 200 A. The resistance of the armature and field windings are 0.06 Ω and 100 Ω respectively. The stray losses are 2000 W. infer the Kw output of prime mover when it is delivering full load and find the load for which the efficiency of the generator is maximum.

Describe briefly the different methods of excitation and characteristics of a DC generators with suitable diagrams.

Derive an expression for the EMF Equation of DC generator.

UNIT-V

PART: B

Describe briefly the various methods of controlling the speed of a DC shunt motor and bring out their merits and demerits. Also, state the situations where each method is suitable

Describe Plugging, dynamic and regenerative braking in DC Motor. 3. A 230 volts DC Shunt motor on no-load runs at a speed of 1200 RPM and draw a current of 4.5 Amperes. The armature and shunt field resistances are 0.3 ohm and 230 ohms respectively. Calculate the back EMF induced and speed, when loaded and drawing a current of 36 Amperes.

Discuss why starting current is high at the moment of starting a DC Motor? Explain the method of limiting the starting current in DC motors and also.

With neat sketch explain three point starter to start the DC Shunt motor.

A DC series motor runs at 500 rpm on 220 V supply drawing a current of50 A. The total resistance of the machine is 0.15Ω, calculate the value of the extra resistance to be connected in series with the motor circuit that will reduce the speed to 300 rpm. The load torque being then half of the previous to the current.

7. (i)A 500V dc shunt motor running at 700 rpm takes an armature current of50A.Its effective armature resistance is 0.4Ω. What resistance must be placed in series with the armature to reduce the speed to 600 rpm, the torque remaining constant?

(ii) Explain briefly the merits and demerits of Hopkinson's test?

Explain the different methods of excitation and characteristics of a DC motors with suitable diagrams.

A 400 Volts DC Shunt motor has a no load speed of 1450 RPM, the line current being 9 Amperes. At full loaded condition, the line current is 75 Amperes. If the shunt field resistance is 200 Ohms and armature resistance is 0.5Ohm. Evaluate the full load speed.

With neat circuit diagram explain the conduction of Swinburne's test and Hopkinson's test.

CS 6456 – OBJECT ORIENTED PROGRAMMING

2 Mark Questions

UNIT – I

OVERVIEW

1. Define Encapsulation.

Encapsulation is the process of combining data and functions into a single unit called class. Using the method of encapsulation, the programmer cannot directly access the data. Data is only accessible through the functions present inside the class.

2. What is Function Overloading?

A function with the same name performing different operations is called function overloading. To achieve function overloading, functions should be declared with the same name but different number and type of arguments.

3. What is Operator Overloading? Give examples.

The same operator performing different operations is called operator overloading. E.g. two matrices cannot be directly overloaded. But the + operator can be overloaded to perform

Addition of two matrices or other user defined data types.

4. Define a Class.

Class is a collection of data members with its member functions. Classes are data types based on which objects are created. Objects with similar properties and methods are grouped together to form a Class.

5. What is an Object?

An object is an instance of a class. Any real world entity such as pen, book, bird, student etc., can be termed as an object. They may also represent user-defined data. They contain both data and code.

What are the differences between Structural and Object Oriented Programming?

STRUCTURAL PROGRAMMING:

Importance is given to functions.

Reusability of code is not possible.

Does not provide abstraction.

OBJECT ORIENTED PROGRAMMING:

Importance is given to data.

Reusability is possible through inheritance.

Provides class level and object level abstraction.

7. What is the use of the specifier 'const'?

The "const" specifier before a variable makes it a constant whose value cannot be changed during the program. Similarly if it is specified before a pointer, the address contained in the pointer cannot be modified.

What does the "volatile" qualifier specify?

A variable or object declared with the volatile keyword may be modified externally from the declaring object. Variables declared to be volatile will not be optimized by the compiler because the compiler must assume that their values can change at any time.

9. What are Static Data Members?

Static variables are normally used to maintain values common to the entire class.

Properties are:

It is initialized to zero when the first object is created and no other initialization is

permitted only one copy of that member is created and shared by all the objects of that class.

- It is visible only within the class, but the life time is the entire program.

What are Friend functions?

The function declaration should be preceded by the keyword friend. A friend function is used for accessing the non-public members from outside of the class. A class can allow non-member functions and other classes to access its own private data, by making them as friends.

11. What are the various access specifiers available in C++?

The various access specifiers available in C++ are private, public and protected.

12. What are methods?

Methods are functions or procedures that can operate upon the data members of a Class.

13. What is a pointer?

A pointer is defined as the variable containing the address of another variable.

What are the advantages of object oriented programming?

It is a natural way of programming.

It allows reusability of code through inheritance.

Writing even large programs is easy.

Testing and managing code are also made easy.

15. What are abstract classes?

Classes containing at least one pure virtual function become abstract classes. Classes inheriting abstract classes must redefine the pure virtual functions; otherwise the derived classes also will become abstract. Abstract classes cannot be instantiated.

16. What is the use of default arguments?

Default arguments are used when a default value is to be supplied when the user does not provide any values. The default values can be overridden by the values supplied by the user. The function assigns a default value to the parameter which does not have a matching argument in the function call.

17. Define OOP.

Object Oriented Programming (OOP) is a method of implementation in which programs are organized as a collection of objects. OOP is a methodology that allows the association of data structures with its operations.

18. What are the features of OOP available?

The features of OOP are classes, objects, encapsulation, data abstraction, inheritance, delegation (or) object composition, polymorphism, dynamic binding, message communication and abstract classes.

19. What is pointer to an object?

Pointer to an object is the process of storing the address of an object into its particular memory location.

20. Define nested class.

A Class which is declared inside the declaration of another one class. That is, an inside class is declared inside the outside class.

21. Define local class.

A Class is to be defined inside the member function is called local class. The function in which the class is defined is not a member function of the class.

Give any four applications of OOP.

- Real time systems.
- Simulation and modeling.
- Object oriented databases.
- AI and expert systems.

What is a scope resolution operator?

A variable declared in an inner block cannot be accessed outside the block.

To resolve this problem the scope resolution operator is used. It can be used to uncover a hidden variable. This operator allows access to the global version of the variable.

Syntax:

return type class name **::** function name

Eg:

void sample **::** get data()

24. What are reference variable ?

A reference variable provides an alternative name (alias) for a previously defined variable. A reference is, exactly as the name suggests, a reference or pointer to another object.

Its **syntax** is given by,

data-type & reference-name = variable-name;

Eg : int &ptr = rupee;

25. What is an inline function?

An inline function is a function that is expanded in line when it is invoked. Here, the compiler replaces the function call with the corresponding function code. The inline function is defined as,

```
inline function-header
{
function body
}
```

UNIT – II

BASIC CHARACTERISITICS OF OOP

1. Define Constructor.

A constructor is a special member function whose task is to initialize the objects of its class. It has the same name as the class. It gets invoked whenever an object is created to that class.

```
Eg: class            //class class name
    sample {
      public:
      sample ();    //  class name(); → constructor declaration
    }
```

List some of the special characteristics of constructor.

Constructors should be declared in the public section.

They are invoked automatically when the objects are created.

They do not have return types

They cannot be inherited.

Give the various types of constructors.

There are four types of constructors. They are,

Default constructors – A constructor that accepts no parameters.

Parameterized constructors – The constructors that can take arguments.

Copy constructor – It takes a reference to an object of the same class as itself as an Argument

What are the ways in which a constructor can be called?

The constructor can be called by two ways. They are,

By calling the constructor explicitly.

e.g., student cricket = student (90, “caching”);

By calling the constructor implicitly.

e.g., student cricket (90, “sachin”);

5. State dynamic initialization of objects.

Class objects can be initialized dynamically. The initial values of an object may be provided during run time. The advantage of dynamic initialization is that various initialization formats can be used.

6. Define destructor.

A destructor is used to destroy the objects that have been created by a constructor. It is a special member function whose name is same as the class and is preceded by a tilde '~' symbol. When an object goes out from object creation, automatically destructor will be executed.

Example:

```
class File {
                          public:
                                  ~File();         //destructor
declaration
                        };
        File::~File()
          {
            close();           // destructor definition
          }
```

List some of the rules for operator overloading.

Only existing operators can be overloaded.

We cannot change the basic meaning of an operator.

The overloaded operator must have at least one operand.

Overloaded operators follow the syntax rules of the original operators.

What are the types of type conversions?

There are three types of conversions. They are

Conversion from basic type to class type – done using constructor

Conversion from class type to basic type – done using a casting operator

Conversion from one class type to another – done using constructor or Casting operator.

What are the conditions should a casting operator satisfy? The conditions that a casting operator should satisfy are,

It must be a class member.

It must not specify a return type.

It must not have any arguments.

10. Define parameterized constructor.

Constructor with arguments is called parameterized constructor.

Eg:

```
Class Student
{
    int m, n;
```

```
Public:
 Student (int x, int y)
  {   m=x; n=y;
   }
};
```

What are the characteristics of destructor?

A destructor must be declared in the public section of a class.

A class cannot have more than one destructor.

It has no return type.

It is incorrect to even declare a void return type.

Define Operator Overloading.

C++ has the ability to provide the operators with a special meaning to an operator is known as Operator Overloading. It is one of the methods of realizing polymorphism.

What are the operators that cannot be overloaded?

Class member access operator (. , .*)

Scope resolution operator (::)

Size operator (size of)

Conditional operator (? :)

Define default constructor.

The constructor with no arguments is called default constructor.

Eg:

```
Class integer
{
     int m, n;
     Public:
            integer ( ); //  default constructor
};
```

15. Define Copy Constructor

A copy constructor is used to declare and initialize an object from another object. It takes a reference to an object of the same class as an argument

Eg: integer i2 (i1);

would define the object i2 at the same time initialize it to the values of i1.

16. What is copy initialization process?

The process of initializing an object through a copy constructor is known as Copy initialization.

Eg: integer i2=i1;

17. Define dynamic constructor.

Allocation of memory to objects at time of their construction is known as dynamic constructor. The memory is allocated with the help of the NEW operator.

```
Eg: Class string
      {
         char *name;
        public: string( )
         {
            name = new char[length +1];
          }
      };
```

Write at least four rules for Operator Overloading.

- Only the existing operators can be overloaded.
- The overloaded operator must have at least one operand that is of user defined data type.
- The basic meaning of the operator should not be changed.
- Overloaded operators follow the syntax rules of the original operators. They cannot be overridden.

Define Binary Operator Overloading.

Binary operator overloading performs its operation by using 2 objects. The first object is passed as an implicit operand and the second object is passed explicitly.

20. Define explicit constructor.

Constructor can be defined explicitly by using the keyword "explicit" is known as an explicit constructor. The explicit constructor will be executed when we call the constructor explicitly.

Ex:

```
explicit brother (string name)
{
 Body of the explicit constructor
}
```

Note: brother is a class name.

21. How will you overload unary and binary operator using Friend Function?

When unary operators are overloaded using friend function, it takes one reference argument (object of the relevant class). When binary operators are overloaded using friend function, it takes two explicit arguments.

22. Define type conversion. What are its two types?

Type conversion is the process of converting one type into another type.It may be

One data type into another data type

One data type into an object (basic type to class type)

An object into data type (class type to basic type)

One class into another class.

What is type casting?

A casting operator is a function. It must be a class member. It must not specify a return type. It must not have any arguments.

The general form of overloaded casting operator is,

```
operator type name ( )
    {
        function statements
    }
```

24. Explain basic type to class type with an example.

Conversion from basic data type to class type can be done in destination class. Using constructors does it. Constructor takes a single argument whose type is to be converted.

Eg: Converting from int type to class complex

```
Complex (int r1, int i1)  // conversion from int data type into  the class complex
        { real = float(r1);
              imag = float(i1);
        }
```

real and imag are the objects of the class complex.

25. Explain class type to basic type with an example.

Using Type Casting operator, conversion from class to basic type conversion can be done. It is done in the source class itself. To perform this type of conversion, the conversion function should be defined in the form of an operator function.

Eg: operator double () // conversion of the objects real and imag to the data type double

```
{
    double s;
    s = double (real) + double (imag);
    return (s);
}
```

UNIT – III

ADVANCED PROGRAMMING

1. Define template.

A template in C++ allows the construction of a family of template functions and classes to perform the same operation on different data types. The template type arguments are called “generic data types”.

2. Define function template.

The templates declared for functions are called function templates. A function template is prefixed with a keyword template and list of template type arguments.

What is the syntax used for writing function template?

```
template <class T>,……….>
class name function name(arguments)
{
 Body of template function
}
```

4. Define class template.

The templates declared for classes are called class templates. Classes can also be declared to operate on different data types. A class template specifies how individual classes can be constructed similar to normal class specification.

5. What is the syntax used for writing class template?

```
template   < class T1, class T2, ……. >
class class name
  {
     data items of template type T1, T2……..
     functions of template arguments T1, T2 …
  };
```

6. Define exception handling process.

The error handling mechanism of C++ is generally referred to as an exception handling. It provides a mechanism for adding error handling mechanism in a program.

What are the two types of an exception? There are two types of an exception.

- Synchronous exception.
- Asynchronous exception.

How many blocks contained in an exception handling model?

Totally three blocks contained in an exception handling process.

Try block

Throw block

Catch block

Define throw construct.

The keyword throw is used to raise an exception when an error is generated in the computation. The throw expression initializes a temporary object of the type T used in throw.

Syntax: throw T // named object, nameless object or by default nothing.

10. Define catch construct.

The exception handler is indicated by the keyword catch. It must be used immediately after the statements marked by the keyword try. Each catch handler will evaluate an exception that matches to the specified type in the argument list.

Syntax:

```
Catch (T) // named object or nameless object
 {
     Actions for handling an exception
 }
```

11. Define try construct.

Try keyword defines a boundary within which an exception can occur. The try keyword is a block of code enclosed by braces. This indicates that the program is prepared to test for the exceptions. If an exception occurs, the program flow is interrupted.

Syntax:

```
try
{
    Code raising an exception
}
catch (type_id1)
{
        Actions for handling an exception
}
…………….
…………….
catch (type_idn)
{
          Actions for handling an exception
}
```

12. What are the tasks performed by an error handling mechanism?

Detect the problem causing an exception(hit the exception)

inform that an error has occurred(throw the exception)

receive the error information(catch the exception)

Take correct actions(handle the exception)

Define exception specification.

It is possible to specify what kind of exception can be thrown by functions, using a specific syntax. We can append the function definition header with throw keyword and possible type of expressions to be thrown in the parenthesis. It is known as exception specification.

What are the two types of an exception specification?

Terminate () function.

Unexpected () function.

Define terminate () function.

Terminate () is the function which calls abort () function to exit the program in the event of runtime error related to the exception.

16. Define unexpected () function.

If a function throws an exception which is not allowed, then a function unexpected () is called which is used to call abort () function to exit the program from its control. It is similar to Terminate () function.

17. Define multiple catch.

Using more than one catch sections for a single try block. At first matching, catch block will get executed when an expression is thrown. If no matching catch block is found, the exception is passed on one layer up in the block hierarchy.

18. Define catch all exception.

It is possible for us to catch all types of exceptions in a single catch section. We can use catch (…) (three dots as an argument) for representing catch all exception.

19. Define an exception.

Exceptions refer to unusual conditions or errors occurred in a program.

20. Define synchronous exception.

This type of an exception occurs during the program execution due to some fault in the input data or technique is known as synchronous exception.

Examples are errors such as out-of-range, overflow, and underflow.

21. Define asynchronous exception.

The exceptions caused by events or faults that are unrelated to the program. Examples are errors such as keyboard interrupts, hardware malfunctions and disk failures.

22. What do you mean by the term 'Generic Programming'?

In the context of C++ (and called Meta programming) it means to write programs that are evaluated at compile time. Templates are generic because the compiler translates the template into actual code.

23. What is an iterator? List out the characteristics of an iterator.

An iterator is any object that pointing to some element in a range of elements (such as an array or a container), has the ability to iterate through the elements of that range using a set of operators (with at least the increment (++) and dereference (*) operators).

24. What is an exception?

An exception is a problem that arises during the execution of a program. An exception is a response to an exceptional circumstance that arises while a program is running, such as an attempt to divide by zero. C++ exception handling is built upon three keywords: try, catch and throw.

25. What is throw() ? What is its use?

A program throws an exception when a problem shows up. This is done using a **throw** keyword**.** Exceptions can be thrown anywhere within a code block using **throw** statements. The operand of the throw statements determines a type for the exception and can be any expression and the type of the result of the expression determines the type of exception thrown.

Following is an example of throwing an exception when dividing by zero condition occurs:

```
double division(int a, int b)
{
  if( b == 0 )
{
    throw "Division by zero condition!";
  }
  return (a/b);
}
```

UNIT- IV

OVERVIEW OF JAVA

1. .What is Java?

Java is a high-level, third generation programming language, like C, FORTRAN, Smalltalk, Perl, and many others. You can use Java to write computer applications that crunch numbers, process words, play games, store data or do any of the thousands of other things computer software can do.

What are the features of Java?

The features of Java are,

Simple.

- Object Oriented.
- Platform Independent.
- Robust.

Multithreaded.

- Secure.

What are the various applications of Java? The various applications of Java are,

Applets

Networking

- Internationalization
- Security

v. Object serialization

Java Database Connectivity (JDBC)

What is meant by virtual machine?

A Java virtual machine (JVM), an implementation of the Java Virtual Machine Specification, interprets compiled Java binary code (called bytecode) for a computer's processor (or "hardware platform") so that it can perform a Java program's instructions

What are the two components of Java platform? The two components of Java platform are,

The Java Virtual Machine

The Java Application Programming Interface (API)

What is bytecode in Java?

Java bytecode is the form of instructions that the Java virtual machine executes. Each bytecode opcode is one byte in length, although some require parameters, resulting in some multi-byte instructions. Not all of the possible 256 opcodes are used.

7. What is an Object?

An object consists of data and functions known as methods which use or change the data. (Methods are similar to procedures or functions in other languages.) Objects of the same kind are said to have the same type or be in the same class. A class defines what data can be in an object, and what operations are performed by the methods. One or more objects can be created or "instantiated" from a class.

8. What is an Object and how do you allocate memory to it?

Object is an instance of a class and it is a software unit that combines a structured set of data with a set of operations for inspecting and manipulating that data. When an object is created using new operator, memory is allocated to it.

What are different types of access modifiers? The
different types of access modifiers are,

public: Any thing declared as public can be accessed from anywhere.

private: Any thing declared as private can‟t be seen outside of its class.

protected: Any thing declared as protected can be accessed by classes in the same package and

subclasses in the other packages.

default modifier: Can be accessed only to classes in the same package.

What is method overloading and method overriding?

Method overloading:

When a method in a class having the same method name with different arguments is said to be method overloading.

Method overriding:

When a method in a class having the same method name with same arguments is said to be method overriding.

11. List out the primitive types in Java.

The seven primitive types are listed in the following table:

Type	Definition
byte	one-byte signed two's complement integer
short	two-byte signed two's complement integer
int	4-byte signed two's complement integer
long	8-byte signed two's complement integer
float	4-byte IEEE 754 single-precision float
double	8-byte IEEE 754 double-precision float
char	2-byte unsigned Unicode character

12. What is String?

A String is a class used to store a sequence of characters in Java. Strings are constant. Their values cannot be changed after they are created. String buffers support mutable strings. Because String objects are immutable they can be shared. Example: String str = "abc";

13. What is an array?

An array is a special object containing a group of contiguous memory locations that have the same name and the same type and a separate variable containing an integer constant equal to the number of array elements. The elements of Java arrays are numbered starting from 0.

Example: double x []; // create an array reference

x = new double [5]; // create array object

What are the methods used in String class?

The methods used in String class are,

charAt(int index)

compareTo(String anotherString)

- concat(String str)
- copyValueOf(char[] data)
- equals(Object anObject)

Why Java use Unicode?

Java use Unicode to represent a character. Unicode defines a fully international character set that can represent all of the characters found in all human languages

16. What is Classpath?

The Classpath is an argument we can set either on the command-line, or through an environment variable that tells the Java Virtual Machine where to look for user defined classes and packages when running Java programs.

17. What is Garbage collection?

The Garbage collection is the process that is used to free the memory of the objects that are no longer in use. When a program stops referencing an object, it is not required any more and can be deleted. The space that is used by the object is released for use by another object

18. What are Nested classes?

A nested class is a class defined as a member of another class. The scope of nested class is bounded by the scope of its enclosing class. The nested class has access to the members of its enclosing class including private members.

19. What do you mean by inheritance?

A subclass inherits variables and methods from its superclass and all of its ancestors. The subclass can use these members as is, or it can hide the member variables or override the methods.

What are the advantages of inheritance?

The advantages of inheritance are,

It permits code reusability.

Reusability saves time in program development.

It encourages the reuse of proven and debugged high-quality software, thus reducing problem after a

system becomes functional.

21. What members does a Subclass inherit?

A subclass inherits all of the members in its superclass that are accessible to that subclass unless the subclass explicitly hides a member variable or overrides a method. Note that constructors are not members and are not inherited by subclasses.

22. What is a byte code?

Byte code is the compiled format for Java programs. Once a Java program has been converted to byte code, it can be transferred across a network and executed by Java Virtual Machine (JVM).

Write the output produced by the following code fragments.

System.out.println("Results=:"+40+30);

System.out.println("Results=:"+(40+30));

Output:

Results=:4030

Results=:70

Define the keyword 'static' in Java.

The static keyword in Java is used for memory management mainly. We can apply java static keyword with variables, methods, blocks and nested classes. The static keyword belongs to the class than instance of the class.

25. Java is robust. Comment.

Java is robust because it is highly supported language. It is portable across many Operating Systems. Java also has feature of Automatic memory management and garbage collection. Strong type checking mechanism of Java also helps in making Java Robust.

UNIT- V

EXCEPTION HANDLING

1. Define stream.

A Stream is a sequence of bytes. It acts either as a source for obtaining the input data or as a destination for sending the output data.

2. What is an input stream?

The source stream that provides the input data to the program is called input stream.

3. What is an output stream?

The destination stream that receives the output data from the program.

What are the unformatted input/output operations are used?

put ()

get ()

getline ()

write ()

5. What is put () function?

put () function is used to display a character to the output device.

Example:

Cout.put ('x'); → displays the character x.

Cout.put (ch); → displays the value of variable ch.

6. What is get () function?

get () function is used to read a character from the input device.

Example:

get (char x) → assign the input character to its argument x.

get void) → it returns the input character.

7. What is get line() function?

get line() function reads a whole line of text that ends with a new line character is transmitted by the return key.

Syntax:

cin.getline(line, size);

This function reads character input into the variable line.

8. What is write () function?

write () function is used to display the name of the string.

Syntax:

cout.write(line ,size);

line → the name of the string to be displayed.

size → number of characters to display.

What is the formatted console input/output operations are used?

width () – to specify the required field size for displaying an output value.

precision () – to display the number of digits to be displayed after the decimal point of a float value.

fill () – to specify a character that is used to fill the unused portions of a field.

setf () – to specify formal flags that can control the form of output display.

*unset () – to clear the flags specified.

10. Define manipulator.

The header file iomanip provides a set of functions called manipulators which can be used to manipulate the output formats. They provide the same features as that of the ios member functions and flags.

What are the manipulators are used?

setw (int width) – sets the field width.

setprecision (int prec) – sets the floating point precision.

setfill (int char) – sets the conversion base.

setiosflags (long flags) – sets the formal flags.

resetiosflags (long flags) – resets the formal flags.

Define files.

A file is a collection of related data stored in a particular area on a disk. Programs can be designed to perform the read and write operations on these files.

What are the methods for opening a file?

Opening files using constructor.

Opening files using open () method.

How to open a file using constructor?

Steps:

- Create a filestream object to manage the stream using the appropriate class.
- Initialize the object with desired file name. Example:

ofstream outfile ("results");

ofstream → the name of an outputstream. outfile → the name of an object. results → the name of the file.

How to open a file using open () method?

File can be opened by using open () function.

Syntax:

```
filestream   streamobject;
streamobject.open("filename");
```

example:

```
ofstream outfile;
outfile.open("data");
…………..
…………..
Outfile.close();          // every file must be closed with close() function.
```

How to detect the end of file?

Example 1: while (fin)

It returns a value of 0 if any error occurs in the file operation including the end of file condition.

Example 2: if (fin1.eof ()! = 0) {exit (1) ;}

eof() is a member function of ios class. It returns a non zero value if the end of file condition is encountered and zero otherwise.

What are the file modes are used in files?

ios :: in → open file for reading only

ios :: out → open file for writing only

ios :: app → append to end of file

ios :: ate → go to end of file on opening

ios :: binary → opens as binary file

ios :: nocreate → open fails if the file does not exist

ios :: noreplace → open fails if the file already exist

ios :: trunk → delete the contents of files if it exist

What are two file pointers?

Input pointer → used for reading the contents of a given file operation.

Output pointer → used for writing to a given file location

What are functions for manipulation of file pointers?

seekg() → moves get pointer to a specified location

seekp() → moves put pointer to a specified location

tellg() → gives the current location of the get pointer

tellp() → gives the current location of the put pointer

What is the use of std namspace.

Std is a name space where the standard library routines are defined. We can use objects like cout or cin without any qualification if we write using namespace std in the beginning of our program. We can use elements of std using qualifier std ::

What are the operations are performed on string objects?

Creating strings

substring operations

Comparison between C++ strings and C strings

What are the ways for creating string objects?

Defining a string object in a normal way

Defining a string object using initialization

Defining a string object using a constructor

What are substring operations?

Find location of a sub string or a character in a given string

find the character at a given location in a given string

insert a specific substring at a specific place replacing specific characters by other characters

append substring to a string

Define Standard Template Library (STL).

Standard Template Library (STL) is collection of generic software components and generic algorithms, glued by objects called iterators.

What are the advantages of using STL containers?

Automatically adjusting growth

Similar behavior to Built-in types

Extensibility

16 MARK QUESTIONS

UNIT I

Describe the basic concepts of Object Oriented Programming and bring out the advantages of OOP.

Explain the declaration and defining a class in C++. How will you define the member functions of a class? Explain.

What is the need for parameterized constructors? Explain the function of constructors with their declaration and definition inside a class.

Illustrate the reserved word inline with two examples.

Explain the constructors and destructors.

Explain the relation between (i) Structured Programming (POP) and (ii)OOPs

Differentiate Object Oriented and Object based Languages.

Explain copy constructor? Explain with a suitable example.

Differentiate Object Oriented and Object based Languages. Give examples for both. List any EIGHT features of Oops. (Detailed explanation or examples are not required).

Explain the following terms with respect to OOPS .Give suitable examples

(i) Dynamic Binding (ii) Message Passing (iii) Reusability (iv) Polymorphism

UNIT II

What is operator overloading? How will you define it? Illustrate unary operator overloading with an example.

Describe the syntax of multiple inheritance. When do we use such an inheritance? Explain with an example.

Define friend class and specify its importance. Explain with suitable example.

Explain the operators used for dynamic memory allocation with examples.

Define functional overloading with example.

Explain in detail the various types of Inheritance with example programs

Explain virtual function in C++. Describe any two applications in which virtual functions may use. For each of these applications, specify the parent classes and derived classes.

What are inline functions? What are their advantages? Give an example .What is the rules to be followed while defining inline functions?

What is the need for and advantages of Templates? What is the difference between function

template and class template?

What are the various type conversions? Explain each with a program.

UNIT III

What is meant by exceptions? How an exception is handled in C++? Bring out the advantage

of using various exception handling mechanisms.

Explain the hierarchy of Stream classes in C++.

Give the hierarchy of console stream classes

Explain in detail about STRINGS in C++, with necessary examples

Explain in detail the various File handling Operations

What are the keywords used in C++ for exception handling? Describe their usage with suitable

example.

What are file modes? Describe various file mode options available in C++.

Explain the use of keywords try, catch and throw in handling exceptions in a program.

Indicate how the control flows in case of occurrence and non-occurrence of exceptions.

a program to implement a stack with appropriate exception handling.

Explain the 4 functions Seekg, Seekp, tellg, tellp used for setting pointers during file operation

and show how they are derived from f stream class.

Write a program to append to the contents of a file.

UNIT IV

1. How is object class created in java environment? Discuss on objects in java

Explain in detail about JAVA VIRTUAL MACHINE

Briefly explain about JAVA Byte-code

What is the purpose of using packages? How to create user-defined package? Give an example.

Explain in detail about java documentation.

Explain about arrays (matrix multiplication) and strings in java

What is meant by Overloading objects? How are related classes used in java?

Differentiate abstract classes and interfaces

Describe the structure of a typical Java program.

10.How can a subclass call a method or a constructor defined in a super class? Illustrate with an

example program.

UNIT V

Define interfaces in java .How interfaces are implemented? How they can be accessed? How to

apply interfaces? What is meant by extension of interfaces? Explain with an example.

What are threads? Synchronization? Explain the life cycle of thread with example

Explain life cycle of applet with example

Explain in detail exception handling in java

Explain in detail about JAVA I/O operations

Explain Inheritance in JAVA

Develop a real-life application program to illustrate the use of multithreads.

Explain with an example how multiple inheritances is achieved in Java.

How is synchronization of threads performed?

An educational institution wishes to maintain its employee's database which is divided into a

number of classes with minimum information as shown in figure. Specify all the classes and

define methods to create the database and retrieve individual information as when required.

EE6402-TRANSMISSION AND DISTRIBUTION

TWO MARKS WIT ANSWER

UNIT I

Why all transmission and distribution systems are 3 phase systems?

A 3 phase a.c circuit using the same size conductors as the single phase circuit can carry three times the power which can be carried by a 1 phase circuit and uses 3 conductors for the 2 phases and one conductor for the neutral. Thus a 3 phase circuit is more economical than a 1 phase circuit in terms of initial cost as well as the losses. Therefore all transmission and distribution systems are 3 phase systems.

Why the transmission systems are mostly overhead systems?

Because of the cost consideration, the transmission systems are mostly overhead systems.

Why all overhead lines use ACSR conductors?

ACSR conductors comprises of hard drawn aluminium wires stranded around a core of single or multiple strand galvanized steel wire. They provides the , necessary conductivity while the steel provides the necessary mechanical strength. Has less corona loss. The breaking load is high and has less weight.

Why transmission lines are 3 phase 3 wire circuits while distribution lines are 3 phase wire circuits?

A Balanced 3 phase circuit does not require the neutral conductor, as the instantaneous sum of the 3 line currents are zero. Therefore the transmission lines and feeders are 3 phase 3 wire circuits. The distributors are 3 phase 4 wire circuits because a neutral wire is necessary to supply the 1 phase loads of domestic and commercial consumers.

Why overhead line conductors are invariably stranded?

They are stranded to make them flexible during erection and while in service.

State the advantages of interconnected systems.

The area fed from one generating station during overload hours can be fed from another power station and thus reserved capacity required is reduced, reliability of supply is increased and efficiency is increased.

What is a ring distributor?

A ring distributor is a distributor which is arranged to form a closed circuit and is fed at one or more than one point.

State any two advantages of ring main system.

Less voltage fluctuations at consumer’s terminals.Less copper is required as each part of the ring carries less current than in radial system.

Mention the disadvantages of a 3 wire system

In 3 wire system a third wire is required .The safety is partially reduced .A balancer is required and therefore cost is increased.

What are the advantages of a 3 wire dc distribution system over a 2 wire dc distribution system?

If 3 wire system is used to transmit the same amount of power over the same distance with same efficiency with same consumer voltage we require 0.3125 times copper as required in 2 wire system.

State kelvin's law.

The annual expenditure on the variable part of the transmission system should be equal to the annual cost of energy wasted in the conductor used in that system.

State any two limitations of kelvin's law.

It is difficult to estimate accurately the annual charge on the capital outlay. It does not give the exact economical size of the conductor.

Define resistance of the transmission line.

It is defined as the loop resistance per unit length of the line in a single phase system. In 3 phase system it is defined as the resistance per phase.

What are the advantages of high voltage ac transmission.

- The power can be generated at high voltages.
- The maintenance of ac substation is easy and cheaper.

Mention the disadvantages of high voltage ac transmission.

- An ac line requires more copper than a dc line.
- The construction of an ac line is more complicated than a dc transmission line.
- Due to skin effect in the ac system the effective resistance of the line is increased.

Mention the limitations of using very high transmission voltage.

- The increased cost of insulating the conductor.
- The increased cost of transformers ,switch gears and other terminal apparatus.

Mention the terminal equipments necessary in HVDC system.

- Converters, mercury arc valves and thyristors.
- Due to absence of charging currents .

Mention the equipments that supply reactive power in HVDC converter stations ?

AC filters Static shunt capacitors Synchronous condensers StaticVAR compensators .

Why dc transmission is economical and preferable over ac transmission for large distances only ?

Because with larger distances ,the saving in cost of dc overhead lines become greater than the additional expenditure on terminal equipment.

20.What is meant by serving of a cable?

Layers of fibrous material permitted with waterproof compound applied to the exterior of the cable is called serving of a cable. pressure cables.

21. Mention the advantages of pvc over paper insulated cables.

Reduced cost and weight, Insulation is resistant to water, Simplified jointing, Increased flexibility No plumbing required.

22. State the merits of paper insulated cables.

High current carrying capacity ,long life and greater reliability

23. State the advantages of polythene insulators.

They are non-hygroscopic, light in weight, low dielectric constant, low loss factor and low thermal resistance.

By what materials cable sheaths are made? Lead sheaths and Aluminium sheaths.

In what way Al sheaths are superior to lead sheaths?

Al sheaths are smaller in weight, high mechanical strength , greater conducyivity, cheap,easy to manufacture and install, withstand the required gas pressure without reinforcement.

UNIT II

1.Define inductance of a line.

It is defined as the loop inductance per unit length of the line .Its unit is henrys per meter.

2.Define capacitance of a line.

It is defined as shunt capacitance between the two wires per unit line length. Its unit is farads per meter.

3.What is skin effect?

The steady current when flowing through the conductor ,does not distribute uniformly, rather it has the tendency to concentrate near the surface of the conductor. This phenomenon is called skin effect.

4.Why skin effect is absent in dc system?

The steady current when flowing through a conductor distributes itself uniformly over the whole cross section of the conductor .That is why skin effect is absent in dc system.

5.What is the effect of skin effect on the resistance of the line?

Due to skin effect the effective area of cross section of the conductor through which current flow is reduced. Hence the resistance of the line is increased when ac current is flowing.

7.Define symmetrical spacing.

In 3 phase system when the line conductors are equidistant from each other then it is called symmetrical spacing**.**

8. Define proximity effect.

The alternating magnetic flux in a conductor caused by the current flowing in a neighbouring conductor gives rise to a circulating current which cause an apparent increase in the resistance of the conductor .This phenomenon is called as proximity effect

9. What is the effect of proximity effect?

It results in the non uniform distribution of current in the cross section, and the increase of resistance.

10. What is a composite conductor?

A conductor which operates at high voltages and composed of 2 or more sub conductors and run electrically in parallel are called composite conductors**.**

11. What is a bundle conductor?

It is a conductor made up of 2 or more sub conductors and is used as one phase conductors.

12. Mention the advantages of using bundled conductors.

Reduced reactance, reduced voltage gradient , reduced corona loss .reduced Interference

13. What is meant by transposition of line conductors?

Transposition means changing the positions of the three phases on the line supports twice over the total length of the line .the line conductors in practice ,are so transposed that each of the three possible arrangements of conductors exit for one-third of the total length of the line .

14. Define voltage regulation.

Voltage regulation is defined as the change in voltage at the receiving (or load) end when the full-load is thrown off, the sending-end (or supply) voltage and supply frequency remaining unchanged.. % voltage regulation= ((Vs-Vr)/Vr)*100 where Vs is the voltage at the sending end Vr is the receiving end voltage.

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17. Define bundled conductors?

The use of more than one conductor per phase is called bundled conductors.

18. What is skin effect?

The phenomenon of concentration of an ac current near the surface of the conductor is known as skin effect.

19. On what factors does the skin effect depends?

The skin effect depends upon the 1, type of the material 2, frequency of the current 3, diameter of conductor& shape of conductor. It increases with the increase of cross-section, permeability and supply frequency.

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UNIT III

1.What is corona?

The phenomenon of violet glow, hissing noise and production o ozone gas in an overhead line is called corona.

2.State any two merits and demerits of corona.

MERITS

Reduces the effects of transients produced by surges .
System performance is improved.

DEMERITS

The transmission efficiency is affected. Corrosion occurs.

3.**Why ACSR conductors are used in lines?**

If the size of the conductor is larger corona effects are reduced and reduces the proximity effect .Hence they are used in lines.

4.Define medium lines.

Lines having length between 60 and 150 km and line voltages between 20 and 100kv are called medium lines.

5. Mention the limitations of end condenser method.

This over estimates the effects of line capacitance .It is assumed to be lumped or concentrated.

6. Explain the term voltage stability.

The ability of the system to maintain the voltage level within its acceptable limits is called as voltage stability.

7. Differentiate between voltage stability and rotor angle stability.

Voltage stability:

- It means load stability.
- It is mainly related to reactive power transfer.
- Here problems arise mainly in the event of faults.

Rotor angle stability:

- It means basically generator stability.
- It is mainly interlinked to real power transfer.
- Here problems arise during and after faults.

8. Mention the significance of Surge impedance loading.

- The voltage and current are equal and are in phase at all points along the line.
  No reactive power is generated or absorbed at the line ends.

9. What is shunt compensation ?

Shunt compensation is the use of shunt capacitors and shunt reactors is the line to avoid voltage instability.

10. Define a synchronous compensator (condenser)?

Synchronous compensator is a synchronous motor with no mechanical output .When it is under exited it operates at lagging p.f (ie it delivers vars) .Thus it operates both as a shunt capacitor and as a shunt reactor .

11. Why series compensation is used in long series ?

To increase transmission capacity -to improve system stability . -to obtain correct load division between parallel circuits.

12. What is end condenser method?

It is a method used for obtaining the performance calculations of medium lines. Here the capacitance of the line is lumped or concentrated at the receiving end.

13. What is power circle diagram?

It is a diagram drawm for the transmission lines network involving the generalized circuit constants and the sending end and receiving end voltage.

14. What are the voltage regulating equipments used in transmission system?

Synchronous motors, tap changing transformers , series and shunt capacitors booster transformers , compound generators and induction regulator.

15. Mention the methods used for voltage control of lines

Tap changing auto- transformer, booster transformer , excitation control and induction regulator.

16. What is sending end power circle diagram?

The circle drawn with sending end true and reactive power as the horizontal and vertical co-ordinates are called sending end power circle diagram.

17. What is receiving end power circle diagram?

The circle drawn with receiving end values are called receiving end power circle diagram.

18. Mention any two advantages of SVS .(Static Var System)

Provides fast control over temporary over voltages. Provides a better control of voltage profile .

19**. State any two comparisons between series compensation and shunt compensation. -**

Series compensation is cheaper than SVS . -Losses are lower than in SVS .

Why series compensation is used in long series ?

-to increase transmission capacity -to improve system stability . -to obtain correct load division between parallel circuits .

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UNIT IV

1.Why cables are not used for long distance transmission?

Cables are not used for long distance transmissions due to their large charging currents.

2.What is the purpose of insulation in a cable?

The insulation or dielectric withstands the service voltage and isolates the conductor with other objects.

5.What is the function of sheath in a cables?

The sheath does not allow the moisture to enter and protects the cable from all external influences like chemical or electrochemical attack fire etc.

4.Define the segmental conductors.

The stranded wires which are compacted by the rollers to minimize the air spaces between the individual wires are called segmented conductors .Here the conductor size is reduced for a given conductance.

5.State the properties of insulating materials.

It should have high insulation resistance ,high dielectric strength ,good mechanical properties ,non-hygroscopic, capable of being operated at high temperatures ,low thermal resistance and low power factor.

Mention the commonly used power cables.

Impregnated paper, Polyvinyl chloride, polyethene

Mention the advantages of pvc over paper insulated cables.

Reduced cost and weight, Insulation is resistant to water, Simplified jointing, Increased flexibility No plumbing required.

8. State the merits of paper insulated cables.

High current carrying capacity ,long life and greater reliability

9. State the advantages of polythene insulators.

They are non-hygroscopic, light in weight, low dielectric constant, low loss factor and low thermal resistance.

10. By what materials cable sheaths are made?

Lead sheaths and Aluminium sheaths.

11. In what way Al sheaths are superior to lead sheaths?

Al sheaths are smaller in weight, high mechanical strength , greater conducyivity, cheap,easy to manufacture and install, withstand the required gas pressure without reinforcement.

12. Where CSA sheath is used in cables ?

Corrugated seamless aluminium sheath is used in high voltage oil filled cables and telephone lines. It is used because it is very flexible and easily by repeated bending the sheath is not distorted and it is not damaged. It has lesser weight and reduced thickness.

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Corrugated seamless aluminium sheath is used in high voltage oil filled cables and telephone lines.

17. Why it is used?

It is used because it is very flexible and easily by repeated bending the sheath is not distorted and it is not damaged. It has lesser weight and reduced thickness.

18. Why protective covering is done in cables?

To protect the cables from mechanical damage , corrosion and electrolytic action when laid direct in the ground the protective covering is made.

19. By what material protective covering is made?

Bitumen &Bituminized materials, pvc and layers of fibrous materials.

20. What is meant by serving of a cable?

Layers of fibrous material permitted with waterproof compound applied to the exterior of the cable is called serving of a cable. pressure cables.

21.Why cables are not used for long distance transmission?

Cables are not used for long distance transmissions due to their large charging currents.

22.Mention the 3 main parts of the cable?

Conductor ,dielectric ,sheath

23.What is the function of conductor?

Conductor provides the conducting path for the current.

24.What is the purpose of insulation in a cable?

The insulation or dielectric withstands the service voltage and isolates the conductor with other objects.

25.What is the function of sheath in a cables?

The sheath does not allow the moisture to enter and protects the cable from all external influences like chemical or electrochemical attack fire etc.

UNIT V

1. Define sag of a line.

The difference in level between the points of supports and the lowest point of the conductor is called as sag.

3. What is the reason for the sag in the transmission line?

While erecting the line , if the conductors are stretched too much between supports then there prevails an excessive tension on the line which may break the conductor. In order to have safe tension in the conductor a sag in the line is allowed.

4. How the capacitance effect is taken into account in a long line?

They have sufficient length and operate at voltage higher than 100 kv the effects of capacitance cannot be neglected. Therefore in order to obtain reasonable accuracy in long lines , the capacitance effects are taken.

5. what is neutral grounding.

Connecting the neutral or star point of any electrical equipment(generator ,transformer etc) to earth.

6. define coefficient of earthing.

(highest rms voltage of healthy line to earth)/(line to line rms voltage) *100 to the power frequency

7. mention 2 disadvantages of ungrounded neutral

- Occurance of insulation breakdown leading to the heavy phase to phase fault condition.
- system cannot be protected from earth fault.
- voltages due to lightning surges do not find path to earth.

8**. Name the various types of grounding**.

- solid grounding
- resistance grounding
- reactance grounding
- resonant grounding

9. define screening coefficient.

Screening coefficient for ‘n’ electrodes in parallel is = (resistance of one electrode)/(resistance of n electrodes in parallel * n)

10.what is a substation.

The assembly of apparatus used to change some characteristic (eg: voltage , A.C toD.C frequency power factor etc) of electric supply is called a substation.-frequency changer substation-converting substationS-industrial substation.

11. What is sending end power circle diagram?

The circle drawn with sending end true and reactive power as the horizontal and vertical co-ordinates are called sending end power circle diagram.

12. What is receiving end power circle diagram?

The circle drawn with receiving end values are called receiving end power circle diagram.

13. what is neutral grounding.

Connecting the neutral or star point of any electrical equipment(generator ,transformer etc) to earth.

14. define coefficient of earthing.

=(highest rms voltage of healthy line to earth)/(line to line rms voltage) *100 to the power frequency

15. mention 2 disadvantages of ungrounded neutral

occurance of insulation breakdown leading to the heavy phase to phase fault condition.
-system cannot be protected from earth fault.
-voltages due to lightning surges do not find path to earth.

16**. Name the various types of grounding**.
-solid grounding -resistance grounding -reactance grounding -resonant grounding

17. give the response of resistance for earth driven rods.

R=□/2 □l *ln(4l/d) Where l – length of the rod d – diameter of the rod resistivity of the rod

for the uniformly current carrying ground driven rod , give the resistance value.

$R=\rho/(2\pi l) * (\ln(8l/d)-1)$ Where ρ - resistivity l – length d – diameter

19. define screening coefficient.

Screening coefficient for ‘n’ electrodes in parallel is = (resistance of one electrode)/(resistance of n electrodes in parallel * n)

20.what is a substation.

The assembly of apparatus used to change some characteristic (eg: voltage , A.C toD.C frequency power factor etc) of electric supply is called a substation.-frequency changer substation-converting substationS-industrial substation

21. Define sag of a line.

The difference in level between the points of supports and the lowest point of the conductor is called as sag.

22.Mention the factors that affect sag in the transmission line.

Weight of the conductor, length of the span , working tensile strength and the temperature.

23. What is the reason for the sag in the transmission line?

While erecting the line , if the conductors are stretched too much between supports then there prevails an excessive tension on the line which may break the conductor. In order to have safe tension in the conductor a sag in the line is allowed.

24. How the capacitance effect is taken into account in a long line?

They have sufficient length and operate at voltage higher than 100 kv the effects of capacitance cannot be neglected. Therefore in order to obtain reasonable accuracy in long lines , the capacitance effects are taken.

25. Mention the limitations of nominal T and pi methods in the line problems.

Generally the capacitance is uniformly distributed over the entire length of the line. But for easy calculations the capacitance is concentrated at one or two points .Due to these effects there are error in the calculations.

EE6402 TRANSMISSION AND DISTRIBUTION
PART-B, QUESTIONS

UNIT-I

(i) Discuss various types of HVDC links.
List out the main components of a HVDC system.
(i) Draw and explain the structure of modern power systems with typical voltage levels.
What is the highest voltage level available in India?

(i) Explain the effect of high voltage on volume of copper and on efficiency.

Explain why the transmission lines are 3 phase 3-wire circuits while distribution lines are 3 phase 4-wire circuits.

(i) Draw the model power system with single line representation. Show its essential constituent sections.

What are the AC transmission and distribution level voltages we have in India?

What are the different kinds of DC links? Draw relevant diagrams.

(i) Explain why EHV transmission is preferred? What are the problems. involved in EHV AC transmission?

With neat schematic, explain the principle of HVDC system operation.

Explain about FACTS with neat diagram **.**

Explain TCSC and SVS systems **.**

Explain with neat diagram about STATCOM and UPFC.

(i) Compare EHVAC and HVDC transmission **.**

Explain the applications of HVDC transmission system.

10.(i)Write short notes on distributed and concentrated loads?

(ii)What are distributors?explain its types in detail**.**

UNIT II

(i)From the fundamentals derive an expression for inductance of a single phase transmission system.

Write short notes on corona discharges.

Derive an expression for capacitances of a single phase transmission system and discuss the effect of earth on capacitance with suitable equation.

Derive an expression for inductance

Of a single-phase overhead line.

A conductor is composed of seven identical copper strands each having a radius r. Find the self-GMD of the conductor**.**

i) Derive an expression for the capacitance between conductors of a Single phase overhead line.

Find the capacitance between the conductors of a single-phase 10 km long line. The diameter of each conductor is 1.213cm. The spacing between conductors is 1.25m. Also find the capacitance of each conductor neutral.

i) Derive the expression for inductance of a two wire 1Φ transmission line

Derive the expression for capacitance of a 1Φ transmission line

i) What are the advantages of bundled conductors?

Derive the expression for capacitance of a double circuit line for hexagonal spacing.

Why is the concept of self GMD is not applicable for capacitance?

i) Explain clearly the skin effect and the proximity effects when referred to overhead lines.

Write a short note on the inductive interference between power and communication lines.

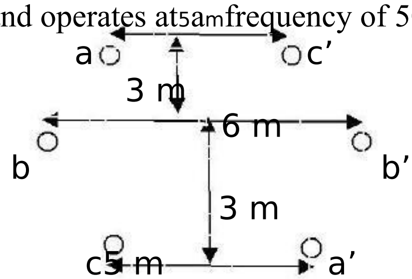
i) Derive the expression for the capacitance per phase of the 3 Φ double circuit line flat vertical spacing with transposition.

A 3 Φ overhead transmission line has its conductors arranged at the corners of an equilateral triangle of 2m side. Calculate the capacitance of each line conductor per km. Given the diameter of each conductor is 1.25cm.

i)Find the capacitance per km per phase of a 3Φ line arrangement in a horizontal plane spaced 8 metres apart. The height of all conductors above the earth is 13 metres. The diameter of each conductor is 2.6 cm. the line is completely transposed and takes the effect of ground into account.

ii). Discuss the concept of GMR and GMD in the calculation of transmission line inductance.

Find the inductance /phase /km of doublecircuit 3phase line shown in fig. the line is completely Transposed and operates at frequency of 50Hz. Radius r = 6mm



UNIT III

Determine the efficiency and regulation of a 3phase, 100Km, 50 Hz transmission line delivering 20 MW at a power factor of 0.8 lagging and 66 kV to a balanced load. The conductors are of copper, each having resistance 1 Ω / Km, 1.5 cm outside dia, spaced equilaterally 2 metres between centres. Use nominal T method.

A three phase 5 km long transmission line, having resistance of 0.5 Ω / km and inductance of 76mH/km is delivering power at 0.8 pf lagging. The receiving end voltage is 32kV. If the supply end voltage is 33 kV, 50 Hz, find line current, regulationand efficiency of the transmission line.

Derive the expressions for sending end voltage in nominal T method and end Condenser method.

What is an equivalent circuit of long line? Derive expression for parameters of this circuit in terms of line parameters.

i) Define regulation of a transmission line and derive the approximate expression for the regulation of a short transmission line.

What is corona loss? How do you determine this loss?

A 220kV, 3Φ transmission line has an impedance per phase of (40+j200)Ω and an admittance of (0+j0.0015) mho. Determine the sending end voltage and sending end current when the receiving end current is 200 A at 0.95 pf lagging. Use nominal method.

Determine the efficiency and regulation of a three phase 200 km, 50Hz transmission line delivering 100MW at a pf of 0.8 lagging and 33kV to a balanced load. The conductors are of copper, each having resistance 0.1 Ω/km, and 1.5cm outside dia, spaced equilaterally 2m between centres. Neglect leakage reactance and use nominal T and π methods.

i) Explain the Ferranti effect with a phasor diagram and its causes.

13. A 50Hz transmission line 300 km long total series impedance of 40+j25 Ω and total shunt admittance of 10-3 mho. The 220 Kv with 0.8 lagging power factor. Find the sending end votage, current, power and power factor using nominal pi method.

14 .i) Explain the classification of lines based on their length of transmission. ii) What are ABCD constants.

UNIT IV

Discuss any two methods to increase the value of string efficiency, with suitable sketches.
Explain any two methods of grading of cables with necessary diagrams.
i) What are different methods to improve string efficiency of an insulator?
In a 3-unit insulator, the joint to tower capacitance is 20% of the capacitance of each unit. By how much should the capacitance of the lowest unit be increased to get a string efficiency of 90%. The remaining two units are left unchanged.
i) Derive the expression for insulator resistance, capacitance and electric stress in a single core cable.Where is the stress maximum and minimum?
A single core 66kv cable working on 3-phase system has a conductor diameter of 2cm and sheath of inside diameter 5.3cm. If two inner sheaths are introduced in such a way that the stress varies between the same maximum and minimum in the three layers find:

- position of inner sheaths
- voltage on the linear sheaths
- maximum and minimum stress

i) Draw the schematic diagram of a pin type insulator and explain its function.
A 3 phase overhead transmission line is being supported by three disc insulators. The potential across top unit (i.e. near the tower) and the middle unit are 8kV and 11kV respectively. Calculate,

The ratio of capacitance between pin and earth to the self capacitance of each unit
Line Voltage
String Efficiency

i) Describe with the neat sketch, the construction of a 3 core belted type cable.

A conductor of 1cm diameter passes centrally through porcelain cylinder of internal diameter 2 cms and external diameter 7cms. The cylinder is surrounded by a tightly fitting metal sheath. The permittivity of porcelain is 5 and the peak voltage gradient in air must not exceed 34kV/cm. Determine the maximum safe working voltage.

i) What are the various properties of insulators? Also briefly explain about suspension type insulators.

Calculate the most economical diameter of a single core cable to be used on 132kV, 3 phase system. Find also the overall diameter of the insulation, if the peak permissible stress does not exceed 60kV/cm. also derive the formula used here.

i) Briefly explain about various types of cables used in underground system.**(8)**

A string of 4 insulator units has a self capacitance equal to 4 times the pin to earth capacitance. Calculate,

- Voltage distribution as a % of total voltage
- String efficiency

i) Give any six properties of a good insulator.

With a neat diagram, explain the strain and stay insulators.

A cable is graded with three dielectrics of permittivities 4, 3 and 2. The maximum permissible potential gradient for all dielectrics is same and equal to 30 kV/cm. The core diameter is 1.5cm and sheath diameter is 5.5cm

i) Explain the constructional features of one LT and HT cable

- Compare and contrast overhead lines and underground cables.

UNIT V

Explain the following:

Neutral grounding

Resistance grounding. 2.Write short notes on AIS. 3.Write short notes on GIS.

Explain various methods of grounding.

An overhead line has a span of 336 m. The line is supported ,at a water Crossing from two towers whose heights are 33.6 m and 29 m above water level. The weight of conductor is 8.33 N/m and tension in the conductor is not to exceed 3.34 × 104 N. Find (i) Clearance between the lowest point on the conductor and water (ii) horizontal distance of this point from the lower support.

a)Derive expressions for sag and tension in a power conductor strung between to supports at equal heights taking into account the wind and ice loading also.

b)An overhead line has a span of 300m. The conductor diameter is 1.953 cm and the conductor weight is 0.844 kg/m. calculate the vertical sag when a wind pressure is 736 N/sq.m of projected area acts on conductor. The breaking strength of conductor is 77990 N and the conductor should not exceed half the breaking strength**.**

A transmission line conductor at a river crossing is supported from two towers at a height of and 80 m above water level. The horizontal distance between the towers is 300 m. if the tension in the conductor is 2000 kg find the clearance between the conductor and water at a point midway between the towers. Weight of conductor/m = 0.844 kg. Derive the formula used.

Derive the expressions for sag and conductor length under bad weather conditions.
Assume Shape of overhead line is a parabola.

Write short notes on

- Explain the design principles of substation grounding system
- Grounding grids

10. With the neat layout explain the design of modern substation with all protecting devices.

UNIT I

INTRODUCTION

What is an LTI system?

A linear time invariant (LTI) system follows two principles: Superposition and Time invariance.

For a linear system, the response due to linear combination of inputs is same as linear combination of corresponding outputs.

Time invariance means shift of time origin of input does not change the response of the system.

A system is characterized by ( ) = ( ) + (−). Is the system linear?

$_1( ) = {}_1( ) + \quad {}_1(-1)$

$_2( ) = {}_2( ) + \quad {}_2(-1)$

Linear combination of two outputs becomes,

$_3( ) = {}_1{}_1( ) + {}_2{}_2( ) = \quad {}_1[{}_1( ) + \quad {}_1(-1)] + {}_2[{}_2( ) + \quad {}_2(-1)]$

Response to linear combination of two inputs will be, $'_3( ) = \{ {}_1{}_1( ) + {}_2{}_2( ) \}$

$_1{}_1( ) + {}_2{}_2( ) + [{}_1{}_1(-1) + {}_2{}_2(-1)]$

$_1{}_1( ) + {}_2{}_2( ) + {}_1{}_1(-1) + {}_2{}_2(-1)$

$_1[{}_1( ) + {}_1(-1)] + {}_2[{}_2( ) + {}_2(-1)]$

Thus $_3( ) = {}'_3( )$, hence this is linear system.

Define impulse response of the system.

When a unit sample δ(n) is applied to the input of the system. The output of the system is called impulse response. It is denoted by h(n). i.e.,

$$h( ) = \{ ( )\}$$

and output of the system to any arbitrary input x(n) is given as,

$$( ) = ( ) * h( )$$

Thus, the convolution of input and impulse response gives output. Impulse response is the characteristics of the system.

Give the formula for discrete convolution.

Discrete convolution is given as,

$$( ) = \sum_{=-\infty}^{=\infty} ( )\, h(-) = \sum_{=-\infty}^{=\infty} h( )\, (-)$$

It is denoted as,

$$( ) = ( ) * h( )$$

Convolution is commutative and associative.

Define system transfer function.

System function is the ratio of Z-transform of the output to Z-transform of the input.

i.e., System function

$$( ) = \frac{( )}{( )} = \frac{ + {}^{-1} + {}^{-2} + \cdots + {}^{-}}{1 + {}^{-1} + {}^{-2} + \cdots + {}^{-}}$$

This is the transfer function of the Nth order LTI system.

What is causal system?

A system is said to be causal if its output depends upon the present input, past inputs or past outputs only. Response of such system begins only when input is applied. For example,

$y(n) = x(n) + x(n-1) - y(n-1)$ is a causal system.

$y(n) = x(n) + x(n+1)$ is a non causal system.

Write the condition for system stability.

A system is said to be BIBO stable, it produces bounded output for bounded input. The stable system produces bounded output for every bounded input. For an LTI system stability is given in terms of impulse response as,

$$\sum_{=-\infty}^{=\infty} h(\ ) < \infty$$

The impulse response must be summable.

What is a shift invariant system? Give an example.

If the input is delayed, then output of the shift invariant system is also delayed by same amount. In other words, input/output relationship of shift invariant system is not affected by changing the time origin of input. For example,

$(\ ) = (\ ) + (\ +1)$

Response of the system to delayed input will be,

$(\ ,\ ) = \{(\ -\ )\} = (\ -\ ) + (\ +1-\ )$

And let us delay itself by same number of samples.

$(\ -\ ) = (\ -\ ) + (\ -\ -1)$

Thus

$(\ ,\ ) = (\ -\ )$. This is shift invariant system.

State the condition for causality and stability of LTI system in Z- domain.

A system is said to be causal if ROC of its system function is exterior of some circle of radius ‘r’ i.e.,

$|\ | >$

The LTI system is stable if ROC of its system function includes the unit circle. i.e.,

< 1

Thus combining the two condition, if

$|\ | >\ \ < 1$, then the system will be causal and stable.

Define a static and a stable system.

Static system: when output at any moment depends on input at that moment only is called static system. Otherwise the system is said to be dynamic.

Stable system: if the system produces bounded output for bounded input, then it is called stable system.

Define commutative and associative law of convolution.

Commutative law: $(\ ) * h(\ ) = h(\ ) * (\ )$

Associative law: $[(\ ) * h_1(\ )] * h_2(\ ) = (\ ) * [h_1(\ ) * h_2(\ )]$

Define sampling theorem.

A continuous time signal can be completely represented in it samples and recovered back if the sampling frequency $F_s \geq 2B$. Here F_s is the sampling frequency and B is the maximum frequency present in the signal.

What is the causality condition for an LTI system?

The LTI system is causal if, h(n) = 0 for n<0

Define linear convolution of two DT signals. The linear convolution of x(n) and h(n) is given as,

$$y(n) = \sum_{k=-\infty}^{k=\infty} x(k)\, h(n-k) = \sum_{k=-\infty}^{k=\infty} h(k)\, x(n-k)$$

Define system function and stability of a DT system.

Stability: The DT system is said to be stable if its impulse response is completely summable.

$$\sum_{=-\infty}^{\infty} |h(\ )| < \infty$$

System function: The ratio of Z-transform of the output to Z-transform of the input is called system function.

$$(\ ) = \frac{(\ )}{(\ )}$$

The transform of impulse response is system function i.e.,

$$h(\ ) \leftrightarrow (\ )$$

Write down the expression for discrete time unit impulse and unit step functions.

DT unit impulse

$$(\ ) = \begin{cases} 1, & = 0 \\ 0, & \neq 0 \end{cases}$$

DT unit step

$$(\ ) = \begin{cases} 1, & \geq 0 \\ 0, & < 0 \end{cases}$$

17. Differentiate between recursive and non recursive difference equations.

S No	Non recursive	Recursive
1.	$(\ ) = \sum_{=0} (\ -\ )$	$(\ ) = -\sum_{=1} (\ -\ ) + \sum_{=0} (\ -\ )$
2.	There is no feedback from output.	There is feedback from output.

18. **Find the fundamental period N_o for** $(\ ) = \ ^{-}(\ \ )$.

$$(\ ) = \ ^{-(}\ \frac{\ }{3}\ )$$

$$= \cos\left(\frac{2}{3}\right) - \left(\frac{2}{3}\right)$$

$$= \cos(2\ \ ) - (2\ \ )$$

Here $= \frac{1}{3} = $ Hence N = N_o = 3 samples.

19. Determine the poles and zeros of ( ), ( ) =

$$(\ ) = \frac{1 - 2^{-1} + 2^{-2}}{(\ -1-\ )(\ -1+\ )} \qquad ^{2} - 2\ + 2$$

Zeros : $z_1 = 0$

Poles : $p_1 = 1+j$, $p_2 = 1-j$

Check for linearity and stability of ( ), ( ) = √ ( ).

Since square root is non linear, the system is non linear.

As long as ( )is bounded, its square root is bounded. Hence this system is stable.

Define shift variant system.

If the input/output characteristics of the system do not change with shift of time origin, such systems are called shift invariant or time invariant systems.

y(n) = x(n) − x(n − 1) is shift invariant
y(n) = n x(n) is shift variant

Define causality.

If the output of the system depends upon the past and present inputs only, then it is called causal system. If output of the system depends upon future inputs, then it is called non-causal system.

y(n) = x(n) − x(n − 1) is causal
y(n) = x(n) − x(n + 1) is non − causal

What is mean by aliasing? How it can be avoided?

Aliasing: When the sampling frequency is less than twice of the highest frequency content of the signal, then aliasing in frequency domain takes place. In aliasing, the high frequencies of the signal mix with lower frequencies and create distortion in frequency spectrum.

To avoid aliasing: Aliasing can be avoided by two ways,

Sampling frequency must be higher than twice of highest frequency present in the signal.

A low pass filter must be used before sampling to bandlimit the signal to some specific frequency.

Check whether the system ( ) = $^{( )}$ is linear.

Since the exponential function is non linear, the system is non linear.

List any two properties of LTI system.

Causality: the LTI system is said to be causal if and only if h( ) = 0 < 0

Stability: the LTI system is stable if its impulse response is absolutely integrable,

$$\sum_{=-\infty}^{\infty} |h(\ )| < \infty$$

Derive the necessary and sufficient condition for an LTI system to be BIBO stable.

Thus bounded input x(n) produces bounded output y(n) in LTI system only if,

$\sum_{=-}^{\infty} |h(\ )| <$ ∞ :For system to be stable.

When this condition is satisfied, the system will be stable. The above condition states that the LTI system is stable if its unit sample resonse is absolutely summable. This is the necessary and sufficient condition for the stability of LTI system.

What are the advantages of DSP?

Flexibility, Accuracy, Easy storage, Mathematical processing, Cost, Repeatability, Adaptability, Universal compatibility, Size and reliability

Consider the analog signal, ( ) =++ , what is the Nyquist rate for the signal?

( ) = 3 cos2000 + 5 sin 6000 + 10cos 12000

3 cos2 1000 + 5 sin 2 3000 + 10 cos2 6000

Above signal contains F_1= 1000Hz, F_2= 3000Hz and F_3= 6000Hz

Here F_{max}= F_3= 6000Hz

Therefore the Nyquist rate = $2F_{max}$= 2 × 6000 = 12000Hz

29. Calculate the minimum sampling frequency required for ( ) = . + so as to avoid aliasing.

( ) = 0.5 sin 50 + 0.25sin 25

= 0.5 sin 2 25 + 0.25sin 2 12.5

Above signal contains F_1= 25Hz and F_2= 12.5Hz. Hence F_{max}= F_1= 25Hz.

Therefore the minimum sampling frequency = $2F_{max}$= 2 × 25Hz = 50Hz.

State any two properties of Auto correlation function.

Autocorrelation is a even function

( ) = (−)

Autocorrelation attains maximum value at zero lag.

| ( )| ≤ (0) =

What is meant by energy and power signals?

Energy signal: The signal is said to be an energy signal if its energy is finite and non zero. i.e.For energy signal, 0<E<∞

Power signal: The signal is said to be power signal if its power is finite and non zero. i.e. For power signal, 0<P<∞

UNIT II
DISCRETE TIME SYSTEM ANALYSIS

What is Z-Transform of ( ) and (−) in terms of X(z).

By differentiation in z-domain property,

) ↔ − ( )

By time shifting property,

(−) ↔ ⁻ ( )

2. Represent the condition satisfied by a stable LTI DT system in the z-domain. What is equivalent condition in time domain?

Z-Domain: LTI system is BIBO stable if and only if the ROC of the system function includes the unit circle.

Time domain: LTI system is BIBO stable if its impulse response is absolutely summable i.e.

$$\sum_{n=-\infty}^{\infty} |h(n)| < \infty$$

3. Determine the ROC of the z-Transform of the sequence ( ) = (− −).

We know that, − (− − 1) ↔ $\frac{1}{1-\ ^{-1}}$ ROC : | | <

Here a= -2, −2 (− − 1) ↔ $\frac{1}{\ }$ ROC : | | < |2|

4. Define transfer function.

The transfer function is the Z-Transform of unit sample response of LTI system. It is given as,

( ) = {h( )}

( )

5. **Define ROC and explain its properties.**

The range of variation of z for which z-transform converges is called region of convergence Of z-transform.

Properties of ROC of Z-Transforms

- ROC of z-transform is indicated with circle in z-plane.
- ROC does not contain any poles.
- If x(n) is a finite duration causal sequence or right sided sequence, then the ROC is entire z-plane except at z = 0.
- If x(n) is a finite duration anti-causal sequence or left sided sequence, then the ROC is entire z-plane except at z = ∞.
- If x(n) is a infinite duration causal sequence, ROC is exterior of the circle with radius a. i.e. |z| > a.
- If x(n) is a infinite duration anti-causal sequence, ROC is interior of the circle with radius a. i.e. |z| < a.

If x(n) is a finite duration two sided sequence, then the ROC is entire z-plane except at z = 0 & z = ∞.

What are the types of Z transform?

1.Unilateral or one sided z-tranform: It is defined as,

X(z)=∑∞ =0 ( )

2.Bilateral:

X(z)=∑∞ =−∞ ( )

7. **Define Z-Transform and its ROC.**

( ) = ∑ ( )

ROC: It is the region where Z-Transform is convergent.

Determine the z-transform and ROC for the signal x(n)=δ(n+k) +δ(n-k).

We know that Z[δ(n)]=1, ROC:entire z- plane

Z[δ(n-k)]=z^{-k}, ROC:entire z-plane except z = 0

Z[δ(n+k)]=z^{k}, ROC:entire z-plane except z = ∞

X(z)=z^{-k} + z^{k} ROC:entire z-plane except z = 0 and z = ∞

9. **Mention the relation between, Z Transform and Fourier transform.**

Fourier series is basically the Z-Transform of the sequence evaluated on unit circle.

( ) = ( ) =

Give any two properties of linear convolution.

Linear convolution is commutative, i.e.,

( ) ∗ h( ) = h( ) ∗ ( )

ii. Linear convolution is distributive, i.e.,

( )[h₁( ) + h₂( )] = ( ) ∗ h₁( ) + ( ) ∗ h₂( )

What are the basic operations involved in convolution process?

The convolution process involves four basic operations, i.Folding i.e. $h(-\ )$
ii.Shifting i.e. $h(-\ )$

iii.Multiplication i.e. $(\ )h(-\ )$ iv.Summation. i.e. $\sum_{=-\infty}^{=+\infty}(\ )h(-\ )$

What is the resultant impulse response of the two system whose impulse response are h1(n) and h2(n) when they are in a) Series b) Parallel.

Series connection: $h(\ ) = h_1(\ ) * h_2(\ )$
Parallel connection: $h(\ ) = h_1(\ ) + h_2(\ )$

13. State the initial and final value theorem of Z-Transform.

Initial value theorem : $(0) = \lim (\ )$

Final value theorem : $(\infty) = \lim(\ -1)(\ )$

What are the different methods of evaluating inverse z-transform?

- Partial fraction expansion
- Power series expansion

Contour integration (Residue method)

Find the convolution for $(\ ) = \{\ ,\ ,\ ,\ \}$ and $(\ ) = \{\ ,\ ,\ \}$.

$x(n) =$		0	1	0	2	
$h(n) =$			2	0	1	
		0	1	0	2	
	0	0	0	0	x	
0	2	0	4	x	x	
0	2	0	5	0	2	

$(\ ) * (\ ) = \{\ \}$

16. How will you perform linear convolution using circular convolution?

Let the length of the x(n) be L, length of h(n) be M, then linear convolution of x(n) and h(n) can be obtained through following steps:

Append x(n) with M-1 zeros. Hence its length will be L+M-1.

Append h(n) with L-1 zeros. Hence its length will be L+M-1.

Perform circular convolution of above sequences. The result is linear convolution of length L+M-1.

Define discrete time Fourier transform pair for a discrete sequence.

DTFT pair is defined as,

$X(\omega) = \sum_{=-\infty}^{\infty}(\ )^{-}$

$x(n) = \frac{1}{2}\int_{-}(\ )\, d\omega$

18. Determine DTFT of a sequence x(n)= an u(n). Discrete Time Fourier Transform (DTFT) is given as,

$$X(\Omega) = \sum_{n=-\infty}^{\infty} x(n)e^{-j\Omega n} = \sum_{n=0}^{\infty} a^n e^{-j\Omega n} \qquad \text{since } x(n) = a^n u(n)$$

$$= \sum_{n=0}^{\infty}(ae^{-j\Omega})^n = \frac{1}{1 - ae^{-j\Omega}}$$

Given a difference equation y[n]=x[n]+3x[n-1]+2y[n-2]. Evaluate the system function H(z).

$$Y[z]=X[z]+3z^{-1}X[z]+2z^{-2}Y[z]$$

$$Y[z]=X\cancel{[z]}[1+3z^{-1}+2z^{-2}]$$

$[\ ] = {}_{[\ ]}^{[\ ]} = 1+3z^{-1}+2z^{-2}$

20. What is the use of Fourier transform?

Fourier transform converts time domain signal to frequency domain. Fourier transform is useful to study the frequency domain nature of periodic as well as non periodic signals. Most of the spectrum and frequency related aspects of the signals are studied with help of Fourier transform.

UNIT III

DISCRETE FOURIER TRANSFORM AND COMPUTATION

State and prove Parseval's relation for DFT.

Parseval's relation for DFT is given as,

$$\sum_{=0}^{-1} |\ (\ )|^2 = \quad 1 \quad \sum_{=0}^{-1} |\ (\ )|^2.$$

Define the properties of convolution.

Convolution is commutative

$$(\ ) * h(\ ) = h(\ ) * (\ )$$

Convolution is associative

$$[\ (\ ) * h_1(\ )] * h_2(\ ) = (\ ) * [h_1(\ ) * h_2(\ )]$$

Convolution is distributive

$$(\ ) * [h_1(\ ) + h_2(\ )] = (\ ) * h_1(\ ) + (\ ) * h_2(\ )$$

3. **What is the relationship between Z-transform and DFT?**

If the Z-transform is evaluated on unit circle at evenly spaced points only, then it becomes DFT. i.e.,

$$X(k)= X(Z)_{z=e}{}^{j2\pi k/N}$$

4. **Distinguish between discrete time Fourier transform and discrete Fourier transform.**

S. No	Discrete Time Fourier Transform	Discrete Fourier Transform
1	Time domain sequence is discrete but frequency domain representation is continuous.	Both time domain sequence and frequency domain representations are discrete.
2	DTFT cannot be evaluated using fast algorithms.	DFT can be evaluated using fast algorithms.
3	DTFT is continuous version of DFT	DFT is discrete version of DTFT.

5. **What is zero padding? What is the purpose of it?**

For the calculation of DFT, the resolution can be increased by increasing 'N'. This means length of the sequence will be less than 'N' but more points of DFT can be taken.

Hence the length of the sequence is increased by putting extra data points as 'zeros' at the end of the sequence. These 'zeros' does not change the meaning of the sequence.Also in the calculation of FFT, the length 'N' should be equal to 2^k, where k is some integer. Hence such length of data sequence can be adjusted by putting 'zeros' at the end. It is called zero padding.

When the length of the sequence is to be increased, zeros are inserted as samples. This does not change meaning. For example,

$x(n) = \{1,1,1\}$ $\qquad$ $N = 3$

$x(n) = \{1,1,1,0,0,0,0,0\}$ $\qquad$ $N = 8$

thus the length of the sequence is increased from 3 to 8.

Zero padding is used in – Calculation of DFT and FFT,Linear convolution.Linear convolution using circular convolution.

6. What is the difference between circular convolution and linear convolution?

S.No	Parameter	Linear convolution	Circular convolution
1.	Shifting of sequences.	Sequences are shifted linearly.	Sequences are shifted circularly.
2.	Convolution sum.	Convolution sum is of infinite length.	Convolution sum is of length 'N'.
3.	Types of sequences.	Sequence are non periodic and must be of finite length.	Sequences are of length 'N' and they are periodic.

7. State the circular frequency shift property of DFT.

If ( ) ↔ X( ) then,

( ) ↔ X(−)$_N$

8. What do you understand by periodic convolution?

Let $x_1(n)$ and $x_2(n)$ be two periodic sequences having Fourier coefficients of $c_1(k)$ and $c_2(k)$ respectively. Let these coefficients be multiplied to give $c_3(k)$. *i.e.,*

$$c_3(k) = c_1(k) \,.\, c_2(k)$$

$$x_3(n) = \sum_{m=0}^{N-1} x_1(m)x_2(n-m)$$

9. Write the analysis and synthesis equation of DFT OR define DFT and inverse DFT OR Define DFT pair OR Write down the pair of DFT equations.

Analysis equation (Transform equation):

( ) = Σ ( ), $\qquad$ = 0,1, … … … … − 1

Synthesis equation:

( ) = 1 Σ ( ) ⁻ , $\qquad$ = 0,1, … …… … − 1

10. How is FFT faster? OR how many multiplication and additions are required to compute N-point DFT using radix-2 FFT?

FFT is aster because it requires less number of complex multiplications and complex additions compared to direct computation of DFT.

Operation	FFT	DFT
Complex Multiplication	— 2 2	N^2

Complex Addition		N^2-N

11. What is FFT?

Special algorithms are developed to compute DFT quickly. These algorithms exploit the periodicity and symmetry properties of twiddle factors (phase factors). Hence DFT is computed fact using such algorithms compared to direct computation. These algorithms are collectively called as Fast Fourier Transform (FFT) algorithms. These algorithms are very efficient in terms of computations.

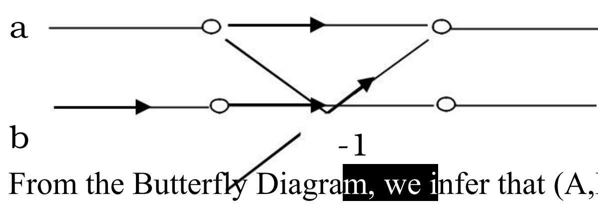
12. What are the advantages of FFT algorithm over direct computation of DFT?

FFT requires less number of multiplications and additions compared to direct computation of DFT.

FFT algorithms can be implemented fast on the DSP Processor.

The calculation of DFT and IDFT both are possible by proper combination of FFT algorithms

What is meant by 'in place' in DIT and DIF algorithms? OR what is inplace computation?



From the Butterfly Diagram, we infer that (A,B) are calculated from (a,b). Hence (A,B) can be stored in place of (a,b) since (a,b) are not required further. This is called in place computation. It reduces the number of memory locations.

Calculate the multiplication reduction factor, α in computing 1024 point DFT, in a radix-2 FFT algorithm.

For direct computation, complex multiplications = $N^2 = (1024)^2$

For radix-2 FFt, complex multiplications = $\frac{N}{2}\log_2 N$

$$= \frac{1024}{2}\log_2 1024 = 5120$$

Multiplication reduction factor, $= \frac{(1024)^2}{5120} = 204.8$

Define the twiddle factor or phase factor of FFT.

The twiddle factor is given as,

Calculate the number of multiplication needed in the calculation of a 512-point radix-2 FFT, when compared to direct DFT.

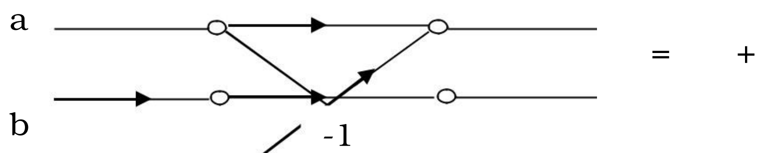
Number of Multiplication required	
Radix-2 FFT	Direct DFT
$\frac{N}{2}\log_2 N$	N^2
$= \frac{512}{2}\log_2 512$	$= (512)^2$
= 2304	= 262144

What do you mean by the term "bit reversal" as applied to FFT?

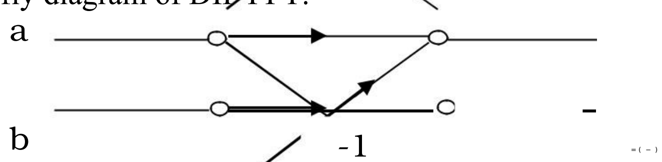
In case of DIT-FFT algorithm, the input sequence x(n) is applied in bit reversed order. For example, if each 'n' is represented by three bits $n_2n_1n_0$, then the bit reversed value of 'n' will be n2n1n0. This means the sequence x(0), x(2), x(4), x(6), x(1), x(3), x(5), x(7) after bit reversing. Similarly, in case of DIF-FFT algorithm, the output DFT X(k) is also shuffled in bit reverse order.

Draw the butterfly diagram for radix 2 DIT-FFT and DIF-FFT OR draw the basic structure of DIT and DIF-FFT flowchart of radix-2.

Butterfly diagram of DIT FFT:



Butterfly diagram of DIF FFT:



Draw the basic butterfly diagram for the computation in the decimation in frequency FFT algorithm and explain.

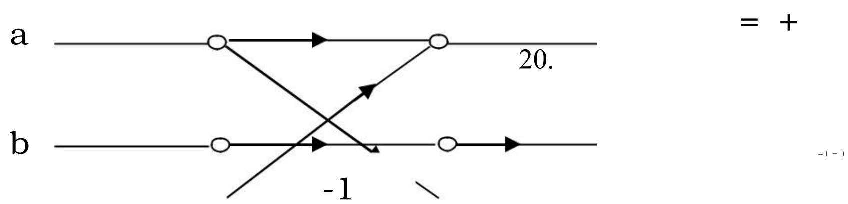


Figure shows the basic operation in DIF-FFT algorithm. Here *a* and *b* are the inputs. A and B are the outputs. is the twiddle factor.

What is meant by radix 2 FFT algorithm?

In radix-2 FFT algorithm, the decimation arrives finally at 2-point sequence. The 2-point DFT is calculated by direct computation.

How many stages of decimations are required in the case of 64 point radix 2 DIT FFT algorithm?

Number of decimation stages are given as,

= 2 = 232 = 1032 = 5 102

Distinguish between DIT and DIF FFT algorithms.

S No	DIT FFT	DIF FFT
1	The time domain sequence is decimated.	The DFT X(k) is decimated.
2	Input sequence is to be given in bit reversed order.	The DFT at the output is in the reversed order.
3	First calculate 2-point DFTs and combines them.	Decimates the sequence step by step to 2-point sequence and calculates DFT.
4	Suitable for calculating inverse DFT.	Suitable for calculating DFT.

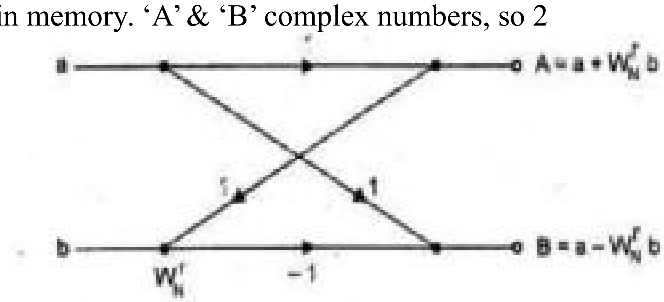
Compare the number of multiplication required to DFT of a 64 point sequence using direct computation and that using FFT.

Number of points N	Direct Computation		DIT FFT algorithm		Improvement in
	Complex multiplications N^2	Complex additions N^2-N	Complex multiplications	Complex additions	processing speed for multiplications
8	64	52	12	24	5.3 times
16	256	240	32	64	8 times
64	4096	4032	192	384	10.67 times
256	65536	65280	1024	2048	6.4 times
1024	1048576	1047552	5120	10240	204.8 times

Explain the symmetry properties of DFT's which provide basis for fast algorithms. Most approaches for improving the efficiency of computation of DFT, exploits the symmetry and periodicity property of W_N^{kn} ie;

What is the advantage of in place computation?

The main advantage of in-place computation is reduction in the memory size in-place computation reduces the memory size. a' & 'b' are inputs and 'A' and 'B' are outputs of butterfly. For anyone input 'a' and 'b' two memory locations are required for each. One memory location to store real part and other memory location to store imagining part. So for both inputs 'a' & 'b' = 2 + 2 = 4 memory location are required. Thus outputs 'A' & 'B' are calculated by using the values 'a' & 'b' stored in memory. 'A' & 'B' complex numbers, so 2



General Butterfly structure

2 = 4 memory location are required.

Once the computation of 'A' & 'B' done then values of 'a' & 'b' are not required. Instead of storing 'A' & 'B' at other memory locations, there values are stored at the same place where 'a' 'b' were stored. That means 'A' & 'B' are stored in the place of 'a' & 'b'. This is called as in-place computation.

Indicate the number stages, the number of complex multiplications at each stage, and the total number of multiplications required to compute 64 point FFT using radix-2 algorithm.

Number of stages $= \log_2 N = \log_2 64 = 6$

Number of complex multiplication $= N_2 \log_2 N = \frac{64}{2} \times 6 = 192$

Total number of multiplications $= N \log_2 N = 64 \times 6 = 384.$

What are the differences and similarities between DIF and DIT algorithms?

Differences:

For DIT the input is bit reversed while the output is in natural order, whereas for DIF the input is in natural order while the output is bit reversed.

The DIF butterfly is slightly different from the DIT butterfly, the difference being that the complex multiplication takes place after the add-subtract operation in DIF.

Similarities:

Both algorithms require same number of operations to compute the DFT. Both algorithms can be done in place and both need to perform bit reversal at some place during the computation.

State the difference between overlap save and overlap add method.

S.No	Overlap save method	Overlap add method
1.	The size of input data block is N=L+M-1	The size of input data block is L
2.	Each data block consist of last M-1 data points of the previous data block followed by L new data points	Each data block is L points and we append M-1 zeros to compute N-point DFT.
3.	Each output block M-1 points are corrupted due to aliasing as circular convolution is employed.	In this no corruption due to aliasing, as linear convolution is performed using circular convolution.

UNIT – IV
DESIGN OF DIGITAL FILTERS

Give the Bilinear Transformation.

Bilinear transformation is given as, $\leftarrow \quad 1 - {}^{-1} = (1 + -1)$

Poles in Right hand of s-plane are mapped outside of the unit circle of z-plane.
Poles in Left hand of s-plane is mapped inside of the unit circle of z-plane.
jΩ axis in s-plane is mapped on unit circle of the z-plane
Bilinear transformation maps poles as well as zeros.

Frequency warping takes place and the frequency relationship is highly non-linear. It is given as

$$= 2\tan^{-1}\ \frac{\Omega}{}$$

What is prewarping?

When bilinear transformation is applied, the discrete time frequency is related to continuous time frequency as,

$$= 2\tan^{-1}\ \frac{\Omega}{}$$

This equation shows that frequency relationship is highly non linear. It is also called frequency warping. This effect can be nullified by applying prewarping. The specifications of equivalent analog filter are obtained by following relationship,

$$= \tan 2$$

This is called prewarping relationship.

What are the limitations of impulse invariant mapping technique?

Frequency mapping is many to one. Therefore aliasing takes place in frequency domain.

Impulse invariant technique is suitable only for lowpass and narrow bandpass filters.

Give the transform relation for converting low pass to band pass in digital domain.

Let the lowpass filter with passband edge frequency ω_p is available. We want bandpass filter with upper band edge ω_u and lower band edge frequency ω_l. then following transformation must be used.

$$^{-1} = - \frac{^{-2} - \ ^{-1} + }{^{-2} - \ ^{-1} + 1}$$

Here

$$= - \frac{2}{+1} \qquad = - \frac{-1}{+1}$$

$$\alpha = \frac{\cos(\frac{+}{2})}{\cos(\frac{-}{2})} \qquad k = \cot(\frac{+}{2})\tan(\frac{}{2})$$

What is frequency wrapping?

The frequency between the continuous analog frequency (Ω) and digital frequency (ω) is given as,

$$\Omega = \frac{2}{}\tan\frac{}{2}$$

For the small values of ω, there exists linear relationship between ω and Ω.But for large values of ω, the relationship is non linear. This non linearity introduces distortion in the frequency axis. This is known as warping effect.

6. What is impulse invariant mapping? What is its limitation?

The impulse invariant mapping is given as,

$$\frac{1}{-} \rightarrow \frac{1}{1 - \ ^{-1}}$$

It maps

Poles in LHS of s-plane inside the unit circle of z-plane .

Poles in RHS of s-plane outside the unit circle of z-plane.

jΩ axis on the unit circle.

Limitations:

The segments $\frac{}{} \le \Omega \le \frac{}{}$ of jΩ axis are all mapped on the unit circle $- \le \le$.

Write the magnitude function of butterworth filter. What is the effect of varying order 'N' on magnitude and phase response?

$$|H(j\Omega)|^2 = \frac{1}{1+\left(\frac{\Omega}{\Omega_c}\right)^{2N}}$$

The transition characteristic of filter changes with order N. the magnitude $|H_a(j\Omega)|$ decays at the rate of 20 N dB per decade.

Mention the two procedures for digitizing the transfer function of an analog filter. The transfer function of an analog filter can be digitized by,

Approximation of derivatives

Impulse invariance

Bilinear transformation

What are the advantages and disadvantages of bilinear transformation? Advantages:

No frequency aliasing

Poles as well as zeros are mapped.

Frequency relationship is highly non linear.

Prewarping is necessary.

Give any two properties of butterworth filter and Chebyshev filter.

Butterworth filters:

The magnitude responses of the Butterworth filter decreases while the frequency Ω increases from 0 to ∞.

The poles of the butterworth filter lie on a circle.

The butterwothfilter have all poles design.

At the cut off frequency Ωc, the magnitude of normalized butterworth filter is $\frac{1}{\sqrt{2}}$.

The filter order N, completely specifies the filter and as the value of N increases the magnitude response approaches the ideal response.

The magnitude response of the Chebyshev filter provides ripple in passband or stopband according to its type.

The poles of the Chebyshev filter lie on the ellipse.

Find the digital transfer function H(z) by using impulse invariant method for the analog transfer function H(s) = $\frac{1}{s+2}$. Assume T= 0.5 sec.

In impulse invariant transformation,

$$\frac{1}{s-p_k} \rightarrow \frac{1}{1-e^{p_kT}z^{-1}}$$

Here p_k = -2 and T= 0.5. Hence we can write,

$$\frac{1}{s+2} \rightarrow \frac{1}{1-e^{-2\times 0.5}z^{-1}} = \frac{1}{1-0.367z^{-1}}$$

Find the digital transfer function H(z) by using impulse invariant method for the analog transfer function H(s) = $\frac{1}{s+2}$. Assume T= 0.1 sec.

In impulse invariant transformation,

$$\frac{1}{s-p_k} \rightarrow \frac{1}{1-e^{p_kT}z^{-1}}$$

Here p_k = -2 and T= 0.1. Hence we can write,

$$\frac{1}{s+2} \rightarrow \frac{1}{1-e^{-2\times 0.1}z^{-1}} = \frac{1}{1-0.818z^{-1}}$$

What is the relationship between analog and digital frequency in impulse invariant transformation?

The relationship between analog and digital frequencies is given as,

= Ω

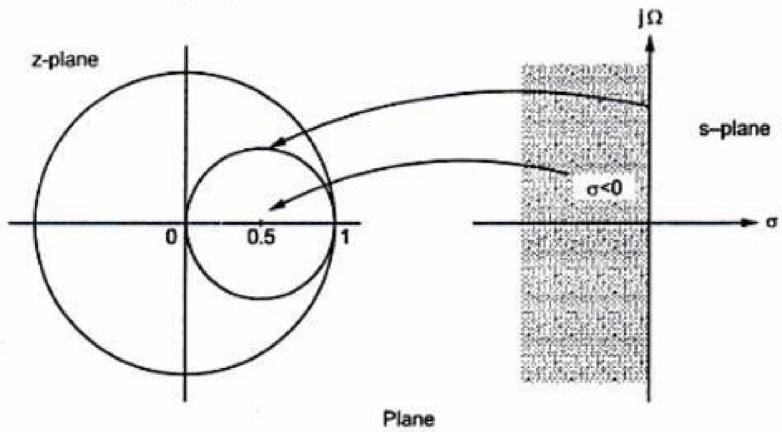
This relationship shows that the mapping of jΩ axis is many to one unit circle. Therefore, there is aliasing in frequency domain.

What is the advantage of Direct form II realization when compared to direct form I realization?

Direct form II realization requires less memory.

Computation time is reduced in Direct form II.

Sketch the mapping of s-plane and z-plane in approximate of derivates.



Plane

What are the properties of impulse invariant transformation.

Only poles of system function are mapped.

There is aliasing in frequency domain due to mapping.

Frequency relationship = Ω is linear.

Stable analog filter is converted to stable digital filter.

Give the steps in the design of a digital filter from analog filters.

Specification of analog filter

↓

Obtain specification of digital filter

↓

Obtain specification of equivalent analog filter as per

Bilinear transform: Ω = 2 tan 2

Impulse invariance: Ω = 2

↓

Obtain order of the filter

↓

Determine cut-off frequency

↓

Determine poles of $H_a(s)$

↓

Determine $H_a(s)$ from/poles in LHS of s-plane

↓

Apply s → z transformation to get H(z)

↓

Apply s → z transformation to get H(z)

18. What are the disadvantages of direct form realization?

Direct form structure is sensitive to parameter quantization. For large value of 'N' the location of poles and zeros is shifted from their actual values in case of direct form realization.

19. Mention the advantages of cascade realization.

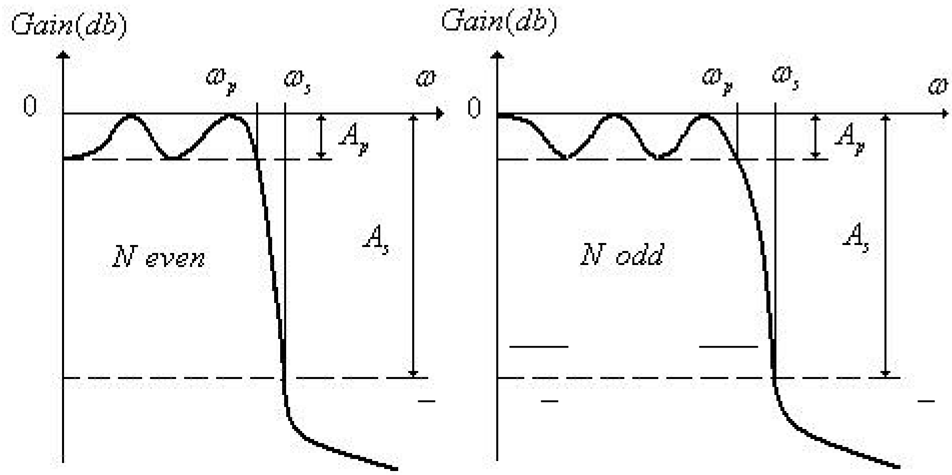
Quantization errors and theirs effects are reduced because of cascade realization.

20. Convert the given analog transfer function into digital by impulse invariant method.

Impulse invariant transformation is given as,

Here p_k = and T= 1sec,

21. Sketch the frequency response of even/odd ordered Chebyshev lowpass filter.



Why impulse invariant method is not preferred in the design of highpass IIR filters?

In impulse invariant method, the segments of $^{2\ -1} \leq \Omega \leq {}^{2\ -1}$ are mapped on the unit circle repeatedly. Hence first set, i.e. $- \leq \Omega \leq$ is mapped correctly. Then $\leq \Omega \leq {}^{3}$ is mapped on the same circle. Thus one point on the circle represents multiple analog frequencies. Hence high frequencies are actually mapped as low frequencies. Therefore all high frequencies are aliased frequencies. Hence impulse invariant technique is not much suitable for highpass filters. But for lowpass filters it is better, since actual mapping takes place.

Why do we go for analog approximation to design a digital filter?

These are effective filter approximation techniques available in analog domain. Using transformation methods a stable analog filter can be converted to stable digital filters. Hence it becomes easier to design IIR filters from analog filters. But such effective approximations are not available in discrete domain.

List the various forms of realizations of IIR system.

Direct form I
Direct form II
Cascade realization

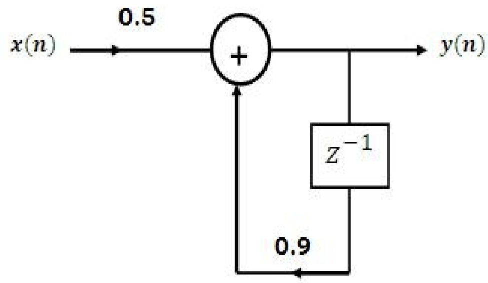
Parallel form realization

Lattice realization

Mention advantages of direct form II and cascade structure.

Direct form II structures require less number of storage locations. Cascade structures are easy to implement, since second order sections are simply cascaded.

Draw the direct form I structure for the system y(n) = 0.5x(n)+0.9 y(n-1).



Write down the expression for the transfer function of a first order butterworth analog filter having lowpass behavior.

The transfer function of the normalized lowpass butterworth filter can be expressed as,

( ) = 1 / + 1

28. What is the main drawback of impulse invariant mapping?

Aliasing is the main drawback of impulse invariant mapping. The segments of 2^{-1} ≤ 2^{-1} are mapped on the unit circle repeatedly. Hence first set, i.e. − ≤ Ω ≤ is mapped correctly. Then ≤ Ω ≤ 3 is mapped on the same circle. Thus one point on the circle represents multiple analog frequencies. This is aliasing.

29. Why IIR filters do not have linear phase?

IIR filters are recursive. These filters use feedback. The present output depends upon previous outputs also. Hence IIR filter have non linear phase. However, it is possible to design IIR filter with piecewise linear phase.

30. What are the properties that are maintained same in the transfer of analog filter into digital filter?

Stability: A stable analog filter is converted to stable digital filter.

Causality: A causal analog filter is converted to causal digital filter.

31. Write down the equation for frequency transformation from lowpass to bandpass filter.

The transformation is given as,

→ (Ω − Ω)

Where Ω_l lower bandedge frequency

Ω_u higher bandedge frequency

Find digital filter equivalent for ( ) = + . Bilinear transformation is given as,

− 1 = (+ 1)

With T = 1, = 2 (+1⁻¹) = 2 (+1⁻¹)

Hence H(z) = H(s)

= 1 / (1 / 2 (1 + ⁻¹) + 8)

$+ \; -1$

Determine the order of the analog butterworth filter that has a -2db passband attenuation at a frequency of 20rad/sec and atleast -10db stopband attenuation at 30 rad/sec.

$A_p = 2$ dB, $A_s = 10$ dB, $\Omega_p = 20$ rad/sec, $\Omega_s = 30$ rad/sec

$$= \frac{1}{2} [\; 10^{0.1} \; -1 \;] \quad \Omega \quad [\; \frac{}{\Omega} \;]$$

$$= \frac{1}{2} [\; 10^{} \; -1 \;] \quad \frac{30}{20}$$

$$= 3.37$$

$$= 4$$

34. By impulse invariant method obtain the digital filter transfer function and differential equation of the analog filter ( ) = + .

$( ) = \frac{1}{+1}$

Impulse invariance transformation is given as,

$$\frac{1}{} \rightarrow \frac{1}{1 - \; ^{-1}}$$

$$( ) = \frac{1}{} , = -1$$

Let T= 1,

$$( ) = \frac{1}{1 - \; ^{-1}} , = \frac{1}{1 - 0.367 \; ^{-1}}$$

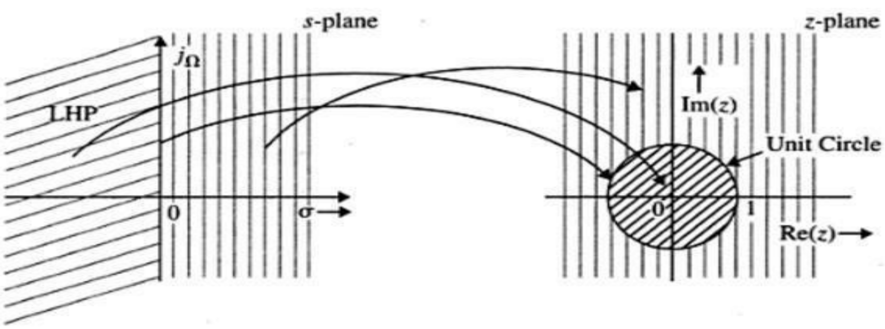
$$\frac{( )}{( )} = \frac{1}{1 - 0.367 \; ^{-1}}$$

$$( ) - 0.367 \; ^{-1} ( ) = ( )$$

Taking inverse z-transform,

$$y(n) - 0.367\, y(n-1) = x(n) \;\; \textbf{or} \;\; y(n) = x(n) + 0.367\, y(n-1)$$

35. Sketch the mapping of s-plane and z-plane in bilinear transformation.



As above shown,

LHS of s-plane is mapped inside the unit circle.

RHS of s-plane is mapped outside the unit circle.

jΩ axis in s plane is mapped on the unit circle.

Convert ( ) = + into a digital filter using approximation of derivatives with T = sec.

Approximation of derivatives is obtained by putting $= 1- {}^{-1}$

H(z) = H(s) = $- \left(\frac{2 \quad -1}{+1} \right) = \frac{1}{1- \quad (\quad)+1} = \frac{1}{1- \quad (\quad)+1}$ 0.01 0.1

$= 1.01 - 2^{-1} + {}^{-2}$

37. What are the requirements for converting a stable analog filter to a stable digital filter?

i) The poles in left hand side of s-plane should be mapped inside the unit circle.
ii) The poles in right hand side of s-plane are mapped outside the unit circle.
iii) The imaginary axis in s-plane should be mapped on unit circle.

38. Convert the analog filter with system function H(s) into a digital IIR filter by means of impulse invariant method:

$( ) = \frac{}{(+ .)(+ .)}$

$( ) = \frac{1}{(+ 0.2)(+ 0.6)} = \frac{}{+ 0.2} - \frac{}{+ 0.6}$

1

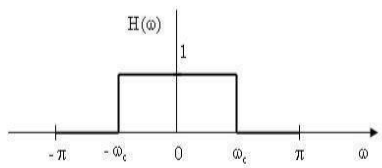
$( ) = \frac{}{1 - \quad -} - \frac{}{1 - \quad -}$

39. Why do we go for analog approximation to design a digital filter?

There are effective filter approximation techniques available in analog domain. Using transformation methods analog filter can be converted to stable digital filters. Hence it becomes easier to design IIR filters from analog filters. But such effective approximations are not available in discrete domain.

What is the effect of having abrupt discontinuity in frequency response of FIR filters?

Consider a low pass filter having frequency response as shown.



Observe that there is abrupt discontinuity in the frequency response at ωc. due to the discontinuity, the impulse response becomes infinite in length i.e.,

$h( ) = \begin{cases} \frac{\sin}{} & \neq 0 \\ \frac{}{} & = 0 \end{cases}$

It is clear from above equation that h(n) is infinite in length. Hence it is necessary to truncate this response.

What are the characteristics features of FIR filters?

FIR filter are all zero filters.
FIR filters are inherently stable filters.
FIR filters can have linear phase.

What is normalized filter?

Normalized filter or prototype filter is the lowpass filter with cutoff frequency of 1 rad/sec.

43. What is canonic structure?

If the number of delays in the structure is equal to order of the difference equation or order of the transfer function, them it is called canonic form realization.

Is bilinear transformation linear or not? What is the merit and demerit of bilinear transformation?

Bilinear transformation is nonlinear.

Merit: Bilinear transformation avoids frequency aliasing while mapping.

Demerit: Mapping is highly nonlinear, hence frequency warping effect takes place.

Write the equation of Bartlett and Hamming window.

Bartlett window: $(\,) = 1 - \frac{2|\; - \frac{-1}{2}\;|}{-1}$

Hamming window: $(\,) = 0.54 - 0.46\cos\frac{2}{-1}$ $, = 0,1,2, \ldots\ldots - 1$

46. Compare bilinear transformation and impulse invariant method of IIR filter design.

S.No	Impulse Invariant Method	Bilinear transform Method
1.	Only poles H(s) are mapped.	Both poles and zeros of H(s) are mapped.
2.	Aliasing of frequencies takes place.	No aliasing since mapping is one to one.
3.	Linear frequency relationship.	Nonlinear frequency relationship.

47. What is meant by linear phase response of a filter?

When the phase shift is directly proportional to frequency, i.e.,

$$\angle(\,) = \quad ,$$

This is called linear phase shift.

48. What is the difference between analog and digital filters?

S.No	**Analog filter**	**Digital filter**
1.	In analog filter both input and output are continuous time signal.	In digital filter both input and output are discrete time signals
2.	It can be constructed using active and passive components.	It can be constructed using adder, multiplier and delay units.
3.	It is defined by linear differential eqn.	It is defined by linear difference eqn
4.	These filters operate in infinite frq. Range, theoretically but in practice it is limited by finite max. operating freq. depending upon the devices used.	freq. range is restricted to half the sampling range and it is also restricted by max. computational speed available for particular application.

Write the transformation equation to convert low pass filter into low pass filter with different cut of frequency and high pass filter.

Cut off frequency = Ω_c

Low pass to Low pass transformatio: Substitute s = s/Ω_c.

Low pass to High pass transformation: Substitute s = Ω_c/s.

State two advantages of bilinear transformation.

Bilinear transformation is one to one mapping

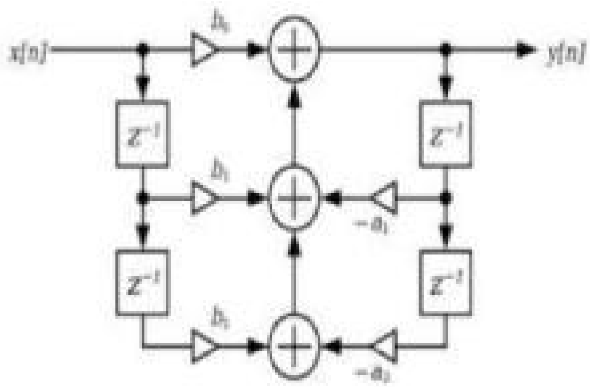
There is no aliasing.

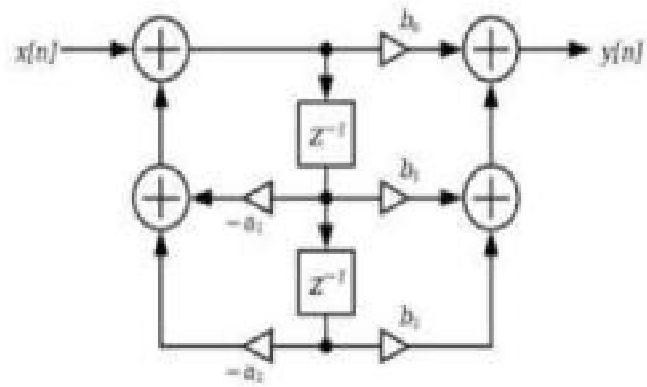
The mapping is highly non-linear producing frequency compression at high frequencies.

Impulse response and phase response of the analog filter is not preserved during bilinear mapping.

Draw the direct form structure of IIR filter.

Direct Form I Structure **Direct Form II structure**





What are the limitations of impulse invariant method of designing digital filters?

In impulse invariant method, the mapping from S-plane to Z-plane is many to one.

Thus, there are an infinite number of poles that map to the same location in the Z-plane. It produces aliasing effect.

Due to spectrum aliasing, the impulse invariance method is inappropriate for designing high pass filters.

What is the relationship between analog and digital frequency in impulse invariant transformation?

The relationship between analog and digital frequency in impulse invariant transformation is $\omega = \Omega T$

What is the basic difference between cascade form and direct form structures for FIR systems?

Cascade form is basically in need of series memory. No of memory space required less in case of direct-2 form of FIR w.r.t. cascade form start use of FIR systems.

What is the importance of Windowing?

The infinite duration impulse response can be converted to a finite duration impulse response by trucating the infinite series at n=±N. But this results in undesirable oscillations in the pass-band and stepband of the digital filter. This is due to slow convergence of the Fourier series near the point of discontinuity. These undesirable oscillations can be reduced by using a set of time limited weighing functions z e referred as windowing function.

The windowing function consists of main lobe which contains most of the energy of window function and side lobes which decay rapidly

A major effect of windowing is that the discontinuities is $H(e^{j\omega})$are converted into transition bands between values on either side of the discontinuity

Window function have side lobes that decrease in energy rapidly as tends to π.

What will happen if length of windows is increased in design of FIR filters?

If length of window is increased in design of FIR filter more coefficients need to be calculated and more memory space used for it.

What are the essential features of a good window for FIR filters?

Features of a good window for FIR filters: 1. Side lobe level should be small. 2. Broaden middle section.3. Attenuation should be more. 4. Smoother magnitude response. 5. The tradeoff between main lobe widths and side lobe level can be adjusted. 6. Smoother ends. 7. If cosine term is used then side lobes are reduced further.

Why FIR digital filters cannot have linear phase?

For FIR filter unit impulse response for symmetric system are given by:

$$h(n) = h(m-1-n) \quad (1),\ n = 0, 1, 2, \ldots\ldots\ldots\ m-1$$

$$n = 0,\ h(0) = h(8-1-0) = h(7)$$

$$n = 1,\ h(1) = h(8-1-1) = h(6)$$

If h (n) is symmetric then filter is symmetric. For antisymmetric sequence.

$$h(n) = -h(m-1-n)\ ;\ n = 0, 1, 2, \ldots\ldots, m-1$$

i.e. condition for linear phase. For FIR filters m is finite i.e. may be odd, symmetric and antisymmetric conditions so in FIR filters m is infinite. So it does not satisfies linear phase condition of eq. (1) and (2).So FIR filters cannot have linear phase.

Define Ripple ratio

The Ripple ratio is defined as , the ratio of maximum sidelobes amplitude to the mainlobe amplitude.

= ( ) 100

What is Gibb's Oscillation? (or) State the effect of having abrupt discontinuity in frequency response of FIR filters.

The truncation of Fourier series is known to introduce the unwanted ripples in the frequency response characteristics H(ω) due to non uniform convergence of Fourier series at a discontinuity These ripples or oscillatory behaviour near the band edge of the filter is known as "Gibb's phenomenon or Gibb's oscillation

What are the methods used to reduce Gibb's phenomenon?

The discontinuity between pass band and stop band in the frequency response is avoided by introducing the transition between the pass band and stop band.

Another technique used for the reduction of Gibb's phenomenon is by using window function that contains a taper which decays towards zero gradually instead abruptly.

What are the necessary and sufficient conditions for linear phase characteristics of a FIR filter?

The necessary and sufficient conditions for linear phase characteristics of a FIR filter is that the phase function should be a linear function of ω,which in turn requires constant phase delay or constant phase and group delay.

For FIR filter to have linear phase, the necessary and sufficient condition is

h( ) = ±h(− 1 −)

where N= duration of the sequence

List the factors that are to be specified in the filter design problem.

• The maximum tolerable passband ripple. •

The maximum tolerable stopband ripple.

• The passband edge frequency ωp •

The stopband edge frequency ωs.

Characteristic features of rectangular window.

The mainlobe width is equal to 4π/N.

The max sidelobe magnitude is -13 dB.

The sidelobe magnitude does not decreases significantly with increasing ω.

65. List features of hanning and hamming window spectrum.

Hanning window:

The mainlobe width is equal to $8\pi/N$.
The max sidelobe magnitude is -31dB.
The sidelobe magnitude decreases with increasing ω.

The mainlobe width is equal to $8\pi/N$.
The max sidelobe magnitude is -41dB.
The sidelobe magnitude remains constant for increasing ω.

What are the advantages of Kaiser window?

- It provides flexibility for the designer to select side lobe level and N
- It has the attractive property that the side level can be varied continuously from the value in the Blackman window to the high value in the rectangular window.

Define Phase delay and Group delay.

- Group delay is defined as derivative of phase with respect to frequency. •

Phase delay is defined as phase divided by frequency.

What are the methods used to design FIR filter?

Window Method: It involves straight forward analytical procedure however in some cases iteration is required to obtain the desired result

Frequency Sampling: A desired frequency response is uniformly sampled and filter coefficients are then determined from these samples using the discrete fourier transform.

Optimal or minimal design: Minimizing the maximum error between the desired and the actual frequency response by spreading the error in PB and SB.

69. Why direct Fourier series method is not used in FIR filter design?

The impulse response h(n) is infinite in duration. The filter is unrealizable since the impulse response begins at- ∞ i.e no finite amount of delay can make the impulse response realizable. Therefore the filter which results from a Fourier series representation of $h(e^{jw})$ is an unrealizable FIR Filter.

What are the advantages and disadvantages of FIR filters

Advantages:

FIR filters have the following advantages over IIR filters.

Fir filter have exact linear phase.
They are always stable.
The design methods are generally linear.
They can be realized efficiently in hardware.
Fir filters can be realized in both recursive and non-recursive structure.

Large storage requirements needed.
For the same filter specifications the order of FIR filter design can be as high as 5 to 10 times that of an IIR design.
Powerful computational facilities are required for the implementation.

71. Discuss the stability of FIR filters.

FIR filter is always stable because all its poles are at the origin. They are all zero filters. FIR filter poles are always inside the unit circle. Hence FIR filters are always stable.

What do you meant by linear phase response?

The phase response of the type

←
( ) = , where K is constant

is called linear phase response. The linear phase filter does not alter the shape of the original signal. In many cases a linear phase characteristic is required throughout the pass band of the filter to preserve the shape of a given signal within the pass band.

State the condition for a digital filter to be causal and stable.

A digital filter is causal if its impulse response

h(n) = 0 for n < 0

A digital filter is stable if its impulse response is absolutely summable

$$\sum_{n=-\infty}^{\infty} |h(n)| < \infty$$

What are the functions of desirable features of a window function?

Features:

The central lobe of the frequency response of the window contains most of energy and it should be narrow

The highest side lobe of the frequency response is very small.

The side lobe of the frequency response is decreased rapidly as ω tends to Π

Write the steps of FIR filter design.

Selective the desired frequency response $H_d(\omega)$.

Take inverse Fourier transform of $H_d(\omega)$ to get $h_d(n)$.

Convert the infinite duration $h_d(n)$ to finite duration sequence h(n).

Take Z transform of h(n) to get the transfer function H(z) of the FIR filter.

Write the conditions for constant phase delay and group delay in linear phase FIR filter.

Phase delay = $\alpha = (M - 1)$ /

2 Group delay = $\beta = \pm \pi / 2$

What is the difference between FIR and IIR?

Sl. No.	FIR filter	IIR filter
1	h(n) is finite duration	h(n) is infinite duration
2	This type of filter is non-recursive and do not use feedback.	They are recursive filter and using feedback from output.
3	Linear phase response	Nonlinear phase response. Linear phase is obtained if $H(z) = \pm z^{-1} H(z^{-1})$
4	There are two design methods: 1. Windowing 2. Frequency sampling method	There are two design methods: 1. Bilinear transform 2. Impulse invariance method.
5	More number of multiplications are required	Less number of multiplications are required

UNIT V
DIGITAL SIGNAL PROCESSOR

1. Write short notes on general purpose DSP processors

General-purpose digital signal processors are basically high speed microprocessors with hard ware architecture and instruction set optimized for DSP operations. These processors make extensive use of parallelism, Harvard architecture, pipelining and dedicated hardware whenever possible to perform time consuming operations.

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Write notes on special purpose DSP processors. There are two types of special; purpose hardware.

Hardware designed for efficient execution of specific DSP algorithms such as digital filter, FFT.

Hardware designed for specific applications, for example telecommunication, digital audio.

Briefly explain about Harvard architecture.

The principal feature of Harvard architecture is that the program and the data memories lie in two separate spaces, permitting full overlap of instruction fetch and execution. Typically these types of instructions would involve their distinct type.

Instruction fetch

Instruction decode

Instruction execute.

Briefly explain about multiplier accumulator.

The way to implement the correlation and convolution is array multiplication Method.

For getting down these operations we need the help of adders and multipliers. The combination of these accumulator and multiplier is called as multiplier accumulator.

What are the types of MAC is available?

There are two types MAC'S available

- Dedicated & integrated
- Separate multiplier and integrated unit

What is meant by pipeline technique?

The pipeline technique is used to allow overall instruction executions to overlap. That is where all four phases operate in parallel. By adapting this technique, execution speed is increased.

What are four phases available in pipeline technique?

The four phases are

Fetch

Decode

Read

Execution

In a non-pipeline machine, the instruction fetch, decode and execute take 30 ns, 45 ns and 25 ns respectively. Determine the increase in throughput if the instruction were pipelined.

Assume a 5ns pipeline overhead in each stage and ignore other delays.

The average instruction time is = 30 ns+45 ns + 25 ns = 100 ns

Each instruction has been completed in three cycles = 45 ns * 3 = 135ns Throughput of the machine =

The average instruction time/Number of M/C per instruction = 100/135 = 0.7407

But in the case of pipeline machine, the clock speed is determined by the speed of the slowest stage plus overheads.

In our case is = 45 ns + 5 ns =50 ns

The respective throughput is = 100/50 = 2.00

The amount of speed up the operation is = 135/50 = 2.7 times

Assume a memory access time of 150 ns, multiplication time of 100 ns, addition time of 100 ns and overhead of 10 ns at each pipe stage. Determine the throughput of MAC

After getting successive addition and multiplications The total time delay is 150 + 100 + 100 + 5 = 355 ns System throughput is = 1/355 ns.

10. Write down the name of the addressing modes.

Direct addressing. Indirect addressing. Bit-reversed addressing. Immediate addressing.
Short immediate addressing.
Long immediate addressing.
Circular addressing.

11. What are the instructions used for block transfer in C5X Processors?

The BLDD, BLDP and BLPD instructions use the BMAR to point at the source or destination space of a block move. The MADD and MADS also use the BMAR to address an operand in program memory for a multiply accumulator operation

12. Briefly explain about the dedicated register addressing modes.

The dedicated-registered addressing mode operates like the long immediate addressing modes, except that the address comes from one of two special-purpose memory-mapped registers in the CPU: the block move address register (BMAR) and the dynamic bit manipulation register (DBMR).

The advantage of this addressing mode is that the address of the block of memory to be acted upon can be changed during execution of the program.

13. Briefly explain about bit-reversed addressing mode?

In the bit-reversed addressing mode, INDX specifies one-half the size of the FFT. The value contained in the current AR must be equal to 2n-1, where n is an integer, and the FFT size is 2n. An auxiliary register points to the physical location of a data value. When we add INDX t the current AR using bit reversed addressing, addresses are generated in a bit-reversed fashion. Assume that the auxiliary registers are eight bits long, that AR2 represents the base address of the data in memory (0110 00002), and that INDX contains the value 0000 10002.

14. Briefly explain about circular addressing mode.

Many algorithms such as convolution, correlation, and finite impulse response (FIR) filters can use circular buffers in memory to implement a sliding window; which contains the most recent data to be processed. The ‘C5x supports two concurrent circular buffer operating via the ARs. The following five memory-mapped registers control the circular buffer operation.

CBSR1- Circular buffer 1 start register.
CBSR2- Circular buffer 2 start Register,
CBER1- Circular buffer 1 end register
CBER2- Circular buffer 2 end register
CBCR - Circular buffer control register.

15.Write the name of various part of C5X hardware.

6.Central arithmetic logic unit (CALU)
7.Parallel logic unit (PLU)
8.Auxiliary register arithmetic unit (ARAU)
9.Memory-mapped registers.
10. Program controller.

16. Write short notes about arithmetic logic unit and accumulator.

The 32-bit general-purpose ALU and ACC implement a wide range of arithmetic and logical functions, the majority of which execute in a single clock cycle. Once an operation is performed in the ALU, the result is transferred to the ACC, where additional operations, such as shifting, can occur. Data that is input to the ALU can be scaled by the prescaler.

The following steps occur in the implementation of a typical ALU instruction:

Data is fetched from memory on the data bus,

Data is passed through the prescaler and the ALU, where the arithmetic is performed, and

The result is moved into the ACC.

The ALU operates on 16-bit words taken from data memory or derived from immediate instructions. In addition to the usual arithmetic instructions, the ALU can perform Boolean operations, thereby facilitating the bit manipulation ability required of high-speed controller. One input to the ALU is always supplied by the ACC. The other input can be transferred from the PREG of the multiplier, the ACCB, or the output of the prescaler. After the ALU has performed the arithmetic or logical operation, the result is stored in the ACC.

17.Write short notes about parallel logic unit.

The parallel logic unit (PLU) can directly set, clear, test, or toggle multiple bits in control/status register pr any data memory location. The PLU provides a direct logic operation path to data memory values without affecting the contents of the ACC or the PREG.

18.What is meant by auxiliary register file?

The auxiliary register file contains eight memory-mapped auxiliary registers (AR0-AR7), which can be used for indirect addressing of the data memory or for temporary data storage. Indirect auxiliary register addressing allows placement of the data memory address of an instruction operand into one of the AR. The ARs are pointed to by a 3-bit auxiliary register pointer (ARP) that is loaded with a value from 0-7, designating AR0-AR7, respectively.

19.Write short notes about circular registers in C5X.

The ‘C5x devices support two concurrent circular buffers operating in conjunction with user-specified auxiliary register. Two 16-bit circular buffer start registers (CBSR1 and CBSR2) indicate the address where the circular buffer starts. Two 16-bit circular buffer end registers (CBER1 and CBER2) indicate the address where the circular buffer ends. The 16-bit circular buffer control register (CBCR) controls the operation of these circular buffers and identifies the auxiliary registers to be used.

20. Mention the four different buses of TMS320C5x and their function.

Program bus:Carries the instruction code and immediate operands from program memory to CPU

Program Address bus:Provides address to program memory space for both read and write.

Data read bus:Interconnects various elements of the CPU to data memory space.

Data read address bus: Provides the address to access the data memory space.

PART –B

UNIT-I
INTRODUCTION

Define energy and power signal? Also examine whether the following signals are energy or power or ~~neith~~er energy nor power signals. (i) $x_1(n)=(1/2)^n$ u(n) (ii) $x_2(n)=\sin(\pi n/6)$ (iii)

$x_3(n)=$ (3 +6)

$x_4(n)=\ ^{(2\ )}\ (\ )$

Describe the concept of Sampling ,quantization and quantization error.

Test the following systems are linear, causal, time invariant, stable, static (i) y(n)=x(2n)

← y(n)=sin[x(n)]

Solve and tell whether the following signals are periodic or not.(i)x(n)=cos(3πn) (ii) x(n) = sin(3n)

Demonstrate which of the following systems are stable (i) y(n)=$\sum_{=-\infty}$ ( )

y(n)=log(1 + | ( )|) (iii) (−1) (iv) y(n)=x^2(n)

Demonstrate which of the following systems are stable (i) y(n)=(| ( )|) (ii) y(n)=$\sum^{\infty}_{=}$ (−)

Explain the classification of discrete signal.

Given y(n)=x[n^2], Test whether the system is linear, time invariant, memoryless and causal.

Test whether the following is an energy signal or power signal (i) x1(n) = cos(2)

$x_2(n)=3[0.5]^n$ u(n)

Summarize from first principles, state and explain sampling theorem both in time domain and in frequency domain.

Demonstrate the response of the following systems to the input signal

x(n)={| | − 3 < < 3}

0

(i) x_1(n)=x(n-2) (− 2) (ii) x_2(n)=x(n+1)u(n-1) (iii)y(n)=$\frac{1}{3}$ [(+ 1) + ( ) + (− 1)] (iv) y(n)=max(x(n+1),x(n),x(n-1)) (v) Find the even and odd components of given x(n).

12. A discrete time systems can be (i)Static or dynamic (ii) Linear or non Linear (iii) Time invariant or time varying (iv) Stable or unstable (v) Causal or noncausal

Examine the following systems with respect to the properties above (i) ( ) = $\sum^{+1}$ ( ) $=-\infty$

13. Test the causality and stability of the systems y(n)=x(-n)+x(n-2)+x(2n-1)
14. Test the system for linearity and time invariance y(n)=(n-1)x^2(n)+c
15. A discrete time system is represented by the following difference equation in which x(n) is input and y(n) is output(n)=3y(n-1)-nx(n)+4x(n)+2x(n+1); and n≥0. Is this system is linear? Shift Invariant? Causal? In each case, quote your answer.
16. Describe the properties of discrete time systems.

UNIT II
DISCRETE TIME SYSTEM AN ALYSIS

1. Calculate the causal signal x(n) whose z-transform is given by

X(Z)= (1+ −1+0.5 −2)

Solve and obtain the z-transform of the signal x(n)=(cosω0n)u(n).

Evaluate the z-transform and ROC of $x(n)=r^n\cos(n\theta)u(n)$

4. Evaluate the Inverse z-transform of X(z) = ———— ROC | | > 1,| | < 1 , 1 < | | < 1

(3 2−4 +1)

5. Find the Z-transform and analyze its associated ROC for the following discrete time signal

−1 n 1 n

u(-n-1)

5

6. Explain the properties of Z-transform.

7. Solve and find the Z transform for the following function.(i) $x(n)=n^2u(n)$ (ii) $x(n)=3^nu(n)$

8. Examine the x(n) for the following X(z) = $\frac{1}{(1-1.5 \;^{-1}+0.5 \;^{-2})}$ ROC;| | > 1, | | < 0.5,0.5 < | | < 1

9. A Linear time-invariant system is characterized by the system function

H(z)= $\frac{3-4 \;^{-1}}{1-3.5 \;^{-1}+1.5 \;^{-2}}$. Specify the ROC of H(z) and Estimate the value of h(n) for the following conditions (1) The system is stable (2) The system is causal (3) The system is anticausal

10. Examine the value of x(n) for the given x(Z) with ROC (1) |z|>2 (2) |z| <2

X(z)= $\frac{1+3 \;^{-1}}{1+3 \;^{-1}+2 \;^{-2}}$ Analyze the impulse response of the system described by the difference equation y(n)=y(n-1)- y(n-2)+x(n)+x(n-1) using Z transform and discuss its stability.

11. Examine the linear convolution of x(n)={2,4,6,8,10} with h(n)={1,3,5,7,9}

12. Evaluate the frequency response of the system described by system function H(z)= $\frac{1}{(1-0.5 \;^{-1})}$

13. Evaluate the impulse response given by difference equation y(n)-3y(n-1)-4y(n-2)=x(n)+2x(n -1).

14. Estimate and predict the frequency response of the LTI system governed by the equation y(n)=a1y(n-1)-a2y(n-2)-x(n)

15. Examine the convolution for the sequence. x(n)={-1,1,2,-2} , h(n)={0.5,1,-1,2,0.75}

↑ ↑

A system is described by the difference equation $y(n)-\frac{1}{2}y(n-1)=5x(n)$. Identify and Determine the solution, when the $x(n)=(\frac{1}{5})^n u(n)$ and the initial condition is given by $y(-1)=1$, using z transform.

Identify and examine the value of DTFT for the given sequence $x(n)=a^n(u(n)-u(n-8)), a< 1$

Quote and prove the linearity and frequency shifting theorems of the DTFT.

UNIT III
DISCRETE FOURIER TRANSFORM & COMPUTATION

Derive and explain the decimation-in time radix-2 FFT algorithm and draw signal flow graph for 8-point sequence.

Using FFT algorithm, Examine the DFT using DIF of x(n)={2,2,2,2,1,1,1,1}.

Explain the following properties of DFT (1) Convolution.(2) Time shifting (3) Conjugate Symmetry.

Examine the 4 point DFT of x(n) ={0,1, 2,3}.

Discuss the Radix 2 DIF - FFT algorithm for 8 point DFT

Estimate the 8 point DFT using DIT - FFT algorithm for x(n)={1,1,1,1,1,1,1,1}

7. An 8-point sequence is given by x(n)={1, 2, 4, 8, 16, 32, 64, 128}. Calculate 8-point DFT of x(n) by radix DIT-FFT method also sketch the magnitude and phase.

Describe the following properties of DFT (1) Time reversal (2) Circular convolution.

Examine the circular convolution of $x_1(n)$= {1, 2, 2, 1} $x_2(n)$ ={1, 2, 3, 1}

Calculate the output y[n] of a filter whose impulse response is h[n]={1,1,1}and input signal x[n]={3,-1,0,1,3,2,0,1,2,1} using overlap save method.

The first five points of the eight point DFT of a real valued sequence are { 0.25 , 0.125 – j0.3018 , 0 , 0.125 - j0.0518 , 0 } Estimate the value of remaining three points

Estimate the eight point DFT of the sequence x=[0,1,2,3,4,5,6,7], using DIF FFT algorithm.

Given x(n)=n+1, and N=8, examine X(K) using DIT, FFT algorithm.

Use 4-point inverse FFT for the DFT result {6,-2+j2,-2,-2-j2} and identify the input sequence.

Examine the 8-point DFT of the sequence x(n)={1,1,1,1,1,1,0,0,0}.

Examine the circular convolution of the sequence using concentric circle method x1={1,1,2,1} and x1={1,2,3,4}

UNIT IV
DESIGN OF DIGITAL FILTERS

Realize a cascade and parallel realization for the system having difference equation
y(n)+0.1y(n-1)-0.2y(n-2)=3x(n)+3.6x(n-1)+0.6x(n-2).

Design a length-5 FIR band reject filter with a lower cut-off frequency of 2KHz, an upper cut-off frequency of 2.4KHz, and a sampling rate of 8000Hz using Hamming window.

Explain the impulse invariant method of designing IIR filter.

Design a second order digital low pass Butterworth filter with a cut-off frequency 3.4 KHz at a sampling rate of 8 KHz using bilinear transformation.

Design an FIR linear phase, digital filter approximating the ideal frequency response

$$H_d(\omega)=\begin{cases} 1 & \omega \le \frac{\pi}{6} \\ 0 & \frac{\pi}{6} \le \omega \le \pi \end{cases}$$

Determine the coefficients of a 25 tap filter based on the window method with a rectangular window

6. Convert the analog filter with system function $H_a(s)=\frac{+0.1}{(\ +0.1)^2+9}$ into a digital IIR filter by means of the impulse invariance method and explain it.
7. Realize the parallel structures for the given difference equation
   y(n)=y(n-1)-0.5y(n-2)+x(n)-x(n-1)+x(n+2).
8. The specification of the desired lowpass filter is

$$\frac{1}{\sqrt{2}} \le |H(w)| \le 1.0 \;;\; 0 \le w \le 0.2\pi$$
$$|H(w)| \le 0.08 \;;\; 0.4\pi \le w \le \pi$$

   Design a Butterworth digital filter using bilinear transformation.

9. The specification of the desired low pass filter is

$$0.9 \le |H(w)| \le 1.0 \;;\; 0 \le w \le 0.25\pi$$
$$|H(w)| \le 0.24 \;;\; 0.5\pi \le w \le \pi$$

   Design a Chebyshev digital filter using impulse invariant transformation.

Design an IIR digital low pass butterworth filter to meet the following requirements:

Pass band ripple (peak to peak): ≤ 0.5dB,

Pass band edge: 1.2 kHz, Stop band attenuation: ≥ 40dB,

Stop band edge: 2.0 kHz,

Sampling rate: 8.0 kHz. Use bilinear transformation technique.

Discuss the limitation of designing an IIR filter using impulse invariant method

Convert the analog filter with system transfer function using bilinear transformation

$H_a(s) = (s+0.3) / ((s+0.3)^2+16)$

The specification of the desired low pass filter is

0.8 ≤ | ( )| ≤ 1.0; 0 ≤ ≤ 0.2
| ( )| ≤ 0.2; 0.32 ≤ ≤ 0.8

Design butterworth digital filter using impulse invariant transformation.

Determine the system function H(z) of the Chebyshev low pass digital filter with the specifications

α_P = 1dB ripple in passband $0 \le \omega \le 0.2\pi$
α_S = 1dB ripple in stopband $0.3\pi \le \omega \le \pi$

Using bilinear transformation (assume T=1sec)

Obtain the direct form I, direct form II, cascade, and parallel form realization for the system y(n)= -0.1y(n-1)+0.2y(n-2)+3x(n)+3.6x(n-1)+0.6x(n-2)

Apply Bilinear Transformation to H(s) =2/(S+2) (S+3) with T=0.1 sec.

Design a band pass filter which approximates the ideal filter with cut off frequencies at 0.2 rad/sec and 0.3 rad/sec. The filter order is M = 7. Use the Hanning window function

UNIT V
DIGITAL SIGNAL PROCESSORS

Draw the block diagram of Hardware architecture and explain each terms.

Discuss the advantages and disadvantages of VLIW architecture.

Describe the following things (i) Memory mapped register addressing (ii) Circular addressing mode (iii) Auxiliary registers

Describe the various addressing modes of a digital signal processor

Discuss in detail the different phases of pipelining

Describe short notes on parallel logic unit and circular registers

Explain about Von Neumann, Harvard architecture and modified Harvard architecture for the computer.

Describe how convolution is performed using a single MAC unit.

What is MAC unit? Describe its functions.

Summarize about pipelining in DSP.

Discuss the addressing modes used in programmable DSP’s

Discuss the architecture of TMS320C50 with a neat diagram

Discuss the Architectural details and features of a DSP processor.

Describe notes on commercial processors

Discuss about bit reversed addressing mode

MEASUREMENTS AND INSTRUMENTATION

UNIT-I INTRODUCTION

Give the methods of obtaining experimental data.

Direct method
Indirect method

List out the dynamic characteristics of any measurement system.

Step change
Linear change
Sinusoidal change
Speed of response
Lag
Fidelity
Dynamic error.

What are the types of errors in measurement?

Instrumental errors, Limiting errors, Environmental errors.

4. What are the static characteristics of an instrument?

The static characteristics of an instrument are considered for instruments which are used to measure an unvarying process condition. All the static performance characteristics are obtained by one form or another o a process called calibration.

5. What is a standard?

A standard is a physical representation of a unit of measurement. A known accurate measure of physical quantity is termed as standard.

Types:

International standards
Primary standards
Secondary standards
Working standards.

6. What is primary sensing element?

The primary sensing element is called as transducer that senses and converts the desired input to a more convenient and practicable form to be handled by the measurement system.

7. What is calibration?

Calibration is the process of checking the accuracy of instrument by comparing the instrument reading with a standard or against a similar meter of known accuracy. It is also defined as a marking the scale of an instrument.

8. Define arithmetic mean.

The average value, or arithmetic mean value, is the most probable value obtained from a series of reading of a given quantity. As a general rule, the more readings, the more closely the computer average represents the most probable value. The average value X is calculated by taking the sum of all the reading and dividing by the number of readings, so that

$$X=\frac{\sum xi}{n} = \frac{x1+x2+x3+\ldots+xn}{n}$$

Where X=the average value or arithmetic mean.

Xi = the value of the i^{th} reading.

N = the number of readings.

9. Define variance.

The variance is the mean square deviation, which is the same as standard deviation, except that square root is not extracted.

$$V = (SD)^2$$

$$= \frac{d2 + d_2^2 + d_3^2 + \dots d_n^2}{N}$$

$$= \frac{\sum d^2}{n}$$

10. Define standard deviation.

The standard deviation or root mean square deviation of a sample is both mathematically more convenient and statistically more meaningful for analyzing grouped data than is the average deviation. By definition, the standard deviation of a sample is given by

$$S = \sqrt{\frac{\sum (X - x_i)^2}{n}} = \sqrt{\frac{\sum d_1^2}{n}}$$

11. Define average deviation.

The mean or average is a measure of how much the data is dispersed, or varies from the average value. The mean D is calculated by adding all the absolute values of deviations of a set of measured values and dividing this sum by the number of observation 'n', so that

$$\overline{D} = \frac{|d| + |d_2| + \dots |d_n|}{n}$$

$$= \frac{\sum |d_i|}{n}$$

Define the terms précising and sensitivity.

Precision:

It is a measure of the consistency or repeatability of a series (successive) of measurements. Although accuracy implies precision, precision does not necessarily accuracy. A precise instrument can be very inaccurate. The precision of a given measurement can be given by

$$\text{Precision} = 1 - \left| \frac{x_i - \overline{X}}{x_i} \right|$$

Where x_i = the value of the ith measurement

X=the ~~av~~erage value of a measurements.

Sensitivity:

It is a measure of the change in reading of an instrument for a given change in the measured quantity.

13. Define static error.

The static error of a measuring instrument is the numerical difference between the value of a quantity and its value as obtained by measurement, i.e., repeated measurement of the same quantity gives different indications.

Distinguish re-predictability and repeatability.

Reproducibility:

It is defined as the degrees of closeness with which a given value may be repeatedly measured. It is specified in terms of unit for a given period of time. Perfect reproducibility means that the instrument has no drift.

Repeatability:

It is defined as variation of scale reading and is random in nature.

15. Define dynamic response of an instrument.

The behaviour of the instrument when inputs very with time (i.e., inputs are dynamic in nature) and so does the output, is called dynamic response of an instrument or system.

What are the different calibration methodologies?

Primary calibration

Secondary calibration

- Direct calibration
- Indirect calibration

Define limiting errors.

Instruments having analog meters are usually guaranteed to be accurate within certain percentage limits, called limiting errors or guarantee errors.

18. Define median.

The middle value of a set of an odd number of readings, if variables are arranged in numerical order, is called median.

Mention any four static characteristics of measurement instruments.

Accuracy.

Precision

Sensitivity

Resolution

Error

20 What is the significance of calibration?

All measuring instruments are to prove themselves their ability to measure reliably and accurately. For this, the results of measurement are to be compared with higher standards which are traceable to national or international standards.

The calibration of a measuring instrument means introducing an accurately known sample of the variable that is to be measured and then observing the system’s response.

21. Why do instruments to be calibrated?

Instruments must be calibrated since it gives the opportunity to check the instrument against a known standard and subsequently to find errors and accuracy.

22. Define the term, accuracy.

It is a measure of the closeness with which an instrument measures the true value of a quantity.

List the main functional elements used in most of the measurement systems.

Primary sensing element.
Variable conversion element.
Variable manipulation element.
Data transmission element.
Data presentation element.

Distinguish between accuracy and precision.

Accuracy	Precision
Accuracy refers to degree of Closeness of the measured value to the true value.	Precision refers to degree of agreement among group of Readings.
Accuracy gives the maximum error that is maximum departure of the final result from its true value.	Precision of a measuring system gives its capability to reproduce a certain reading with a given Accuracy

UNIT-II ELECTRICAL AND ELECTRONIC INSTRUMENTS

1. What is the working principle of wattmeter employed in measuring equipment?

A wattmeter consists of two coils namely, current coil and pressure coil. The current coil is connected in series with the load and it will measures the current flowing through the load whereas the voltage coil is connected across the load and it will measures voltage across the load. The deflecting torque of the moving coil (pressure coil) is directly proportional to the current flowing through the load, voltage across the load and the power factor of the load

Single phase power, $P=VI \cos \Phi$

Where, $V \rightarrow$ Voltage across the load.

$I \rightarrow$ Current flowing through the load

$\cos\Phi \rightarrow$ Power factor of the load.

How are the analog instruments classified on the basis of method used for comparing the known quantity?

Electrostatic type instruments.
Electromagnetic type instruments.
Instruments using magnetic effect.
Instruments using heating effect.
Instruments using Hall Effect.

Give the advantages of moving iron meters.

The same instrument can be universally used for both A,C and D,C measurement.

Torque/weight ratio is very high and hence errors due to friction are very small.

As the single moving system can be used for wide range of measurement, this instrument, are cheaper than other types of instruments.

These instruments are robust and simple in construction as there are no current carrying moving parts.

These instruments are highly accurate.

They are available with 240˚circular scales.

What are the different methods of measurement of frequency in the power frequency range?

- Mechanical resonance type/vibrating reed type frequency meter.
- Electrical resonance type/ferrodynamic type frequency meter.

Weston type frequency meter.

Why it is necessary to make the potential coil circuit purely resistive in wattmeters?

The inductance of the potential coil can cause error in wattmeter measurement. Hence the inductance of the potential coil is compensated by a capacitance thus making the potential coil circuit purely resistive in nature.

What are the advantages of digital instruments over analog instruments?

Highly accurate reading can be taken

Better resolution

High input impedance

Digital display eliminates observational errors, interpolation errors and parallax errors committed by operators.

Reading speed is very high.

Digital output can be directly recorded.

Portable.

They can be used for the measurement of quantities like current, impedance, capacitance, temperature, pressure etc.

How are resistors checked using digital multimeters?

For the measurement of various ranges of resistances, ohms convertor is used which is nothing but a low current source. A known current from the low current source is passed through the unknown resistance and the voltage drop across the resistance is measured. This voltage drop gives the direct indication of the unknown resistance.

8. What is auto ranging?

Auto ranging is the process of changing the range of the digital voltmeter for getting a reading with the optimum resolution under all the circumstances.

Define resolution of DVM? Resolution of DVM is given by,

$R=1/10^n$

Where, R=resolution of DVM.

N=number of full digits in a digital display

For a 4.5 digit display, n=4 and hence

$R=1/10^4=0.0001$ or 0.001%

10. What is volt ampere hour and watt-hour?

Volt ampere hour is the reactive power consumed by the load whereas watt-hour is the real power consumed by the load.

11. What is the need to evaluate phase-angle error is instrument transformers?

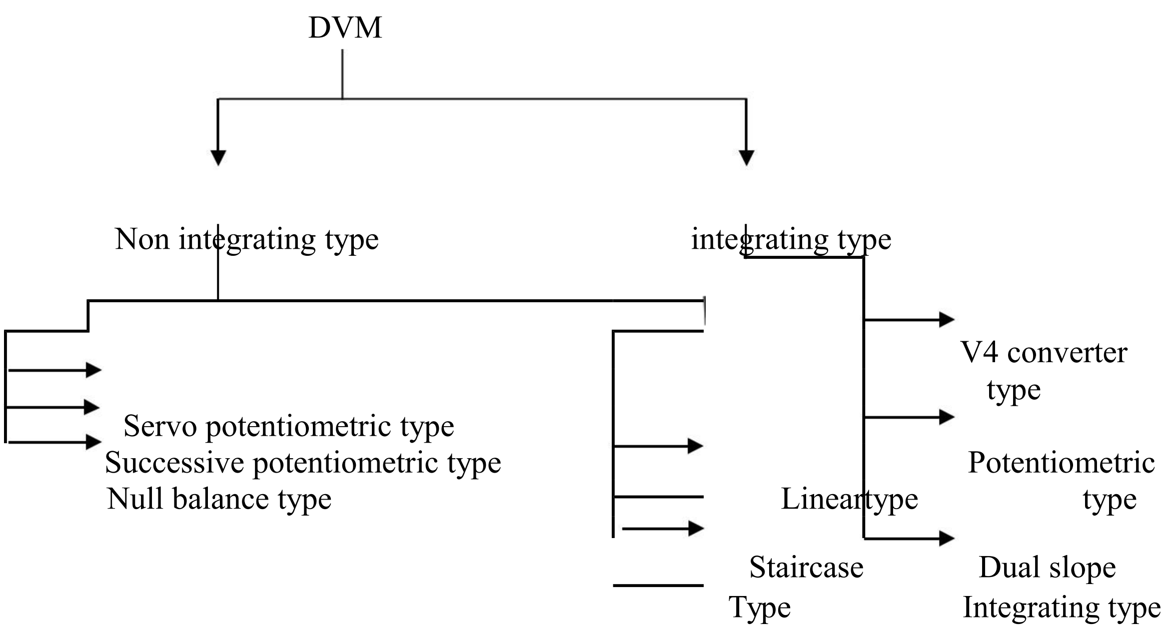
The main condition to be satisfied by the instrument transformer is, that the phase angle of the secondary (V or I) must be displaced exactly by 180˚ from that of primary Parameter (V or I). If this condition is not satisfied, heavy phase angel error will occur in the measurement using instrument transformer, Hence this phase angle error should be reduced to minimum and its value should be evaluated and included in the measuring quantity to get highly accurate measurement.

12. What is the purpose of instrument transformers?

The transformers used in conjunction with the measuring instruments for measuring very large values of current or voltage which cannot be directly measured are called instrument transformers. Two types of instrument transformers are

Current transformer and
Voltage transformer/potential transformer.

What are the various types of Digital Voltmeters?



14. Give the importance of iron loss measurement.

Iron loss is the loss of power due to hysteresis and eddy currents. Knowledge of iron loss in ferromagnetic materials is important to design an apparatus made up of ferromagnetic material.

15. What is the reason for using MI instruments on both A.C and D.C?

MI instruments can be used for both A.C and D.C measurement because, whatever may be the direction of the current through the coil in the instrument, the iron vanes get magnetized and there will be a force of attraction in the attraction type instrument will be a force of repulsion in the repulsion type instrument.

16. What is the precaution to be followed while using current transformer?

A transfer instrument is one that may be calibrated with a d.c source and then used to measure A.C without any modification. Example for transfer instrument is electrodynameter type instruments.

17. Why the PMMC instruments are not used for a.c measurements?

When the PMMC instruments are connected to a.c, the torque reverses as the current reverses and the pointer cannot follow the rapid reversals. Hence the deflection corresponding to mean torque is zero thus making the PMMC instruments not suitable for a.c. measurements.

18. State the principle of digital phase meter.

When two signals of same frequency, whose phase difference is to be measured are applied to the phase meter, the signal are converted to a square waveform without changing the phase relationship using two separate preamplifier and attenuator block. The converted square pulses are fed to the flip flop. The function of two flip flops is that, one flip flip enables the AND gate while the other disables it. The number of pulses allowed to pass during enabling and disabling the gate are counted which is proportional to the phase difference between the two signals.

19. Which torque is absent in energy meter? why?

In energy meter, there is no controlling torque, as the driving torque alone is enough to cause continuous revolution of the disc.

What are the sources of errors in D.C voltage measurement?

Sources of errors in D.C voltage measurements are

(i)Change in resistance with time and

(ii) Change in resistance with temperature.

21. Define ratio $Kn = \dfrac{\text{rated primary current}}{\text{Rated Secondary current}}$ For C.T and

$$Kn = \frac{\text{rated primary voltage}}{\text{Rated Secondary voltage}}$$ For P.T and

UNIT – III COMPARISION METHOD OF INSTRUMENTS

1. What are the characteristics of a DC amplifier?

It may need balanced differential inputs giving a high common mode rejection ratio (CMRR)

It should have an extremely good thermal and long term stability.

List the merits and demerits of a DC amplifier. It is easy to calibrate at low frequencies.

It is able to recover from an overload condition unlike it's AC counterpart.

Give the purpose of bridge circuits. What is the different type?

The bridge circuits are used in instrumentation systems for the measurement of resistance , inductance and capacitance.

Types:

DC type and

AC type.

What are the 2 types of Wheatstone bridge?
Null type bridge
Deflection Type Bridge.

What are the different types of AC bridges?
- AC bridge using push-pull transducers
- AC bridge with push-pull inductive transducers
- Inductive transducer Blumlein bridge
- Capacitive transducer Blumlein bridge

Define slew rate
Slew rate is defined as the maximum output voltage change per unit time.

List the requirements of an instrumentation amplifier
Low drift
High i/p impedance
High linearity
High CMRR
High noise rejection capability

Give few applications of instrumentation amplifier.
The instrumentation amplifier finds increasing application in the amplification of the output signals obtained from thermocouples, strain gauge bridge and biological electrode.

9. What is a filter?
A filter is often a frequency selective circuit that passes a specified band of frequencies and blocks or attenuated signal of frequencies outside this band.

List the different types of filters.
Analog or digital filters
Passive or active filters
Audio (AF) or radio (RF) filters.

Specify the advantages of an active filter
Gain and frequency adjustment flexibility
No loading problem
Low cost

What is frequency scaling?
The procedure of converting a cutoff frequency to a new cutoff frequency is called frequency scaling.

13. What is quality factor?
The ratio of resonant frequency to bandwidth is known as the quality factor Q.

14. What is acquisition time of S/H circuit?
Acquisition time is the time required for the capacitor to charge up to the value of the input voltage after the switch is first started.

15. Define transformer Ratio Bridge.

Transformer Ratio Bridge uses a ratio transformer which is highly accurate and versatile. Instead of conventional bridges it can replaced and it act as a ideal transformer which is having following properties

No flux linkage

No core loss.

No resistance of winding.

16. What is mean by Electromagnetic interference?.

The interference caused by the electromagnetic waves is called Electromagnetic interference.

17. What is mean by common mode and series mode voltages?.

The common mode voltages are those voltages which appear on both sides of a signal into a common reference point which is generally ground.

The voltage is considered to be in series with transducer voltage is called series voltage.

18. Define input guarding.

The complete measuring or input circuit unit of a differential amplifier is placed inside a metallic guard.

19. What are all basic requirement of A.C potentiometers?

While comparing two voltages in A.C potentiometer it is necessary to measure the potentiometer voltage accurately as A.C reference is not available in the circuit.

At all the instants of time both the voltages being compared must be equal to magnitude and phase both.

Define sensitivity of Wheatstone bridge.

Sensitivity S=Deflection D/Current I

UNIT-IV STORAGE AND DISPLAY DEVICES

What are the different types of amplifiers used for CRO's?

Vertical amplifier 2. Horizontal amplifier

Give the principle of LCD type display device.

LCDs are passive type display devices used for display of numeric and alphanumeric character in dot-matrix and segmental display.

When the cell is not activated, the transmittive type cell simply transmits the light through the cell in the straight lines. In this condition, the cell will not appear bright.

When the cell is activated, the incident light is scattered forward, as the cell appears quite bright even under high intensity light conditions.

Write two advantages of LED on electronic displays.

Low power consumption.

Very fast action.

Very small size and weight.

Extremely long life.

State the features of ink-jet printers.

They can print from two or four pages per minute.
Resolution is about 360 dots per inch; therefore better printing quality is achieved.
The operating cost is quite low; the only needs replacement is the ink cartridge.
Colour ink jet printers have flour ink nozzles with colour cyan, magneter, and yellow.

What are the various methods or recording data?

Direct recording.
Frequency modulated (FM recording).
Pulse duration modulation recording.

Differentiate between LED and LCD.

LED	LCD
Consumes more power (in.mW)	Consumes less power (in mW).
Capable of generating its own light.	Requires an external of internal light source.
High cost.	Low cost.
More life time.	Less life time.
Colour depends on the materials	Monochrome in nature.

In what ways line printer are advantageous over dot matrix printer?

Prints one line at a time.
Printing speed is better.
Printing quality is better.

What are the different types of magnetic recording?

Direct recording.
Frequency Modulated (FM) recording.
Pulse duration modulation recording.

What are the different materials used in LED? Also name the colours emitted.

Materials:

Gallium Arsenide phosphate
Gallium Arsenide (GaAs)
Gallium Phosphide (GaP)

Colours emitted:

Infrared, red, yellow, green.

10. What are data loggers?

The data loggers are used to automatically make a record of the readings of instruments located at different parts of the plant.

11. What are the functions of a data logger?

The data logger is used to automatically make a record of the readings of instruments located at different parts of the plant.

The main function is to measure electrical output from virtually any type of transducer and log the value automatically.

List out the advantages and disadvantages of LCD.

Advantages:

Low cost.
Low power consumption.
It requires very low voltage.

Life time is very less compared with LED.
Response time is more compared with LED.
The occupy large area.
Reliability is quite low.

What are the advantages of LCD over LED?

Low cost
Low power consumption.
It requires very low voltage.

What is the sweeper in oscilloscope?

Triggered sweep.
Delayed sweep.

What is the basic operating of digital tape recording?

Digital data can be recorded and stored in magnetic tapes using a variety of techniques. The basic principle used is to modulate the digital data in some form and then record this modulated data in the tape.

Discuss the advantages and disadvantages of PDM recording.

Advantages:

Capable to record information simultaneously from a large number of channels.
Has high s/w ratio
Has high accuracy due the fact that it can be SD/f calibreated.

Limited frequency response.
Highly complicated electronic circuitry and therefore the reliability of system is low.

Define the deflection sensitivity of CRT.

The deflection sensitivity of a CRT is defined as the deflection of the screen per unit deflection voltage.

List out the main parts of cathode ray tube.

Electron gun assembly.
Deflection plates assemble.
Fluorescent screen.
Glass envelope.

What is recorder? How they are classified?

A recorder is a device that records electrical and non-electrical quantities as a function of time classifications.

-Analog recorders
Graphic recorders
Magnetic tape recorder
Oscilloscope recorder
-Digital recorder

20. What is magnetic recorder?

Magnetic recorder is a recorder which records analog data in such a manner that they can be retrieved or reproduced in electrical form again.

What are the basic components of a tape recorder?

Recording head.2. Magnetic tape.
Reproducing head. 4. Tape transport mechanism.
Conditioning devices.

UNIT-V TRANSDUCERS AND DATA ACQUISITION SYSTEMS

1. What is transducer?

A transducer is defined as a device that receives energy from one system and transmits it to another, often in a different form. A transducer, in general form, may be defined as a device which energy from one form to another.

Mention some advantages of electrical transducers.

Electrical amplification and attenuation can be easily done.
The effects of friction are minimized.
Mass-intertia effects are minimized.
Very small power is required for controlling the electrical or electronic system.
The electrical output can be amplified to any desired level.

How the transducers are classified?

on the basis of transduction form used.
As primary and secondary transducers.
As active and passive transducers.
As analog and digital transducers.
as transducers and inverse transducers.

What is an active transducer?

An active transducer generates an electrical signal directly in response to the physical parameter and does not require an external power source for its operation. Such transducers draw energy from the system under measurement. Active transducers are also called self generating type transducers.

5. Mention some example for active transducer.

Typical examples of active transducers are tacho-generators used for measurement of angularvelocity, thermocouples used for measurement of temperature, piezoelectric crystal used for measurement of force.

6. What is a passive transducer?

Transducers, in which electrical parameters i.e. resistance, inductance or capacitance changes with the change in input signal, are called the passive transducers. These transducers require external power source for energy conversion. These transducers may draw some energy from the system under measurement. Typical examples are strain gauges, thermistors etc.

7. What is an analog transducer?

Analog transducer converts input signal into output signal, which is a continuous function of time such as thermistors, strain gauge, LVDT, thermocouple.

8. What is a digital transducer?

Digital transducer converts input signal into the output signal. Which is in the form of pulses i.e.it gives discrete output.

9. What is an inverse transducer?

It is defined as a device which converts an electrical quantity into a non-electrical quantity.

Give the factors to be considered in selecting a transducer.

Operating range
Sensitivity.
Electrical output characteristics
Environmental conditions.
Errors
Accuracy

Define strain gauges.

The strain guage is an example of a passive transducer that uses the variation in electrical resistance in wires to sense the strain produced by a force on the wires.

12. Define gauge factor.

The gauge factor is defined as the ratio of per unit change in resistance to per unit change in length.

Gauge factor $Gf = \frac{R/R}{L/L}$

13. Mention the different types of strain gauges.

The strain gauges are mainly of four types namely
Wire strain gauges
Foil strain gauges
Thin film strain gauges
Semiconductor strain gauges

What are thermistors?

Thermistors (thermally sensitive res IS TOR) are non-metallic resistors (semiconductor material) made by sintering mixtures of metallic oxides such as manganese, nickel, cobalt, copper and uranium. Thermistors have a Negative temperature coefficient (NTC) i.e.,resistance decreases as temperature rises.

Mention some applications of thermistors.
Applications of Thermistors:

Measurement of temperature.
Control of temperature.

Temperature compensation.
Measurement of power at high frequencies.
Measurement of thermal conductivity, level, flow, pressure of liquids, vacuum, composition of gases.

What is an inductive transducer?

An inductive transducer is a device that converts physical motion into a change in inductance. Transducers on the variable inductance type work upon one of the following principles.

Number of turns
Geometric configuration
Permeability of the magnetic material or magnetic circuits.

Mention some advantages of LVDT.

The output of LVDT is practically linear for displacements up to 5 mm. The LVDTs have a very high range of measurement of displacement.

LVDT has infinite resolution as it gives step less output and it has got no mechanical element to change output in discrete steps. Now –a-days transducers are available with the resolution up to 1 micron.

LVDT has high sensitivity. It usually varies from 10 mv/mm to 40 v/mm.

The LVDT gives a high output and many a times there is no need for amplification.

Mention the applications of LVDT.

LVDTs are used to measure
Displacement
Force
Weight
Pressure
Position

What is piezoelectric effect?

A piezoelectric material is one in which an electric potential appears across certain surfaces of a crystal if the dimensions of the crystal are changes by the application of a mechanical force. This potential is produced by the displacement of charges. The effect is reversible also i.e. if a varying potential is applied to the proper axis of the crystal, it will change the dimensions of the crystal thereby deforming it. This phenomenon is known as piezoelectric effect.

20. What are the materials used for piezoelectric transducers?

Common piezoelectric materials include Rochelle salt, ammonium dihydrogen phosphate (ADP), quartz and ceramics made with barium titanate, dipotassium tartrate, potassium dihydrogen phosphate and lithium sulfate are used in real applications.

List out the applications of DAS.

Aerospace application.
Biomedical field.
Telemetry industries.
Industries.

Mention any four types of analog to digital converter.

Direct type

Indirect type

Flash (comparator) type converter.

Staircase type converter.

Tracking or servo converter.

Successive approximation type converter.

Indirect types are classified as:

Charge balancing analog to digital converter.

Dual slope analog to digital converter.

What are the types of digital to analog converters?

1. Binary weighted resistors DAC 2. R-2Rladder 3. Inverted R-2R ladder

24. What is smart sensor?

Smart sensors are sensors with integrated electronics that can perform one or more of the following functions.

Logic functions,

Two-way communication,

Make decisions.

25. Mention some applications of smart sensor.

Smart sensor also enhances the following applications:

Self calibration

Computation

Communication

Multisensing

******

16 MARK QUESTIONS

UNIT I

Explain the block diagram and functional elements of measurement system with neat diagram.

Classify and explain the different types of standards and errors of measurements.

With suitable illustrations, elaborate the significance of calibration.

Write a technical note on static and dynamic characteristics of instrumentation systems.

(i) Define accuracy and reproducibility of an instrument and explain.
Describe the primary and secondary standards in instruments.

A circuit was tuned for resonance by eight different students and the values of resonant frequency in kHZ were recorded as 532,548,543,535,531,543,and 536.Calculate (i) Arithmetic mean (ii) Deviation (iii) Average deviation (iv) Standard deviation.

Describe the various modes of statistical evaluation of measurement data.

What are the three categories of systematic errors in the instrument and explain in detail.

Explain the normal or Gaussian curve of errors in the study of random effects.

Describe the functional elements of an instrument with its block diagram. Also illustrate them with pressure gauge, pressure thermometer and D'Arsonval galvanometer.

With circuit and phasor diagram, explain the working of single phase ac energy meter.

(i) Obtain B-H curve of a ring specimen.
Describe how to obtain iron loss of a ring specimen using wattmeter.

With a neat sketch explain the working principle of PMMC instrument. Also derive the expression for deflection.

Explain the construction and its working principle of electrodynamometer type wattmeter.

Discuss in detail, about the working principle and characteristics of CT with its phasor diagram.

Write short notes on (i) Current transformer (ii) Weston frequency meter.

Explain the functions three phase wattmeter.

Draw and explain the circuit diagram of digital frequency meter.

Explain with a neat diagram the working of successive approximation type digital voltmeter.

Describe the construction functioning of mechanical type (vibrating reed type) frequency meter.

UNIT III

With a circuit diagram, explain the principle of operation of Duo-range DC Potentiometer.

Draw a neat diagram of Kelvin double bridge and explain how to measure low resistance.

Obtain an expression for measurement of inductance using Maxwell's Inductance bridge with a neat circuit diagram.

Explain the theory and working principle of Wheatstone's bridge . Derive the relation for finding unknown resistance.

Describe any one method for the measurement of high resistance.

Explain how the inductance is measured in terms of known capacitance using Maxwell's bridge. Derive the condition for balance.

Explain the following (i) Grounding techniques (ii) Causes of electromagnetic interferences in measurements.

Explain how wein bridge used for frequency measurement wit neat circuit diagram. Also derive the suitable expression.

Discuss the effects of electrostatic and electromagnetic interference in instruments.

Explain the construction of Anderson's bridge. Derive the unknown quantities at balance conditions. Also write it's advantages and disadvantages.

Describe the construction and working of magnetic tape recorder.

With a block diagram, explain the working of digital CRO.

Draw a neat block diagram of X-Y recorder and describe its working.

Explain the principle and working of CRT display with a neat diagram.

Compare and contrast the construction, working principle and applications of LED and LCDs.

Write a detailed note on Data loggers. Explain how they differ from Data Acquisition systems.

Write a detailed technical note on dot matrix display.

Explain in detail, how the data is stored in a magnetic disk and tape?

Describe the pulse duration modulation as used in magnetic tape recording and explain its merits and demerits.

Describe the mesh storage technique used in storage oscilloscope.

UNIT V

Describe the construction and working of potentiometer type resistance transducer for measuring linear displacement.

Explain the working of D/A converter with a neat diagram.

What is called Piezo electric transducer? Explain its working with a diagram.

Explain how to measure pressure using capacitive type transducer.

Describe in detail, the working principle of capacitive microphones.

Write a detailed note on smart sensors. Explain also the various built in features of them compared to conventional sensors.

Detail the construction and working principle of linear variable differential transformer (LVDT).

Discuss the successive approximation type ADC with its characteristics.

Explain the principle of the following transducers (i) Thermistors (ii) LVDT

How is differential output taken from an inductive transducer? Discuss in detail.