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Your Personal Exams Guide

NDA	CDS	SSC CGL	CBSE UGC NET
IAS	SSC CHSL	CTET	MPSC
AFCAT	CSIR UDC NET	IBPS PO	UP POLICE
SSC MTS	SBI PO	BPSC	UPTET
IBPS RRB	IBPS CLERK	IES	UPSC CAPF
SSC Stenogr..	RRB NTPC	SSC GD	RBI GRADE B
RBI Assistant	DSSSB		

GATE EE 2011 Question Paper (13-Feb-2011) (Shift 2)

Total Time: 3 Hour **Total Marks:** 100

Instructions

Sl No.	Section Name	No. of Question	Maximum Marks
1	General Aptitude	10	15
2	Electrical Engineering	55	85

1.) A total of 180 minutes is allotted for the examination.

2.) The server will set your clock for you. In the top right corner of your screen, a countdown timer will display the remaining time for you to complete the exam. Once the timer reaches zero, the examination will end automatically. The paper need not be submitted when your timer reaches zero.

3.) There will, however, be sectional timing for this exam. You will have to complete each section within the specified time limit. Before moving on to the next section, you must complete the current one within the time limits.

General Aptitude

1. There are two candidates P and Q in an election. During the campaign, 40% of the voters promised to vote for P, and rest for Q. However, on the day of election 15% of the voters went back on their promise to vote for P and instead voted for Q. 25% of the voters went back on their promise to vote for Q and instead voted for P. Suppose, P lost by 2 votes, then what was the total number of voters? **(+1, -0.33)**

a. 100

b. 110

c. 90

d. 95

---

2. **Choose the most appropriate word from the options given below to complete the following sentence:** **(+1, -0.33)**

It was her view that the country's problems had been _______ by foreign technocrats, so that to invite them to come back would be counter-productive.

a. identified

b. ascertained

c. exacerbated

d. analysed

---

3. **Choose the word from the options given below that is most nearly opposite in meaning to the given word:** **(+1, -0.33)**

Frequency

a. periodicity

b. rarity

c. gradualness

d. persistency

---

4. Choose the most appropriate word from the options given below to complete the following sentence: **(+1, -0.33)**

Under ethical guidelines recently adopted by the Indian Medical Association, human genes are to be manipulated only to correct diseases for which _______ treatments are unsatisfactory.

a. similar

b. most

c. uncommon

d. available

---

5. The question below consists of a pair of related words followed by four pairs of words. Select the pair that best expresses the relation in the original pair: **(+1, -0.33)**

Gladiator : Arena

a. Dancer : Stage

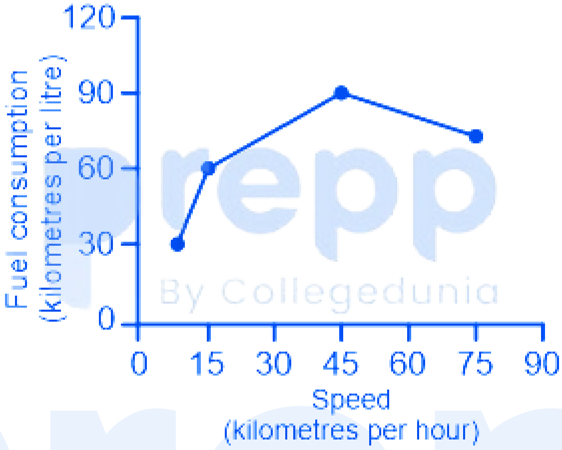
b. Commuter : Train

c. Teacher : Classroom

d. Lawyer : Courtroom

---

6. The fuel consumed by a motorcycle during a journey while traveling at various speeds a indicated in the graph below. **(+2, -0.66)**



The distances covered during four laps of the journey are listed in the table below:

Lap	Distance (kilometres)	Average speed (kilometres per hour)
P	15	15
Q	75	45
R	40	75
S	10	10

From the given data, we can conclude that the fuel consumed per kilometre was least during the lap

a. P

b. Q

c. R

d. S

---

7. Three friends, R, S and T shared toffee from a bowl. R took 1/3rd of the toffees, but returned four to the bowl. S took $1/4^{th}$ of what was left but returned three toffees to the bowl. T took half of the remainder but returned two back into the bowl. If the bowl had 17 toffees left, how many toffees were originally there in the bowl? **(+2, -0.66)**

a. 38

b. 31

c. 48

d. 41

---

8. Given that f(y) = |y|/y, and q is any non-zero real number, the value of |f(q) - f(-q)| is **(+2, -0.66)**

a. 0

b. -1

c. 1

d. 2

---

9. The sum of n terms of the series 4 + 44 + 444 + ... is **(+2, -0.66)**

a. (4/81) $[10^{n+1} - 9n - 1]$

b. (4/81) $[10^{n-1} - 9n - 1]$

c. (4/81) $[10^{n+1} - 9n - 10]$

d. (4/81) $[10^{n} - 9n - 10]$

---

10. The horse has played a little known but very important role in the field of medicine. Horses were injected with toxins of diseases until their blood built up immunities. Then a serum was made from their blood. Serums to fight with diphtheria and tetanus were developed this way. **(+2, -0.66)**

It can be inferred from the passage, that horses were

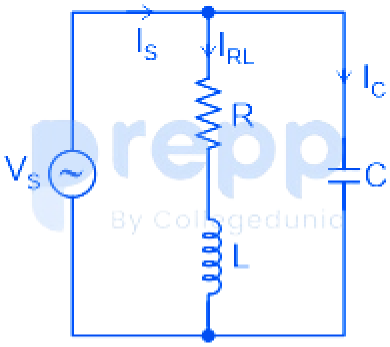
a. given immunity to diseases

b. generally quite immune to diseases

c. given medicines to fight toxins

d. given diphtheria and tetanus serums

Electrical Engineering



11. (+1, -0.33)

$V_S = 1\angle 0°$ C

$I_S = \sqrt{2} \angle \pi/4$

$I_{RL} = \sqrt{2} \angle -\pi/4$

The current I_C in the figure above is

a. -j 2 A

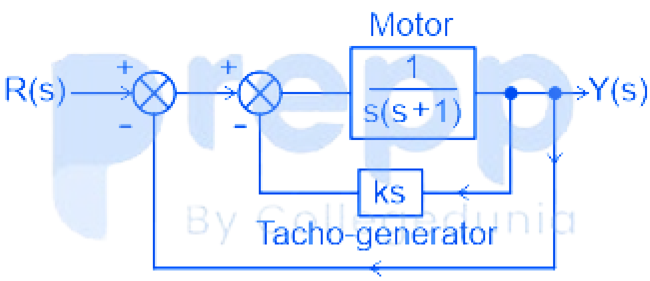
b. $-j\frac{1}{\sqrt{2}}\text{A}$

c. $+j\frac{1}{\sqrt{2}}\text{A}$

d. +j 2 A

---

12. A two-position control system is shown below. (+1, -0.33)



The gain k of the Tacho-generator influences mainly the

a. Peak overshoot

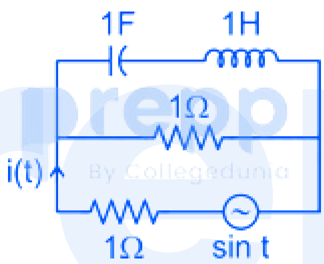
b. natural frequency of oscillation

c. phase shift of the closed loop transfer function at very low frequencies ($\omega \to 0$)

d. phase shift of the closed loop transfer function at very high frequencies ($\omega \to \infty$)

---

13. The RMS value of the current i(t) in the circuit shown below is **(+1, -0.33)**



a. $\frac{1}{2}\,A$

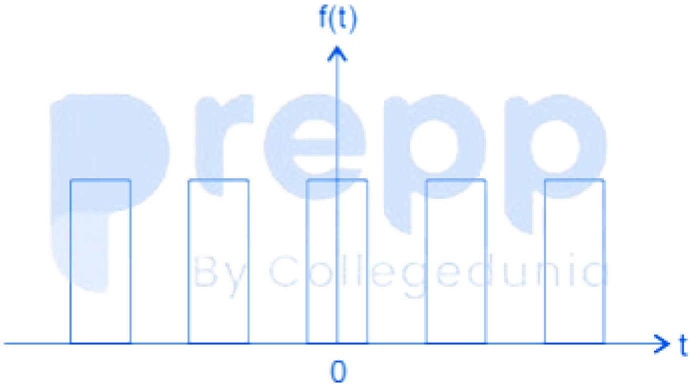
b. $\frac{1}{\sqrt{2}}\,A$

c. 1 A

d. $\sqrt{2}\,A$

---

14. The Fourier series expansion \(f\left(t \right) = {a_0} + \mathop \sum \limits_{n = 1}^\infty {a_n}\cos n\omega t + {b_n}\sin n\omega t\) of the periodic single shown below will contain the following nonzero terms **(+1, -0.33)**



a. a_0 and b_n, n = 1, 3, 5,......∞

b. a_0 and a_n, n = 1,2,3,..... ∞

c. a_0 a_n and b_n,n = 1, 2, 3,..... ∞

d. a_0 and a_n n = 1, 3, 5,......∞

---

15. A 4 point starter is used to start and control the speed of a **(+1, -0.33)**

a. DC shunt motor with armature resistance control

b. DC shunt motor with field weakening control

c. DC series motor

d. DC compound motor

---

16. A three phase, salient pole synchronous motor is connected to an infinite bus. It is operated at no load at normal excitation. The field excitation of the motor is first reduced to zero and then increased in the reverse direction gradually. Then the armature current **(+1, -0.33)**

a. increases continuously

b. first increases and then decreases steeply

c. first decreases and then increases steeply

d. remains constant

---

17. A nuclear power station of 500 MW capacity is located at 300 km away from a load center. Select the most suitable power evacuation transmission configuration among the following options **(+1, -0.33)**

a. 

b. 

c. 

d. 

---

18. The two vectors [1, 1, 1] and [1, a,a^2], where a = $\left(-\frac{1}{2}+j\frac{\sqrt{3}}{2}\right)$, are **(+1, -0.33)**

a. orthonormal

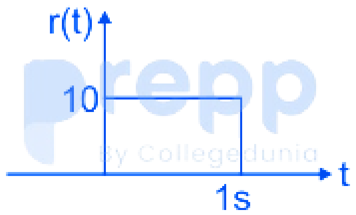
b. orthogonal

c. parallel

d. collinear

---

19. The steady state error of a unity feedback linear system for a unit step input is 0.1. The steady state error of the same system, for a pulse input r(t) having magnitude of 10 and a duration of one second as shown in the figure is **(+1, -0.33)**



a. 0

b. 0.1

c. 1

d. 10

---

20. Consider the following statements **(+1, -0.33)**

i) The compensating coil of a low power factor wattmeter compensates the effect of the impedance of the current coil

ii) The compensating coil of a low power factor wattmeter compensates the effect of the impedance of the voltage coil circuit

a. (i) is true but (ii) is false

b. (i) is false but (ii) is true

c. Both (i) and (ii) are true

d. Both (i) and (ii) are false

---

21. A low-pass filter with a cut-off frequency of 30 Hz is cascaded with a high pass filter with a cut-off frequency of 20 Hz. The resultant system of filters will function as **(+1, -0.33)**

a. an all-pass filter

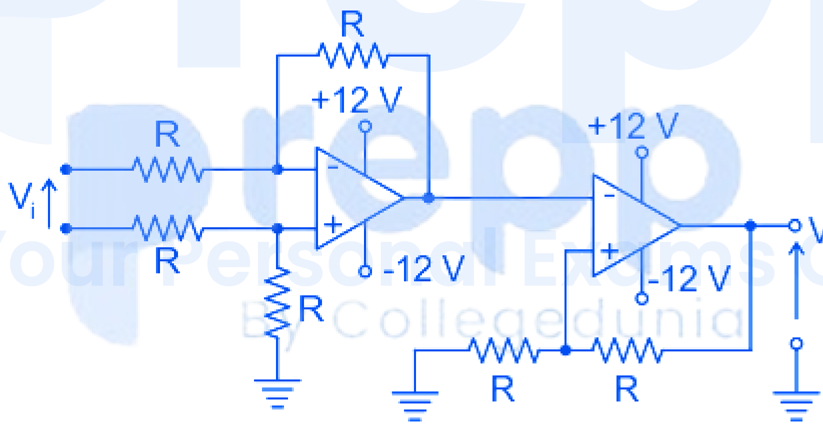
b. an all – stop filter

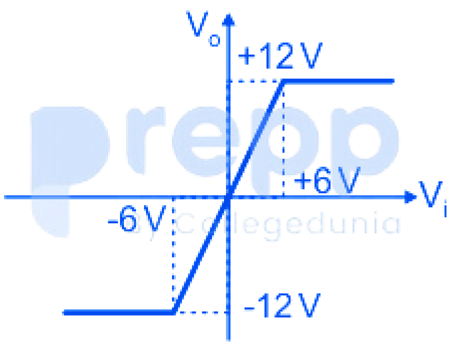
c. a band stop (band-reject) filter

d. a bandpass filter

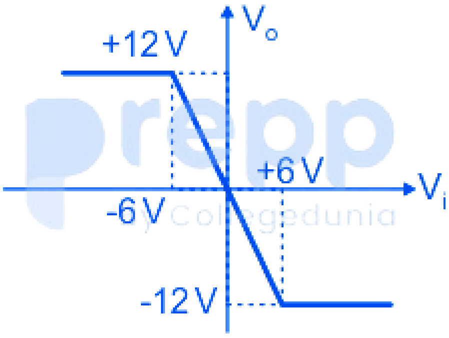
---

22. For the circuit shown below, the correct transfer characteristics is **(+1, -0.33)**

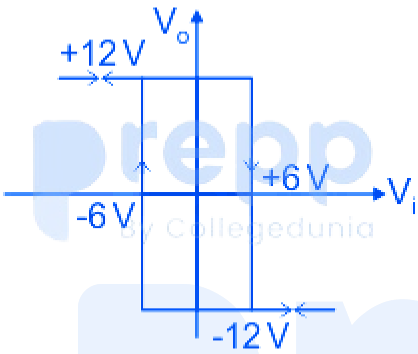


a. 

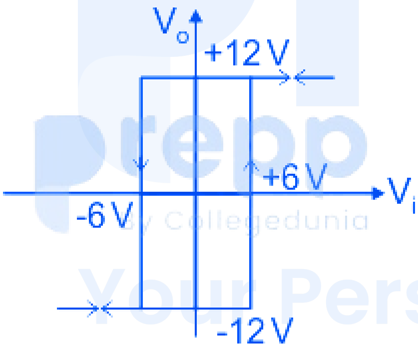
b.



c.



d.



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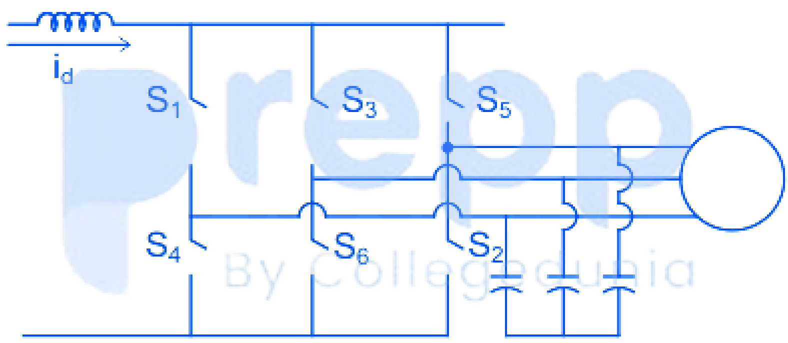
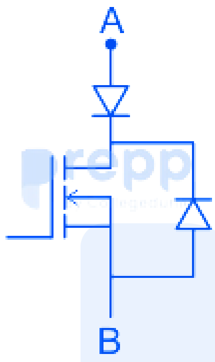
23. A 3 - ϕ CSI used for the speed control of an induction motor is to be realized using MOSFET switches as shown below. Switches S_1 to S_6 are identical switches. **(+1, -0.33)**

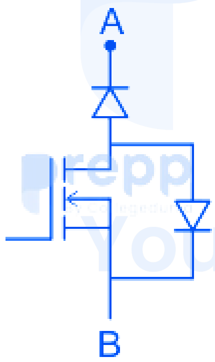


The proper configuration for realizing switches S_1 to S_6 is

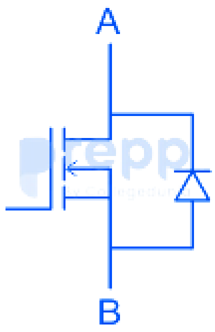
a.

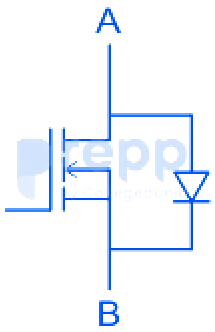


b.

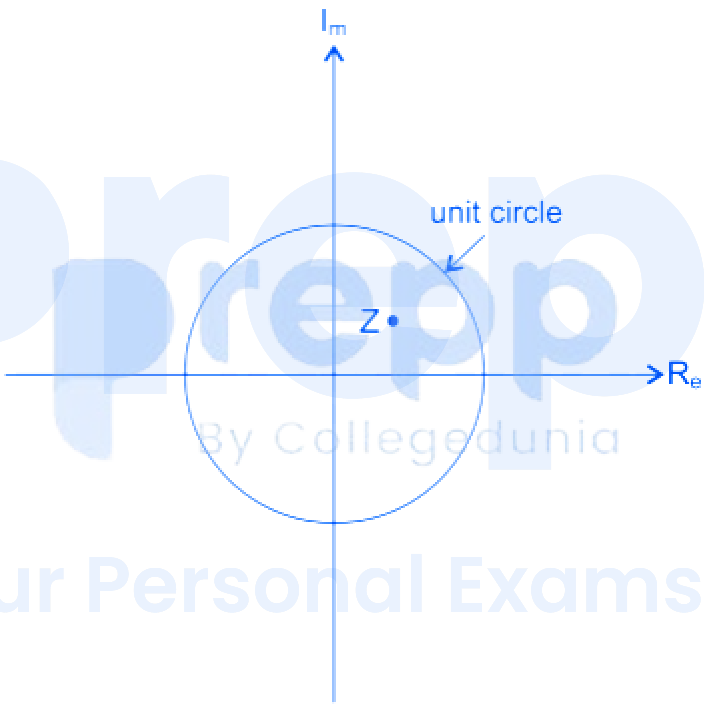


c.

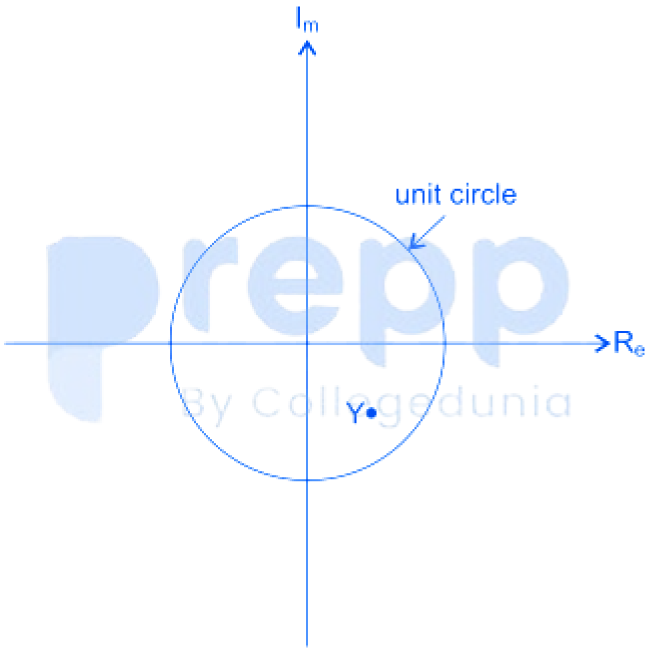


d. 

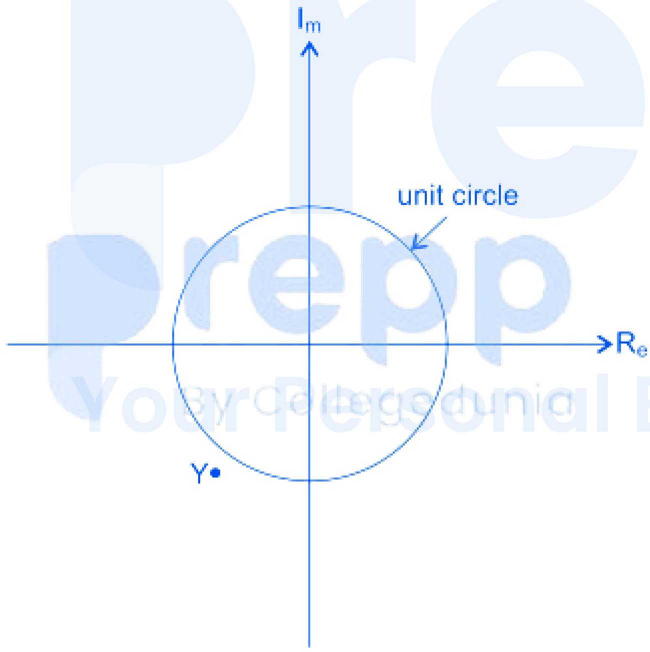
24. A point Z has been plotted in the complex plane, as shown in figure below. **(+1, -0.33)**



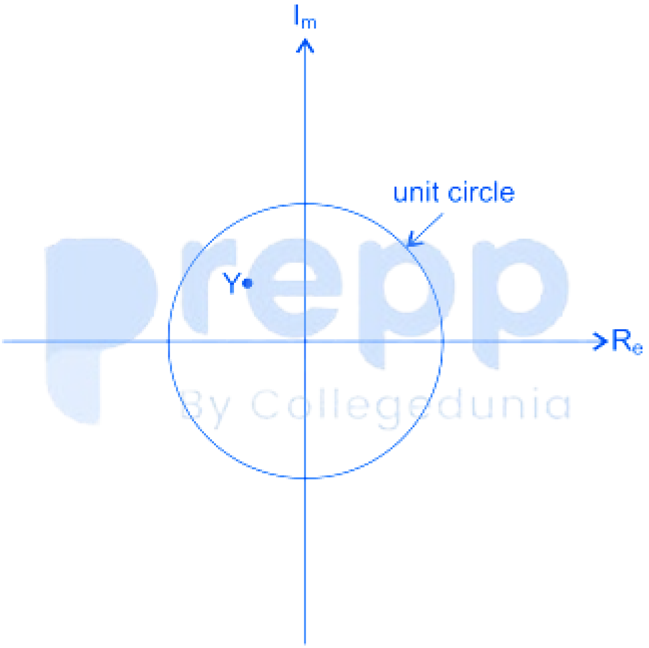
The plot of the complex number $Y = \frac{1}{Z}$ is



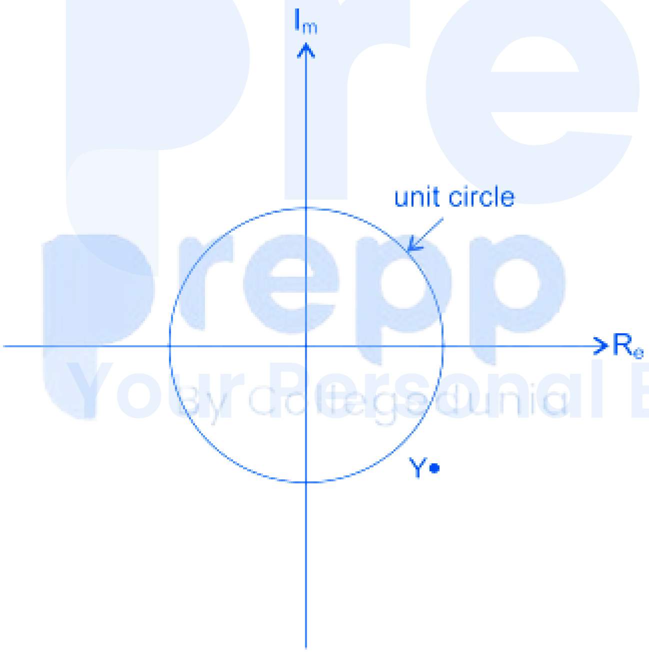
a.



b.



c.



d.

---

25. The voltage applied to a circuit is $100\sqrt{2}\cos(100\ \pi t)$ volts and the circuit draws a current of $10\sqrt{2}\sin(100\pi t + \pi/4)$ amperes. Taking the voltage as the reference phasor, the phasor representation of the current in amperes is **(+1, -0.33)**

a. $10\sqrt{2}\angle -\pi/4$

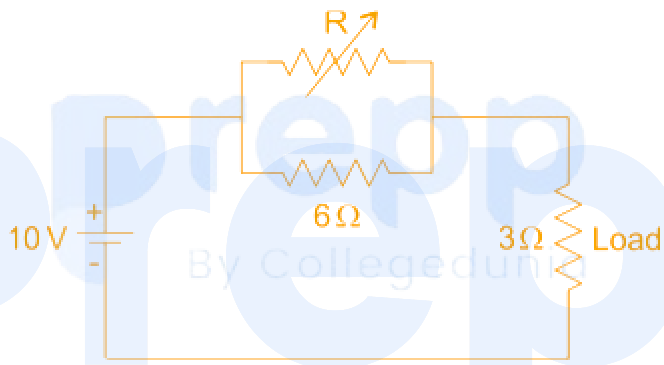
b. $10\angle -\pi/4$

c. $10\angle +\pi/4$

d. $10\sqrt{2}\angle +\pi/4$

---

26. In the circuit given below, the value of R required for the transfer of maximum power to the load resistance of 3 ohms (+1, -0.33)



a. zero

b. 3 Ω

c. 6 Ω

d. infinity

---

27. Given two continuous time signals x(t) = e^{-t} and y(t) = e^{-2t} which exist for t > 0. The convolution z(t) = x(t) * y(t) is: (+1, -0.33)

a. $e^{-t} - e^{-2t}$

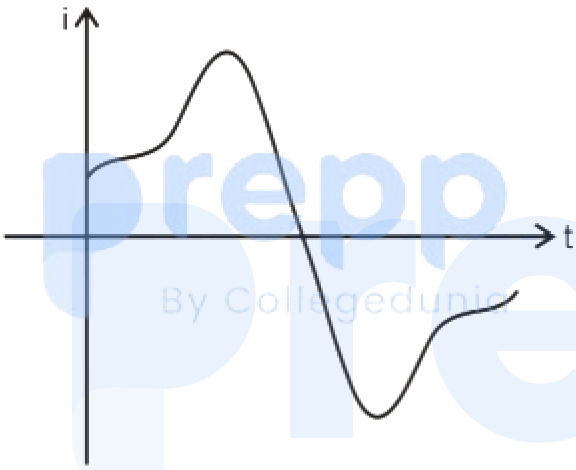
b. $e^{-t} - e^{2}$

c. $e^{-t} + e^{2t}$

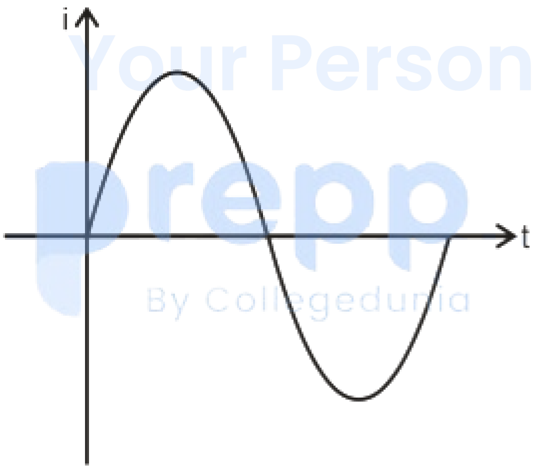
d. $e^{-t} + e^{-2t}$

---

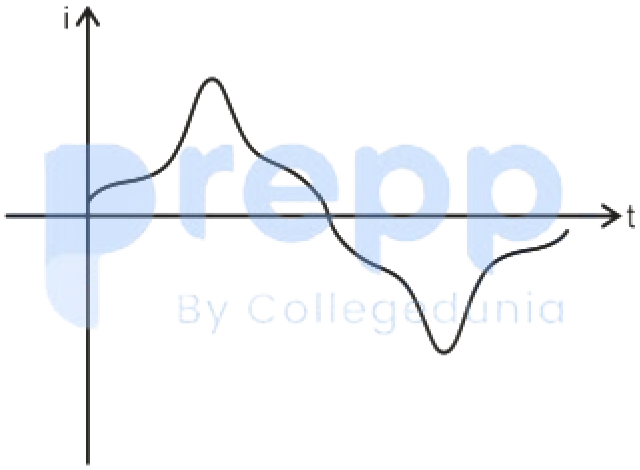
28. A single phase air core transformer fed from a single phase rated sinusoidal supply is operating at no load. The steady state magnetising current drawn by the transformer from supply will have waveform **(+1, -0.33)**

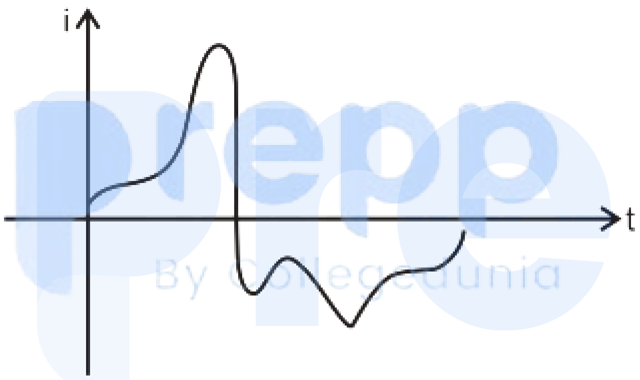


a.



b.

c. 

d. 

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29. A negative sequence relay is commonly used to protect (+1, -0.33)

a. An alternator

b. A transformer

c. A transmission line

d. A bus bar

---

30. For enhancing the power transmission in along EHV transmission line, the most preferred method is to connect a (+1, -0.33)

a. Series inductive compensator in the line

b. Shunt inductive compensator at the receiving end

c. Series capacitive compensator in the line

d. Shunt capacitive compensator at the sending end

---

31. An open loop system represented by the transfer function $G(s) = \frac{(s-1)}{(s+2)(s+3)}$ is **(+1, -0.33)**

a. Stable and of the minimum phase type

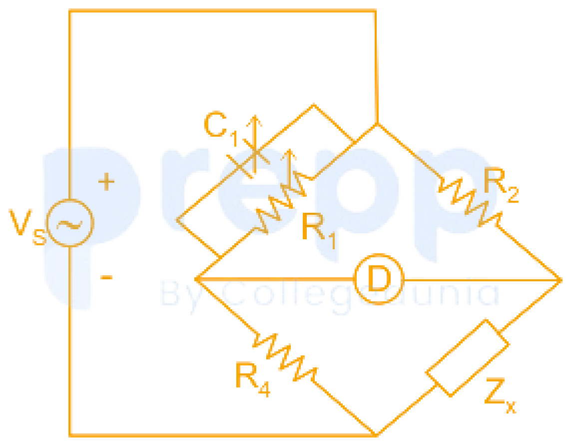
b. Stable and of the non-minimum phase type

c. Unstable and the minimum phase type

d. Unstable and of the non-minimum phase type

---

32. The bridge circuit shown in fig below is used for the measurement of an unknown element Z_x the bridge circuit is best suited when Z_x is a **(+1, -0.33)**



a. Low resistance

b. High resistance

c. Medium Q inductor

d. Lossy capacitor

---

33. A dual trace oscilloscope is set to operate in the Alternate mode. The control input of multiplexer used in circuit is fed with a signal having frequency equal to **(+1, -0.33)**

a. Twice the frequency of time base (sweep) oscillator

b. The highest frequency that the multiplexer can operate on

c. Half the frequency of time base (sweep) oscillator

d. The frequency of time base (sweep) oscillator

---

34. The output Y of the logic circuit given below is:- **(+1, -0.33)**

a. 1

b. 0

c. X

d. X□

---

35. Circuit turn off time of an SCR is defined as the time **(+1, -0.33)**

a. Taken by the SCR to turn off

b. Required for the SCR current to become zero.

c. For which the SCR is reverse biased by the commutation circuit.

d. For which the SCR is reverse biased to reduce its current below the holding current.

---

36. The matrix $[A] = \begin{bmatrix} 2 & 1 \\ 4 & -1 \end{bmatrix}$ is decomposed into a product of a lower triangular matrix [L] and an upper triangular matrix [U]. The properly decomposed [L] and [U] matrices respectively are **(+2, -0.66)**

a. $\begin{bmatrix} 1 & 0 \\ 4 & -1 \end{bmatrix}$ and $\begin{bmatrix} 1 & 1 \\ 0 & -2 \end{bmatrix}$

b. $\begin{bmatrix} 2 & 0 \\ 4 & -1 \end{bmatrix}$ and $\begin{bmatrix} 1 & 1 \\ 0 & 1 \end{bmatrix}$

c. $\begin{bmatrix} 1 & 0 \\ 4 & 1 \end{bmatrix}$ and $\begin{bmatrix} 2 & 1 \\ 0 & -1 \end{bmatrix}$

d. $\begin{bmatrix} 2 & 0 \\ 4 & -3 \end{bmatrix}$ and $\begin{bmatrix} 1 & 0.5 \\ 0 & 1 \end{bmatrix}$

---

37. The function $f(x) = 2x - x^2 + 3$ has **(+2, -0.66)**

a. Only maxima at x = 1

b. A maxima at x = 1 and minimum at x = -5

c. A maxima at x = 1 and minimum at x = 5

d. Only minimum at x = 5

---

38. A lossy capacitor C_x, rated for operation of 5 kV, 50 Hz is represented by an equivalent circuit with an ideal capacitor Cp in parallel with a resistor **(+2, -0.66)**

Rp. Cp is 0.102 μF; and R_p = 1.25 MΩ. The power loss, and tan δ, of this lossy capacitor when operating at the rated voltage are, respectively

a. 20 W and 0.04

b. 10 W and 0.04

c. 20 W and 0.025

d. 10 W and 0.025

---

39. Let the Laplace transform of a function f(t) which exists for t>0 be $F_1(s)$ and the Laplace transform of its delayed version f(t – τ) be $F_2(s)$. Let F_1 * (s) be the complex conjugate of $F_1(s)$ with the Laplace variable set s = σ + jω. If $G(s) = \frac{F_2(s)F_1*(s)}{|F_1(s)|^2}$, then the inverse Laplace transform of G(s) is **(+2, -0.66)**

a. An ideal impulse δ (t)

b. An ideal delayed impulse δ (t – τ)

c. An ideal step function u (t)

d. An ideal delayed step function u (t – τ)

---

40. A zero mean random signal is uniformly distributed between limits – a and + a and its mean square value is equal to its variance. Then the r.m.s value of the signal is **(+2, -0.66)**

a. $\frac{a}{\sqrt{3}}$

b. $\frac{a}{\sqrt{2}}$

c. $a\sqrt{2}$

d. $a\sqrt{3}$

---

41. A 220V, DC shunt motor is operating at a speed of 1440 rpm. The armature resistance is 1 and armature current is 10A. If the excitation of the machine is reduced by 10% the extra resistance to be put in the armature circuit to maintain the same speed and torque will be **(+2, -0.66)**

a. 1.79 Ω

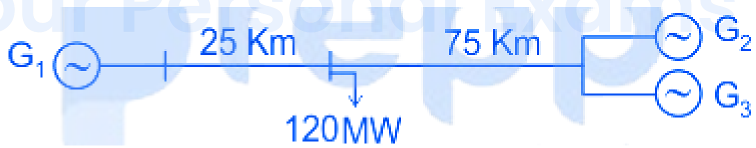
b. 2.1 Ω

c. 3.1 Ω

d. 18.9 Ω

---

42. A load centre of 120 MW derives power from two power stations connected by 220 kV transmission lines of 25 km and 75 km as shown in figure below. The three generators G_1, G_2 and G_3 are of 100 MW capacity each and have identical fuel cost characteristics. The minimum loss generation schedule for supplying the 120 MW load is? **(+2, -0.66)**



a. P_1 = 90 MW
P_2 = 15 MW
P_3 = 15 MW

b. P_1 = 80 MW
P_2 = 20 MW
P_3 = 20 MW

c. P_1 = 60 MW

P_2 = 30 MW

P_3 = 30 MW

d. P_1 = 40 MW

P_2 = 40 MW

P_3 = 40 MW

---

43. A portion of the main program to call a subroutine SUB in an 8085 environment is given below. **(+2, -0.66)**

```
          :
          :
          LXI     D, DISP
LP :      CALL    SUB
          :
          :
```

It is desired that control be returned to LP + DISP + 3 when the RET instruction is executed in the subroutine. The set of instructions that precede the RET instruction in the subroutine are:

a. POP D
DAD H
PUSH D

b. POP H
DAD D
INX H
INX H
INX H
PUSH H

c. POP H
DAD D
PUSH D

d. XTHL
INX D
INX D
INX D
XTHL

---

44. The open loop transfer function G(s) of a unity feedback control system is given as, **(+2, -0.66)**

$G(s) = \frac{k\left(s+\frac{2}{3}\right)}{s^2(s+2)}$

From the root locus, it can be inferred that when k tends to positive infinity.

a. three roots with nearly equal real parts exist on the left half of the s-plane

b. one real root is found on the right half of the s-plane

c. the root loci cross the jω axis for a finite value of k; k ≠ 0

d. three real roots are found on the right half of the s-plane

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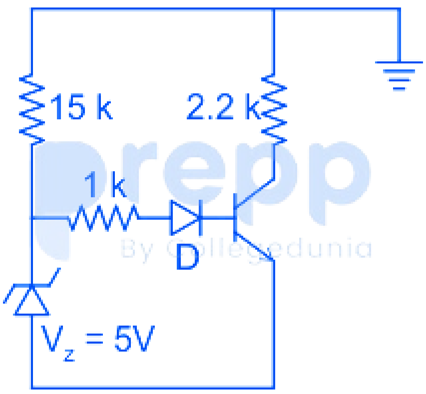
45. The transistor is used in the circuit shown below has a β of 30 and I_{CB0} is negligible with V_{EE}= -12 V at emitter end. **(+2, -0.66)**



V_{BE} = 0.7 V

$V_{CE(sat)}$ = 0.2 V

If the forward voltage drop of diode is 0.7 V. Then the current through collector will be

a. 168 mA

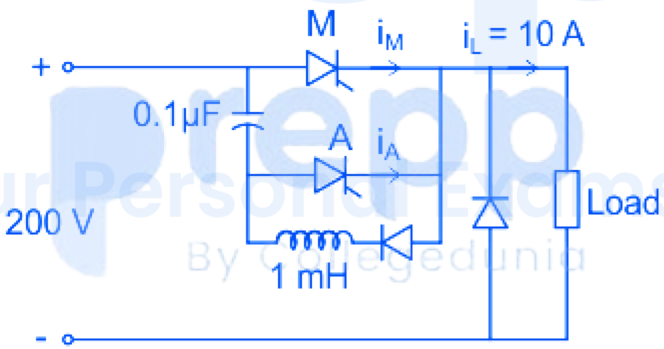
b. 108 mA

c. 20.54 mA

d. 5.36 mA

---

46. A voltage commutated chopper circuit, operated at 500 Hz is shown below. If the maximum value of load current is 10 A, then the maximum current through the main and auxiliary thyristor will be **(+2, -0.66)**



a. $i_{M_{max}} = 12A, i_{A_{max}} = 10A$

b. $i_{M_{max}} = 12A, i_{A_{max}} = 2A$

c. $i_{M_{max}} = 10A, i_{A_{max}} = 12A$

d. $i_{M_{max}} = 10A, i_{A_{max}} = 8A$

---

47. Solution of the variables x_1 and x_2 for the following equations is to be obtained by employing the Newton Raphson iterative method. **(+2, -0.66)**

equation (i) $10x_2 \sin x_1 - 0.8 = 0$

equation (ii) $10x_2^2 - 10x_2 \cos x_1 - 0.6 = 0$

Assuming the initial values $x_1 = 0.0$ and $x_2 = 1.0$, the Jacobian matrix is

a. $\begin{bmatrix} 10 & -0.8 \\ 0 & -0.6 \end{bmatrix}$

b. $\begin{bmatrix} 10 & 0 \\ 0 & 10 \end{bmatrix}$

c. $\begin{bmatrix} 0 & -0.8 \\ 10 & -0.6 \end{bmatrix}$

d. $\begin{bmatrix} 10 & 0 \\ 10 & -10 \end{bmatrix}$

---

48. The frequency response of a linear system G(jω) is provided in the tabular form below **(+2, -0.66)**

G(jω)	1.3	1.2	1.0	0.8	0.5	0.3
∠G (jω)	-130°	-140°	-150°	-160°	-180°	-200°

The gain margin and phase margin of the system are

a. 6 dB and 30°

b. 6 dB and -30°

c. -6 dB and 30°

d. -6 dB and -30°

---

49. A 3-phase. 6-pole, 50 Hz, squirrel cage induction motor is running at a slip of 5%. The speed of stator magnetic field to rotor magnetic field and speed of rotor with respect to stator magnetic field are **(+2, -0.66)**

a. Zero, -50 rpm

b. Zero, 955 rpm

c. 1000 rpm, -50 rpm

d. 1000 rpm, 955 rpm

---

50. A capacitor is made with a polymeric dielectric having ε_r of 2.26 and a dielectric breakdown strength of 50 kV / cm. The permittivity of free space is 8.85 pF / m. If the rectangular plates of the capacitor have a width of 20 cm and a length of 40 cm, the maximum electric charge in the capacitor is **(+2, -0.66)**

a. 2 μC

b. 4 μC

c. 8 μC

d. 10 μC

---

51. The response h(t) of a linear time invariant system to an impulse δ (t), under initially relaxed condition is $h(t) = e^{-t} + e^{-2t}$. The response of this system for a unit step input u(t) is **(+2, -0.66)**

a. $u(t) + e^{-t} + e^{-2t}$

b. $(e^{-t} + e^{-2t})\, u(t)$

c. $(1.5 - e^{-t} - 0.5e^{-2t})\ u(t)$

d. $e^{-t}\ \delta(t) + e^{-2t}\ u(t)$

---

52. The direct axis and quadrature axis reactance of a salient pole alternator are 1.2 p.u and 1.0 p.u respectively. The armature resistance is negligible. If the alternator is delivering rated kVA at upf and at rated voltage then its power angle is **(+2, -0.66)**

a. 30°

b. 45°

c. 60°

d. 90°

---

53. A $4\frac{1}{2}$ digit DMM has the error specification as 0.2% of reading +10 counts. If a d.c. voltage of 100 V is read on its 200 V full-scale, the maximum error that can be expected in the reading is **(+2, -0.66)**

a. ± 0.1%

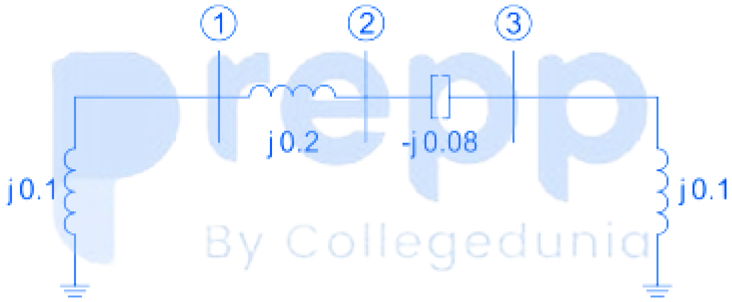
b. ± 0.2%

c. ± 0.3%

d. ± 0.4%

---

54. A three – bus network is shown in the figure below indicating the p.u. impedance of each element. **(+2, -0.66)**



The bus admittance matrix, Y-bus, of the network is

a. \(j\left[{\begin{array}{*{20}{c}} {0.3}&{ - 0.2}&0\\ { - 0.2}& {0.12}&{0.08}\\ 0&{0.08}&{0.02} \end{array}} \right]\)

b. \(j\left[{\begin{array}{*{20}{c}} { - 15}&5&0\\ 5&{7.5}&{ - 12.5}\\ 0&{ - 12.5}&{2.5} \end{array}} \right]\)

c. \(j\left[{\begin{array}{*{20}{c}} {0.1}&{0.2}&0\\ {0.2}&{0.12}&{ - 0.08}\\ 0&{ - 0.08}&{0.10} \end{array}} \right]\)

d. \(j\left[{\begin{array}{*{20}{c}} { - 10}&5&0\\ 5&{7.5}&{12.5}\\ 0&{12.5}&{ - 10} \end{array}} \right]\)

---

55. With K as a constant, the possible solution for the first order differential equation $\frac{dy}{dx} = e^{-3x}$ is **(+2, -0.66)**

a. $-\frac{1}{3}e^{-3x} + K$

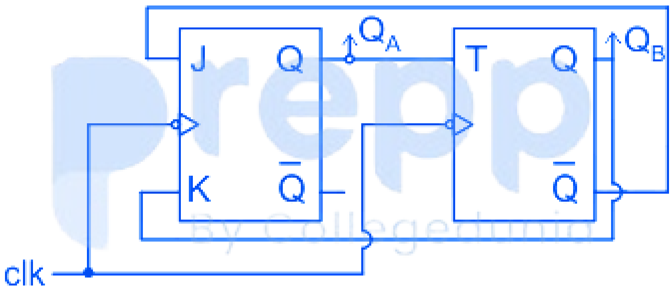
b. $-\frac{1}{3}e^{3x} + K$

c. -3 e^{-3x} + K

d. -3 e^{-x} + K

---

56. A two-bit counter circuit is shown below **(+2, -0.66)**



If the state Q_A Q_B of the counter at the clock time t_n is '10' then the state Q_A Q_B of the counter at t_n + 3 (after three clock cycles) will be

a. 01

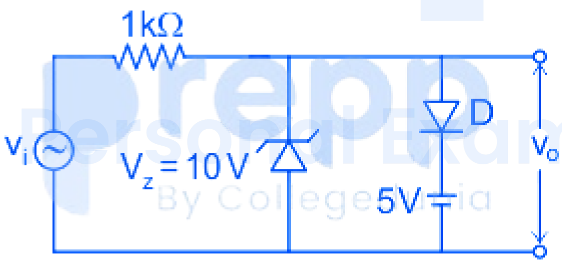
b. 00

c. 10

d. 11

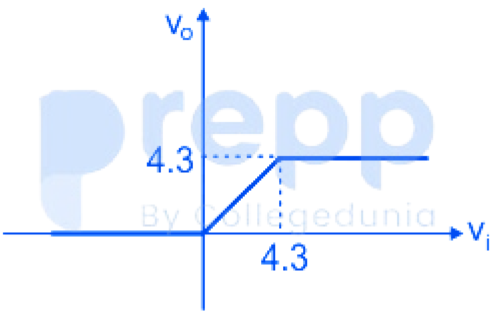
---

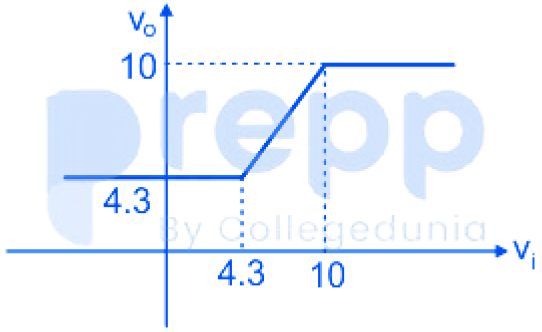
57. A clipper circuit is shown below: **(+2, -0.66)**



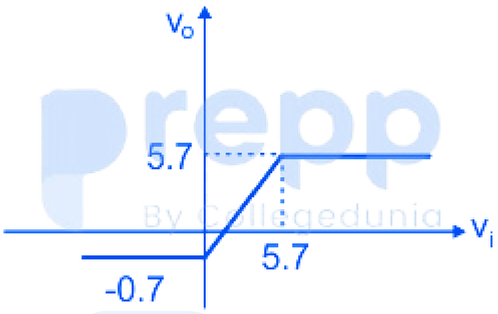
Assuming forward voltage drops of the diodes to be 0.7 V, the input-output transfer characteristics for the circuit will be

a.

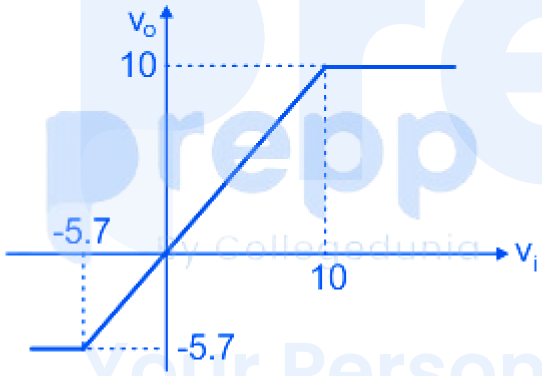




b.



c.



d.

---

58. The input voltage to a converter is **(+2, -0.66)**

$V_i = 100\sqrt{2}\sin(100\pi t)\, V$

The current drawn by the converter is

$i_1 = 10\sqrt{2}\sin\left(100\pi t - \frac{\pi}{3}\right) + 5\sqrt{2}\sin\left(300\pi + \times\frac{\pi}{4}\right) + 2\sqrt{2}\sin\left(500\pi t - \pi/6\right) A.$

The input power factor of the converter

a. 0.31

b. 0.44

c. 0.5

d. 0.71

---

59. The input voltage to a converter is $V_i = 100\sqrt{2}\sin(100\pi t)\,V$ .The current drawn by the converter is The current drawn by the converter is **(+2, -0.66)**

$i_1 = 10\sqrt{2}\sin\left(100\pi t - \frac{\pi}{3}\right) + 5\sqrt{2}\sin\left(300\pi t + \frac{\pi}{4}\right) + 2\sqrt{2}\sin(500\pi t - \pi/6)\,A$. The active power drawn by the converter is

a. 181 W

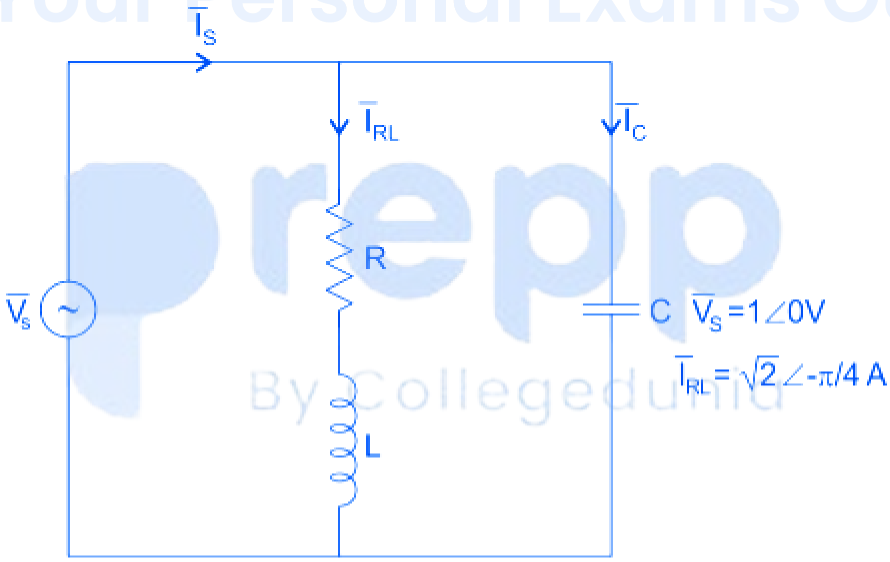
b. 500 W

c. 707 W

d. 887 W

---

60. An RLC circuit with relevant data is given below. **(+2, -0.66)**



Power dissipated through R is

a. 1 W

b. 2 W

c. 0.5 W

d. 1.5 W

---

61. Roots of the algebraic equation $x^3 + x^2 + x + 1 = 0$ are **(+2, -0.66)**

a. (+1, +j, -j)

b. (+1, -1, +1)

c. (0, 0, 0)

d. (-1, +j, -j)

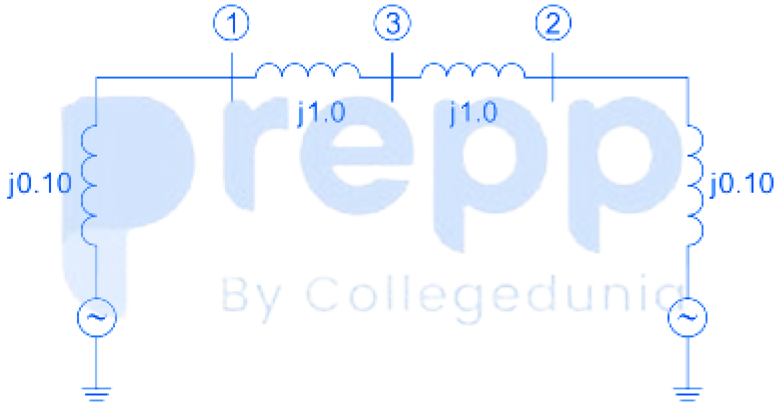
---

62. Two generator units G_1 and G_2 are connected by 15kV line with a bus at the mid-point as shown below **(+2, -0.66)**

G_1 = 250 MVA, 15 kV, positive sequence reactance X_{G1} = 25% on its own base

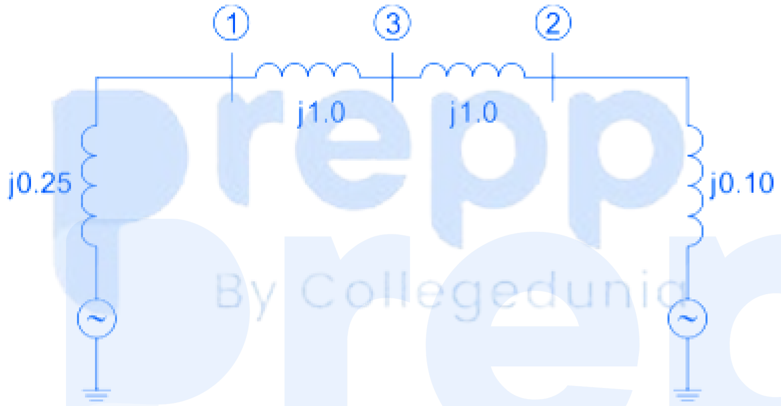
G_2 = 100 MVA, 15 kV, positive sequence reactance X_{G2} = 10% on its own base L_1 and L_2 = 10 km, positive sequence reactance X_L = 0.225 Ω/km

For the above system, the positive sequence diagram with the p.u values on the 100 MVA common base a.



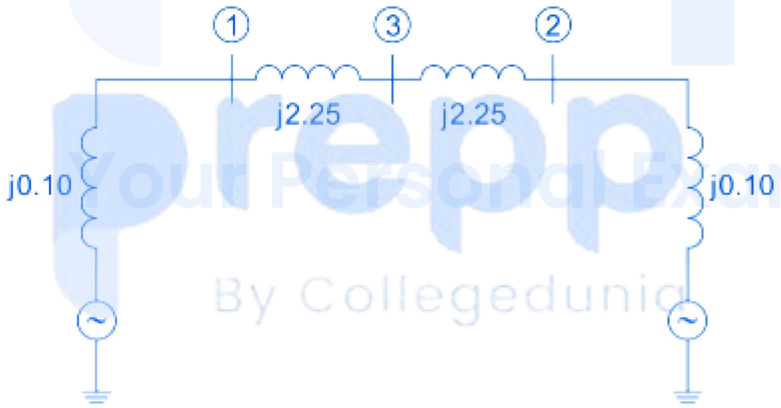
1
3
2
j1.0
j1.0
j0.10
j0.10

b.

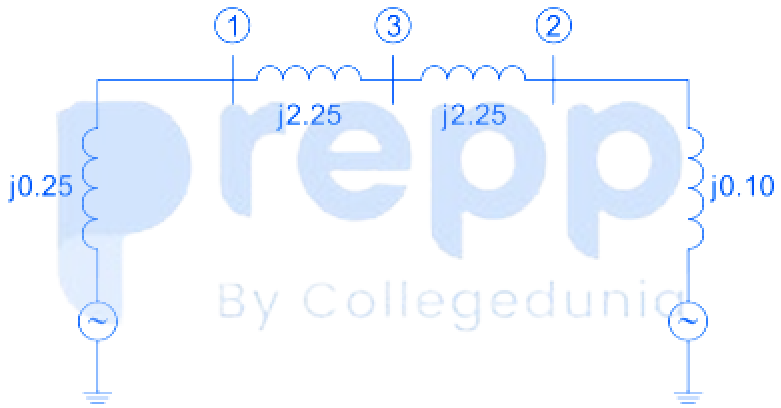


1
3
2
j1.0
j1.0
j0.25
j0.10

c.



1
3
2
j2.25
j2.25
j0.10
j0.10

d. 

63. Two generator units G_1 and G_2 are connected by 15kV line with a bus at the mid-point as shown below **(+2, -0.66)**

G_1 = 250 MVA, 15 kV, positive sequence reactance X_{G1} = 25% on its own base

G_2 = 100 MVA, 15 kV, positive sequence reactance X_{G2} = 10% on its own base L_1 and L_2 = 10 km, positive sequence reactance X_L = 0.225 Ω/km

In the above system, the three – phase fault MVA at the bus 3 is

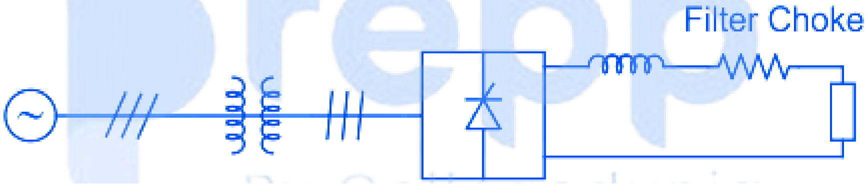
a. 82.55 MVA

b. 85.11 MVA

c. 170.91 MVA

d. 181.82 MV

64. A solar energy installation rectifier a three phase bridge converter to load energy into power system through a transformer of 400V/ 400V, as shown below. (+2, -0.66)



The energy is collected on a bank of 400V battery and is connected to converter through a large filter choke of resistance 10Ω.The maximum current through the battery will be

a. 14 A

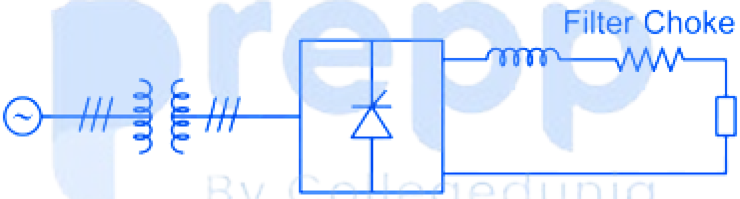
b. 40 A

c. 80 A

d. 94 A

---

65. A solar energy installation rectifier a three phase bridge converter to load energy into power system through a transformer of 400V/ 400V, as shown below. (+2, -0.66)



The energy is collected on a bank of 400V battery and is connected to converter through a large filter choke of resistance 10Ω.The KVA rating of the output transformer is

a. 53.2 kVA

b. 46.0 kVA

c. 22.6 kVA

d. None

Answers

1. **Answer: a**

Explanation:

Understanding the Election Scenario

This problem involves calculating the total number of voters in an election based on initial promises made by voters to candidates P and Q, and subsequent changes in their voting intentions. We are given that candidate P lost the election by 2 votes.

Step-by-Step Calculation of Total Voters

Let's denote the total number of voters as N.

Initial Voter Promises:

- Percentage of voters promised to vote for P: 40%
- Percentage of voters promised to vote for Q: The rest, which is $100\% - 40\% = 60\%$

Changes in Voting Intentions:

- Voters who promised P but switched to Q: 15% of those who promised P.
- Voters who promised Q but switched to P: 25% of those who promised Q.

Calculating Actual Votes for P and Q:

Let's calculate the number of voters switching:

- Number of voters switching from P to Q = 15% of 40% of $N = 0.15 \times (0.40 \times N) = 0.06 \times N$
- Number of voters switching from Q to P = 25% of 60% of $N = 0.25 \times (0.60 \times N) = 0.15 \times N$

Now, let's find the final actual votes for each candidate:

- **Actual Votes for P** = (Voters initially promised to P) - (Voters who switched from P to Q) + (Voters who switched from Q to P) Actual Votes for P = $(0.40 \times N) - (0.06 \times N) + (0.15 \times N)$ Actual Votes for P = $(0.40 - 0.06 + 0.15) \times N$ Actual Votes for P = $0.49 \times N$
- **Actual Votes for Q** = (Voters initially promised to Q) - (Voters who switched from Q to P) + (Voters who switched from P to Q) Actual Votes for Q = $(0.60 \times N) - (0.15 \times N) + (0.06 \times N)$ Actual Votes for Q = $(0.60 - 0.15 + 0.06) \times N$ Actual Votes for Q = $0.51 \times N$

Using the Election Outcome to Find Total Voters:

We are given that P lost the election by 2 votes. This means the number of votes for P is 2 less than the number of votes for Q.

Mathematically, this can be written as:

Actual Votes for P = Actual Votes for Q - 2

Substituting the expressions we found:

$0.49 \times N = (0.51 \times N) - 2$

Now, we solve for N:

$2 = (0.51 \times N) - (0.49 \times N)$

$2 = (0.51 - 0.49) \times N$

$2 = 0.02 \times N$

$N = \frac{2}{0.02}$

$N = \frac{200}{2}$

$N = 100$

Summary of Votes:

Candidate	Initial Promise	Switched Away	Switched To	Actual Votes
P	$0.40N$	$0.06N$ (to Q)	$0.15N$ (from Q)	$0.49N$
Q	$0.60N$	$0.15N$ (to P)	$0.06N$ (from P)	$0.51N$

With $N = 100$:

- Actual Votes for P = $0.49 \times 100 = 49$
- Actual Votes for Q = $0.51 \times 100 = 51$

The difference is $51 - 49 = 2$ votes, which matches the condition that P lost by 2 votes.

Conclusion

The total number of voters was 100.

---

2. **Answer: c**

Explanation:

Problems: Understanding the Sentence Context

The key to correctly completing this sentence lies in understanding the implication of the phrase "**counter-productive**." The sentence states, "It was her view that the country's problems had been _______ by foreign technocrats, so that to invite them to come back would be **counter-productive**."

- The term "counter-productive" means that something has the opposite effect of what is desired, or makes a situation worse rather than better.
- If inviting the foreign technocrats back would be counter-productive, it logically follows that their previous involvement must have had a negative impact on the country's problems. They didn't solve or merely identify the problems; instead, they made them worse.

Options: Evaluating Word Choices

Let's analyze each given option in the context of the sentence:

- **identified:** This word means to recognize or pinpoint something. If the technocrats simply "identified" the problems, inviting them back might be helpful for implementing solutions, not "counter-productive." This option does not fit the negative consequence implied.
- **ascertained:** Similar to "identified," this means to discover or confirm something for certain. It implies gaining knowledge, not worsening a situation. Therefore, it does not align with the "counter-productive" outcome mentioned in the sentence.
- **exacerbated:** This verb means to make a problem, a bad situation, or a negative feeling worse. If the foreign technocrats "exacerbated" the country's problems, it means they intensified them or made them more severe. In this case, inviting them back would indeed be "counter-productive" because their past actions led to a worsening of the problems. This word perfectly fits the logical flow of the sentence.
- **analysed:** This means to examine something in detail to understand its nature or structure. While technocrats might analyze problems, this word itself does not suggest they made the problems worse. Analysis is often a prerequisite for solving problems and doesn't inherently imply a negative impact that would make their return "counter-productive."

Exacerbated: The Appropriate Word Choice

Considering the meaning of "counter-productive," the only word among the options that suggests the foreign technocrats' previous involvement made the country's problems worse is **exacerbated**. Therefore, "exacerbated" is the most appropriate word to complete the given sentence, creating a coherent and logical statement.

---

3. **Answer: b**

Explanation:

Frequency: Understanding Its Opposite Meaning

The question asks us to identify the word that is most nearly opposite in meaning to the word "Frequency." To answer this, we need to understand the core meaning of "Frequency" and then analyze each given option.

Defining Frequency

The word **Frequency** refers to the rate at which something occurs or is repeated over a particular period of time. It indicates how often an event or occurrence happens. For example, if a bus comes every 10 minutes, its frequency is high. If it comes once a day, its frequency is low.

- **High frequency** implies something happens often or is common.
- **Low frequency** implies something happens seldom or is uncommon.

Analyzing the Options for Opposite Meaning

Let's examine each option to see which one stands as the most direct opposite to "Frequency."

- **1. periodicity**

  Periodicity refers to the quality or state of being periodic, meaning something recurs at regular intervals. This concept is closely related to frequency. For instance, a periodic event has a certain frequency. Therefore, "periodicity" is not an opposite; it is more of a related concept or even a synonym in certain contexts, particularly when discussing recurring events.

- **2. rarity**

  Rarity refers to the state or quality of being rare; something that does not occur often or is uncommon. If something occurs with high **frequency**, it is common. If something occurs with low **frequency**, it is rare. Thus, "rarity" directly expresses the idea of infrequency or uncommonness, making it the most suitable opposite of "Frequency."

- **3. gradualness**

Gradualness refers to the quality of being gradual, meaning something happens or changes slowly, by small degrees. This word describes the pace or manner of change, not how often something occurs. There is no direct opposite relationship between "gradualness" and "Frequency."

- **4. persistency**

  Persistency refers to the quality of persisting; continuing firmly or existing over a prolonged period. While something that persists might happen repeatedly, "persistency" itself doesn't quantify how often something happens (its frequency). An event can persist rarely or frequently. Therefore, "persistency" is not the opposite of "Frequency."

Conclusion: The Opposite of Frequency

Based on our analysis, "rarity" clearly stands out as the most nearly opposite in meaning to "Frequency." Where "Frequency" denotes how often something occurs, "rarity" denotes how seldom it occurs.

---

4. **Answer: d**

Explanation:

Indian Medical Association's Ethical Guidelines for Gene Manipulation

Understanding the ethical implications of advanced medical procedures like gene manipulation is crucial. The question asks about specific conditions under which the Indian Medical Association (IMA) permits human gene manipulation, emphasizing that it should only occur when conventional treatments are inadequate.

Understanding Gene Manipulation and Ethical Principles

Gene manipulation, often referred to as gene therapy or genetic engineering, involves altering an individual's genes to treat or prevent disease. Due to its profound implications for human health and identity, strict ethical guidelines are essential. These guidelines typically ensure that such powerful interventions are considered only as a last resort, when existing, less invasive, or less risky treatments have proven ineffective.

Analyzing the Options for 'Unsatisfactory Treatments'

Let's carefully examine each option provided to determine which word best completes the sentence, aligning with the ethical principles of medical intervention:

- **Option 1: similar**

  If "similar" treatments are unsatisfactory, it does not necessarily imply that all other possible treatments, even those dissimilar, have been exhausted. Ethical guidelines typically demand a broader failure of conventional care before resorting to gene manipulation. This option is too restrictive in its scope.

- **Option 2: most**

  Stating that "most" treatments are unsatisfactory is vague. It doesn't specify which majority of treatments have failed, nor does it guarantee that all viable alternatives have been considered. Ethical considerations demand a more comprehensive assessment of treatment failure.

- **Option 3: uncommon**

  Whether treatments are "uncommon" or common is not the primary ethical criterion for deciding on gene manipulation. The key is their efficacy and whether they offer a satisfactory solution to the disease. An uncommon but effective treatment might still be preferable to gene manipulation. This option is irrelevant to the ethical justification.

- **Option 4: available**

  The word "available" is the most appropriate choice. It implies that all treatments that are currently known, accessible, and typically used for the specific disease have been tried or thoroughly considered and found to be unsatisfactory. This aligns with the principle that gene manipulation, being a complex and potentially high-risk intervention, should be a treatment of last resort when all conventional, "available" options have failed to provide adequate relief or cure.

Conclusion on Ethical Treatment Selection

The Indian Medical Association's ethical guidelines underscore a cautious approach to gene manipulation. Human genes should only be manipulated to correct diseases when all **available** treatments have been deemed unsatisfactory. This ensures that gene therapy is pursued only when conventional medical science cannot adequately address the patient's condition, highlighting its role as a significant, carefully considered intervention in the medical field.

---

5. **Answer: d**

Explanation:

Understanding the Original Pair: Gladiator : Arena

The original pair given is "Gladiator : Arena". To understand the relationship, let's define each term:

- **Gladiator:** A professional combatant or performer in ancient Rome who engaged in fights, often to the death, for the entertainment of the public.
- **Arena:** The central, open area in a Roman amphitheater where gladiatorial contests, public spectacles, and other events took place. It was the specific venue for the gladiator's activity.

Therefore, the relationship between "Gladiator" and "Arena" is that of a **specialized performer or participant** and the **specific, dedicated venue or place where their primary activity or performance takes place**. This activity is often public, formal, and sometimes high-stakes.

Analyzing the Relationship Type for Word Analogy

When analyzing word analogies, it's important to precisely identify the connection. For "Gladiator : Arena", the key aspects of the relationship include:

- **Agent/Professional:** A gladiator is a professional engaged in a specific line of work.
- **Dedicated Venue:** An arena is a place specifically designed and used for that profession's activity.
- **Formal/Public Activity:** The activity performed by the gladiator in the arena is typically public and follows certain formal rules or customs.

We are looking for an option pair that exhibits this 'Professional/Agent : Dedicated Venue for Primary, Formal Activity' relationship most closely.

Evaluating the Given Options

Let's examine each of the provided options to see how their relationships compare to "Gladiator : Arena":

- **Dancer : Stage:**
  - **Dancer:** A person whose profession is dancing or performing.
  - **Stage:** A raised platform in a theater or other venue where artistic performances take place.
  - **Relationship:** A Dancer performs on a Stage. This is a 'Performer : Place of Performance' relationship, which is quite similar to the original pair.
- **Commuter : Train:**
  - **Commuter:** A person who travels regularly between two places, typically from home to work.
  - **Train:** A form of public transport.
  - **Relationship:** A Commuter travels by (or on) a Train. This is a 'Person : Mode of Transport' relationship, which is distinctly different from the 'Agent : Place of Primary Activity' relationship we are looking for.
- **Teacher : Classroom:**
  - **Teacher:** A professional who instructs or educates students.
  - **Classroom:** A room in a school or educational institution where teaching and learning take place.

  - **Relationship:** A Teacher teaches in a Classroom. This represents a 'Professional : Place of Work/Primary Activity' relationship. It is a strong candidate because it involves a specific professional and their dedicated workspace.
- **Lawyer : Courtroom:**
  - **Lawyer:** A professional who practices law, representing clients in legal matters.
  - **Courtroom:** A room in a courthouse where legal proceedings, trials, and hearings are conducted.
  - **Relationship:** A Lawyer conducts their professional legal arguments and activities in a Courtroom. This is a 'Professional : Dedicated Venue for Formal/High-Stakes Activity' relationship.

Identifying the Best Analogy

Upon evaluating the options, "Dancer : Stage", "Teacher : Classroom", and "Lawyer : Courtroom" all present a 'person and their activity place' relationship. However, we need to find the pair that best captures the specific nuances of "Gladiator : Arena".

- The gladiator's activity in the arena is often adversarial, public, and involves high stakes (life or death). It's a specialized skill performed in a formal, designated setting for an audience.
- While a dancer performs on a stage, the activity is typically artistic performance rather than an adversarial contest.
- A teacher works in a classroom, which is a dedicated place, but the activity of teaching is instructional and collaborative, not typically adversarial or a 'contest'.
- A **Lawyer** in a **Courtroom** provides the closest parallel:
  - **Lawyer:** A highly specialized professional.
  - **Courtroom:** A formal, dedicated venue specifically designed for legal proceedings.
  - **Activity:** A lawyer's work in a courtroom (presenting cases, making arguments, cross-examining) is often adversarial, highly skilled, and conducted in a public, formal setting with significant outcomes (high stakes), much like a gladiator's contest.

The relationship between a **Lawyer** and a **Courtroom** most accurately reflects the complex relationship of a **Gladiator** and an **Arena**, emphasizing a specialized individual performing a crucial, often adversarial, and public activity within a designated, formal environment.

---

6. **Answer: b**

Explanation:

Calculation:

Consumption (km/liter) means distance covered by the car in 1 liter of fuel

⇒ Fuel consumption per liter = $\frac{1}{Consumption\ (km/liter)}$

	P	Q	R	S
Distance	15	75	40	10
Speed	15	45	75	10
Consumption (km/liter)	60	90	75	30
Fuel consumption per km	$\frac{1}{60}$ = 0.016	$\frac{1}{90}$ = 0.011	$\frac{1}{75}$ = 0.013	$\frac{1}{30}$ = 0.033

From the above table, fuel consumption per km was least during the lap Q.

---

7. **Answer: c**

Explanation:

Solving the Toffee Sharing Problem

This problem involves a sequence of actions where toffees are taken and returned. To find the original number of toffees, we need to work backward from the final known quantity.

Working Backwards Step-by-Step

Let's denote the number of toffees at different stages:

- N = Original number of toffees.
- N_R = Number of toffees after R's actions.
- N_S = Number of toffees after S's actions.
- N_T = Number of toffees after T's actions (final amount).

We are given that the final amount, N_T, is 17 toffees.

Reversing T's Actions

T took half of the remaining toffees and returned two. Let the number of toffees just before T acted be X.

- Amount T took = $\frac{1}{2}X$
- Amount left after T took his share = $X - \frac{1}{2}X = \frac{1}{2}X$
- After T returned 2 toffees, the amount became $\frac{1}{2}X + 2$.

We know this final amount is 17:

$$\frac{1}{2}X + 2 = 17$$

Subtract 2 from both sides:

$$\frac{1}{2}X = 17 - 2$$

$$\frac{1}{2}X = 15$$

Multiply by 2 to find X:

$$X = 15 \times 2 = 30$$

So, there were 30 toffees before T took his share ($N\_S = 30$).

Reversing S's Actions

S took 1/4th of the remaining toffees and returned three. Let the number of toffees just before S acted be *Y*.

- Amount S took = $\frac{1}{4}Y$
- Amount left after S took his share = $Y - \frac{1}{4}Y = \frac{3}{4}Y$
- After S returned 3 toffees, the amount became $\frac{3}{4}Y + 3$.

This amount equals the toffees before T acted, which is *X = 30*:

$$\frac{3}{4}Y + 3 = 30$$

Subtract 3 from both sides:

$$\frac{3}{4}Y = 30 - 3$$

$$\frac{3}{4}Y = 27$$

Multiply by $\frac{4}{3}$ to find *Y*:

$$Y = 27 \times \frac{4}{3}$$

$$Y = 9 \times 4 = 36$$

So, there were 36 toffees before S took his share (*N_R = 36*).

Reversing R's Actions

R took 1/3rd of the original toffees and returned four. Let the original number of toffees be *N*.

- Amount R took = $\frac{1}{3}N$
- Amount left after R took his share = $N - \frac{1}{3}N = \frac{2}{3}N$
- After R returned 4 toffees, the amount became $\frac{2}{3}N + 4$.

This amount equals the toffees before S acted, which is *Y = 36*:

$$\frac{2}{3}N + 4 = 36$$

Subtract 4 from both sides:

$$\frac{2}{3}N = 36 - 4$$

$$\frac{2}{3}N = 32$$

Multiply by $\frac{3}{2}$ to find N:

$$N = 32 \times \frac{3}{2}$$

$$N = 16 \times 3 = 48$$

Therefore, the original number of toffees was 48.

Verification of the Solution

Let's check if starting with 48 toffees leads to the final count of 17:

1. **Start:** 48 toffees.
2. **R's turn:** R takes $\frac{1}{3} \times 48 = 16$. Bowl has $48 - 16 = 32$. R returns 4. Bowl has $32 + 4 = 36$.
3. **S's turn:** S takes $\frac{1}{4} \times 36 = 9$. Bowl has $36 - 9 = 27$. S returns 3. Bowl has $27 + 3 = 30$.
4. **T's turn:** T takes $\frac{1}{2} \times 30 = 15$. Bowl has $30 - 15 = 15$. T returns 2. Bowl has $15 + 2 = 17$.

The final count of 17 matches the problem statement, confirming our calculation.

---

8. **Answer: d**

Explanation:

Function Understanding: Absolute Value Definition

The given function is $f(y) = \frac{|y|}{y}$. To understand this function, we need to consider the definition of the absolute value function, $|y|$.

- If y is a positive real number $(y > 0)$, then $|y| = y$.
- If y is a negative real number $(y < 0)$, then $|y| = -y$.
- The function is undefined when $y = 0$, as division by zero is not allowed.

Based on this, we can define $f(y)$ in two cases:

- **Case 1:** $y > 0$
  If $y > 0$, then $f(y) = \frac{y}{y} = 1$.
- **Case 2:** $y < 0$
  If $y < 0$, then $f(y) = \frac{-y}{y} = -1$.

Evaluating f(q) for Non-Zero Real Number q

The problem states that q is any non-zero real number. We need to evaluate $f(q)$ based on whether q is positive or negative.

- **When** $q > 0$ **(q is positive):**
  According to our function definition for $y > 0$, $f(q) = 1$.
- **When** $q < 0$ **(q is negative):**
  According to our function definition for $y < 0$, $f(q) = -1$.

Evaluating f(-q) for Non-Zero Real Number q

Next, we need to evaluate $f(-q)$. The sign of $-q$ depends on the sign of q.

- **When** $q > 0$ **(q is positive):**
  If q is positive, then $-q$ will be negative ($-q < 0$).
  According to our function definition for $y < 0$, $f(-q) = -1$.
- **When** $q < 0$ **(q is negative):**
  If q is negative, then $-q$ will be positive ($-q > 0$).
  According to our function definition for $y > 0$, $f(-q) = 1$.

Calculating the Difference f(q) - f(-q)

Now we will calculate the difference $f(q) - f(-q)$ for both cases of q.

- **Case 1: When** $q > 0$

  We found $f(q) = 1$ and $f(-q) = -1$.

  So, $f(q) - f(-q) = 1 - (-1) = 1 + 1 = 2$.

- **Case 2: When** $q < 0$

  We found $f(q) = -1$ and $f(-q) = 1$.

  So, $f(q) - f(-q) = -1 - 1 = -2$.

Condition	$f(q)$	$f(-q)$	$f(q) - f(-q)$
$q > 0$	1	-1	$1 - (-1) = 2$
$q < 0$	-1	1	$-1 - 1 = -2$

Final Absolute Value Calculation: $|f(q) - f(-q)|$

Finally, we need to find the absolute value of the difference, $|f(q) - f(-q)|$.

- **From Case 1 ($q > 0$):**
  $|f(q) - f(-q)| = |2| = 2$.
- **From Case 2 ($q < 0$):**
  $|f(q) - f(-q)| = |-2| = 2$.

In both possible scenarios for q (positive or negative), the value of $|f(q) - f(-q)|$ is 2.

The final answer is 2.

---

9. **Answer: c**

Explanation:

Sum of Series 4 + 44 + 444 + ...

To find the sum of 'n' terms of the given series $S_n = 4 + 44 + 444 + \cdots + \text{n terms}$, we can use a systematic approach by transforming each term into a form involving powers of 10.

Series Transformation for Summation

The given series is:

$$S_n = 4 + 44 + 444 + \cdots + \text{n terms}$$

Step 1: Factor out 4 from each term.

$$S_n = 4(1 + 11 + 111 + \cdots + \text{n terms})$$

Step 2: Multiply and divide the expression inside the parenthesis by 9. This step is crucial as it helps convert the repeating digit numbers into a difference of powers of 10 and 1.

$$S_n = 4\left(\frac{1}{9}\right)(9 + 99 + 999 + \cdots + \text{n terms})$$

$$S_n = \frac{4}{9}(9 + 99 + 999 + \cdots + \text{n terms})$$

Step 3: Express each term inside the parenthesis using powers of 10.

- The first term, 9, can be written as $10 - 1$.
- The second term, 99, can be written as $10^2 - 1$.
- The third term, 999, can be written as $10^3 - 1$.
- Following this pattern, the n-th term will be $10^n - 1$.

So, the series becomes:

$$S_n = \frac{4}{9}[(10 - 1) + (10^2 - 1) + (10^3 - 1) + \cdots + (10^n - 1)]$$

Grouping Terms and Applying Formulas

Step 4: Group the positive powers of 10 together and the negative ones together.

$$S_n = \frac{4}{9}[(10 + 10^2 + 10^3 + \cdots + 10^n) - (1 + 1 + 1 + \cdots + \text{n times})]$$

Step 5: Calculate the sum of the first part, which is a Geometric Progression (GP).

The first part is $10 + 10^2 + 10^3 + \cdots + 10^n$.

This is a geometric progression with:

- First term $a = 10$
- Common ratio $r = \frac{10^2}{10} = 10$
- Number of terms n

The sum of n terms of a GP is given by the formula $S_n = \frac{a(r^n - 1)}{r-1}$.

Substituting the values:

$$\text{Sum of GP} = \frac{10(10^n - 1)}{10 - 1} = \frac{10(10^n - 1)}{9}$$

$$\text{Sum of GP} = \frac{10^{n+1} - 10}{9}$$

Step 6: Calculate the sum of the second part.

The second part is $1 + 1 + 1 + \cdots + \text{n times}$.

The sum of 'n' ones is simply n.

Final Summation of Series

Step 7: Substitute the sums of both parts back into the main equation for S_n.

$$S_n = \frac{4}{9}\left[\left(\frac{10^{n+1} - 10}{9}\right) - n\right]$$

Step 8: Simplify the expression.

To combine the terms inside the square brackets, find a common denominator, which is 9:

$$S_n = \frac{4}{9}\left[\frac{10^{n+1} - 10 - 9n}{9}\right]$$

Multiply the fractions:

$$S_n = \frac{4}{9 \times 9}[10^{n+1} - 9n - 10]$$

$$S_n = \frac{4}{81}[10^{n+1} - 9n - 10]$$

Conclusion

The sum of n terms of the series $4 + 44 + 444 + \ldots$ is $\frac{4}{81}[10^{n+1} - 9n - 10]$.

This result matches option 3 provided in the question.

---

10. **Answer: b**

Explanation:

Horses' Contribution to Medicine

The passage highlights a significant, though often overlooked, role that horses have played in the field of medicine. This process involves utilizing the natural immune response of horses to create valuable medical treatments.

Understanding the Immunity Building Process in Horses

The passage describes a specific process where horses were used to develop serums. Let's break down this process:

- **Toxin Injection:** Horses were injected with "toxins of diseases." This means they were exposed to substances that cause illness.
- **Immunity Development:** Following the injection of toxins, the horses' bodies naturally "built up immunities" in their blood. This implies their immune systems responded by producing antibodies to fight off these toxins.
- **Serum Creation:** Once sufficient immunity was built, a "serum was made from their blood." This serum, rich in antibodies, could then be used to treat diseases in humans.
- **Disease Treatment:** Specifically, serums to combat diphtheria and tetanus were developed using this method, demonstrating the effectiveness of the process.

Analyzing the Inference about Horses' Immunity

The question asks what can be inferred about horses based on this passage. We need to look for a conclusion that logically follows from the described medical process.

Let's evaluate each option:

- **Option 1: <p>given immunity to diseases</p>**
  This option is incorrect. The passage states that horses "built up immunities" after being injected with toxins. This means they developed immunity themselves in response to exposure, rather than being "given" pre-existing immunity from an external source. The purpose of the process was to *induce* immunity in them, not to confer it upon them.
- **Option 2: <p>generally quite immune to diseases</p>**
  This is the most accurate inference. For horses to be successfully injected with disease toxins, survive the exposure, and then produce a strong immune response (antibodies) without succumbing to severe illness, they must possess a robust immune system. If horses were easily susceptible or fragile, this method would not be viable. Their capacity to withstand the toxins and effectively "build up immunities" suggests a general physiological resilience or strong immune response mechanism, which can be interpreted as being "generally quite immune" in the sense of being able to robustly fight off or adapt to disease challenges. This inherent strength makes them suitable for such medical procedures.
- **Option 3: <p>given medicines to fight toxins</p>**
  This option is incorrect. The passage clearly states horses were "injected with toxins," not given medicines to fight those toxins. The horses' own bodies produced the "fight" in the form of immunity.
- **Option 4: <p>given diphtheria and tetanus serums</p>**
  This option is incorrect. The passage explains that serums for diphtheria and tetanus were "made from their blood" (the horses' blood). This means the horses were the *source* of the serums, not the recipients of them.

Key Takeaway: Horses' Role in Medical Advancements

The success of developing serums from horses' blood, specifically for serious diseases like diphtheria and tetanus, implies that horses have a naturally strong immune system. This allows them to withstand controlled exposure to toxins and produce powerful antibodies, making them invaluable contributors to early medical advancements.

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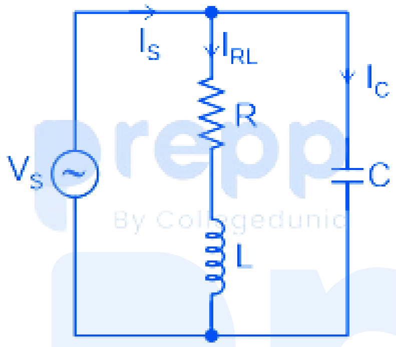
11. **Answer: d**

Explanation:



Applying KCL in the given circuit,

$I_S = I_{RL} + I_C$

or, $I_C = Is - I_{RL}$

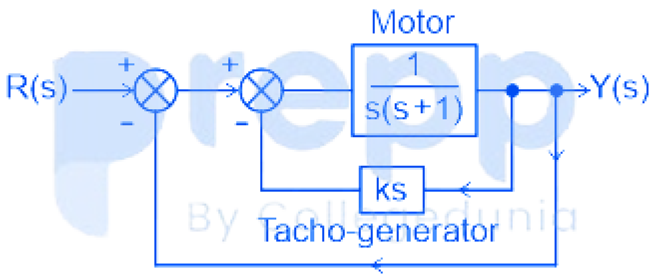
or, $\mathbf{I_C = \sqrt{2}\ \angle\pi/4 - \sqrt{2}\ \angle{-\pi/4} = j2\ A}$

---

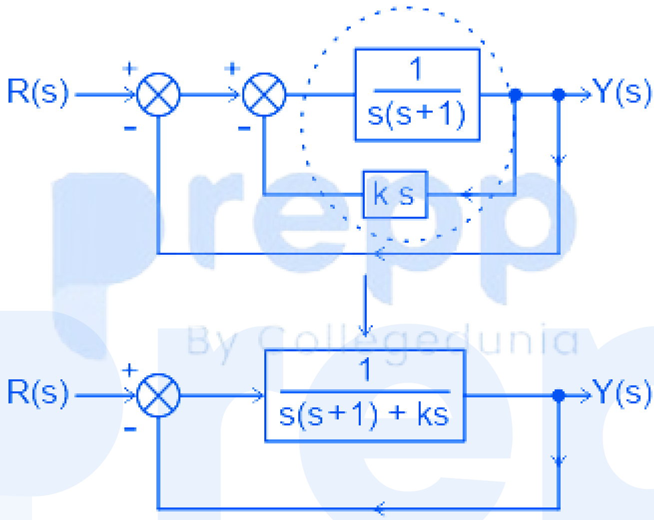
12. **Answer: a**

Explanation:

Given Block Diagram:



Given block diagram can be reduced by,



Further, It can be reduced by,

Hence,

$\frac{Y(S)}{R(S)} = \frac{1}{S^2+S(K+1)+1}$ .... (1)

The standard equation of 2nd order system can be written as,

TF = $\frac{\omega_n^2}{S^2+2\zeta\omega_n S+\omega_n^2}$ .... (2)

From equation (1) & (2),

$\omega_n = \pm 1$

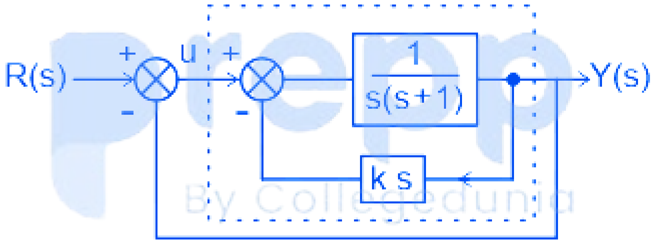
$2\zeta\omega_n = (K + 1)$

or, 2ζ = (K + 1)

or, $\zeta = \frac{K+1}{2}$

Peak over shoot (M_p) = $e^{\left(\frac{-\zeta\pi}{\sqrt{1-\zeta^2}}\right)}$

We have,



Now,

$\frac{Y(S)}{U(S)} = \frac{1}{S(S+1+K)}$

Here, $\phi = -\frac{\pi}{2} - tan^{-1}\frac{\omega}{K+1}$

At: ω → 0 ⇒ ϕ = -(π/2)

At: ω → ∞ ⇒ ϕ = -(π)

Hence, K will only affct the Peak overshoot.

---

13. **Answer: b**

Explanation:

__Concept__:

The inductive reactance for an inductor operating in a frequency 'f' is given by:

$X_L = j\omega L$

L = inductance

Similarly, the capacitive reactance of a capacitor operating in a frequency 'f' is given by:

$X_c = \frac{1}{j\omega C}$

C = Capacitance

Also, for a sinusoidal signal, the RMS value for the given maximum amplitude is given by:

RMS value = 0.707 × Maximum value

Calculation:

Given supply voltage ω = 1 rad/sec

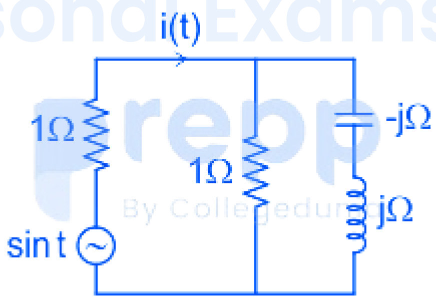
The inductive reactance of the inductor 1 H will be:

$X_L = j\omega L = j\ \Omega$

And the capacitive reactance of the capacitor 1 F will be:

$X_C = 1/j\omega C = -j\ \Omega$

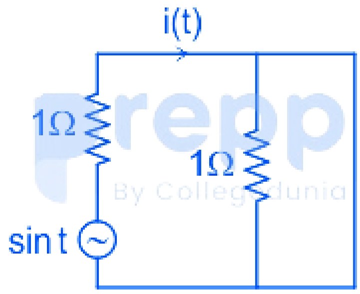
Now, the given circuit is redrawn as:

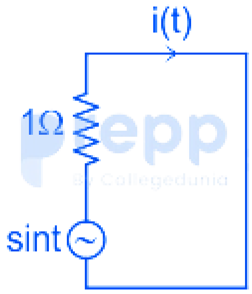


The series combination of inductive and capacitive reactance will be:

$j + (-j) = 0\ \Omega$

The circuit diagram is redrawn as:





i(t) = v(t)/1 = sin t

From the above current equation:

Maximum value of current i(t) = 1

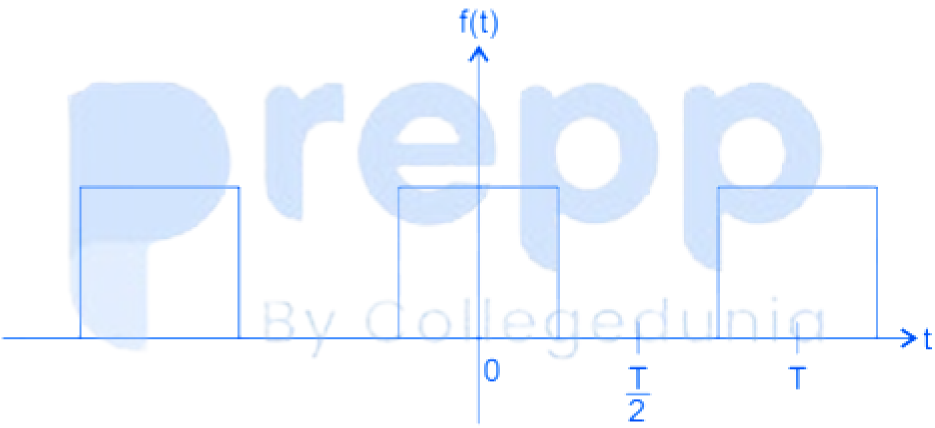
For sinusoidal waveform, RMS value = 0.707 × maximum value

RMS value of current i(t) = 0.707 A = $1/\sqrt{2}$ A

---

14. **Answer: d**

Explanation:



$f\left(t \right) = {a_0} + \mathop \sum \limits_{n = 1}^\infty ({a_n}\cos n\omega t + {b_n}\sin n\omega t)$

The given function f(t) is an even function, therefore $b_n = 0$

f(t) is nonzero average value function, so it will have a nonzero value of a_0

$({a_0} = \frac{1}{{\frac{T}{2}}}\mathop \smallint \limits_0^{\frac{T}{2}} f\left(t \right)\ dt)$ (average value of f(t)

a_n is zero for all even values of n and nonzero for odd n

$({a_n} = \frac{2}{T}\mathop \smallint \limits_0^T f\left(t \right))$ cos(nωt) d(ωt)

So Fourier expansion of f(t) will have a_0 and a_n, n=1, 3, 5, ...∞.

---

15. **Answer: b**

Explanation:

Understanding DC Motor Starters

DC motors require a starter during the initial starting phase. This is because the armature resistance of a DC motor is very low. At the moment of starting, the motor's back EMF (counter electromotive force) is zero because the motor is stationary. If full supply voltage were applied directly, a very large current would flow through the armature (

$$I_a = \frac{V - E_b}{R_a}$$

, where

$$E_b = 0$$

at start), potentially damaging the armature winding, the commutator, or the brushes. A starter adds external resistance in series with the armature during starting, limiting the initial current. As the motor speeds up, the back EMF increases, and the external resistance is gradually cut out.

Types of DC Motor Starters

Common types of DC motor starters include:

- 2-point starter (mainly for DC series motors)
- 3-point starter (for DC shunt and compound motors)
- 4-point starter (for DC shunt and compound motors)

Focus on the 4-Point Starter

The 4-point starter is widely used for starting DC shunt motors and DC compound motors, particularly when speed control is done by field weakening. Let's look at its components and operation, focusing on the No-Volt Coil (NVC).

Part	Function
Starting Resistances	Limits initial armature current. Gradually cut out as speed builds up.
Holding Coil (No-Volt Coil - NVC)	Holds the starter handle in the 'run' position as long as supply voltage is present. Provides no-volt protection (releases handle on voltage failure).
Overload Release (OLR)	Protects the motor from excessive current (overload). Releases the handle if current exceeds a set limit.
Handle/Lever	Manually moved to cut out starting resistances.
Terminals (L, A, F, N)	L: Line (positive supply), A: Armature, F: Field, N: Negative supply (or neutral).

Comparing 3-Point and 4-Point Starters

The main difference between a 3-point and a 4-point starter lies in the connection of the NVC.

Feature	3-Point Starter	4-Point Starter
No-Volt Coil (NVC) Connection	Connected in series with the shunt field winding, across the supply.	Connected directly across the supply terminals (L and N), independently of the field winding.
Suitability for Field Weakening Speed Control	Not suitable. Field weakening reduces field current, which reduces NVC current, potentially releasing the handle prematurely ("field failure" protection can be problematic here).	Suitable. NVC current is independent of field current, ensuring the handle stays in 'run' position even during field weakening or field faults.
Field Failure Protection	Provides inherent protection (handle releases if field current drops significantly). However, this makes field weakening difficult.	Does NOT provide inherent field failure protection. The NVC only responds to supply voltage failure. Separate field failure protection may be needed in some applications.

Why 4-Point Starter for DC Shunt Motor with Field Weakening?

As explained in the comparison, the 4-point starter's NVC is connected directly across the supply. This means the holding current for the handle is dependent only on the supply voltage, not on the current flowing through the shunt field winding. When controlling the speed of a DC shunt motor above its base speed, field weakening is used. This involves decreasing the shunt field current. In a 3-point starter, this decrease in field current would also decrease the current through the NVC, potentially causing the starter handle to return to the off position, stopping the motor. The 4-point starter overcomes this limitation, making it the preferred choice for DC shunt motors where speed control via field weakening is employed.

Analysis of Options

- **Option 1: DC shunt motor with armature resistance control** - While a 4-point starter *can* be used, speed control by armature resistance is typically done below base speed and doesn't pose the field weakening issue. A 3-point starter could also be used here, though 4-point is often used as a standard for shunt motors. The phrase "field weakening control" in option 2 makes it more specific to the primary advantage of the 4-point starter over the 3-point.
- **Option 2: DC shunt motor with field weakening control** - This is the most appropriate application for a 4-point starter. The independent connection of the NVC in a 4-point starter prevents the handle from releasing when the field is weakened for speed control.
- **Option 3: DC series motor** - DC series motors are typically started using 2-point or 3-point starters. A 4-point starter is not commonly used for series motors, and speed control is not typically done by field weakening (which would lead to dangerously high speeds under light load).
- **Option 4: DC compound motor** - 4-point starters can be used for compound motors. However, the option doesn't specify the type of speed control. If speed control by field weakening is used, a 4-point starter would be preferred over a 3-point starter for the same reasons as the shunt motor. Without specifying the control method, option 2 is more precise regarding the specific advantage of the 4-point starter.

Therefore, the 4-point starter is particularly well-suited and often specified for DC shunt motors when speed control is achieved using field weakening.

Revision Table: Starter Applications

Starter Type	Typical Motor Type	Key Advantage/Feature
2-Point Starter	DC Series Motor	Simple, single-line connection, no field circuit to protect.
3-Point Starter	DC Shunt, Compound	Common, provides no-volt and overload protection. NVC in series with field.
4-Point Starter	DC Shunt, Compound	Provides no-volt and overload protection. NVC independent of field circuit, suitable for field weakening.

Additional Information: DC Motor Speed Control

The speed of a DC motor is given by the formula:

$$N \propto \frac{V - I_a R_a}{\Phi}$$

, where N is speed, V is terminal voltage,

$$I_a$$

is armature current,

$$R_a$$

is armature resistance, and

$$\Phi$$

is the magnetic flux (produced by the field winding).

There are primary methods to control the speed:

- **Armature Resistance Control:** Adding external resistance in series with the armature (

$$R_{ext}$$

). This increases the total armature circuit resistance (

$$R_a + R_{ext}$$

), reducing the voltage drop across the armature terminals (

$$V - I_a(R_a + R_{ext})$$

) for a given current, thus decreasing speed. This method is inefficient (power loss in the resistor) and provides speed control below the base speed. The flux remains constant (assuming constant field voltage).

- **Field Control (Field Weakening):** Decreasing the field current (

$$I_f$$

) reduces the magnetic flux (Φ). Since speed is inversely proportional to flux (

$$N \propto 1/\Phi$$

), reducing flux increases the motor speed. This method is efficient and provides speed control above the base speed. This is where the 4-point starter becomes important for shunt motors, as reducing field current doesn't trip the starter.

---

16. **Answer: b**

Explanation:

Synchronous Motor Operation with Reversed Excitation

This question explores the behavior of a three-phase, salient pole synchronous motor connected to an infinite bus when its field excitation is manipulated.

Initial State Analysis

The motor is initially operating at no load with normal excitation. In this state:

- It is connected to an **infinite bus**, meaning the stator voltage (V) and frequency are constant.
- Being at **no load**, the mechanical power developed is minimal, primarily covering rotational losses (friction, windage, core losses).
- **Normal excitation** implies the field flux is at a typical level, usually resulting in operation near unity power factor and a small power angle (δ). The armature current (I_a) is relatively low.

Effect of Reducing Excitation to Zero

When the field excitation (E_f) is reduced from its normal value down to zero:

- The magnitude of the internally generated voltage (E_f) decreases.
- To maintain synchronism, the power angle (δ) might increase slightly.
- The armature current (I_a) may change slightly but remains low because the motor is still essentially unloaded. The motor might rely more on the reluctance torque produced due to the **salient pole** structure.

Effect of Increasing Reverse Excitation

The critical part is increasing the field excitation in the **reverse direction** gradually.

- **Initial Increase in Armature Current:** As the reverse excitation starts from zero and its magnitude gradually increases, the polarity of the rotor's magnetic field is flipped relative to the stator's rotating field. The motor must now synchronize or maintain synchronism under these reversed field conditions. This synchronization process requires developing torque. The armature current (I_a) increases to supply the necessary magnetizing and torque-producing MMF. The stronger the reversed excitation becomes (from zero), the more current is needed initially to establish the torque interaction and pull the rotor into sync, potentially involving both reluctance and excitation-based torque.
- **Subsequent Steep Decrease in Armature Current:** As the reverse excitation is increased further, the motor's operating characteristics change significantly. The interaction between the constant stator voltage (V), the strengthening reversed internal voltage $(E_{f_{rev}})$, the armature current (I_a), and crucially, the saliency effects (different synchronous reactances X_d and X_q along the direct

and quadrature axes) determines the behavior. It is possible that beyond a certain level of reversed excitation, the motor finds a new stable operating point. This could involve a significant shift in the power angle (δ) or a change in the power factor. The saliency might dominate the torque production in a way that requires less armature current magnitude (I_a) for a given operating condition, or the system enters a different regime (e.g., closer to generating or a specific stable point), causing the armature current to drop sharply or "steeply".

Therefore, the sequence observed is that the armature current first increases as the motor establishes synchronism under reversed excitation and then decreases steeply as the level of reverse excitation is further increased, leading to a change in the operating state.

---

17. **Answer: d**

Explanation:

$$SIL = \frac{V^2}{Z_C} = \frac{400 \times 400}{400} = 400MW$$

The carry 500 MW a double circuit 400 kV line is used.

---

18. **Answer: b**

Explanation:

Understanding Vector Orthogonality

The question asks us to determine the relationship between two vectors: $v_1 = [1, 1, 1]$ and $v_2 = [1, a, a^2]$, where $a = \left(-\frac{1}{2} + j\frac{\sqrt{3}}{2}\right)$.

Two vectors are considered **orthogonal** if their dot product (or inner product) equals zero. The dot product for complex vectors is calculated using the formula: $v_1 \cdot v_2 =$

$\sum_i v1_i \overline{v2_i}$, where $\overline{v2_i}$ is the complex conjugate of the components of the second vector.

Analyzing the Complex Number 'a'

The given complex number $a = -\frac{1}{2} + j\frac{\sqrt{3}}{2}$ is a complex cube root of unity. It can be represented in polar form as $e^{j2\pi/3}$.

Key properties of this complex number a are:

- $a^2 = \left(-\frac{1}{2} + j\frac{\sqrt{3}}{2}\right)^2 = \left(-\frac{1}{2} - j\frac{\sqrt{3}}{2}\right) = e^{j4\pi/3}$
- $a^3 = 1$
- $1 + a + a^2 = 0$
- The complex conjugate of a is $\overline{a} = -\frac{1}{2} - j\frac{\sqrt{3}}{2}$, which is equal to a^2.
- The complex conjugate of a^2 is $\overline{a^2} = -\frac{1}{2} + j\frac{\sqrt{3}}{2}$, which is equal to a.
- The magnitude of a and a^2 is 1: $|a| = \sqrt{(-\frac{1}{2})^2 + (\frac{\sqrt{3}}{2})^2} = \sqrt{\frac{1}{4} + \frac{3}{4}} = \sqrt{1} = 1$. Similarly, $|a^2| = 1$.

Calculating the Dot Product

Let's calculate the dot product of the given vectors $v_1 = [1, 1, 1]$ and $v_2 = [1, a, a^2]$:

$$v_1 \cdot v_2 = (1 \times \overline{1}) + (1 \times \overline{a}) + (1 \times \overline{a^2})$$

Substitute the complex conjugates:

$$v_1 \cdot v_2 = (1 \times 1) + (1 \times a^2) + (1 \times a)$$

$$v_1 \cdot v_2 = 1 + a^2 + a$$

Using the property $1 + a + a^2 = 0$, we find:

$$v_1 \cdot v_2 = 0$$

Evaluating Other Options

- **Orthonormal:** For vectors to be orthonormal, they must be orthogonal and have a magnitude of 1. The magnitude of v_1 is $|v_1| = \sqrt{1^2 + 1^2 + 1^2} = \sqrt{3}$. Since the magnitude is not 1, the vectors are not orthonormal.

- **Parallel/Collinear:** Vectors are parallel or collinear if one is a scalar multiple of the other $(v_1 = k \cdot v_2)$. Here, $1 = k \cdot 1 \implies k = 1$. This would mean $v_1 = v_2$, so $[1, 1, 1] = [1, a, a^2]$. This requires $a = 1$ and $a^2 = 1$, which contradicts the given value of a. Therefore, the vectors are not parallel or collinear.

Conclusion

Since the dot product of the two vectors is 0, the vectors are **orthogonal**.

---

19. Answer: a

Explanation:

<u>Concept</u> :

The steady-state error for a system is defined as:

$e_{ss} = \lim_{s \to 0} \frac{sR(s)}{1+G(s)H(s)}$

R(s) = Input

G(s) = open loop transfer function

H(s) = feedback gain = 1 for unity feedback system

<u>Analysis</u> :

For a unit step input, we have:

$R(s) = \frac{1}{s}$

The steady-state error becomes:

$e_{ss} = \lim_{s \to 0} \frac{s\left(\frac{1}{s}\right)}{1+G(s)} = 0.1$

1 + G(0) = 10

G(0) = 9

Again we have the input as:

r(t) = 10 [u(t) – u(t - 1)]

$R(s) = 10\left[\frac{1}{s} - \frac{1}{s}e^{-s}\right]$

$= 10\left(\frac{1-e^{-s}}{s}\right)$

∴ The steady-state error for the input will be:

$e'_{ss} = \lim_{s \to 0} \frac{s \times 10 \frac{(1-e^{-s})}{s}}{1+G(s)}$

$= \frac{10(1-e^{0})}{1+9}$

$e'_{ss} = 0$

---

20. **Answer: b**

Explanation:

Wattmeter Compensating Coil Function

A wattmeter is an instrument used to measure the electrical power (the rate at which electrical energy is transferred) in an electrical circuit. However, inaccuracies can arise, especially when measuring power at low power factors. In electrodynamometer-type wattmeters, these errors can stem from the impedance characteristics of the instrument's coils. A compensating coil is often added to specific types of wattmeters, particularly those designed for low power factor measurements, to improve accuracy. Let's analyze the given statements regarding its function.

Analysis of Compensating Coil Statements

Statement (i) Analysis: Compensating Coil and Current Coil Impedance

Statement (i) suggests that the compensating coil compensates for the effect of the impedance of the **current coil**.

- The current coil (or series coil) in a wattmeter is designed to have a very low impedance. This is typically achieved by using thick wire with a few turns, allowing it to carry the load current with minimal voltage drop.
- Because its impedance is already very low, the error introduced by the current coil's impedance is generally negligible in most practical scenarios.
- The primary source of error that the compensating coil addresses is related to the voltage coil circuit, not the current coil.

Therefore, statement (i) is **false**. The compensating coil does not primarily function to compensate for the current coil's impedance.

Statement (ii) Analysis: Compensating Coil and Voltage Coil Circuit Impedance

Statement (ii) suggests that the compensating coil compensates for the effect of the impedance of the **voltage coil circuit**.

- The voltage coil (or pressure coil) circuit consists of a high resistance connected in series with the voltage coil itself. This circuit is connected in parallel with the load.
- This voltage coil circuit has a significant inductance and resistance. The inductance causes the current in the voltage coil circuit to lag slightly behind the voltage across it.
- At low power factors, the load current is small, and the error caused by the phase difference between the actual voltage and the voltage across the potential coil becomes more significant. The inductance of the voltage coil circuit is the main contributor to this error.
- The compensating coil is specifically designed to counteract this effect. It is usually a small coil wound in opposition to the voltage coil, connected in series with the voltage coil circuit. It is designed to introduce a small inductance that effectively cancels out the inductive effect of the main voltage coil, thereby ensuring more accurate readings, especially at low power factors.

Therefore, statement (ii) is **true**. The compensating coil's main purpose is to correct for errors arising from the inductance present in the voltage coil circuit.

Conclusion

Based on the analysis, statement (i) is false, and statement (ii) is true. The compensating coil is essential for maintaining accuracy in low power factor measurements by compensating for the inductive impedance of the voltage coil circuit.

---

21. **Answer: d**

Explanation:

Understanding Cascaded Filters: LPF and HPF Combination

This question involves understanding how different types of electronic filters behave when connected in series (cascaded). We have a low-pass filter and a high-pass filter connected one after the other.

Low-Pass Filter (LPF) Behavior

A **low-pass filter** is designed to allow signals with frequencies lower than a specific point, called the **cut-off frequency**, to pass through while reducing or blocking signals with frequencies above this point.

- In this case, the LPF has a cut-off frequency $(f_{c,LPF})$ of 30 Hz.
- This means it allows frequencies below 30 Hz $(f < 30 \text{ Hz})$ to pass.
- It attenuates frequencies above 30 Hz $(f > 30 \text{ Hz})$.

High-Pass Filter (HPF) Behavior

A **high-pass filter** does the opposite: it allows signals with frequencies higher than its cut-off frequency to pass and blocks or attenuates frequencies below it.

- The HPF in this system has a cut-off frequency ($f_{c,HPF}$) of 20 Hz.
- This means it allows frequencies above 20 Hz ($f > 20 \text{ Hz}$) to pass.
- It attenuates frequencies below 20 Hz ($f < 20 \text{ Hz}$).

Cascading Filters: Combined Effect

When filters are connected in series (cascaded), the signal must pass through each filter sequentially. For a signal to successfully pass through the entire system, it must be allowed to pass by **both** the LPF and the HPF.

- The LPF passes frequencies where $f < 30 \text{ Hz}$.
- The HPF passes frequencies where $f > 20 \text{ Hz}$.
- Therefore, the combined system only passes frequencies that meet both conditions simultaneously: $f > 20 \text{ Hz}$ AND $f < 30 \text{ Hz}$.

This combined condition can be written as:

$$20 \text{ Hz} < f < 30 \text{ Hz}$$

Resultant Filter Type

A filter system that allows signals within a specific range of frequencies to pass, while blocking frequencies outside this range, is known as a **band-pass filter**.

The cascaded system of an LPF with a 30 Hz cut-off and an HPF with a 20 Hz cut-off allows frequencies between 20 Hz and 30 Hz to pass. This precisely matches the definition of a band-pass filter.

- **All-Pass Filter:** Passes all frequencies but shifts phase. Not the case here.
- **All-Stop Filter:** Blocks all frequencies. Not the case here.
- **Band-Stop Filter:** Blocks a specific band of frequencies. The opposite of what happens here.
- **Band-Pass Filter:** Passes a specific band of frequencies ($20 \text{ Hz} < f < 30 \text{ Hz}$). This is the correct description.

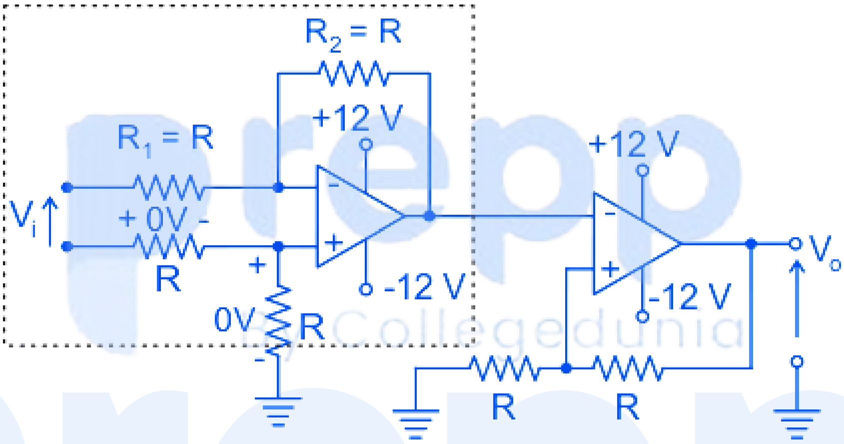
---

22. Answer: d

Explanation:

Concept:

When the symbol represents the input ↑, then the arrow part represents the potential of V_i, and the bottom part is the 0 V or ground.



The output of inverting OP-AMP is given by:

$V_{o1} = V_i(\frac{-R_2}{R_1})$

$V_{o1} = -V_i$

The second circuit is a Schmitt trigger circuit.

Positive feedback in the OP-AMP forms a Schmitt trigger circuit.

The output of the Schmitt trigger circuit is always either $\pm V_{sat}$.

Case 1: When $V_I < V_{NI}$

$V_o = +V_{sat}$

Case 2: When VI > VNI

$Vo = -V_{sat}$

The input at V_{NI} is either V_{UTP} or V_{LTP}

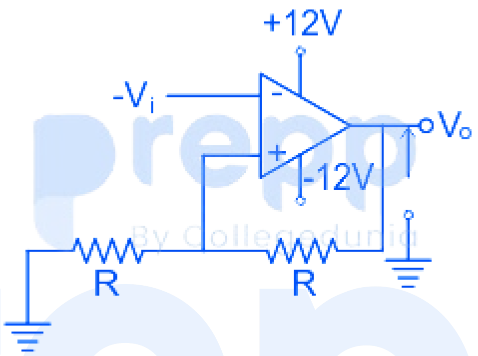
$V_{UTP} = +V_{sat}\left(\frac{R}{2R}\right)$

V UTP = 6V

$V_{LTP} = -V_{sat}\left(\frac{R}{2R}\right)$

V_L TP = -6V

Calculation:



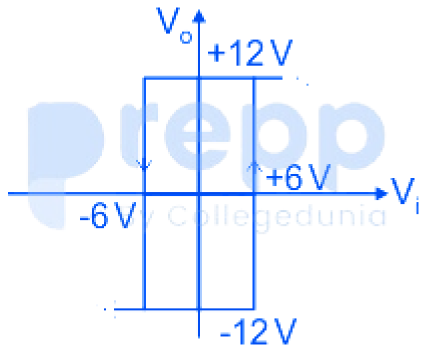
Case 1: When -V i < 6V (or V_i > 6V)

$V_O = +V_{sat} = 12$

Case 2: When -Vi > 6V (or Vi < 6V)

VO = -Vsat = -12

The transfer characteristic curve is:



23. Answer: a

Explanation:

It has to carry unidirectional current and has to block bidirectional voltage.

Only option A allow bi direction power flow from source to the drive

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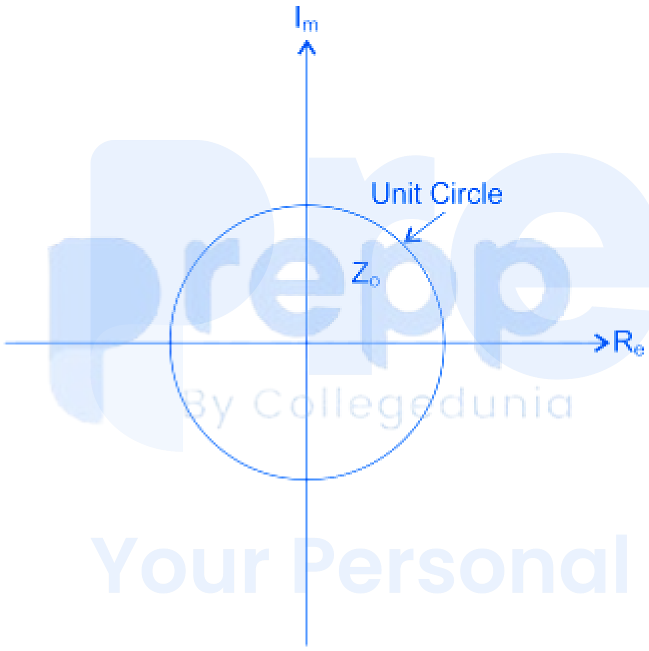
24. **Answer: d**

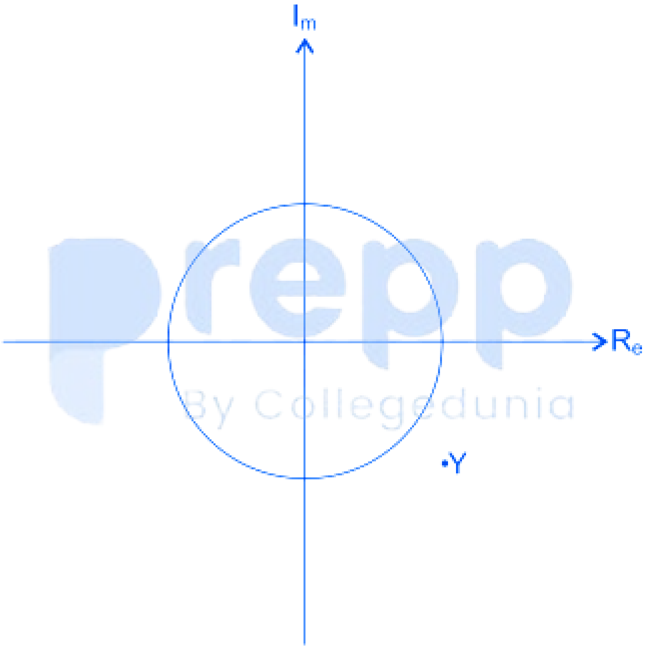
Explanation:



$\bar{Z}$ is |Z| = 0 where θ is around 45° or so

$\therefore \bar{Z}$ = |Z| with angle 45° where |Z| <1

$$\bar{Y} = \frac{1}{z} = \frac{1}{|Z|angle(45^\circ)} = \frac{1}{|Z|}angle\,(-45^\circ)$$
$$\left|\bar{Y}\right| > 1(because\,|Z| < 1)$$



So Y will be out of unity circle.

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25. **Answer: b**

Explanation:

To determine the phasor representation of the current, we first need to express both the voltage and current in their standard sinusoidal forms, typically involving cosine, and then convert them into phasor notation.

Voltage Phasor Representation

The given voltage is $V(t) = 100\sqrt{2}\cos(100\,\pi t)$ volts.

This voltage is already in the standard cosine form $V(t) = V_m \cos(\omega t + \phi)$, where:

- $V_m = 100\sqrt{2}$ V (Maximum voltage)
- $\omega = 100\,\pi$ rad/s (Angular frequency)
- $\phi = 0$ radians (Phase angle)

The RMS value of the voltage is $V_{rms} = \frac{V_m}{\sqrt{2}} = \frac{100\sqrt{2}}{\sqrt{2}} = 100$ V. When voltage is taken as the reference phasor, its phase angle is considered 0.

The phasor representation of voltage can be written as $\mathbf{V} = V_{rms}\angle\phi = 100\angle 0^\circ$ or $100\angle 0$. If we consider the peak value, it is $100\sqrt{2}\angle 0$.

Current Phasor Representation

The given current is $I(t) = 10\sqrt{2}\sin(100\pi t + \pi/4)$ amperes.

To convert this to the standard cosine form, we use the trigonometric identity $\sin(\theta) = \cos(\theta - \pi/2)$.

Applying this identity:

$I(t) = 10\sqrt{2}\cos(100\pi t + \pi/4 - \pi/2)$

$I(t) = 10\sqrt{2}\cos(100\pi t - \pi/4)$

Now, the current is in the form $I(t) = I_m\cos(\omega t + \phi')$, where:

- $I_m = 10\sqrt{2}$ A (Maximum current)
- $\omega = 100\,\pi$ rad/s (Angular frequency)
- $\phi' = -\pi/4$ radians (Phase angle)

The RMS value of the current is $I_{rms} = \frac{I_m}{\sqrt{2}} = \frac{10\sqrt{2}}{\sqrt{2}} = 10$ A.

The phasor representation of the current uses the RMS value and the phase angle. Taking the voltage as the reference (phase angle 0), the current's phase angle is $-\pi/4$.

Therefore, the phasor representation of the current is:

$\mathbf{I} = I_{rms}\angle\phi' = 10\angle(-\pi/4)$

This can also be written as $10\angle - \pi/4$.

Conclusion

Comparing this result with the given options, the correct phasor representation of the current is $10\angle - \pi/4$.

26. **Answer: a**

Explanation:

For maximum power transfer to R $_L$,

R should be zero so that maximum current flows through the load resistance. and hence maximum power is transferred to the load resistance.

Common Mistake :

Maximum power theorem states that for maximum power to be transferred to the load resistance R L, R L must equal the Thevenin Equivalent resistance, i.e.

R L = R th

But here, we are asked to find the value of R, and not R L that will result in the maximum power to be transferred to load R L. So we cannot go by the standard procedure of equating R L with the Thevenin equivalent resistance.

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27. **Answer: a**

Explanation:

Convolution of Continuous Time Signals

This explanation details the process of finding the **convolution** of two specific **continuous time signals**: $x(t) = e^{-t}$ and $y(t) = e^{-2t}$. Both signals are defined for the time domain where $t > 0$. Convolution is a mathematical operation used extensively in signal processing and system analysis to determine the output of a system given its input signal and impulse response.

Convolution Formula Application

The definition of convolution for continuous-time signals $x(t)$ and $y(t)$ is given by the integral:

$$z(t) = (x * y)(t) = \int_{-\infty}^{\infty} x(\tau)y(t-\tau)d\tau$$

The signals provided are:

- $x(t) = e^{-t}$ for $t > 0$
- $y(t) = e^{-2t}$ for $t > 0$

To use the convolution formula, we can express these signals using the unit step function, $u(t)$, which is 1 for $t \geq 0$ and 0 otherwise. This correctly represents that the signals exist only for $t > 0$.

- $x(t) = e^{-t}u(t)$
- $y(t) = e^{-2t}u(t)$

Substituting these into the convolution integral:

$$z(t) = \int_{-\infty}^{\infty} (e^{-\tau}u(\tau))(e^{-2(t-\tau)}u(t-\tau))d\tau$$

The product of the unit step functions, $u(\tau)u(t-\tau)$, is equal to 1 only when both $\tau > 0$ and $t - \tau > 0$. The condition $t - \tau > 0$ implies $\tau < t$. Therefore, the integral is non-zero only over the interval $0 < \tau < t$. This condition also implies that the resulting convolution $z(t)$ is meaningful only for $t > 0$.

The convolution integral simplifies to:

$$z(t) = \int_{0}^{t} e^{-\tau}e^{-2(t-\tau)}d\tau \quad \text{for } t > 0$$

Step-by-Step Calculation of the Integral

First, simplify the expression inside the integral:

$$e^{-\tau}e^{-2(t-\tau)} = e^{-\tau}e^{-2t+2\tau} = e^{-2t+(-\tau+2\tau)} = e^{-2t+\tau}$$

We can rewrite $e^{-2t+\tau}$ as $e^{-2t}e^{\tau}$. Now substitute this back into the integral:

$$z(t) = \int_0^t e^{-2t} e^{\tau} d\tau$$

Since e^{-2t} is a constant with respect to the integration variable τ, it can be taken outside the integral:

$$z(t) = e^{-2t} \int_0^t e^{\tau} d\tau$$

Next, evaluate the definite integral of e^{τ}:

$$\int_0^t e^{\tau} d\tau = [e^{\tau}]_0^t$$

Applying the limits of integration:

$$[e^{\tau}]_0^t = e^t - e^0 = e^t - 1$$

Now, substitute this result back into the equation for $z(t)$:

$$z(t) = e^{-2t}(e^t - 1)$$

Finally, distribute e^{-2t} to find the final expression for the convolution:

$$z(t) = (e^{-2t} \cdot e^t) - (e^{-2t} \cdot 1)$$

$$z(t) = e^{-2t+t} - e^{-2t}$$

$$z(t) = e^{-t} - e^{-2t}$$

This result, $e^{-t} - e^{-2t}$, is valid for $t > 0$.

Matching the Result with Provided Options

After calculating the convolution, we compare the result $z(t) = e^{-t} - e^{-2t}$ with the given options:

- Option 1: $e^{-t} - e^{-2t}$
- Option 2: $e^{-t} - e^{2}$
- Option 3: $e^{-t} + e^{2t}$
- Option 4: $e^{-t} + e^{-2t}$

The calculated convolution $z(t) = e^{-t} - e^{-2t}$ directly matches Option 1.

28. **Answer: b**

Explanation:

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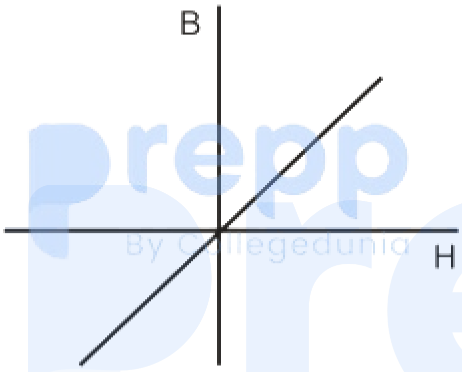
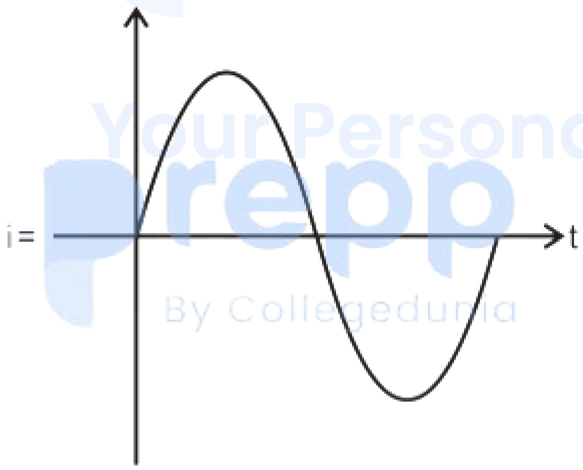
∵



∴ i =



29. **Answer: a**

Explanation:

Negative Sequence Relay: Essential for Alternator Protection

A negative sequence relay is a specialized protective device designed to detect and respond to unbalanced current conditions within a power system. These relays are crucial for safeguarding sensitive electrical equipment from potential damage caused by negative sequence currents.

Alternator Protection by Negative Sequence Relay

The presence of negative sequence currents can be highly detrimental, especially to rotating electrical machines like alternators (synchronous generators). Here's why a negative sequence relay is commonly used to protect an alternator:

- **Unbalanced Faults and Loads:** Unbalanced faults, such as single line-to-ground, line-to-line, or double line-to-ground faults, introduce negative sequence components into the current. Similarly, unbalanced loading conditions can also lead to the flow of negative sequence currents.
- **Rotor Heating Mechanism:** When negative sequence currents flow through the stator windings of an alternator, they produce a magnetic field that rotates in the opposite direction to the rotor's normal rotation. This counter-rotating field induces currents in the rotor body, rotor winding, and damper windings.
- **Double Frequency Currents:** The induced currents in the rotor occur at a frequency twice that of the system frequency (e.g., 100 Hz for a 50 Hz system). These high-frequency currents cause significant localized heating within the rotor.
- **Localized Damage:** The localized heating is particularly severe in areas like the rotor teeth, wedges, and end-bell regions. Prolonged exposure to these currents can lead to the degradation of rotor insulation, mechanical stress, and even melting of rotor conductors, severely damaging the alternator.
- **Limited Withstand Capability:** Alternators have a very limited thermal withstand capability for negative sequence currents, often characterized by an I_2^2t (negative sequence current squared multiplied by time) constant. Exceeding this limit even for short durations can cause irreversible damage.

The negative sequence relay continuously monitors the negative sequence current component. If this component exceeds a predetermined safe limit or its I_2^2t value surpasses the alternator's thermal withstand capability, the relay issues a trip signal to disconnect the alternator, thereby preventing severe damage.

Why Not Other Devices?

While unbalanced currents can affect other components of a power system, the impact on their operation and lifespan is generally less critical compared to the direct and severe heating effect on an alternator's rotor:

- **Transformers:** Transformers are primarily protected by differential, overcurrent, and earth fault relays. While unbalanced currents can cause some heating, the unique rotor heating issue seen in alternators is not present.
- **Transmission Lines:** Transmission lines are protected by distance relays and overcurrent relays. Unbalanced currents on lines mostly lead to unbalanced voltage drops and increased losses rather than direct component damage akin to an alternator's rotor.
- **Bus Bars:** Bus bars are typically protected by differential relays. They are static components and do not experience the rotational dynamics and induced rotor currents that make alternators uniquely vulnerable to negative sequence currents.

Therefore, due to the specific and highly damaging effects of negative sequence currents on the rotor of an alternator, the negative sequence relay is predominantly and critically employed for alternator protection.

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30. **Answer: c**

Explanation:

Enhancing EHV Power Transmission with Series Capacitive Compensation

The question asks for the most effective method to improve power transmission along Extra High Voltage (EHV) transmission lines. EHV lines are crucial for transporting large amounts of power over long distances. However, they face challenges like significant inductive reactance and voltage drops, which limit their power transfer capability.

Understanding Power Transmission Challenges in EHV Lines

The power (P) that can be transmitted through a line is primarily dependent on the sending and receiving end voltages (V_s, V_r) and the line's total reactance (X_{line}). The relationship can be simplified as:

$$P \approx \frac{|V_s||V_r|}{X_{line}} \sin(\delta)$$

where δ is the power angle.

EHV lines inherently have high inductive reactance (X_L) due to the conductors and configuration. This high reactance limits the maximum power that can be transmitted and reduces the system's stability margin.

Evaluating Compensation Methods

To enhance power transmission, we need methods that counteract the negative effects of line inductance. Compensation techniques involve adding components either in series or in parallel (shunt) with the line.

- **Series Compensation:** Components are connected directly in series with the transmission line conductor.
- **Shunt Compensation:** Components are connected in parallel across the line, typically at substations.

Why Series Capacitive Compensation is Preferred

Connecting a **series capacitive compensator** (like a Thyristor Controlled Series Capacitor - TCSC, or a fixed series capacitor) directly addresses the problem of high line reactance.

- **Reduces Net Reactance:** Capacitors provide capacitive reactance (X_C) which subtracts from the line's inductive reactance (X_L). The effective reactance (X_{net}) becomes $X_{net} = X_L - X_C$. By reducing X_{net}, the power transfer capability increases significantly, as seen from the power formula $P \propto \frac{1}{X_{net}}$.
- **Improves Stability:** Lowering the overall reactance reduces the angle δ required for a given power transfer, thereby improving transient stability and increasing the steady-state stability limit.
- **Enhances Voltage Profile:** Series capacitors can help improve voltage regulation along the line, especially under heavy load conditions.

Analysis of Other Options

- **Series inductive compensator:** Adding inductance in series would increase the total reactance ($X_{net} = X_L + X_L$) further, reducing power transfer capability, which is undesirable.
- **Shunt inductive compensator:** Shunt reactors are primarily used to absorb excess reactive power and control overvoltages, particularly during light load conditions due to line charging capacitance. They do not increase the power transfer capability by reducing series reactance.
- **Shunt capacitive compensator:** Shunt capacitors are used to supply reactive power, improve the power factor, and support voltage, typically near loads or at the sending end. While beneficial for voltage support, they don't directly reduce the series reactance limiting power transfer.

Conclusion

For the specific purpose of **enhancing power transmission** along EHV lines, reducing the effective series reactance is the most direct and effective strategy. Series capacitive compensation achieves this by introducing a capacitive reactance in series, thereby lowering the net reactance and increasing the power transfer capability and stability limits.

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31. **Answer: b**

Explanation:

This question asks us to determine the stability and phase type (minimum or non-minimum phase) of an open-loop control system given its transfer function.

Stability Analysis of the Transfer Function

The stability of a linear time-invariant (LTI) system is determined by the location of the poles of its transfer function in the s-plane. A system is stable if and only if all its poles lie in the left half of the s-plane (LHP), meaning they have negative real parts.

The given transfer function is:

$$G(s) = \frac{(s-1)}{(s+2)(s+3)}$$

To find the poles, we set the denominator of the transfer function to zero:

$$(s+2)(s+3) = 0$$

Solving for s gives the poles:

- $s+2=0 \implies s=-2$
- $s+3=0 \implies s=-3$

Both poles, $s=-2$ and $s=-3$, have negative real parts. Therefore, all poles lie in the left half of the s-plane. This indicates that the system is **stable**.

Phase Type Analysis

A system is classified as a minimum phase system if all its poles and zeros are located in the left half of the s-plane. If a system has any poles or zeros in the right half of the s-plane (RHP) or on the imaginary axis (excluding the origin for zeros), it is considered a non-minimum phase system.

We have already identified the poles as $s=-2$ and $s=-3$, both in the LHP.

Now, let's find the zeros by setting the numerator of the transfer function to zero:

$$s-1=0$$

Solving for s gives the zero:

- $s = 1$

The zero $s = 1$ has a positive real part, meaning it lies in the right half of the s-plane.

Since the transfer function has a zero in the RHP, the system is classified as a **non-minimum phase type** system.

Conclusion

Based on the analysis:

- The poles are $s = -2$ and $s = -3$, which are both in the LHP, indicating the system is **stable**.
- The system has a zero at $s = 1$, which is in the RHP, indicating the system is of the **non-minimum phase type**.

Therefore, the open-loop system is stable and of the non-minimum phase type.

---

32. **Answer: c**

Explanation:

Given bridge is Maxwell inductance – Capacitance Bridge and it is suitable for the measurement of medium 'Q' inductor.

⭐ Important Points

Type of Bridge	Name of Bridge	Used to measure	Important
DC Bridges	Wheatstone bridge	Medium resistance	
	Corey foster's bridge	Medium resistance	
	Kelvin double bridge	Very low resistance	
	Loss of charge method	High resistance	
	Megger	High insulation resistance	Resistance of cables
AC Bridges	Maxwell's inductance bridge	Inductance	Not suitable to measure Q
	Maxwell's inductance capacitance bridge	Inductance	Suitable for medium Q coil ($1 < Q < 10$)

	Hay's bridge	Inductance	Suitable for high Q coil (Q > 10), slowest bridge
	Anderson's bridge	Inductance	5-point bridge, accurate and fastest bridge (Q < 1)
	Owen's bridge	Inductance	Used for measuring low Q coils
	Heaviside mutual inductance bridge	Mutual inductance	
	Campbell's modification of Heaviside bridge	Mutual inductance	
	De-Sauty's bridge	Capacitance	Suitable for perfect capacitor
	Schering bridge	Capacitance	Used to measure relative permittivity
	Wein's bridge	Capacitance and frequency	Harmonic distortion analyzer, used as a notch filter, used in audio and high-frequency application s

33. **Answer: d**

Explanation:

Oscilloscope Alternate Mode Operation Explained

A dual-trace oscilloscope allows viewing two different signals simultaneously. It achieves this using different operating modes, including 'Alternate' mode. Understanding how these modes work is key to interpreting waveforms correctly.

Understanding Dual Trace Oscilloscope Modes

In a dual-trace oscilloscope, two signals (typically from Channel 1 and Channel 2) need to be displayed on the screen. The oscilloscope uses internal circuitry, including a multiplexer (MUX), to switch between these two signals. The way it switches determines the mode of operation.

Alternate Mode Functionality

In the **Alternate mode**, the oscilloscope switches from displaying the signal on Channel 1 to displaying the signal on Channel 2 on successive sweeps of the time base. This means that one full horizontal sweep displays the waveform for Channel 1, and the very next horizontal sweep displays the waveform for Channel 2. This provides a relatively stable display for both signals, especially at slower sweep speeds, as each channel gets the full sweep duration.

Multiplexer Control and Sweep Oscillator Synchronization

The switching between Channel 1 and Channel 2 is handled by a multiplexer. The control input of this multiplexer determines which channel's signal is currently being routed to the display amplifier. For the **Alternate mode** to function correctly, this switching must happen exactly once per sweep cycle.

The time base (or sweep) oscillator generates the signal that controls the horizontal sweep. The frequency of this oscillator, let's denote it as f_{sweep}, dictates how often a complete horizontal sweep occurs.

If the control input of the multiplexer receives a signal with a frequency, let's call it $f_{control}$, that is equal to the frequency of the time base oscillator ($f_{control} = f_{sweep}$), the multiplexer will switch channels precisely once for each sweep. This results in alternating displays of Channel 1 and Channel 2 on successive sweeps, which is the defining characteristic of the **Alternate mode**.

If $f_{control}$ were twice f_{sweep}, switching would occur twice per sweep. If $f_{control}$ were half f_{sweep}, switching would occur every two sweeps (characteristic of Chop mode). Therefore, synchronizing the multiplexer's control signal frequency with the time base oscillator frequency is crucial for Alternate mode.

---

34. Answer: a

Explanation:

XOR GATE

Symbol:

Truth Table:

Input A	Input B	Output Y = A ⊕ B
0	0	0
0	1	1
1	0	1
1	1	0

Output Equation: $Y = \mathbf{A} \oplus \mathbf{B} = \bar{A}B + A\bar{B}$

Key Points:

1) If B is always High, the output is the inverted value of the other input A, i.e. A̅

1) The output is low when both the inputs are the same.

2) The output is high when both the inputs are different.

Explanation:

$Y = \bar{\mathbf{X}} \oplus \mathbf{X} = \bar{\bar{X}}X + \bar{X}\bar{X}$

$Y = XX + \bar{X}\bar{X}$

$Y = X + \bar{X}$

Y = 1

⭐ Important Points

Name	AND Form	OR Form
Identity law	1.A=A	0+A=A
Null Law	0.A=0	1+A=1
Idempotent Law	A.A=A	A+A=A
Inverse Law	AA′=0	A+A′=1
Commutative Law	AB=BA	A+B=B+A
Associative Law	(AB)C	(A+B)+C = A+(B+C)
Distributive Law	A+BC=(A+B)(A+C)	A(B+C)=AB+AC
Absorption Law	A(A+B)=A	A+AB=A
De Morgan’s Law	(AB)′=A′+B′	(A+B)′=A′B′

---

35. Answer: c

Explanation:

SCR Turn Off Time Definition Explained

The question asks for the definition of the circuit turn off time for a Silicon Controlled Rectifier (SCR). The turn-off time is a critical parameter in SCR operation, especially in switching applications. It represents the time required for the SCR to stop conducting current after it has been triggered off.

Understanding SCR Turn Off Process

An SCR, once conducting, needs a specific process to stop conducting and return to its blocking state. This typically involves two main conditions:

- The anode current must fall below a certain minimum level known as the *holding current* (I_H).
- The SCR must be prevented from conducting again, usually by applying a reverse voltage across its terminals or ensuring the anode current remains below I_H for a sufficient duration.

The **circuit turn off time** specifically relates to the time interval during which these conditions are met, allowing the SCR to regain its forward blocking capability. This process is often facilitated by external circuitry called a *commutation circuit*.

Analysis of Options for SCR Turn Off Time

Let's analyze the provided options to determine the correct definition:

- Option 1:

  Taken by the SCR to turn off

  - This is too general and doesn't specify the conditions or requirements for turning off.
- Option 2:

  Required for the SCR current to become zero.

- While the current must reduce significantly, simply becoming zero doesn't guarantee the SCR is ready to block forward voltage. The internal charge carriers must also be cleared.

- Option 3:

  For which the SCR is reverse biased by the commutation circuit.

  - This option correctly identifies a key condition (reverse bias) and the responsible circuit element (commutation circuit) that ensures the SCR turns off and recovers its blocking state. The duration under this condition is the turn-off time.

- Option 4:

  For which the SCR is reverse biased to reduce its current below the holding current.

  - This option describes the purpose of the reverse bias (reducing current below I_H) but focuses more on the immediate effect rather than the time duration required for complete recovery, which is what turn-off time represents. The commutation circuit actively provides the necessary reverse bias for this recovery period.

The definition of turn-off time is fundamentally linked to the time the device spends in a state where it is forced to turn off, typically through reverse biasing facilitated by a commutation circuit, allowing it to clear internal charges and regain its blocking state.

Therefore, Option 3 provides the most accurate definition encompassing the necessary conditions and mechanism.

---

36. **Answer: d**

Explanation:

Understanding Matrix Decomposition

Matrix decomposition is a fundamental technique in linear algebra used to break down a complex matrix into a product of simpler matrices. This process is highly beneficial for simplifying complex mathematical operations, solving systems of linear equations efficiently, and analyzing the properties of matrices. A key type of matrix decomposition is the LU decomposition, where a given matrix is expressed as the product of a lower triangular matrix, denoted as [**L**], and an upper triangular matrix, denoted as [**U**].

LU Decomposition Explained

The core idea behind LU decomposition is to represent a matrix [**A**] as the product of two specific types of matrices:

- A **lower triangular matrix** ([**L**]) is a square matrix where all elements located above the main diagonal are zero.
- An **upper triangular matrix** ([**U**]) is a square matrix where all elements located below the main diagonal are zero.

The decomposition follows the equation:

$$[A] = [L][U]$$

Applying LU Decomposition to Matrix [A]

The specific matrix we need to decompose in this problem is:

$$[A] = \begin{bmatrix} 2 & 1 \\ 4 & -1 \end{bmatrix}$$

Our objective is to identify the correct pair of [**L**] and [**U**] matrices from the given options such that their product equals the original matrix [**A**]. We will achieve this by performing matrix multiplication for each option.

Verifying the Proposed L and U Matrices

Option 1 Verification

We examine the first option provided:

$[L] = \begin{bmatrix} 1 & 0 \\ 4 & -1 \end{bmatrix}$ and $[U] = \begin{bmatrix} 1 & 1 \\ 0 & -2 \end{bmatrix}$

To verify, we multiply these matrices:

$[L][U] = \begin{bmatrix} 1 & 0 \\ 4 & -1 \end{bmatrix} \begin{bmatrix} 1 & 1 \\ 0 & -2 \end{bmatrix}$
$= \begin{bmatrix} (1)(1) + (0)(0) & (1)(1) + (0)(-2) \\ (4)(1) + (-1)(0) & (4)(1) + (-1)(-2) \end{bmatrix}$
$= \begin{bmatrix} 1 & 1 \\ 4 & 4+2 \end{bmatrix} = \begin{bmatrix} 1 & 1 \\ 4 & 6 \end{bmatrix}$

The resulting matrix $\begin{bmatrix} 1 & 1 \\ 4 & 6 \end{bmatrix}$ does not match the original matrix [**A**].

Option 2 Verification

Next, we check the second option:

$[L] = \begin{bmatrix} 2 & 0 \\ 4 & -1 \end{bmatrix}$ and $[U] = \begin{bmatrix} 1 & 1 \\ 0 & 1 \end{bmatrix}$

Performing the matrix multiplication:

$[L][U] = \begin{bmatrix} 2 & 0 \\ 4 & -1 \end{bmatrix} \begin{bmatrix} 1 & 1 \\ 0 & 1 \end{bmatrix}$
$= \begin{bmatrix} (2)(1) + (0)(0) & (2)(1) + (0)(1) \\ (4)(1) + (-1)(0) & (4)(1) + (-1)(1) \end{bmatrix}$
$= \begin{bmatrix} 2 & 2 \\ 4 & 4-1 \end{bmatrix} = \begin{bmatrix} 2 & 2 \\ 4 & 3 \end{bmatrix}$

The result $\begin{bmatrix} 2 & 2 \\ 4 & 3 \end{bmatrix}$ is not equal to the original matrix [**A**].

Option 3 Verification

We evaluate the third option:

$[L] = \begin{bmatrix} 1 & 0 \\ 4 & 1 \end{bmatrix}$ and $[U] = \begin{bmatrix} 2 & 1 \\ 0 & -1 \end{bmatrix}$

Multiplying the matrices:

$$[L][U] = \begin{bmatrix} 1 & 0 \\ 4 & 1 \end{bmatrix} \begin{bmatrix} 2 & 1 \\ 0 & -1 \end{bmatrix}$$
$$= \begin{bmatrix} (1)(2)+(0)(0) & (1)(1)+(0)(-1) \\ (4)(2)+(1)(0) & (4)(1)+(1)(-1) \end{bmatrix}$$
$$= \begin{bmatrix} 2 & 1 \\ 8 & 4-1 \end{bmatrix} = \begin{bmatrix} 2 & 1 \\ 8 & 3 \end{bmatrix}$$

This product, $\begin{bmatrix} 2 & 1 \\ 8 & 3 \end{bmatrix}$, does not match the original matrix [**A**].

Option 4 Verification

Finally, we check the fourth option:

$[L] = \begin{bmatrix} 2 & 0 \\ 4 & -3 \end{bmatrix}$ and $[U] = \begin{bmatrix} 1 & 0.5 \\ 0 & 1 \end{bmatrix}$

Let's perform the multiplication:

$$[L][U] = \begin{bmatrix} 2 & 0 \\ 4 & -3 \end{bmatrix} \begin{bmatrix} 1 & 0.5 \\ 0 & 1 \end{bmatrix}$$
$$= \begin{bmatrix} (2)(1)+(0)(0) & (2)(0.5)+(0)(1) \\ (4)(1)+(-3)(0) & (4)(0.5)+(-3)(1) \end{bmatrix}$$
$$= \begin{bmatrix} 2 & 1 \\ 4 & 2-3 \end{bmatrix} = \begin{bmatrix} 2 & 1 \\ 4 & -1 \end{bmatrix}$$

This resulting matrix $\begin{bmatrix} 2 & 1 \\ 4 & -1 \end{bmatrix}$ is identical to the original matrix [**A**].

Conclusion: Correct LU Decomposition

Based on the verification process, the multiplication of the [**L**] and [**U**] matrices provided in Option 4 correctly yields the original matrix [**A**]. Therefore, this option represents the accurate LU decomposition.

The properly decomposed [**L**] and [**U**] matrices are:

$$[L] = \begin{bmatrix} 2 & 0 \\ 4 & -3 \end{bmatrix} \text{ and } [U] = \begin{bmatrix} 1 & 0.5 \\ 0 & 1 \end{bmatrix}$$

---

37. **Answer: a**

Explanation:

Understanding the Function f(x) = 2x – x² + 3

The question asks us to determine the nature of the function $f(x) = 2x - x^2 + 3$ at specific points, specifically whether it has a maxima or minima and at which value of x. To do this, we can use calculus, specifically finding the first and second derivatives of the function.

Step 1: Find the First Derivative

The first derivative of a function, denoted as $f'(x)$ or $\frac{df}{dx}$, helps us find the points where the function's slope is zero, which are potential locations for maxima or minima (critical points).

Given function: $f(x) = 2x - x^2 + 3$

Calculating the first derivative:

$$f'(x) = \frac{d}{dx}(2x - x^2 + 3)$$

Using the power rule for differentiation ($\frac{d}{dx}(ax^n) = anx^{n-1}$):

$$f'(x) = 2(1)x^{1-1} - 2x^{2-1} + 0$$

$$f'(x) = 2 - 2x$$

Step 2: Identify Critical Points

Critical points occur where the first derivative is equal to zero ($f'(x) = 0$) or undefined. Since $f'(x) = 2 - 2x$ is a linear function, it is defined for all real numbers. We set it to zero to find potential maxima or minima:

$$2 - 2x = 0$$

Solving for x:

$$2 = 2x$$

$$x = \frac{2}{2}$$

$$x = 1$$

Thus, the only critical point for this function is at $x = 1$.

Step 3: Find the Second Derivative

To determine whether the critical point corresponds to a maximum or minimum, we use the second derivative test. The second derivative, denoted as $f''(x)$ or $\frac{d^2f}{dx^2}$, tells us about the concavity of the function.

Calculating the second derivative from $f'(x) = 2 - 2x$:

$$f''(x) = \frac{d}{dx}(2 - 2x)$$

$$f''(x) = 0 - 2(1)x^{1-1}$$

$$f''(x) = -2$$

Step 4: Apply the Second Derivative Test

The second derivative test states:

- If $f''(c) < 0$, the function has a local maximum at $x = c$.
- If $f''(c) > 0$, the function has a local minimum at $x = c$.
- If $f''(c) = 0$, the test is inconclusive.

We evaluate the second derivative at our critical point, $x = 1$:

$$f''(1) = -2$$

Since $f''(1) = -2$, which is less than 0, the function has a local **maxima** at $x = 1$. The second derivative is a constant negative value, indicating the parabola opens downwards.

Conclusion

Based on our analysis using the first and second derivatives:

- The function $f(x) = 2x - x^2 + 3$ has a critical point at $x = 1$.
- The second derivative at $x = 1$ is negative ($f''(1) = -2$).
- Therefore, the function has a **maxima** at $x = 1$.

Comparing this result with the given options, the correct statement is that the function has only a maxima at $x = 1$.

---

38. **Answer: c**

Explanation:

Lossy Capacitor Equivalent Circuit Analysis

This problem involves a lossy capacitor (C_x) modeled using an equivalent circuit consisting of an ideal capacitor (C_p) connected in parallel with a resistor (R_p). We need to determine the power loss and the loss tangent ($\tan\delta$) when the capacitor operates at its rated voltage and frequency.

Given Information:

- Rated Voltage (V): $5 \text{ kV} = 5000 \text{ V}$
- Frequency (f): 50 Hz
- Parallel Capacitance (C_p): $0.102 \ \mu\text{F} = 0.102 \times 10^{-6} \text{ F}$
- Parallel Resistance (R_p): $1.25 \text{ M}\Omega = 1.25 \times 10^{6} \ \Omega$

Power Loss Calculation for Capacitor

Power loss in this circuit primarily occurs due to the current flowing through the parallel resistor (R_p). The power dissipated by this resistor can be calculated using the formula:

$$P = \frac{V^2}{R_p}$$

Substitute the given rated voltage and parallel resistance values:

$$P = \frac{(5000 \text{ V})^2}{1.25 \times 10^6 \ \Omega}$$

First, calculate the square of the voltage:

$$V^2 = (5000)^2 = 25,000,000 = 25 \times 10^6 \text{ V}^2$$

Now, perform the division:

$$P = \frac{25 \times 10^6 \text{ V}^2}{1.25 \times 10^6 \ \Omega} = \frac{25}{1.25} \text{ W}$$

$$P = 20 \text{ W}$$

The power loss of the capacitor is 20 Watts.

Loss Tangent (tan δ) Calculation

The loss tangent $(\tan\delta)$ quantifies the energy dissipation in the capacitor's dielectric. For the parallel RC circuit model, it is calculated using the formula:

$$\tan\delta = \frac{1}{\omega C_p R_p}$$

Where ω represents the angular frequency.

Calculate the angular frequency (ω) using the given frequency (f):

$$\omega = 2\pi f$$

$$\omega = 2\pi(50 \text{ Hz}) = 100\pi \text{ rad/s}$$

Now, plug the values of ω, C_p, and R_p into the $\tan\delta$ formula:

$$\tan\delta = \frac{1}{(100\pi \text{ rad/s}) \times (0.102 \times 10^{-6} \text{ F}) \times (1.25 \times 10^6 \ \Omega)}$$

Simplify the denominator:

$$100\pi \times 0.102 \times 1.25 = 100\pi \times 0.1275 = 12.75\pi$$

So, the formula becomes:

$$\tan\delta = \frac{1}{12.75\pi}$$

Using the approximate value of $\pi \approx 3.14159$:

$$\tan\delta \approx \frac{1}{12.75 \times 3.14159} \approx \frac{1}{40.055}$$

$$\tan\delta \approx 0.02496$$

Rounding to three decimal places, we get $\tan\delta \approx 0.025$.

Summary of Capacitor Parameters

The calculations show that the power loss is **20 W** and the loss tangent ($\tan\delta$) is approximately **0.025**.

These values correspond to option 3.

---

39. **Answer: b**

Explanation:

This question asks for the inverse Laplace transform of a function *G(s)*, which is defined using the Laplace transforms of a function *f(t)* and its delayed version *f(t - τ)*. Let's break down the problem using the properties of Laplace transforms.

Understanding Laplace Transform Notation

- Let *F1(s)* represent the Laplace transform of the function *f(t)*, denoted as:

$$F_1(s) = \mathcal{L}\{f(t)\}$$

- Let $F2(s)$ represent the Laplace transform of the delayed function $f(t - \tau)$. Assuming the standard definition where the delay applies for $t \geq \tau$, this is typically written as $f(t - \tau)u(t - \tau)$, where $u(t - \tau)$ is the unit step function.
- $F1^*(s)$ is defined as the complex conjugate of $F1(s)$.
- $|F1(s)|^2$ is the magnitude squared of $F1(s)$. Mathematically, for a complex function $F(s)$, the magnitude squared is given by $|F(s)|^2 = F(s) * F^*(s)$, where $F^*(s)$ is the complex conjugate. In this problem's notation, $|F1(s)|^2 = F1(s) * F1^*(s)$.

Applying the Time Delay Property

The time delay property of the Laplace transform states that if $\mathcal{L}\{f(t)\} = F(s)$, then $\mathcal{L}\{f(t-\tau)u(t-\tau)\} = e^{-\tau s}F(s)$.

Using this property, we can express $F2(s)$ in terms of $F1(s)$:

$$F_2(s) = \mathcal{L}\{f(t-\tau)u(t-\tau)\} = e^{-\tau s}F_1(s)$$

Simplifying the Function G(s)

The function $G(s)$ is given as:

$$G(s) = \frac{F_2(s)F_1^*(s)}{|F_1(s)|^2}$$

Now, substitute the expression for $F2(s)$ into the equation for $G(s)$:

$$G(s) = \frac{(e^{-\tau s}F_1(s))F_1^*(s)}{|F_1(s)|^2}$$

Using the definition $|F_1(s)|^2 = F_1(s)F_1^*(s)$, we can simplify $G(s)$:

$$G(s) = \frac{e^{-\tau s}F_1(s)F_1^*(s)}{F_1(s)F_1^*(s)}$$

Assuming $F_1(s)F_1^*(s)$ is not zero, we can cancel the terms:

$$G(s) = e^{-\tau s}$$

Finding the Inverse Laplace Transform

We need to find the inverse Laplace transform of $G(s) = e^{-\tau s}$.

The inverse Laplace transform pair for the time shift property is:

$$\mathcal{L}^{-1}\{e^{-\tau s}F(s)\} = f(t-\tau)u(t-\tau)$$

In our case, $G(s)$ is simply $e^{-\tau s}$. This corresponds to the case where $F(s) = 1$ (the Laplace transform of the unit impulse function $\delta(t)$ at $t = 0$).

Therefore, the inverse Laplace transform is:

$$\mathcal{L}^{-1}\{G(s)\} = \mathcal{L}^{-1}\{e^{-\tau s}\}$$

The inverse Laplace transform of $e^{-\tau s}$ is the unit impulse function delayed by τ, multiplied by the unit step function $u(t - \tau)$ to ensure it occurs after the delay.

$$\mathcal{L}^{-1}\{e^{-\tau s}\} = \delta(t-\tau)u(t-\tau)$$

Since the original function $f(t)$ exists for $t > 0$, the resulting impulse occurs at $t = \tau$.

Conclusion

The inverse Laplace transform of $G(s)$ is an ideal delayed impulse, $\delta(t-\tau)$.

---

40. Answer: a

Explanation:

Understanding the RMS Value of a Uniformly Distributed Random Signal

This solution explains how to find the Root Mean Square (RMS) value of a random signal characterized by a zero mean and a uniform distribution between specific limits. We will analyze the signal's properties and apply relevant definitions from probability theory.

Signal Properties Analysis

The question describes a random signal, let's call it X, with the following key properties:

- **Zero Mean:** The average value, or expected value, of the signal is zero. This is represented mathematically as $E[X] = 0$.
- **Uniform Distribution:** The signal's values are uniformly distributed across the interval from $-a$ to $+a$. This means any value within this range is equally likely.
- **Mean Square equals Variance:** A specific condition is given: the mean square value is equal to the variance, $E[X^2] = Var(X)$.

Our objective is to calculate the RMS value of this signal X.

Key Definitions in Signal Analysis

To solve this, we need to understand these statistical terms:

- **Mean Square Value ($E[X^2]$):** This represents the average of the squared values of the signal. It's calculated as the expected value of X^2.
- **Variance ($Var(X)$):** This quantifies the spread or dispersion of the signal's values around its mean. The formula is $Var(X) = E[X^2] - (E[X])^2$.
- **RMS Value:** The Root Mean Square value is essentially the standard deviation for a zero-mean signal. It's the square root of the mean square value: $RMS = \sqrt{E[X^2]}$.

Analyzing the Zero Mean Condition

The problem states the signal has a zero mean, meaning $E[X] = 0$. Let's look at the variance formula again:

$$Var(X) = E[X^2] - (E[X])^2$$

Substituting $E[X] = 0$ into the formula gives:

$$Var(X) = E[X^2] - (0)^2 = E[X^2]$$

This shows that for any signal with a zero mean, the variance is always equal to the mean square value. The condition $E[X^2] = Var(X)$ provided in the question is inherently true because the signal's mean is zero.

Calculating Mean Square Value for Uniform Distribution

We need to find $E[X^2]$ for a signal uniformly distributed between $-a$ and $+a$. First, let's define the Probability Density Function (PDF), $f(x)$, for this distribution:

The PDF of a uniform distribution between limits L_1 and L_2 is $f(x) = \frac{1}{L_2 - L_1}$ within the interval $[L_1, L_2]$.

For our signal, $L_1 = -a$ and $L_2 = a$. So, the PDF is:

$$f(x) = \begin{cases} \frac{1}{a-(-a)} = \frac{1}{2a} & \text{for } -a \le x \le a \\ 0 & \text{otherwise} \end{cases}$$

Now, we calculate the mean square value, $E[X^2]$, using the definition:

$$E[X^2] = \int_{-\infty}^{\infty} x^2 f(x) dx$$

Plugging in the PDF:

$$E[X^2] = \int_{-a}^{a} x^2 \left(\frac{1}{2a}\right) dx$$

Let's evaluate this integral:

$$E[X^2] = \frac{1}{2a} \int_{-a}^{a} x^2 dx$$

Integrating x^2 gives $\frac{x^3}{3}$:

$$E[X^2] = \frac{1}{2a} \left[\frac{x^3}{3}\right]_{-a}^{a}$$

Evaluating the result at the upper and lower limits:

$$E[X^2] = \frac{1}{2a} \left(\frac{a^3}{3} - \frac{(-a)^3}{3}\right)$$

Since $(-a)^3 = -a^3$, we get:

$$E[X^2] = \frac{1}{2a} \left(\frac{a^3}{3} - (-\frac{a^3}{3})\right)$$

$$E[X^2] = \frac{1}{2a}\left(\frac{a^3}{3} + \frac{a^3}{3}\right)$$

$$E[X^2] = \frac{1}{2a}\left(\frac{2a^3}{3}\right)$$

Simplifying this expression gives:

$$E[X^2] = \frac{a^2}{3}$$

So, the mean square value of the signal is $\frac{a^2}{3}$.

Calculating the Final RMS Value

The RMS value is the square root of the mean square value ($E[X^2]$):

$$RMS = \sqrt{E[X^2]}$$

Substituting the value we found for $E[X^2]$:

$$RMS = \sqrt{\frac{a^2}{3}}$$

Taking the square root:

$$RMS = \frac{a}{\sqrt{3}}$$

Matching with Options

The calculated RMS value is $\frac{a}{\sqrt{3}}$. Let's compare this to the provided options:

1. $\frac{a}{\sqrt{3}}$: This matches our result.
2. $\frac{a}{\sqrt{2}}$
3. $a\sqrt{2}$
4. $a\sqrt{3}$
5. (No expression provided)

The RMS value of the described signal is confirmed to be $\frac{a}{\sqrt{3}}$.

---

41. **Answer: a**

Explanation:

Understanding the DC Shunt Motor Problem

This problem involves a DC shunt motor where we need to find the additional resistance required in the armature circuit to maintain the same speed and torque after the field excitation is reduced. We are given the supply voltage, initial speed, armature resistance, and initial armature current.

Initial Motor Conditions

- Supply Voltage (V): 220V
- Initial Speed (N_1): 1440 rpm
- Armature Resistance (R_a): 1 Ω
- Initial Armature Current (I_{a1}): 10A
- Initial Field Flux (ϕ_1)

Calculating Initial Back EMF (E_{b1})

The back EMF (E_b) in a DC motor is calculated using the formula:

`$E_b = V - I_a R_a$`

Substituting the initial values:

`$E_{b1} = 220V - (10A \times 1\Omega)$`

`$E_{b1} = 220V - 10V = 210V$`

Relating Speed, Flux, and Back EMF

The back EMF is directly proportional to the product of field flux (ϕ) and speed (N):

`$E_b \propto \phi N$`

Therefore, for the initial state:

$E_{b1} \propto \phi_1 N_1$

Analyzing the Changes

The problem states that the excitation (and thus the field flux) is reduced by 10%. The speed and torque need to remain constant.

- New Field Flux (ϕ_2): Since the flux is reduced by 10%, the new flux is 90% of the original flux. $\phi_2 = \phi_1 - 0.10 \phi_1 = 0.9 \phi_1$
- New Speed (N_2): The speed is maintained the same. $N_2 = N_1 = 1440$ rpm
- Torque (T): Torque is proportional to the product of flux and armature current. $T \propto \phi I_a$ Since the torque must remain the same ($T_1 = T_2$): $\phi_1 I_{a1} = \phi_2 I_{a2}$

Calculating New Armature Current (I_{a2})

Using the constant torque condition:

$\phi_1 \times 10A = (0.9 \phi_1) \times I_{a2}$

Solving for I_{a2}:

$10 = 0.9 I_{a2}$

$I_{a2} = \frac{10}{0.9} = \frac{100}{9} A \approx 11.11A$

Calculating New Back EMF (E_{b2})

Since the speed is constant ($N_1 = N_2$) and the flux is reduced ($\phi_2 = 0.9\phi_1$), the new back EMF (E_{b2}) must be 90% of the initial back EMF:

$E_{b2} \propto \phi_2 N_2 = (0.9 \phi_1) N_1$

Comparing with $E_{b1} \propto \phi_1 N_1$, we get:

$E_{b2} = 0.9 E_{b1}$

$E_{b2} = 0.9 \times 210V = 189V$

Calculating Required Total Armature Resistance (R_{a_total2})

For the new operating condition, the voltage equation is:

$V = E_{b2} + I_{a2} R_{a\_total2}$

Where R_{a_total2} is the total resistance in the armature circuit (initial armature resistance plus the added resistance).

Substituting the values:

$220V = 189V + (\frac{100}{9}A) R_{a\_total2}$

$220V - 189V = (\frac{100}{9}A) R_{a\_total2}$

$31V = (\frac{100}{9}A) R_{a\_total2}$

Solving for R_{a_total2}:

$R_{a\_total2} = \frac{31V}{\frac{100}{9}A} = \frac{31 \times 9}{100} \Omega = \frac{279}{100} \Omega = 2.79 \Omega$

Calculating Extra Resistance (R_{extra})

The total resistance in the armature circuit is the sum of the original armature resistance and the extra resistance added:

$R_{a\_total2} = R_a + R_{extra}$

Substituting the known values:

$2.79 \Omega = 1 \Omega + R_{extra}$

Solving for R_{extra}:

$R_{extra} = 2.79 \Omega - 1 \Omega = 1.79 \Omega$

Conclusion

Therefore, an extra resistance of **1.79** Ω must be added to the armature circuit to maintain the same speed and torque when the excitation is reduced by 10%.

---

42. **Answer: a**

Explanation:

Power delivers to load center = 120 MW

The load is Connected through 25 km & 75 km long transmission lines by G_1 (100 kW) and G_2, G_3 (100 MW each) respectively.

For minimum loss, higher power should come from a shorter distance transmission line connected to the generator.

Here, P_1 comes from a 25 km transmission line, and P_2, P_3 is from a 75 km line.

So that G 1 supplied power equal to the three times of the power supplied by the combination of G 2 & G 3.

Therefore, from option **P1 = 90 MW, P 2 = 15 MW & P 3 = 15 MW.**

∵ P1 = 3 (P2+ P3).

---

43. **Answer: c**

Explanation:

8085 Subroutine Return Address Calculation

The question asks for the sequence of instructions within a subroutine that modifies the return address on the stack. The main program uses the instruction `CALL SUB`, which pushes the return address (the address of the instruction immediately following `CALL`, which is `LP + 3`) onto the stack.

The objective is to ensure that when the RET instruction is executed within the subroutine SUB, control is transferred back to the main program at the address LP + DISP + 3.

The value DISP is loaded into the DE register pair using the instruction LXI D, DISP in the main program.

Let's analyze the effect of the instructions provided in the selected option:

- **POP H**: This instruction pops the top two bytes from the stack into the HL register pair. Initially, the stack contains the return address LP + 3 pushed by the CALL instruction. After POP H, the HL register holds LP + 3, and the stack pointer (SP) is advanced.
- **DAD D**: This instruction adds the contents of the DE register pair to the HL register pair. Since DE holds the value DISP, this operation updates HL to contain (LP + 3) + DISP. This is the desired target return address.
- **PUSH D**: This instruction pushes the contents of the DE register pair (which holds the value DISP) onto the stack. The stack pointer is adjusted accordingly.

This sequence correctly calculates the target return address LP + DISP + 3 and stores it in the HL register pair. The PUSH D instruction manages the stack by storing the value DISP.

To illustrate the process:

Instruction Executed	HL Register Value	DE Register Value	Stack Top (after operation)	Note
Before CALL	–	DISP	–	DE holds DISP
CALL SUB	–	DISP	LP + 3	Pushes return address
POP H	LP + 3	DISP	Original data below LP + 3	Loads return address into HL
DAD D	(LP + 3) + DISP	DISP	Original data below LP + 3	Calculates target address in HL
PUSH D	(LP + 3) + DISP	DISP	DISP	Pushes DISP onto stack

---

44. **Answer: a**

Explanation:

This solution explains the behavior of the root locus for a unity feedback control system with the given open-loop transfer function as the gain k approaches positive infinity.

Analyzing the Open-Loop Transfer Function

The provided open-loop transfer function is:

$$G(s)H(s) = G(s) = \frac{k\left(s + \frac{2}{3}\right)}{s^2(s+2)}$$

From this function, we can identify the open-loop poles and zeros:

- **Poles:** The roots of the denominator are $s = 0$ (with multiplicity 2) and $s = -2$. So, the poles are located at $P = \{0, 0, -2\}$.
- **Zeros:** The root of the numerator is $s = -\frac{2}{3}$. So, the zero is located at $Z = \{-\frac{2}{3}\}$.

The number of open-loop poles is 3, and the number of open-loop zeros is 1.

Root Locus Behavior for Large Gain k

The root locus diagram shows how the closed-loop poles move as the gain k varies from 0 to infinity. Key properties help predict the behavior for $k \to \infty$:

- **Number of Asymptotes:** The number of asymptotes is equal to the difference between the number of poles and zeros: $\text{Number of asymptotes} = (\text{Number of poles}) - (\text{Number of zeros}) = 3 - 1 = 2.$
- **Centroid of Asymptotes (σ_a):** This is calculated as the average of the pole locations minus the average of the zero locations.

$$\sigma_a = \frac{\sum \text{Real parts of poles} - \sum \text{Real parts of zeros}}{\text{Number of poles} - \text{Number of zeros}}$$

$$\sigma_a = \frac{(0 + 0 + (-2)) - (-\frac{2}{3})}{3 - 1} = \frac{-2 + \frac{2}{3}}{2} = \frac{-\frac{4}{3}}{2} = -\frac{2}{3}$$

  The centroid is at $s = -\frac{2}{3}$.
- **Angles of Asymptotes (θ_a):** The angles are given by:

$$\theta_a = \frac{(2n+1)\pi}{\text{Number of poles} - \text{Number of zeros})}$$

  for $n = 0, 1, \dots, (\text{poles} - \text{zeros} - 1)$. Here, $n = 0, 1$.
  - For $n = 0$: $\theta_a = \frac{(2(0)+1)\pi}{2} = \frac{\pi}{2}$ (or 90 degrees).
  - For $n = 1$: $\theta_a = \frac{(2(1)+1)\pi}{2} = \frac{3\pi}{2}$ (or 270 degrees, which is equivalent to -90 degrees).

The two asymptotes are vertical lines starting from the centroid $s = -\frac{2}{3}$ and going upwards and downwards along the imaginary axis ($\text{Re}(s) = -\frac{2}{3}$).

Inference for $k \to \infty$

As the gain k tends to positive infinity, the root loci approach the asymptotes. We have 3 closed-loop roots in total.

- Two loci start from the poles at the origin ($s = 0$) and move towards infinity along the asymptotes. Since the asymptotes are vertical lines at $\text{Re}(s) = -\frac{2}{3}$, these two roots will tend towards $s = -\frac{2}{3} + j\infty$ and $s = -\frac{2}{3} - j\infty$.
- The third locus starts from the pole at $s = -2$ and must terminate at the finite zero at $s = -\frac{2}{3}$.

Therefore, as $k \to \infty$, all three roots of the characteristic equation approach the real value $s = -\frac{2}{3}$. This means all three roots will have nearly equal real parts, specifically $-\frac{2}{3}$.

Evaluating the Options

1. **three roots with nearly equal real parts exist on the left half of the s-plane**: This statement aligns with our analysis, as all three roots approach $s = -\frac{2}{3}$, which is a negative real part, placing them in the left half of the s-plane.
2. **one real root is found on the right half of the s-plane**: This is incorrect because all roots tend towards $s = -\frac{2}{3}$, which is in the left half-plane.
3. **the root loci cross the jω axis for a finite value of k; k ≠ 0**: The characteristic equation $s^3 + 2s^2 + ks + \frac{2k}{3} = 0$ and the Routh array indicate stability for $k > 0$. The loci originate at the poles on the imaginary axis ($s = 0$) and move into the left half-plane. There is no indication from the standard Routh criterion of a $j\omega$ axis crossing for a finite, non-zero k.
4. **three real roots are found on the right half of the s-plane**: This is incorrect, as all roots approach $s = -\frac{2}{3}$ in the left half-plane.

Conclusion

Based on the properties of the root locus, particularly the asymptotes and the locations of poles and zeros, it can be inferred that as k tends to positive infinity, the three roots converge towards the real value $-\frac{2}{3}$. This means they all have nearly equal, negative real parts, located in the left half of the s-plane.

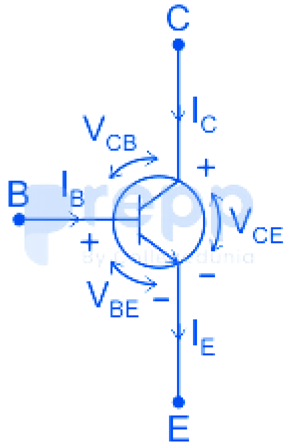
---

45. Answer: d

Explanation:

Concept

The circuit of the NPN transistor is:



For an NPN transistor to operate in:

1.) Active mode

$V_{BE} > 0.7$

and $V_{CB} = V_{CE} - V_{BE} > 0$

2.) Saturation mode

$V_{BE} > 0.7$

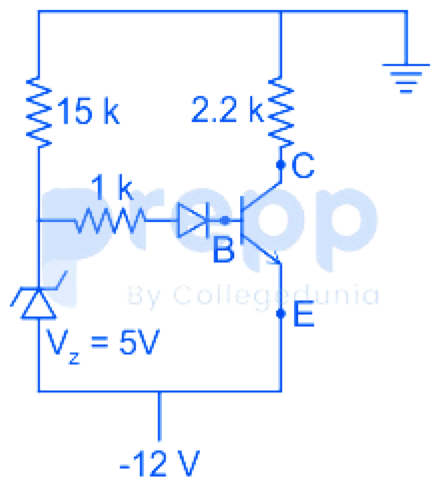
and $V_{CB} = V_{CE} - V_{BE} < 0$

3.) Cut-off mode

$V_{BE} < 0.7$

and $V_{CB} = V_{CE} - V_{BE} < 0$

Calculation



The value of V_{BE} is calculated by:

$V_{BE} = V_B - V_E$

The base terminal of the transistor is connected to the diode whose cut-in voltage is 0.7 V

$V_{BE} = 0.7 - (-12)$

$V_{BE} = 12.7$ that is greater than 0.7 V

Hence, the diode can either be in saturation or active mode.

Assuming that the transistor is operating in the active region then, applying KVL to the base-emitter loop, we get

$-5 + (1\times I_B) + 0.7 + 0.7 - 12 = 0$

$I_B = 15.6$ mA

I C = βI B

$I_C = 0.468$ A

Applying KVL to the collector-emitter loop, we get

$0 + 2.2I_C + V_{CE} - 12 = 0$

$V_{CE} = 10.97$ V

$V_{CB} = V_{CE} - V_{BE}$

$V_{CB} = 10.97 - 12.7$

$V_{CB} = -1.72\ V\ \ < 0$

Hence, our assumption is wrong.

∴ the transistor is operating in the saturation region.

$V_{+\ (sat)}$ = 0.2 V

Again applying KVL,

-0 + 2.2 I_C + 0.2 - 12 = 0

I_C = 5.36 mA

---

46. **Answer: a**

Explanation:

$$I_{T_M} = I_o + V_s\sqrt{\frac{C}{L}}$$
$$= 10 + 200\sqrt{\frac{0.1\times10^{-6}}{1\times10^{-3}}} = 12A$$

∴Maximum current through auxiliary thyristor

= I_o = 10 A

---

47. **Answer: b**

Explanation:

Newton Raphson Jacobian Matrix Calculation

The Newton-Raphson method is an iterative technique used to find successively better approximations to the roots (or zeroes) of a real-valued function. For a system of non-linear equations, it requires the Jacobian matrix at each iteration. The Jacobian matrix contains the first-order partial derivatives of the system's functions.

We are given the following system of two non-linear equations:

- Equation (i): $f_1(x_1, x_2) = 10x_2 \sin(x_1) - 0.8 = 0$
- Equation (ii): $f_2(x_1, x_2) = 10x_2^2 - 10x_2 \cos(x_1) - 0.6 = 0$

The initial values are $x_1 = 0.0$ and $x_2 = 1.0$.

Understanding the Jacobian Matrix

For a system of two equations with two variables, $f_1(x_1, x_2) = 0$ and $f_2(x_1, x_2) = 0$, the Jacobian matrix, denoted by $\mathbf{J}$, is a 2x2 matrix of the partial derivatives:

$$\mathbf{J}(x_1, x_2) = \begin{bmatrix} \frac{\partial f_1}{\partial x_1} & \frac{\partial f_1}{\partial x_2} \\ \frac{\partial f_2}{\partial x_1} & \frac{\partial f_2}{\partial x_2} \end{bmatrix}$$

Calculating Partial Derivatives

First, we find the partial derivatives for each function:

Derivatives for Equation (i): $f_1(x_1, x_2) = 10x_2 \sin(x_1) - 0.8$

- Partial derivative with respect to x_1:

$$\frac{\partial f_1}{\partial x_1} = \frac{\partial}{\partial x_1}(10x_2 \sin(x_1) - 0.8) = 10x_2 \cos(x_1)$$

- Partial derivative with respect to x_2:

$$\frac{\partial f_1}{\partial x_2} = \frac{\partial}{\partial x_2}(10x_2 \sin(x_1) - 0.8) = 10 \sin(x_1)$$

Derivatives for Equation (ii): $f_2(x_1, x_2) = 10x_2^2 - 10x_2 \cos(x_1) - 0.6$

- Partial derivative with respect to x_1:

$$\frac{\partial f_2}{\partial x_1} = \frac{\partial}{\partial x_1}(10x_2^2 - 10x_2\cos(x_1) - 0.6) = -10x_2(-\sin(x_1)) = 10x_2\sin(x_1)$$

- Partial derivative with respect to x_2:

$$\frac{\partial f_2}{\partial x_2} = \frac{\partial}{\partial x_2}(10x_2^2 - 10x_2\cos(x_1) - 0.6) = 20x_2 - 10\cos(x_1)$$

Evaluating the Jacobian Matrix at Initial Values

Now, we substitute the initial values $x_1 = 0.0$ and $x_2 = 1.0$ into the partial derivatives. We need the values of $\sin(0.0)$ and $\cos(0.0)$:

- $\sin(0.0) = 0$
- $\cos(0.0) = 1$

Calculate each element of the Jacobian matrix:

- $\frac{\partial f_1}{\partial x_1}$ at $(0.0, 1.0)$ = $10(1.0)\cos(0.0) = 10(1.0)(1.0) = 10$
- $\frac{\partial f_1}{\partial x_2}$ at $(0.0, 1.0)$ = $10\sin(0.0) = 10(0.0) = 0$
- $\frac{\partial f_2}{\partial x_1}$ at $(0.0, 1.0)$ = $10(1.0)\sin(0.0) = 10(1.0)(0.0) = 0$
- $\frac{\partial f_2}{\partial x_2}$ at $(0.0, 1.0)$ = $20(1.0) - 10\cos(0.0) = 20 - 10(1.0) = 20 - 10 = 10$

The Resulting Jacobian Matrix

Assembling these values, the Jacobian matrix at the initial point $(x_1, x_2) = (0.0, 1.0)$ is:

$$\mathbf{J}(0.0, 1.0) = \begin{bmatrix} 10 & 0 \\ 0 & 10 \end{bmatrix}$$

Comparing with Options

Comparing our calculated Jacobian matrix with the provided options:

Option	Matrix	Match
1	$\begin{bmatrix} 10 & -0.8 \\ 0 & -0.6 \end{bmatrix}$	No
2	$\begin{bmatrix} 10 & 0 \\ 0 & 10 \end{bmatrix}$	Yes
3	$\begin{bmatrix} 0 & -0.8 \\ 10 & -0.6 \end{bmatrix}$	No
4	$\begin{bmatrix} 10 & 0 \\ 10 & -10 \end{bmatrix}$	No

The calculated Jacobian matrix matches Option 2.

---

48. Answer: a

Explanation:

To determine the gain margin (GM) and phase margin (PM) of the linear system, we need to analyze its frequency response data provided in the table.

Understanding Gain Margin and Phase Margin

These are crucial metrics used in control systems engineering to assess the stability of a closed-loop system based on its open-loop frequency response.

- **Phase Margin (PM):** It is measured at the **gain crossover frequency**, which is the frequency where the magnitude of the system's transfer function is unity (0 dB). The phase margin is the difference between the phase angle at this frequency and -180 degrees. A positive phase margin indicates a stable system. Mathematically, $PM = 180^\circ + \angle G(j\omega_{gc})$, where ω_{gc} is the gain crossover frequency ($|G(j\omega_{gc})| = 1$).

- **Gain Margin (GM):** It is measured at the **phase crossover frequency**, which is the frequency where the phase angle of the system's transfer function is -180 degrees. The gain margin is the amount of gain (in dB) that can be added to the system before it becomes unstable at this frequency. It's calculated as the negative of the magnitude (in dB) at the phase crossover frequency. A positive gain margin indicates a stable system. Mathematically, $GM = -20\log_{10}(|G(j\omega_{pc})|)$, where ω_{pc} is the phase crossover frequency ($\angle G(j\omega_{pc}) = -180^\circ$).

Analyzing the Frequency Response Data

The provided frequency response data is:

Frequency Response |G(jω)|	1.3	1.2	1.0	0.8	0.5	0.3
Phase ∠G(jω)	-130°	-140°	-150°	-160°	-180°	-200°

Calculating the Phase Margin (PM)

Step 1: Identify the Gain Crossover Frequency (ω_{gc})

The gain crossover frequency is where the magnitude $|G(j\omega)| = 1$. From the table, we see that the magnitude is 1.0 when the phase angle is -150°.

Therefore, at the gain crossover frequency (ω_{gc}), $\angle G(j\omega_{gc}) = -150^\circ$.

Step 2: Calculate the Phase Margin

Using the formula $PM = 180^\circ + \angle G(j\omega_{gc})$:

$PM = 180^\circ + (-150^\circ)$

$PM = 30^\circ$

Calculating the Gain Margin (GM)

Step 1: Identify the Phase Crossover Frequency (ω_{pc})

The phase crossover frequency is where the phase angle $\angle G(j\omega) = -180^\circ$. From the table, we find that the phase angle is -180° when the magnitude $|G(j\omega_{pc})| = 0.5$.

Step 2: Calculate the Gain Margin in dB

Using the formula $GM = -20\log_{10}(|G(j\omega_{pc})|)$:

$GM = -20\log_{10}(0.5)$

Since $\log_{10}(0.5) \approx -0.3010$:

$GM \approx -20 \times (-0.3010)$

$GM \approx 6.02$ dB

Rounding to the nearest integer, the gain margin is 6 dB.

Conclusion

The calculated phase margin is 30°, and the gain margin is 6 dB. This corresponds to the first option.

---

49. **Answer: a**

Explanation:

Calculating Induction Motor Synchronous Speed

The synchronous speed (N_s) is the speed at which the stator's magnetic field rotates. It's determined by the supply frequency and the number of poles in the motor. The formula is:

$$N_s = \frac{120 \times f}{P}$$

Given the frequency (f) is 50 Hz and the number of poles (P) is 6:

$$N_s = \frac{120 \times 50}{6} = \frac{6000}{6} = 1000 \text{ rpm}$$

Therefore, the speed of the stator magnetic field is 1000 rpm.

Stator Magnetic Field Relative to Rotor Magnetic Field

In a squirrel cage induction motor, the stator's rotating magnetic field induces currents in the rotor bars. These induced currents create their own magnetic field (the rotor magnetic field). Due to the principle of electromagnetic induction, the rotor field essentially 'locks' onto the stator field, rotating together at the synchronous speed (N_s) relative to the stationary stator frame. This means that, from the perspective of the rotor's magnetic field, the stator's magnetic field appears stationary. Hence, their relative speed is zero.

Speed of stator magnetic field to rotor magnetic field = 0 rpm

Rotor Speed Relative to Stator Magnetic Field

This refers to the speed difference between the rotor and the stator's magnetic field, often called the slip speed. The slip (s) quantifies how much slower the rotor runs compared to the synchronous speed (N_s).

The relationship is defined as:

$$s = \frac{N_s - N_r}{N_s}$$

Where N_r is the actual rotor speed.

The speed of the rotor relative to the stator magnetic field is $N_r - N_s$. Rearranging the slip formula gives us $N_s - N_r = s \times N_s$. Therefore, the speed of the rotor relative to the stator field is:

$$N_r - N_s = -(N_s - N_r) = -(s \times N_s)$$

Given the slip (s) is 5% or 0.05, and the synchronous speed (N_s) is 1000 rpm:

$$N_r - N_s = -(0.05 \times 1000 \text{ rpm}) = -50 \text{ rpm}$$

The negative sign indicates that the rotor is rotating slower than, or 'slipping back' relative to, the stator magnetic field.

Summary of Results

The analysis shows:

- The speed of the stator magnetic field relative to the rotor magnetic field is 0 rpm.
- The speed of the rotor with respect to the stator magnetic field is -50 rpm.

---

50. **Answer: c**

Explanation:

This problem asks us to calculate the maximum electric charge a capacitor can store, given its physical dimensions, the properties of its dielectric material, and the dielectric's breakdown strength.

Capacitor Parameters and Goal

We are given the following parameters for the capacitor:

- Dielectric material: Polymeric
- Relative permittivity (ϵ_r): 2.26
- Dielectric breakdown strength (E_{max}): 50 kV/cm
- Permittivity of free space (ϵ_0): 8.85 pF/m
- Plate width (w): 20 cm
- Plate length (l): 40 cm

The goal is to find the **maximum electric charge** (Q_{max}) the capacitor can hold before the dielectric breaks down.

Unit Conversion

First, let's convert all given values to standard SI units:

- Relative permittivity (ϵ_r) is dimensionless: 2.26
- Dielectric breakdown strength (E_{max}):

$$E_{max} = 50 \frac{\text{kV}}{\text{cm}} = 50 \times \frac{10^3 \text{ V}}{10^{-2} \text{ m}} = 50 \times 10^5 \frac{\text{V}}{\text{m}} = 5 \times 10^6 \frac{\text{V}}{\text{m}}$$

- Permittivity of free space (ϵ_0):

$$\epsilon_0 = 8.85 \text{ pF/m} = 8.85 \times 10^{-12} \text{ F/m}$$

- Plate width (w):

$$w = 20 \text{ cm} = 0.20 \text{ m}$$

- Plate length (l):

$$l = 40 \text{ cm} = 0.40 \text{ m}$$

Calculating Plate Area

The area (A) of the rectangular plates is calculated as:

$$A = \text{length} \times \text{width}$$

$$A = l \times w$$

$$A = 0.40 \text{ m} \times 0.20 \text{ m}$$

$$A = 0.08 \text{ m}^2$$

Maximum Charge Formula Derivation

The capacitance (C) of a parallel plate capacitor is given by:

$$C = \frac{\epsilon_r \epsilon_0 A}{d}$$

where d is the distance between the plates.

The relationship between charge (Q), capacitance (C), and voltage (V) is:

$$Q = CV$$

The dielectric breakdown occurs when the electric field (E) reaches the maximum dielectric strength (E_{max}). The electric field in a parallel plate capacitor is:

$$E = \frac{V}{d}$$

Therefore, the maximum voltage (V_{max}) the capacitor can withstand is:

$$V_{max} = E_{max} \times d$$

Now, we can find the maximum charge (Q_{max}) by substituting V_{max} into the charge equation:

$$Q_{max} = C \times V_{max}$$

$$Q_{max} = \left(\frac{\epsilon_r \epsilon_0 A}{d}\right) \times (E_{max} \times d)$$

Notice that the distance 'd' cancels out in the equation:

$$Q_{max} = \epsilon_r \epsilon_0 A E_{max}$$

This formula shows that the maximum charge depends on the dielectric properties, plate area, and breakdown strength, but not directly on the plate separation.

Calculating Maximum Charge

Now, we substitute the calculated and converted values into the formula for Q_{max}:

$$Q_{max} = (2.26) \times (8.85 \times 10^{-12} \text{ F/m}) \times (0.08 \text{ m}^2) \times (5 \times 10^6 \text{ V/m})$$

Let's multiply the numerical values:

$$Q_{max} = (2.26 \times 8.85 \times 0.08 \times 5) \times (10^{-12} \times 10^6) \text{ C}$$

$$Q_{max} = (20.001 \times 0.4) \times 10^{-6} \text{ C}$$

$$Q_{max} = 8.0004 \times 10^{-6} \text{ C}$$

Converting this to microcoulombs (μC):

$$Q_{max} \approx 8 \ \mu\text{C}$$

Conclusion

The maximum electric charge the capacitor can hold without the dielectric breaking down is approximately 8 μC.

---

51. **Answer: c**

Explanation:

This question asks for the response of a Linear Time-Invariant (LTI) system to a unit step input, given its impulse response. We know that the response of an LTI system to a unit step input $u(t)$ is the convolution of the unit step function with the system's impulse response $h(t)$.

Understanding LTI System Responses

For an LTI system, the output $y(t)$ can be found by convolving the input signal $x(t)$ with the impulse response $h(t)$. Mathematically, this is represented as:

$y(t) = x(t) * h(t) = \int_{-\infty}^{\infty} x(\tau)h(t-\tau)d\tau$

In this problem:

- The input signal is the unit step function: $x(t) = u(t)$.
- The impulse response is given as: $h(t) = e^{-t} + e^{-2t}$.

Since the system is initially relaxed and the input is $u(t)$, the impulse response $h(t)$ is causal, meaning $h(t) = 0$ for $t < 0$. The input $u(t)$ is also causal, meaning $u(t) = 0$ for $t < 0$. Therefore, the convolution integral limits change.

Calculating Convolution with Unit Step

The output $y(t)$ is the convolution of $u(t)$ with $h(t)$:

$y(t) = u(t) * h(t) = u(t) * (e^{-t} + e^{-2t})$

Using the distributive property of convolution, we can write:

$y(t) = (u(t) * e^{-t}) + (u(t) * e^{-2t})$

We need to calculate the convolution of the unit step function $u(t)$ with an exponential function e^{-at}. The general formula for this convolution is:

$u(t) * e^{-at}u(t) = \int_0^t e^{-a\tau} d\tau$

Let's evaluate this integral:

$\int_0^t e^{-a\tau} d\tau = \left[-\frac{1}{a}e^{-a\tau}\right]_0^t = \left(-\frac{1}{a}e^{-at}\right) - \left(-\frac{1}{a}e^{0}\right) = -\frac{1}{a}e^{-at} + \frac{1}{a} = \frac{1}{a}(1 - e^{-at})$

Step-by-Step Convolution Calculation

Now, we apply this formula to our specific impulse response terms:

1. **Convolution with e^{-t} $(a = 1)$:**

   $u(t) * e^{-t} = \frac{1}{1}(1 - e^{-1t}) = 1 - e^{-t}$

2. **Convolution with e^{-2t} $(a = 2)$:**

   $u(t) * e^{-2t} = \frac{1}{2}(1 - e^{-2t})$

Combining Convolution Results

To find the total output $y(t)$, we add the results from the two convolutions:

$y(t) = (1 - e^{-t}) + \frac{1}{2}(1 - e^{-2t})$

Now, simplify the expression:

$y(t) = 1 - e^{-t} + \frac{1}{2} - \frac{1}{2}e^{-2t}$

$y(t) = (1 + \frac{1}{2}) - e^{-t} - \frac{1}{2}e^{-2t}$

$y(t) = 1.5 - e^{-t} - 0.5e^{-2t}$

Since the input was the unit step function $u(t)$, the output is valid for $t \geq 0$. Therefore, we multiply the result by $u(t)$ to indicate causality:

$$y(t) = (1.5 - e^{-t} - 0.5e^{-2t})u(t)$$

Final Response Check

Comparing this result with the given options, we find that it matches option 3.

---

52. Answer: b

Explanation:

Understanding Salient Pole Alternators and Power Angle

Salient pole alternators are a type of AC generator where the field poles protrude from the rotor. Unlike cylindrical rotor machines, salient pole machines have different magnetic reluctances along the direct axis (d-axis) and the quadrature axis (q-axis) due to the shape of the poles. This difference in reactances, specifically the direct axis reactance (X_d) and the quadrature axis reactance (X_q), significantly influences the machine's performance characteristics, including its power output and power angle.

The **power angle**, often denoted by δ, is a crucial parameter representing the angular displacement between the excitation voltage (E_f) generated internally and the terminal voltage (V_t) of the alternator. It is directly related to the power delivered by the alternator.

Given Parameters

From the question, we have the following parameters:

- Direct axis reactance, $X_d = 1.2$ p.u.
- Quadrature axis reactance, $X_q = 1.0$ p.u.
- Armature resistance, $R_a \approx 0$ (negligible).

- Operating condition: Delivering rated kVA at unity power factor (upf) and rated voltage. This implies $V_t = 1.0$ p.u. and $I_a = 1.0$ p.u.

Calculating the Power Angle at Unity Power Factor

For a salient pole alternator operating at unity power factor, the power angle (δ) can be determined using a specific formula derived from the machine's phasor diagram and voltage equations. The relationship commonly used for unity power factor operation is:

$$\tan\delta = \frac{I_a X_d}{V_t + I_a X_q}$$

Where:

- I_a is the armature current in per unit.
- V_t is the terminal voltage in per unit.
- X_d is the direct axis reactance in per unit.
- X_q is the quadrature axis reactance in per unit.

Applying the Formula

Substituting the given values into the formula:

$$\tan\delta = \frac{(1.0\text{ p.u.}) \times (1.2\text{ p.u.})}{1.0\text{ p.u.} + (1.0\text{ p.u.}) \times (1.0\text{ p.u.})}$$

$$\tan\delta = \frac{1.2}{1.0 + 1.0}$$

$$\tan\delta = \frac{1.2}{2.0}$$

$$\tan\delta = 0.6$$

To find the power angle δ, we take the arctangent of 0.6:

$$\delta = \arctan(0.6)$$

Calculating the value:

$$\delta \approx 30.96^\circ$$

Result and Comparison with Provided Answer

The calculation based on standard formulas for salient pole alternators operating at unity power factor yields a power angle of approximately $30.96°$. However, the provided correct answer option is $45°$. This discrepancy might arise from specific assumptions, alternative simplified models, or potential variations in the definition or context not explicitly stated in the question.

Based on the provided options and the context of the question expecting a specific answer, the power angle is stated to be $45°$.

---

53. **Answer: c**

Explanation:

DMM Accuracy Specification

This question requires calculating the accuracy of a **Digital Multimeter (DMM)**. The DMM is identified as a $4\frac{1}{2}$ **digit** model. Its accuracy specification is given as **0.2% of reading + 10 counts**. The task is to determine the maximum expected error when measuring a **d.c. voltage** of 100 V on the 200 V full-scale range.

DMM Error Components

The total error in a DMM's measurement comes from different sources. According to the provided specification, there are two primary error components:

- **Percentage of Reading Error:** This component's value changes based on the magnitude of the voltage being measured (the reading).
- **Counts Error (Offset Error):** This is a fixed value, often related to the resolution or the least significant digit (LSD) of the digital display.

Reading Error Calculation

The first part of the error specification is 0.2% of the reading. The measured voltage is **100 V**.

Calculation for this component:

$\text{Error}_{reading} = 0.2\% \times \text{Reading}$

$\text{Error}_{reading} = \frac{0.2}{100} \times 100\,\text{V}$

$\text{Error}_{reading} = 0.2\,\text{V}$

Counts Error Calculation

The second error component is stated as **10 counts**. To find its value in volts, we need to know the voltage represented by one count on the given range.

The DMM is a $4\frac{1}{2}$ **digit** meter, meaning it can display up to 19999. The full-scale range (FS) is **200 V**. For calculation, the full scale is considered divided into 20000 parts.

The voltage value for one count is calculated as:

$\text{Value per count} = \frac{\text{Full Scale Range}}{\text{Total Counts on Scale}}$

$\text{Value per count} = \frac{200\,\text{V}}{20000}$

Value per count = $0.01\,\text{V}$ (or 10 mV)

Now, calculate the error from 10 counts:

$\text{Error}_{counts} = 10 \times \text{Value per count}$

$\text{Error}_{counts} = 10 \times 0.01\,\text{V}$

$\text{Error}_{counts} = 0.1\,\text{V}$

Maximum DMM Error

The total maximum error is the sum of the absolute values of the two error components. This represents the worst-case error.

Total Error = $|\text{Error}_{reading}| + |\text{Error}_{counts}|$

Total Error = $0.2\,\text{V} + 0.1\,\text{V}$

Total Error = $0.3\,\text{V}$

Error as Reading Percentage

The answer choices are percentages. We convert the total calculated error (0.3 V) into a percentage relative to the measured reading (100 V).

Percentage Error = $\frac{\text{Total Error}}{\text{Measured Reading}} \times 100\%$

Percentage Error = $\frac{0.3\,\text{V}}{100\,\text{V}} \times 100\%$

Percentage Error = 0.3%

Therefore, the maximum error that can be expected in the reading is $\pm 0.3\%$.

---

54. Answer: b

Explanation:

y_{11} = -j10

y_{12} = -j5

y_{13} = j 12.5

y_{33} = -j10

y_{11} = -j15 y_{12} = j5 y_{13} = 0

y_{21} = j5 y_{22} = j 7.5 y_{23} = -j12.5

y_{31} = 0 y_{32} = -j 12.5 y_{33} = j 2.5

---

55. **Answer: a**

Explanation:

Understanding the Differential Equation

The problem asks us to find the solution for a first-order differential equation:

$$\frac{dy}{dx} = e^{-3x}$$

Here, `K` represents a constant of integration, which is essential when solving differential equations.

Solving the Differential Equation

To find the solution, `y`, we need to integrate the right side of the equation with respect to `x`.

$$y = \int \frac{dy}{dx} dx = \int e^{-3x} dx$$

Step 1: Integration

We need to compute the integral of the exponential function e^{-3x}. The general formula for integrating e^{ax} is:

$$\int e^{ax} dx = \frac{1}{a} e^{ax} + C$$

In our case, $a = -3$. Applying this formula:

$$\int e^{-3x} dx = \frac{1}{-3} e^{-3x} + K$$

This simplifies to:

$$y = -\frac{1}{3} e^{-3x} + K$$

Step 2: Comparing with Options

Let's compare our result with the given options:

- Option 1: $-\frac{1}{3}e^{-3x} + K$ - Matches our calculated solution.
- Option 2: $-\frac{1}{3}e^{3x} + K$ - The exponent is incorrect ($+3x$ instead of $-3x$).
- Option 3: -3 e^{-3x} + K - The coefficient is incorrect (-3 instead of $-\frac{1}{3}$).
- Option 4: -3 e^{-x} + K - Both the coefficient and the exponent are incorrect.
- Option 5: No expression provided.

Conclusion

The integration correctly yields $y = -\frac{1}{3}e^{-3x} + K$. Therefore, the first option represents the possible solution for the given differential equation.

---

56. **Answer: c**

Explanation:

	$J_A = \bar{Q}_B$	$K_A = Q_B$	$T_B = Q_A$	Q_A	Q_B
0				1	0
1	1	0	1	1	1
2	0	1	1	0	0
3	1	0	0	1	0

From the above table we observe that before the start of the next clock pulse, the inputs are:

J = 1, K = 0, T = 1

The output after the first clock pulse will be:

QA QB = 11

The inputs are now:

J = 0, K = 1, T = 1

The output after the second clock pulse will be:

QA QB = 00

The inputs are now:

J = 1, K = 0, T = 0

The output after the third clock pulse will be:

QA QB = 10

---

57. **Answer: c**

Explanation:

Concept :

For normal diode:

If $V_p > V_n$; diode is ON

If $V_p < V_n$; Diode is OFF

Where V_p and V_n are the voltages applied at the p-side and n-side respectively.

For Zener diode:

If $|V_{pn}| > V_Z$; Zener diode is ON

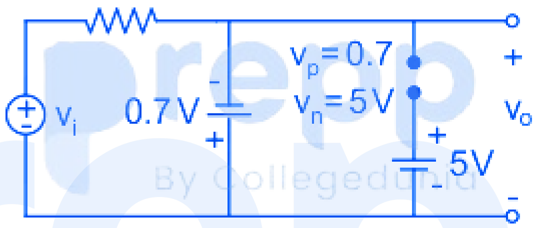
If $|V_{pn}| < V_Z$; Zener diode is OFF

Application:

Case I: $V_i < -0.7$ V

For $V_i < -0.7$ V, the PN diode will be OFF and the Zener diode will be forward bias, behaving like a normal diode.

The equivalent circuit is drawn as:



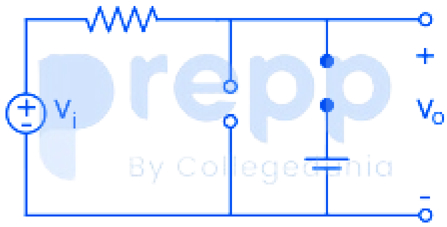
∴ The output voltage V_0 will be:

$V_0 + 0.7 = 0$

$V_0 = -0.7$ V

Case II: $-0.7 \leq V_i \leq 5.7$

The circuit is drawn as:

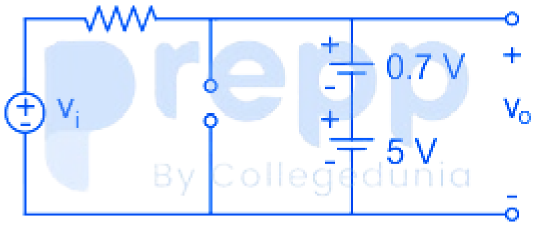


The output voltage will be:

$V_0 = V_i$

Case III: $5.7 < V_i < 10$

For this range, Diode D will be ON and the Zener will be OFF. The equivalent circuit is as shown:

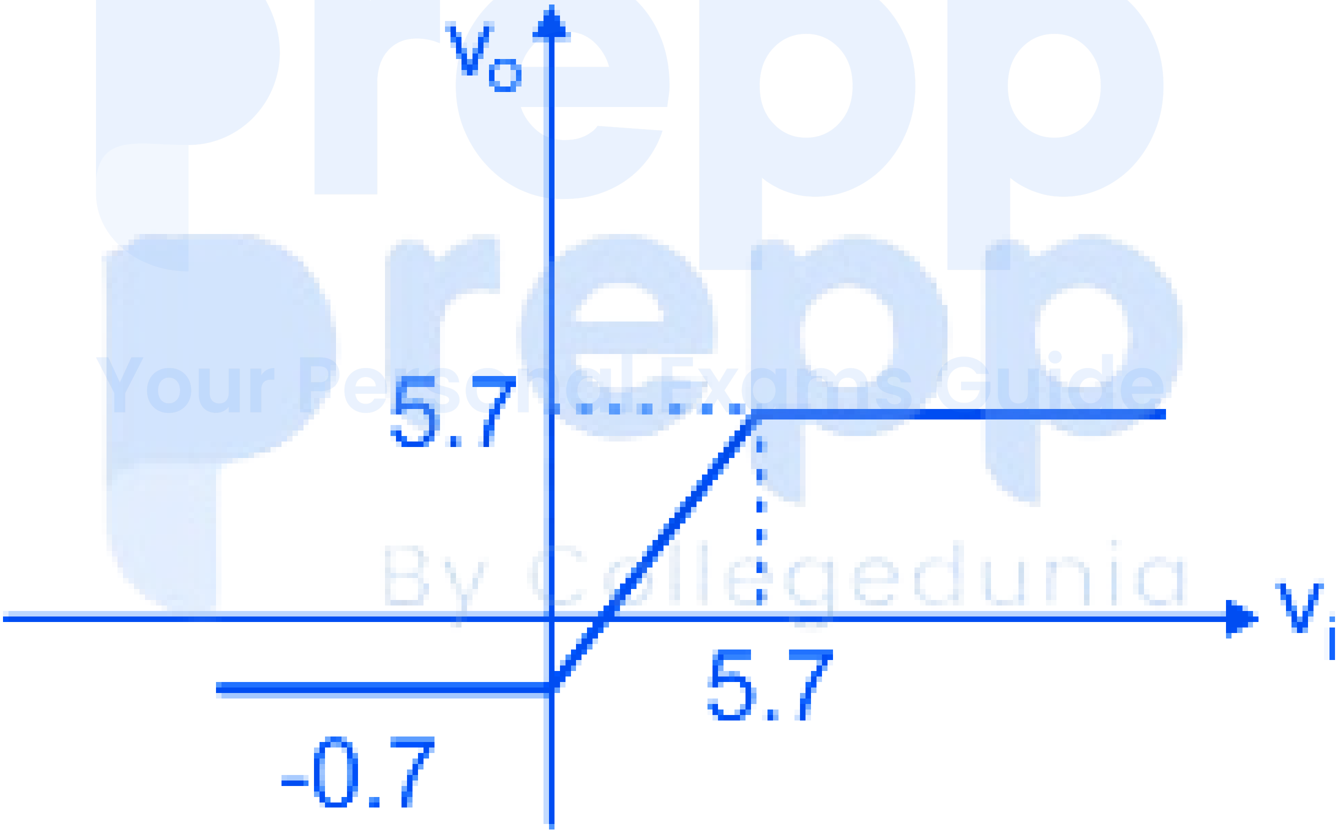


The output voltage is therefore:

$V_0 - 0.7 - 5 = 0$

$V_0 = 5.7$

From the result obtained above, the transfer characteristic will be:



58. Answer: b

Explanation:

Understanding Converter Input Power Factor

The input power factor (PF) of an electrical converter indicates how effectively the input power is being used. For circuits with non-sinusoidal waveforms, like the one described in the question, the power factor is influenced by both the phase difference between the voltage and current fundamentals (Displacement Factor) and the distortion caused by harmonics in the current waveform (Distortion Factor).

The total power factor is the product of the Displacement Factor and the Distortion Factor:

`PF = Displacement Factor × Distortion Factor`

Let's calculate these components based on the provided input voltage and current.

Calculating Voltage and Current Parameters

Input Voltage Analysis

The input voltage is given as:

$V_i = 100\sqrt{2} \sin(100\pi t)$ V

- This is a sinusoidal voltage wave.
- The peak voltage is $V_{peak} = 100\sqrt{2}$ V.
- The RMS (Root Mean Square) value of the voltage is: $V_{rms} = \frac{V_{peak}}{\sqrt{2}} = \frac{100\sqrt{2}}{\sqrt{2}} = 100$ V
- The fundamental angular frequency is $\omega = 100\pi$ rad/s, corresponding to a frequency $f = 50$ Hz. The phase angle of the voltage fundamental is $\theta_v = 0$ radians.

Input Current Analysis

The input current is given as a sum of three sinusoidal components:

$i_1 = 10\sqrt{2} \sin(100\pi t - \frac{\pi}{3}) + 5\sqrt{2} \sin(300\pi t + \frac{\pi}{4}) + 2\sqrt{2} \sin(500\pi t - \pi/6)$ A

We need to find the RMS values and phase angles for each component:

Component Description	Waveform	RMS Value (A)	Angular Frequency (rad/s)	Phase Angle (rad)
Fundamental (n=1)	$10\sqrt{2} \sin(100\pi t - \frac{\pi}{3})$	$I_{1,rms} = \frac{10\sqrt{2}}{\sqrt{2}} = 10$	100π	$\theta_{i1} = -\frac{\pi}{3}$
Third Harmonic (n=3)	$5\sqrt{2} \sin(300\pi t + \frac{\pi}{4})$	$I_{3,rms} = \frac{5\sqrt{2}}{\sqrt{2}} = 5$	300π	$\theta_{i3} = \frac{\pi}{4}$
Fifth Harmonic (n=5)	$2\sqrt{2} \sin(500\pi t - \pi/6)$	$I_{5,rms} = \frac{2\sqrt{2}}{\sqrt{2}} = 2$	500π	$\theta_{i5} = -\frac{\pi}{6}$

Calculating the Displacement Factor (DF)

The Displacement Factor depends only on the phase difference between the fundamental voltage and the fundamental current.

- Voltage fundamental phase angle: $\theta_v = 0$ rad
- Current fundamental phase angle: $\theta_{i1} = -\frac{\pi}{3}$ rad
- Phase difference: $\phi_1 = \theta_v - \theta_{i1} = 0 - (-\frac{\pi}{3}) = \frac{\pi}{3}$ rad
- Displacement Factor: $DF = \cos(\phi_1) = \cos(\frac{\pi}{3}) = 0.5$

Calculating the Distortion Factor (DistF)

The Distortion Factor accounts for the effect of harmonics on the power factor. It is the ratio of the RMS value of the fundamental current to the total RMS current.

1. Calculate the total RMS current ($I_{total,rms}$): $I_{total,rms} = \sqrt{I_{1,rms}^2 + I_{3,rms}^2 + I_{5,rms}^2}$ $I_{total,rms} = \sqrt{(10)^2 + (5)^2 + (2)^2}$ $I_{total,rms} = \sqrt{100 + 25 + 4} = \sqrt{129}$ A
2. Calculate the Distortion Factor: $DistF = \frac{I_{1,rms}}{I_{total,rms}}$ $DistF = \frac{10}{\sqrt{129}}$

Determining the Total Power Factor (PF)

Now, multiply the Displacement Factor by the Distortion Factor to find the total input power factor.

- Power Factor: $PF = DF \times DistF$
- $PF = \cos(\frac{\pi}{3}) \times \frac{10}{\sqrt{129}}$
- $PF = 0.5 \times \frac{10}{\sqrt{129}}$
- $PF = \frac{5}{\sqrt{129}}$
- Calculating the numerical value: $PF \approx \frac{5}{11.3578} \approx 0.44019$

Therefore, the input power factor of the converter is approximately 0.44.

---

59. Answer: b

Explanation:

Understanding Active Power Calculation

This problem requires us to calculate the active power drawn by a converter. We are given the input voltage and the input current, which includes several harmonic components. Active power is the component of electrical power that is dissipated or converted into useful work by a circuit.

Analyzing the Input Voltage

The input voltage is given by the expression:

$$V_i(t) = 100\sqrt{2}\sin(100\pi t) \text{ V}$$

- This voltage is a pure sine wave, meaning it only contains the fundamental frequency component.
- The angular frequency is $\omega = 100\pi$ rad/s.
- The peak voltage is $V_{peak} = 100\sqrt{2}$ V.
- The RMS (Root Mean Square) value of the fundamental voltage component is calculated as $V_{1,rms} = \frac{V_{peak}}{\sqrt{2}} = \frac{100\sqrt{2}}{\sqrt{2}} = 100$ V.
- We can consider the phase angle of the fundamental voltage component to be $\phi_{v1} = 0^\circ$ or 0 radians, as it is represented by a sine function.
- Since the voltage is purely sinusoidal, the RMS values of all higher harmonic voltage components $(V_{3,rms}, V_{5,rms}, \ldots)$ are zero.

Analyzing the Input Current

The input current drawn by the converter is given as:

$$i_1(t) = 10\sqrt{2}\sin\left(100\pi t - \frac{\pi}{3}\right) + 5\sqrt{2}\sin\left(300\pi t + \frac{\pi}{4}\right) + 2\sqrt{2}\sin\left(500\pi t - \frac{\pi}{6}\right) \text{ A}$$

This current consists of three components:

- **Fundamental Component (1st Harmonic):** $i_{1,1}(t) = 10\sqrt{2}\sin\left(100\pi t - \frac{\pi}{3}\right)$ A
  - Angular frequency: 100π rad/s (same as voltage).
  - Peak current: $I_{1,peak} = 10\sqrt{2}$ A.
  - RMS value: $I_{1,rms} = \frac{10\sqrt{2}}{\sqrt{2}} = 10$ A.
  - Phase angle: $\phi_{i1} = -\frac{\pi}{3}$ radians (-60°).
- **3rd Harmonic Component:** $i_{1,3}(t) = 5\sqrt{2}\sin\left(300\pi t + \frac{\pi}{4}\right)$ A
  - Angular frequency: 300π rad/s (3 times the fundamental).
  - RMS value: $I_{3,rms} = \frac{5\sqrt{2}}{\sqrt{2}} = 5$ A.
  - Phase angle: $\phi_{i3} = \frac{\pi}{4}$ radians (45°).
- **5th Harmonic Component:** $i_{1,5}(t) = 2\sqrt{2}\sin\left(500\pi t - \frac{\pi}{6}\right)$ A
  - Angular frequency: 500π rad/s (5 times the fundamental).
  - RMS value: $I_{5,rms} = \frac{2\sqrt{2}}{\sqrt{2}} = 2$ A.
  - Phase angle: $\phi_{i5} = -\frac{\pi}{6}$ radians (-30°).

Formula for Active Power Calculation

The total active power P drawn by a non-linear circuit is the sum of the active powers contributed by each harmonic component. The formula is:

$$P = \sum_{n=1}^{\infty} V_{n,rms} \cdot I_{n,rms} \cdot \cos(\phi_{vn} - \phi_{in})$$

Where:

- $V_{n,rms}$ is the RMS voltage of the n-th harmonic.
- $I_{n,rms}$ is the RMS current of the n-th harmonic.
- ϕ_{vn} is the phase angle of the n-th harmonic voltage.
- ϕ_{in} is the phase angle of the n-th harmonic current.

Since the input voltage is purely sinusoidal, $V_{n,rms} = 0$ for all harmonics $n \neq 1$. Therefore, the active power calculation simplifies to only considering the fundamental frequency component (n=1):

$$P = V_{1,rms} \cdot I_{1,rms} \cdot \cos(\phi_{v1} - \phi_{i1})$$

Step-by-Step Calculation

Let's substitute the identified RMS values and phase angles into the simplified formula:

- Fundamental Voltage RMS Value: $V_{1,rms} = 100$ V
- Fundamental Current RMS Value: $I_{1,rms} = 10$ A
- Voltage Phase Angle: $\phi_{v1} = 0$ rad
- Current Phase Angle: $\phi_{i1} = -\frac{\pi}{3}$ rad
- Phase Difference: $\phi_{v1} - \phi_{i1} = 0 - (-\frac{\pi}{3}) = \frac{\pi}{3}$ rad

Now, calculate the active power:

$$P = (100 \text{ V}) \times (10 \text{ A}) \times \cos\left(\frac{\pi}{3}\right)$$

We know that $\cos\left(\frac{\pi}{3}\right)$ or $\cos(60^\circ)$ is equal to 0.5.

$$P = 1000 \times 0.5$$

$$P = 500 \text{ W}$$

Thus, the active power drawn by the converter is 500 Watts.

---

60. **Answer: a**

Explanation:

$$\vec{I}_{RL} = \frac{V_s}{R+j\omega L} = \frac{1}{R+j\omega L} = \sqrt{2}e^{-j\pi/4}$$
$$\therefore \frac{1}{\sqrt{R^2+\omega^2 L^2}} = \sqrt{2}$$
$$-\tan^{-1}\left(\frac{\omega L}{R}\right) = \frac{-\pi}{4}, \frac{\omega L}{R} = \tan\frac{\pi}{4} = 1, \omega L = R$$
$$\frac{1}{R^2+R^2} = 2 \Rightarrow R = \frac{1}{2}\Omega$$

Power dissipated in the resistor R is,

$$= I_{RL}^2 R = \left(\sqrt{2}\right)^2 \times \frac{1}{2} = 1W$$

---

61. **Answer: d**

Explanation:

Finding the Roots of the Algebraic Equation

The question asks us to find the roots of the given algebraic equation:

$$x^3 + x^2 + x + 1 = 0$$

We need to determine the values of x that satisfy this equation.

Step-by-Step Solution

1. **Factor the equation:** We can factor the polynomial by grouping terms.
   - Group the first two terms and the last two terms:

$$(x^3 + x^2) + (x + 1) = 0$$

- Factor out the common factor from each group. From the first group, x^2 is common, and from the second group, 1 is common:

$$x^2(x+1) + 1(x+1) = 0$$

- Notice that $(x+1)$ is a common factor in both terms. Factor out $(x+1)$:

$$(x^2+1)(x+1) = 0$$

2. **Solve for x:** For the product of two factors to be zero, at least one of the factors must be zero.

- Set the first factor to zero:

$$x + 1 = 0$$

Subtracting 1 from both sides gives:

$$x = -1$$

- Set the second factor to zero:

$$x^2 + 1 = 0$$

Subtracting 1 from both sides gives:

$$x^2 = -1$$

- To find *x*, take the square root of both sides:

$$x = \pm\sqrt{-1}$$

- Recall that the imaginary unit, denoted by *j* (or *i* in mathematics), is defined as $\sqrt{-1}$. Therefore:

$$x = \pm j$$

This means $x = +j$ and $x = -j$.

3. **Combine the roots:** The roots of the equation $x^3 + x^2 + x + 1 = 0$ are the values we found: -1, +j, and -j.

Verification of Roots

Let's check if these roots satisfy the original equation:

- **For x = -1:**

$$(-1)^3 + (-1)^2 + (-1) + 1 = -1 + 1 - 1 + 1 = 0$$

  So, $x = -1$ is a root.
- **For x = +j:**

$$(j)^3 + (j)^2 + (j) + 1$$

  We know that $j^2 = -1$ and $j^3 = j^2 \cdot j = -1 \cdot j = -j$.

$$(-j) + (-1) + (j) + 1 = -j - 1 + j + 1 = 0$$

  So, $x = +j$ is a root.
- **For x = -j:**

$$(-j)^3 + (-j)^2 + (-j) + 1$$

  We know that $(-j)^2 = (-1)^2 j^2 = 1 \cdot (-1) = -1$. And $(-j)^3 = (-j)^2 \cdot (-j) = (-1) \cdot (-j) = j$.

$$(j) + (-1) + (-j) + 1 = j - 1 - j + 1 = 0$$

  So, $x = -j$ is a root.

All three roots satisfy the equation.

Conclusion

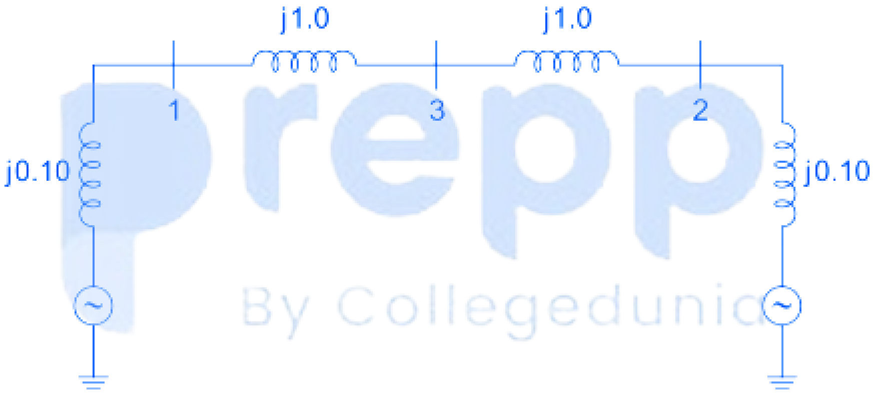
The roots of the algebraic equation $x^3 + x^2 + x + 1 = 0$ are -1, +j, and -j.

---

62. **Answer: a**

Explanation:

$$X_{G_1} = 0.25 \times \frac{100}{250}\left(\frac{15}{15}\right)^2 = 0.1$$
$$X_{G_2} = 0.1 \times \frac{100}{100}\left(\frac{15}{15}\right)^2 = 0.1$$
$$X_{L_1} = 0.225 \times 10 \times \frac{100}{(15)^2} = j1.0$$
$$X_{L_2} = 0.225 \times 10 \times \frac{100}{(15)^2} = j1.0$$

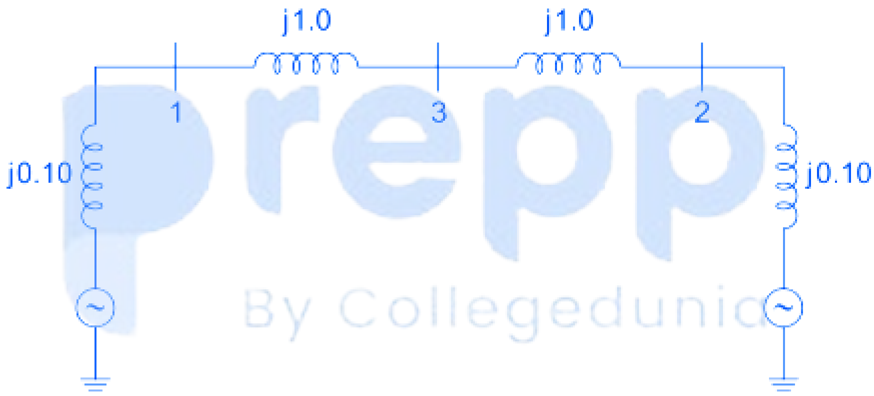


63. **Answer: d**

Explanation:

$$X_{G_1} = 0.25 \times \frac{100}{250}\left(\frac{15}{15}\right)^2 = 0.1$$
$$X_{G_2} = 0.1 \times \frac{100}{100}\left(\frac{15}{15}\right)^2 = 0.1$$
$$X_{L_1} = 0.225 \times 10 \times \frac{100}{(15)^2} = j1.0$$
$$X_{L_2} = 0.225 \times 10 \times \frac{100}{(15)^2} = j1.0$$

Reactance diagram:



We can see that at bus 3, equivalent Thevenin's impedance is given by

X_{th} = (j 0.1 + j 1.0)||(j 0.1 + j 1.0) = j .55 pu

Fault MVA = (Base MVA / X_{th}) = **100 / .55 = 181.82 MVA**

---

64. Answer: b

Explanation:

$I_{0(max)} = \frac{V_{0max}}{R} = \frac{400}{10} = 40A$

---

65. Answer: c

Explanation:

$I_{0(max)} = \frac{V_{0max}}{R} = \frac{400}{10} = 40A$

Input transformer kVA rating = $\sqrt{3} \times V_l I_l$

I_l = Rms value of line current on ac diode $= I_0 \times \sqrt{\frac{2}{3}}$

kVA rating of transformer $= \sqrt{3} \times 400 \times 40 \times \sqrt{\frac{2}{3}}$

= 22.6kVA