

Diffusion Models Presentation

추성재, DGIST



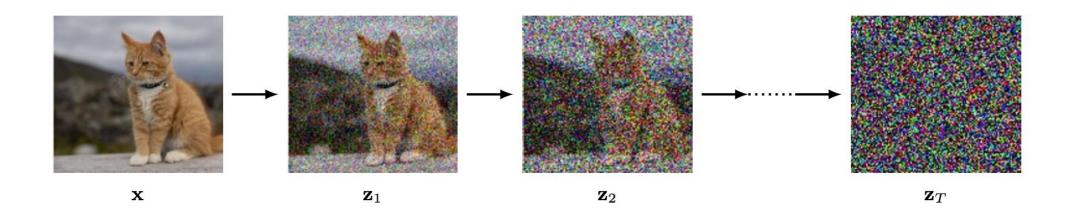
What is Diffusion Model?

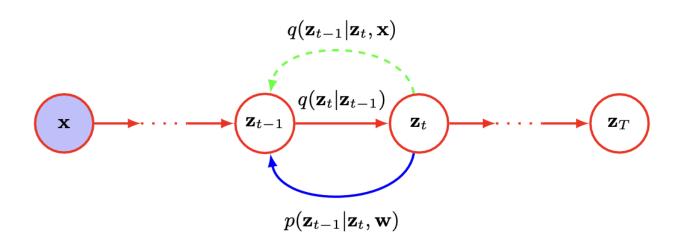
Diffusion Model or Denoising Diffusion Probabilistic Model (DDPM)

- Forward Encoder
 - corrupt training image using multi-step noising process
 - training image turn into sample of Gaussian distribution
- Reverse Decoder
 - deep neural network is trained to invert noising process
 - trained network can generate new images with new sample of Gaussian distribution



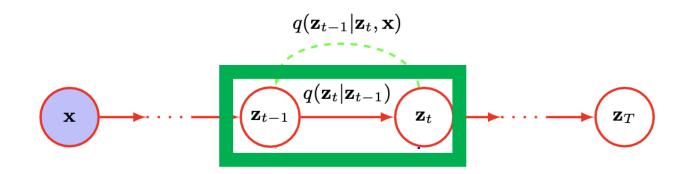












- Forward process step
 - Markov chain current state only influenced by previous state

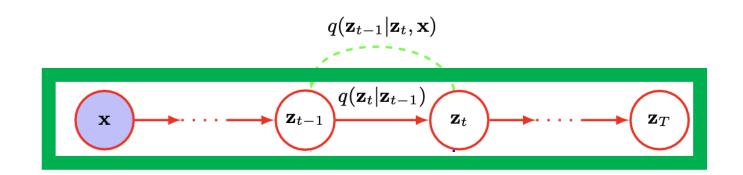
• step:
$$\mathbf{z}_t = \sqrt{1 - \beta_t} \mathbf{z}_{t-1} + \sqrt{\beta_t} \epsilon_t$$

where $\epsilon_t \sim \mathcal{N}(\epsilon_t | \mathbf{0}, \mathbf{I}), \beta_t \in (0, 1), \beta_1 < \beta_2 < ... < \beta_T$

• can be rewrite in : $q(\mathbf{z}_t|\mathbf{z}_{t-1}) = \mathcal{N}(\mathbf{z}_t|\sqrt{1-\beta_t}\mathbf{z}_{t-1},\beta_t\mathbf{I})$



20.1 Forward Encoder – Diffusion kernel



Diffusion kernel:

$$q(\mathbf{z_t}|\mathbf{x}) = \mathcal{N}(\mathbf{z_t}|\sqrt{\alpha_t}\mathbf{x}, (1-\alpha_t)\mathbf{I}$$

where $\alpha_t = \Pi_{\tau=1}^t (1-\beta_{\tau})$
can be written like $\mathbf{z}_t = \sqrt{\alpha_t}\mathbf{x} + \sqrt{1-\alpha_t}\epsilon_t$
if $T \to \infty$, $q(\mathbf{z_t}|\mathbf{x}) = \mathcal{N}(\mathbf{z}_T|\mathbf{0}, \mathbf{I})$



20.1 Forward Encoder – Pytorch example

```
def cal_alpha(beta, step):
    result = torch.tensor(1).float()

    for i in range(0, step+1):
        tmp = beta[0, i]
        result *= (1-tmp)

    return result
```

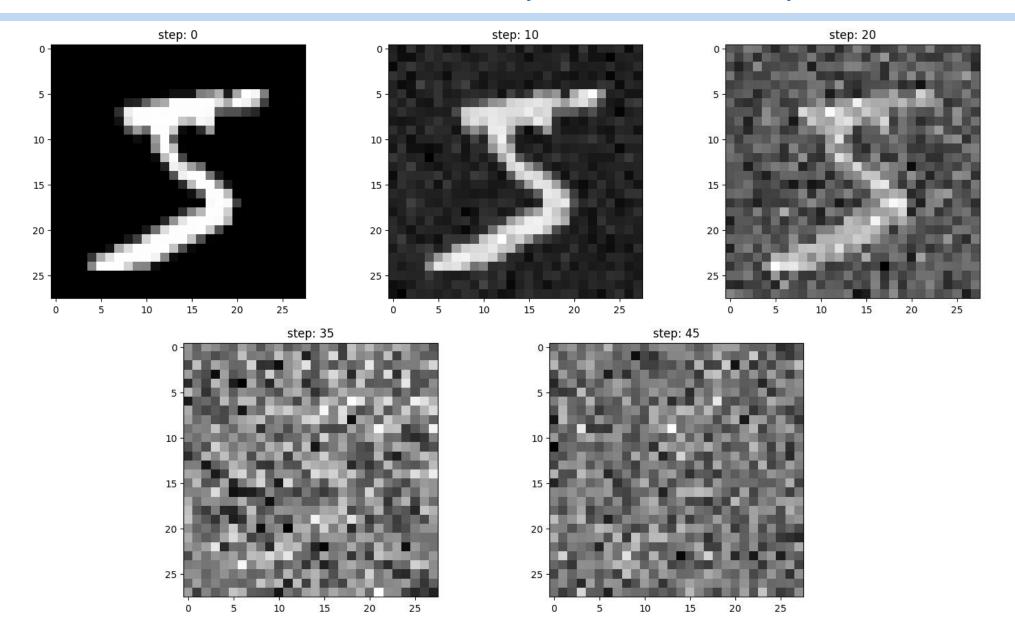
```
# 점진적 forward 과정

for t in range(0, terminal_step, 5):
    noise_t = torch.randn(28,28)
    alpha_t = cal_alpha(beta, t)

# forward 과정
    z_t = forward_image * torch.sqrt(alpha_t) + noise_t * (1-torch.sqrt(alpha_t))
    plt.figure()
    plt.imshow(z_t, cmap='gray')
```

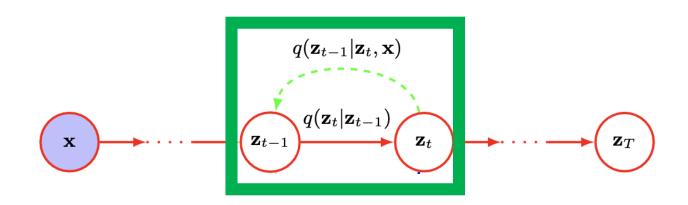


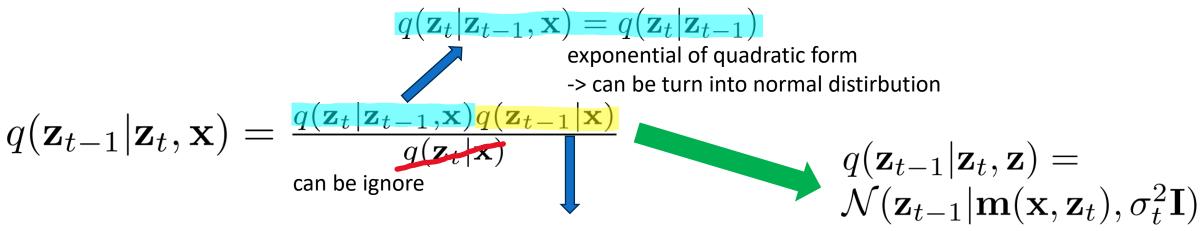
20.1 Forward Encoder – Pytorch example





20.1 Forward Encoder – Conditional distribution

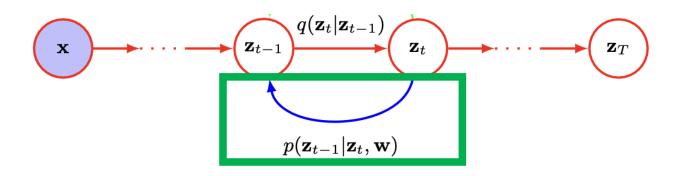




diffusion kernel: normal distribution





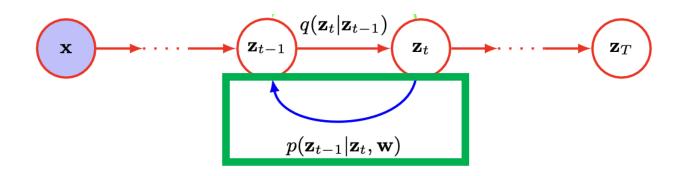


- Instead of solving $q(\mathbf{z}_{t-1}|\mathbf{z}_t)$, we learn approximation of reverse distribution: $p(\mathbf{z}_{t-1}|\mathbf{z}_t,\mathbf{w}) = \mathcal{N}(\mathbf{z}_{t-1}|\mu(\mathbf{z}_t,\mathbf{w},t),\beta_t\mathbf{I})$
- $\mu(\mathbf{z}_t, \mathbf{w}, t)$ is deep neural network with parameter \mathbf{w}
- Main assumption :

$$\beta_t \ll 1$$
 so that $q(\mathbf{z}_{t-1}|\mathbf{z}_t)$ approximately Gaussian distribution over \mathbf{z}_{t-1}







- $\mu(\mathbf{z}_t, \mathbf{w}, t)$ can invert total step of forward process
- input dimension and output dimension of $\mu(\mathbf{z}_t, \mathbf{w}, t)$ have to be same, so U-net architecture is commonly used.
- total denoising process:

$$p(\mathbf{x}, \mathbf{z}_1, \dots, \mathbf{z}_T | \mathbf{w}) = p(\mathbf{z}_T) \left\{ \prod_{t=2}^T p(\mathbf{z}_{t-1} | \mathbf{z}_t, \mathbf{w}) \right\} p(\mathbf{x} | \mathbf{z}_1, \mathbf{w}).$$
(20.19)

- DEVIST
- Total denoising process is intractable: process involve integrating over the neural network functions
- We use Evidence Lower Bound (ELBO)
 - for all distribution $q(\mathbf{z})$, following relation always holds

$$\ln p(\mathbf{x}|\mathbf{w}) = \mathcal{L}(\mathbf{w}) + \mathrm{KL}\left(q(\mathbf{z})||p(\mathbf{z}|\mathbf{x},\mathbf{w})\right)$$

• Where Kullback-Leibler divergence term is always bigger than 0, we can get following relation

$$\ln p(\mathbf{x}|\mathbf{w}) \ge \mathcal{L}(\mathbf{w})$$

Why KL-divergence is always bigger than 0?

$$p(\mathbf{z}|\mathbf{x}, \mathbf{w})$$
 and $q(z)$ is different.
so $\mathrm{KL}(q(z)||p(\mathbf{z}|\mathbf{x}, \mathbf{w})) \geq 0$

• Our objective : train neural network to maximize ELBO

Evidence Lower Bound (ELBO)는 계산이 용이한 임의의 가우시안 분포 q(z)를 가지고 계산이 불가능한 $\log p_{\theta}(x)$ (Evidence)의 lower bound를 최대 화하는 방식으로 $\log p_{\theta}(x)$ 을 근사하는 변분 추론 방법입니다.

$$egin{aligned} \log p_{ heta}(x) &\geq \sum_z q(z) \log rac{p_{ heta}(x,z)}{q(z)} \ &= \sum_z q(z) \log p_{ heta}(x,z) - \sum_z q(z) \log q(z) \ &= \sum_z q(z) \log p_{ heta}(x,z) + H(q) \end{aligned}$$

$$\therefore ELBO(q) = \sum_z q(z) \log p_{ heta}(x,z) + H(q)$$

출처 : 생성모델 입문서 - WikiDocs

KL-Divergence

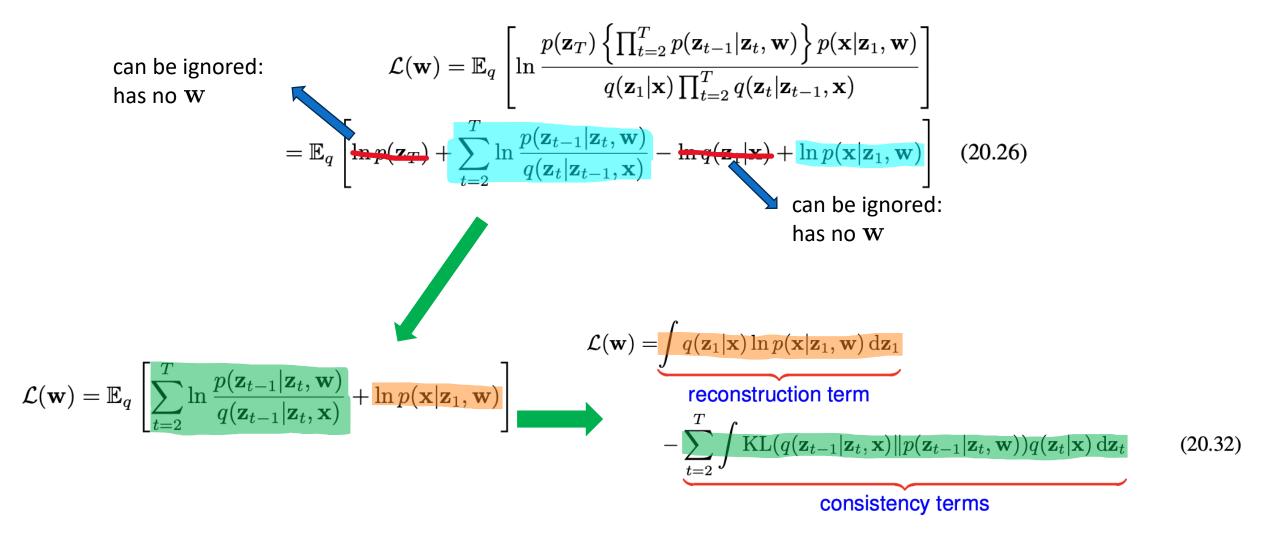
임의의 두 확률 분포 p(x)와 q(x)에 대해 KL-Divergence는 다음 성질을 만족합니다.

$$D_{KL}(p||q) = \sum_x p(x) log rac{p(x)}{q(x)}$$

- ullet $D_{KL}(p||q) \geq 0$: 두 분포간 차이는 존재해야 합니다 (extstyle < 0)
- $D_{KL}(p||q)=0$ if p=q: 두 분포가 같다면 KLD는 0입니다.
- $D_{KL}(p||q)
 eq D_{KL}(q||p)$: 교환법칙은 성립하지 않습니다.

출처 : <u>생성모델 입문서 - WikiDocs</u>







$$\mathcal{L}(\mathbf{w}) = \int q(\mathbf{z}_{1}|\mathbf{x}) \ln p(\mathbf{x}|\mathbf{z}_{1}, \mathbf{w}) d\mathbf{z}_{1}$$
reconstruction term
$$-\sum_{t=2}^{T} \int \frac{\mathrm{KL}(q(\mathbf{z}_{t-1}|\mathbf{z}_{t}, \mathbf{x}) || p(\mathbf{z}_{t-1}|\mathbf{z}_{t}, \mathbf{w})) q(\mathbf{z}_{t}|\mathbf{x}) d\mathbf{z}_{t}}{\mathrm{consistency terms}}$$
(20.32)

$$KL(q(\mathbf{z}_{t-1}|\mathbf{z}_t, \mathbf{x}) || p(\mathbf{z}_{t-1}|\mathbf{z}_t, \mathbf{w}))$$

$$= \frac{1}{2\beta_t} ||\mathbf{m}_t(\mathbf{x}, \mathbf{z}_t) - \boldsymbol{\mu}(\mathbf{z}_t, \mathbf{w}, t)||^2 + \text{const}$$

- KL-divergence term acts like squared-loss funtion
- minimizing squared error to maximize ELBO



20.2 Reverse Decoder – Predicting the noise

• We introduce new neural network $\mathbf{g}(\mathbf{z}_t, \mathbf{w}, t)$, that predicts total noise added to \mathbf{x} to generate \mathbf{z}_t

$$\mu(\mathbf{z}_t, \mathbf{w}, t) = \frac{1}{\sqrt{1 - \beta_t}} \left\{ \mathbf{z}_t - \frac{\beta_t}{\sqrt{1 - \alpha_t}} \mathbf{g}(\mathbf{z}_t, \mathbf{w}, t) \right\}$$

with several process, we can get

$$\mathcal{L}(\mathbf{w}) = -\sum_{t=1}^{T} \left\| \mathbf{g}(\sqrt{\alpha_t} \mathbf{x} + \sqrt{1 - \alpha_t} \boldsymbol{\epsilon}_t, \mathbf{w}, t) - \boldsymbol{\epsilon}_t \right\|^2$$

Algorithm 20.1: Training a denoising diffusion probabilistic model

return w



20.2 Reverse Decoder – Generating new sample

Algorithm 20.2: Sampling from a denoising diffusion probabilistic model

```
Input: Trained denoising network \mathbf{g}(\mathbf{z}, \mathbf{w}, t)
              Noise schedule \{\beta_1, \ldots, \beta_T\}
Output: Sample vector x in data space
\mathbf{z}_T \sim \mathcal{N}(\mathbf{z}|\mathbf{0},\mathbf{I}) // Sample from final latent space
for t \in T, \ldots, 2 do
     \alpha_t \leftarrow \prod_{\tau=1}^t (1-eta_{	au}) // Calculate alpha
       // Evaluate network output
      \mu(\mathbf{z}_t, \mathbf{w}, t) \leftarrow \frac{1}{\sqrt{1-\beta_t}} \left\{ \mathbf{z}_t - \frac{\beta_t}{\sqrt{1-\alpha_t}} \mathbf{g}(\mathbf{z}_t, \mathbf{w}, t) \right\}
      oldsymbol{\epsilon} \sim \mathcal{N}(oldsymbol{\epsilon} | \mathbf{0}, \mathbf{I}) // Sample a noise vector
      \mathbf{z}_{t-1} \leftarrow \boldsymbol{\mu}(\mathbf{z}_t, \mathbf{w}, t) + \sqrt{\beta_t} \boldsymbol{\epsilon} // Add scaled noise
end for
\mathbf{x} = \frac{1}{\sqrt{1-eta_1}} \left\{ \mathbf{z}_1 - \frac{eta_1}{\sqrt{1-lpha_1}} \mathbf{g}(\mathbf{z}_1, \mathbf{w}, t) \right\} // Final denoising step
return x
```



20.3 Score Matching

 Diffusion model is highly related with "score matching" deep generative model

• "score matching" model uses "score function" or "Stein score"

$$\mathbf{s}(\mathbf{x}) = \nabla_{\mathbf{x}} \ln p(\mathbf{x})$$

• s(x) has same dimensionality as x



20.3 Score Matching – score loss function

• loss function to learn s(x, w)

$$J(\mathbf{w}) = \frac{1}{2} \int \|\mathbf{s}(\mathbf{x}, \mathbf{w}) - \nabla_{\mathbf{x}} \ln p(\mathbf{x})\|^2 p(\mathbf{x}) d\mathbf{x}$$

- Problem of this loss function is we do not know $\nabla_{\mathbf{x}} \ln p(\mathbf{x})$
- So, we need to use data points "smear out" from the noise kernel (Parzen estimator)

$$q_{\sigma}(\mathbf{z}) = \int q(\mathbf{z}|\mathbf{x}, \sigma)p(\mathbf{x}) d\mathbf{x}$$
 (20.47)

where $q(\mathbf{z}|\mathbf{x}, \sigma)$ is the *noise kernel*. A common choice of kernel is the Gaussian

$$q(\mathbf{z}|\mathbf{x},\sigma) = \mathcal{N}(\mathbf{z}|\mathbf{x},\sigma^2\mathbf{I}). \tag{20.48}$$



20.3 Score Matching – score loss function

Change the loss function:

$$J(\mathbf{w}) = \frac{1}{2} \iint \|\mathbf{s}(\mathbf{z}, \mathbf{w}) - \nabla_{\mathbf{z}} \ln q(\mathbf{z}|\mathbf{x}, \sigma)\|^2 q(\mathbf{z}|\mathbf{x}, \sigma) p(\mathbf{x}) \, d\mathbf{z} \, d\mathbf{x} + \text{const.}$$

- If we choose noise kernel as Gaussian kernel, $\nabla_{\mathbf{z}} \ln q(\mathbf{z}|\mathbf{x},\sigma) = -\frac{1}{\sigma}\epsilon$
- Then, loss function measures difference between neural network and noise
- Therefore, this loss function has same minimum as diffusion model's loss function : $\mathbf{s}(\mathbf{x},\mathbf{w}) \sim -\frac{1}{\sigma}\mathbf{g}(\mathbf{z},\mathbf{w})$
- Langevin dynamics is used to sample from score function



20.3 Score Matching – Noise variance

- Score function has three problems
 - If probability density is zero, score function is undefined
 - If data density is low, estimation of score function is inaccurate
 - If data distribution is mixture of disjoint distribution, Langevin procedure might not sample properly
- All problems can be solved with choosing large noise variance σ^2 , but this might cause distortion of original distribution
- This can be handled with sequence of noise variance which follows

$$\sigma_1^2 < \sigma_2^2 < \dots < \sigma_T^2$$

score network is modified to get noise variance

$$\mathbf{s}(\mathbf{x},\mathbf{w},\sigma^2)$$



20.4 Guided Diffusion

- In many applications, we want diffusion models to generate image with certain "conditions"
 - conditions could be label or text description (prompt)

- most simple approach is to add conditioning variable ${\bf c}$ inside denoising neural network ${\bf g}({\bf z},{\bf w},t,{\bf c})$ and train network with $\{{\bf x}_n,{\bf c}_n\}$
- In this approach, we need "guidance" which is weight given to conditioning variable



20.4 Guided Diffusion – Classifier guidance

• If there are trained classifier $p(\mathbf{c}|\mathbf{x})$, and there are diffusion model using score function, there are relation that

$$\nabla_{\mathbf{x}} \ln p(\mathbf{x}|\mathbf{c}) = \nabla_{\mathbf{x}} \ln \left\{ \frac{p(\mathbf{c}|\mathbf{x})p(\mathbf{x})}{p(\mathbf{c})} \right\} \qquad \nabla_{\mathbf{x}} \ln p(\mathbf{c}) = 0$$
 diffusion model
$$= \nabla_{\mathbf{x}} \ln p(\mathbf{x}) + \nabla_{\mathbf{x}} \ln p(\mathbf{c}|\mathbf{x}) \qquad \text{makes denoising process}$$
 using score function
$$= \nabla_{\mathbf{x}} \ln p(\mathbf{x}) + \nabla_{\mathbf{x}} \ln p(\mathbf{c}|\mathbf{x}) \qquad \text{maximize probability of } \mathbf{c}$$

• add hyperparameter λ (guidance scale), score function becomes

$$score(\mathbf{x}, \mathbf{c}, \lambda) = \nabla_{\mathbf{x}} \ln p(\mathbf{x}) + \lambda \nabla_{\mathbf{x}} \ln p(\mathbf{c}|\mathbf{x}).$$

• drawback: need to train separate classifier $p(\mathbf{c}|\mathbf{x})$



20.4 Guided Diffusion – Classifier-free guidance

• If we change score function of "Classifier guidance"

$$\operatorname{score}(\mathbf{x}, \mathbf{c}, \lambda) = \lambda \nabla_{\mathbf{x}} \ln p(\mathbf{x}|\mathbf{c}) + (1 - \lambda) \nabla_{\mathbf{x}} \ln p(\mathbf{x})$$

- We can avoid training separate networks. ($p(\mathbf{x}) = p(\mathbf{x}|\mathbf{c} = \mathbf{0})$)
- This approach is used in many guided diffusion models
 - Text-guided diffusion model: using prompt (text sequence) as conditioning variable
 - image super-resolution : denoising high-resolution image with using low-resolution image as conditioning variable



20.4 Guided Diffusion – Stable Diffusion

