

Diffusion Models: Theory and Applications

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Overview

1. Diffusion Model **Theories**
2. Diffusion Models in Recent **Conferences** (CVPR'22/23)
3. Real-world **Applications** with Diffusion Models
4. Possible Applications on SmartPhone with Diffusion Models

Diffusion Models: The Theory

Before Diffusion Models: Score Matching

- Song et al. *Generative Modeling by Estimating Gradients of the Data Distribution*, NeurIPS 2019
- <https://score-based-methods-workshop.github.io/>
- <https://yang-song.net/blog/2021/score/>

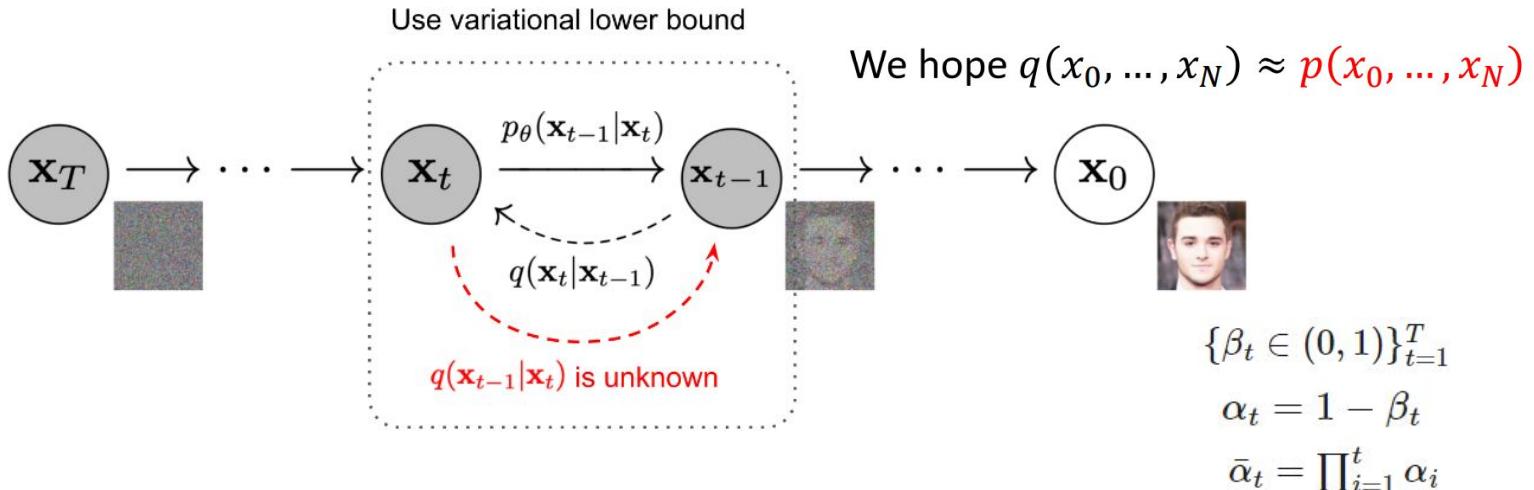
Diffusion Models:

- Ho et al. *Denoising diffusion probabilistic models (DDPM)*, NeurIPS 2020
<https://hojonathanho.github.io/diffusion/>
- Song et al. *Denoising Diffusion Implicit Models (DDIM)*, ICLR 2021
- Rombach et al. *High-Resolution Image Synthesis with Latent Diffusion Models (SD)*, CVPR 2022

Useful Links:

- <https://lilianweng.github.io/posts/2021-07-11-diffusion-models/>
-

Denoising Diffusion Probabilistic Models (DDPM)

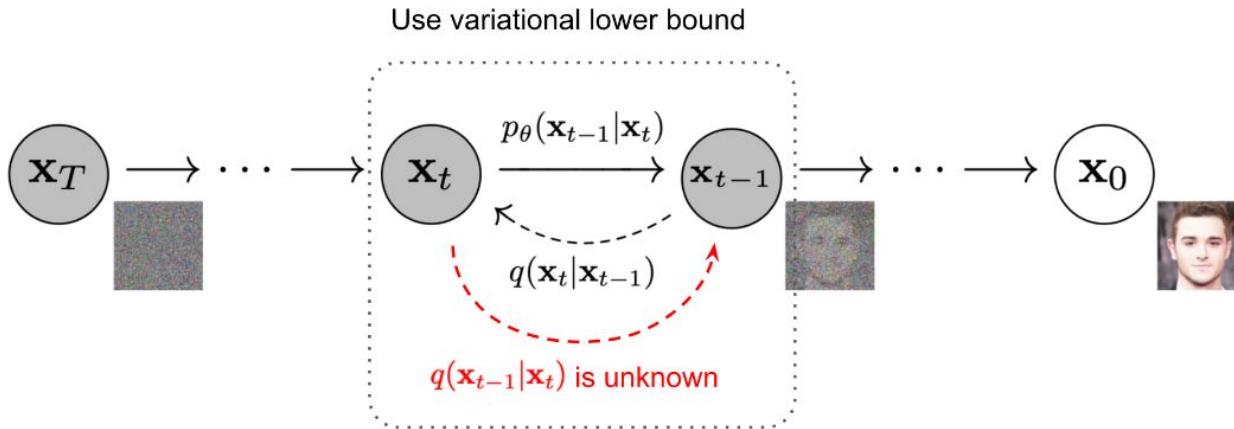


Forward process (add noise): $q(\mathbf{x}_t | \mathbf{x}_{t-1}) = \mathcal{N}(\mathbf{x}_t; \sqrt{1 - \beta_t} \mathbf{x}_{t-1}, \beta_t \mathbf{I})$ $q(\mathbf{x}_{1:T} | \mathbf{x}_0) = \prod_{t=1}^T q(\mathbf{x}_t | \mathbf{x}_{t-1})$

$$q(\mathbf{x}_t | \mathbf{x}_0) = \mathcal{N}(\mathbf{x}_t; \sqrt{\bar{\alpha}_t} \mathbf{x}_0, (1 - \bar{\alpha}_t) \mathbf{I})$$

Reverse process (denoise): $p_\theta(\mathbf{x}_{0:T}) = p(\mathbf{x}_T) \prod_{t=1}^T p_\theta(\mathbf{x}_{t-1} | \mathbf{x}_t)$ $p_\theta(\mathbf{x}_{t-1} | \mathbf{x}_t) = \mathcal{N}(\mathbf{x}_{t-1}; \boldsymbol{\mu}_\theta(\mathbf{x}_t, t), \boldsymbol{\Sigma}_\theta(\mathbf{x}_t, t))$

Denoising Diffusion Probabilistic Models (DDPM)



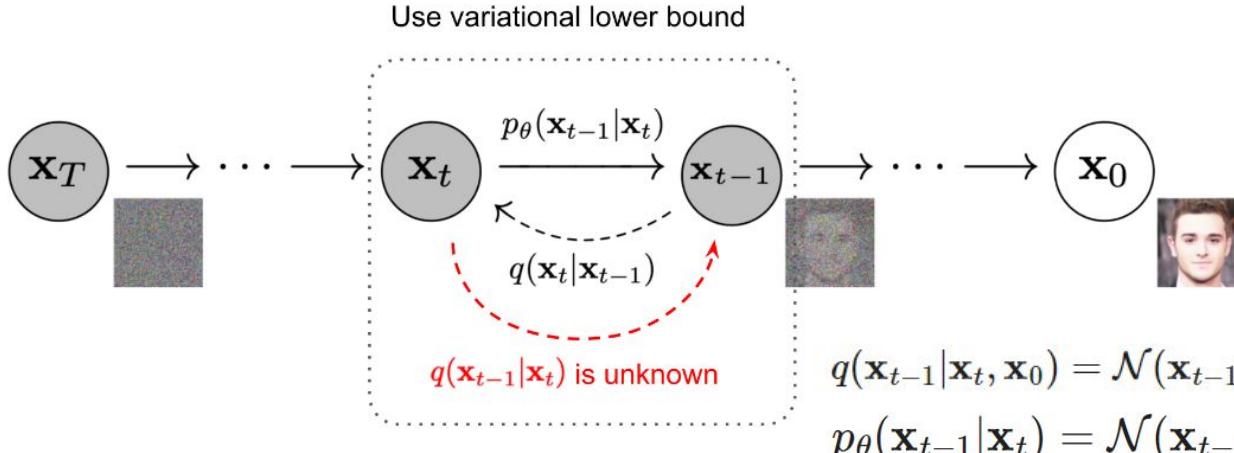
$$\begin{aligned}-\log p_\theta(\mathbf{x}_0) &\leq -\log p_\theta(\mathbf{x}_0) + D_{\text{KL}}(q(\mathbf{x}_{1:T}|\mathbf{x}_0) \| p_\theta(\mathbf{x}_{1:T}|\mathbf{x}_0)) \\&= -\log p_\theta(\mathbf{x}_0) + \mathbb{E}_{\mathbf{x}_{1:T} \sim q(\mathbf{x}_{1:T}|\mathbf{x}_0)} \left[\log \frac{q(\mathbf{x}_{1:T}|\mathbf{x}_0)}{p_\theta(\mathbf{x}_{0:T})/p_\theta(\mathbf{x}_0)} \right] \\&= -\log p_\theta(\mathbf{x}_0) + \mathbb{E}_q \left[\log \frac{q(\mathbf{x}_{1:T}|\mathbf{x}_0)}{p_\theta(\mathbf{x}_{0:T})} + \log p_\theta(\mathbf{x}_0) \right] \\&= \mathbb{E}_q \left[\log \frac{q(\mathbf{x}_{1:T}|\mathbf{x}_0)}{p_\theta(\mathbf{x}_{0:T})} \right]\end{aligned}$$

$$\text{Let } L_{\text{VLB}} = \mathbb{E}_{q(\mathbf{x}_{0:T})} \left[\log \frac{q(\mathbf{x}_{1:T}|\mathbf{x}_0)}{p_\theta(\mathbf{x}_{0:T})} \right] \geq -\mathbb{E}_{q(\mathbf{x}_0)} \log p_\theta(\mathbf{x}_0)$$

Denoising Diffusion Probabilistic Models (DDPM)

$$\begin{aligned}
L_{\text{VLB}} &= \mathbb{E}_{q(\mathbf{x}_{0:T})} \left[\log \frac{q(\mathbf{x}_{1:T} | \mathbf{x}_0)}{p_\theta(\mathbf{x}_{0:T})} \right] \\
&= \mathbb{E}_q \left[\log \frac{\prod_{t=1}^T q(\mathbf{x}_t | \mathbf{x}_{t-1})}{p_\theta(\mathbf{x}_T) \prod_{t=1}^T p_\theta(\mathbf{x}_{t-1} | \mathbf{x}_t)} \right] \\
&= \mathbb{E}_q \left[-\log p_\theta(\mathbf{x}_T) + \sum_{t=1}^T \log \frac{q(\mathbf{x}_t | \mathbf{x}_{t-1})}{p_\theta(\mathbf{x}_{t-1} | \mathbf{x}_t)} \right] \\
&= \mathbb{E}_q \left[-\log p_\theta(\mathbf{x}_T) + \sum_{t=2}^T \log \frac{q(\mathbf{x}_t | \mathbf{x}_{t-1})}{p_\theta(\mathbf{x}_{t-1} | \mathbf{x}_t)} + \log \frac{q(\mathbf{x}_1 | \mathbf{x}_0)}{p_\theta(\mathbf{x}_0 | \mathbf{x}_1)} \right] \\
&= \mathbb{E}_q \left[-\log p_\theta(\mathbf{x}_T) + \sum_{t=2}^T \log \left(\frac{q(\mathbf{x}_{t-1} | \mathbf{x}_t, \mathbf{x}_0)}{p_\theta(\mathbf{x}_{t-1} | \mathbf{x}_t)} \cdot \frac{q(\mathbf{x}_t | \mathbf{x}_0)}{q(\mathbf{x}_{t-1} | \mathbf{x}_0)} \right) + \log \frac{q(\mathbf{x}_1 | \mathbf{x}_0)}{p_\theta(\mathbf{x}_0 | \mathbf{x}_1)} \right] \\
&= \mathbb{E}_q \left[-\log p_\theta(\mathbf{x}_T) + \sum_{t=2}^T \log \frac{q(\mathbf{x}_{t-1} | \mathbf{x}_t, \mathbf{x}_0)}{p_\theta(\mathbf{x}_{t-1} | \mathbf{x}_t)} + \sum_{t=2}^T \log \frac{q(\mathbf{x}_t | \mathbf{x}_0)}{q(\mathbf{x}_{t-1} | \mathbf{x}_0)} + \log \frac{q(\mathbf{x}_1 | \mathbf{x}_0)}{p_\theta(\mathbf{x}_0 | \mathbf{x}_1)} \right] \\
&= \mathbb{E}_q \left[-\log p_\theta(\mathbf{x}_T) + \sum_{t=2}^T \log \frac{q(\mathbf{x}_{t-1} | \mathbf{x}_t, \mathbf{x}_0)}{p_\theta(\mathbf{x}_{t-1} | \mathbf{x}_t)} + \log \frac{q(\mathbf{x}_T | \mathbf{x}_0)}{q(\mathbf{x}_1 | \mathbf{x}_0)} + \log \frac{q(\mathbf{x}_1 | \mathbf{x}_0)}{p_\theta(\mathbf{x}_0 | \mathbf{x}_1)} \right] \\
&= \mathbb{E}_q \left[\log \frac{q(\mathbf{x}_T | \mathbf{x}_0)}{p_\theta(\mathbf{x}_T)} + \sum_{t=2}^T \log \frac{q(\mathbf{x}_{t-1} | \mathbf{x}_t, \mathbf{x}_0)}{p_\theta(\mathbf{x}_{t-1} | \mathbf{x}_t)} - \log p_\theta(\mathbf{x}_0 | \mathbf{x}_1) \right] \\
&= \mathbb{E}_q \underbrace{[D_{\text{KL}}(q(\mathbf{x}_T | \mathbf{x}_0) \| p_\theta(\mathbf{x}_T))]}_{L_T} + \sum_{t=2}^T \underbrace{[D_{\text{KL}}(q(\mathbf{x}_{t-1} | \mathbf{x}_t, \mathbf{x}_0) \| p_\theta(\mathbf{x}_{t-1} | \mathbf{x}_t)) - \log p_\theta(\mathbf{x}_0 | \mathbf{x}_1)]}_{L_{t-1}}
\end{aligned}$$

Denoising Diffusion Probabilistic Models (DDPM)



$$\begin{aligned}\tilde{\boldsymbol{\mu}}_t(\mathbf{x}_t, \mathbf{x}_0) &= \left(\frac{\sqrt{\alpha_t}}{\beta_t} \mathbf{x}_t + \frac{\sqrt{\bar{\alpha}_{t-1}}}{1 - \bar{\alpha}_{t-1}} \mathbf{x}_0 \right) / \left(\frac{\alpha_t}{\beta_t} + \frac{1}{1 - \bar{\alpha}_{t-1}} \right) \quad \tilde{\boldsymbol{\mu}}_t = \frac{\sqrt{\alpha_t}(1 - \bar{\alpha}_{t-1})}{1 - \bar{\alpha}_t} \mathbf{x}_t + \frac{\sqrt{\bar{\alpha}_{t-1}}\beta_t}{1 - \bar{\alpha}_t} \frac{1}{\sqrt{\alpha_t}} (\mathbf{x}_t - \sqrt{1 - \bar{\alpha}_t} \boldsymbol{\epsilon}_t) \\ &= \left(\frac{\sqrt{\alpha_t}}{\beta_t} \mathbf{x}_t + \frac{\sqrt{\bar{\alpha}_{t-1}}}{1 - \bar{\alpha}_{t-1}} \mathbf{x}_0 \right) \frac{1 - \bar{\alpha}_{t-1}}{1 - \bar{\alpha}_t} \cdot \beta_t \\ &= \frac{\sqrt{\alpha_t}(1 - \bar{\alpha}_{t-1})}{1 - \bar{\alpha}_t} \mathbf{x}_t + \frac{\sqrt{\bar{\alpha}_{t-1}}\beta_t}{1 - \bar{\alpha}_t} \mathbf{x}_0\end{aligned}$$

Reparameterization

$$\boldsymbol{\mu}_\theta(\mathbf{x}_t, t) = \frac{1}{\sqrt{\alpha_t}} \left(\mathbf{x}_t - \frac{1 - \alpha_t}{\sqrt{1 - \bar{\alpha}_t}} \boldsymbol{\epsilon}_\theta(\mathbf{x}_t, t) \right)$$

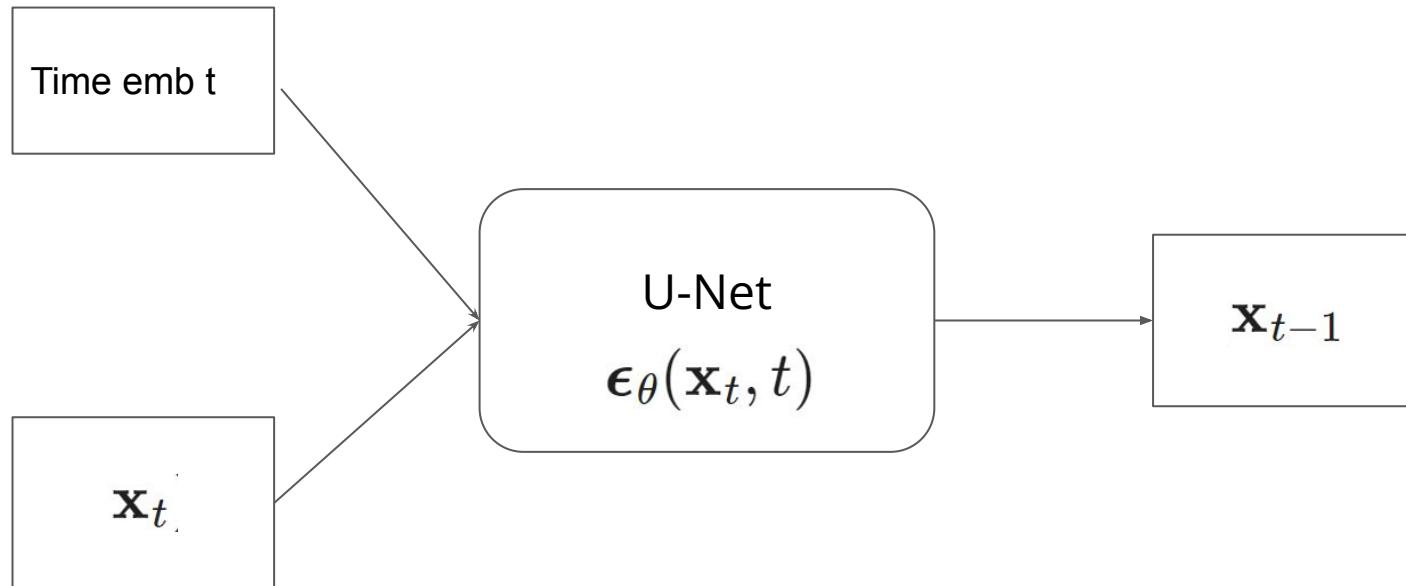
Denoising Diffusion Probabilistic Models (DDPM)

$$\begin{aligned} L_t &= \mathbb{E}_{\mathbf{x}_0, \epsilon} \left[\frac{1}{2\|\Sigma_\theta(\mathbf{x}_t, t)\|_2^2} \|\tilde{\mu}_t(\mathbf{x}_t, \mathbf{x}_0) - \mu_\theta(\mathbf{x}_t, t)\|^2 \right] \\ &= \mathbb{E}_{\mathbf{x}_0, \epsilon} \left[\frac{1}{2\|\Sigma_\theta\|_2^2} \left\| \frac{1}{\sqrt{\alpha_t}} \left(\mathbf{x}_t - \frac{1-\alpha_t}{\sqrt{1-\bar{\alpha}_t}} \epsilon_t \right) - \frac{1}{\sqrt{\alpha_t}} \left(\mathbf{x}_t - \frac{1-\alpha_t}{\sqrt{1-\bar{\alpha}_t}} \epsilon_\theta(\mathbf{x}_t, t) \right) \right\|^2 \right] \\ &= \mathbb{E}_{\mathbf{x}_0, \epsilon} \left[\frac{(1-\alpha_t)^2}{2\alpha_t(1-\bar{\alpha}_t)\|\Sigma_\theta\|_2^2} \|\epsilon_t - \epsilon_\theta(\mathbf{x}_t, t)\|^2 \right] \\ &= \mathbb{E}_{\mathbf{x}_0, \epsilon} \left[\frac{(1-\alpha_t)^2}{2\alpha_t(1-\bar{\alpha}_t)\|\Sigma_\theta\|_2^2} \|\epsilon_t - \epsilon_\theta(\sqrt{\bar{\alpha}_t}\mathbf{x}_0 + \sqrt{1-\bar{\alpha}_t}\epsilon_t, t)\|^2 \right] \end{aligned}$$

$$\begin{aligned} L_t^{\text{simple}} &= \mathbb{E}_{t \sim [1, T], \mathbf{x}_0, \epsilon_t} \left[\|\epsilon_t - \epsilon_\theta(\mathbf{x}_t, t)\|^2 \right] \\ &= \mathbb{E}_{t \sim [1, T], \mathbf{x}_0, \epsilon_t} \left[\|\epsilon_t - \epsilon_\theta(\sqrt{\bar{\alpha}_t}\mathbf{x}_0 + \sqrt{1-\bar{\alpha}_t}\epsilon_t, t)\|^2 \right] \end{aligned}$$

Denoising Diffusion Probabilistic Models (DDPM)

Network Architecture: Conditional U-Net

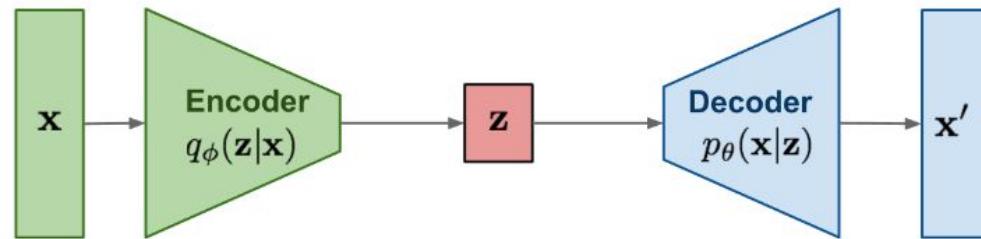


Diffusion Model vs. VAE

Detailed explanation:

https://zhuanlan.zhihu.com/p/563543020?utm_source=wechat_session&utm_medium=social&utm_source_id=675020504734371840

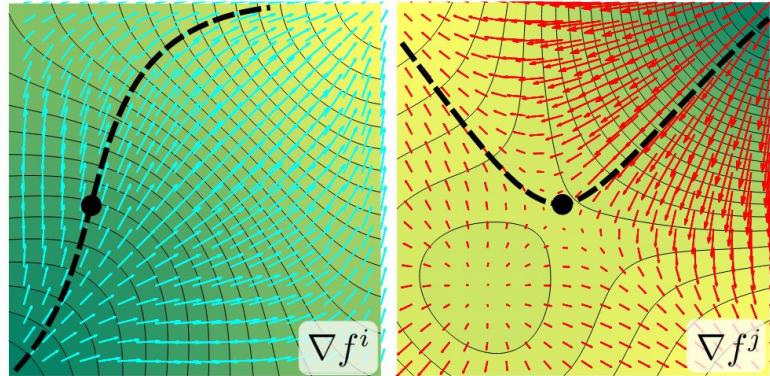
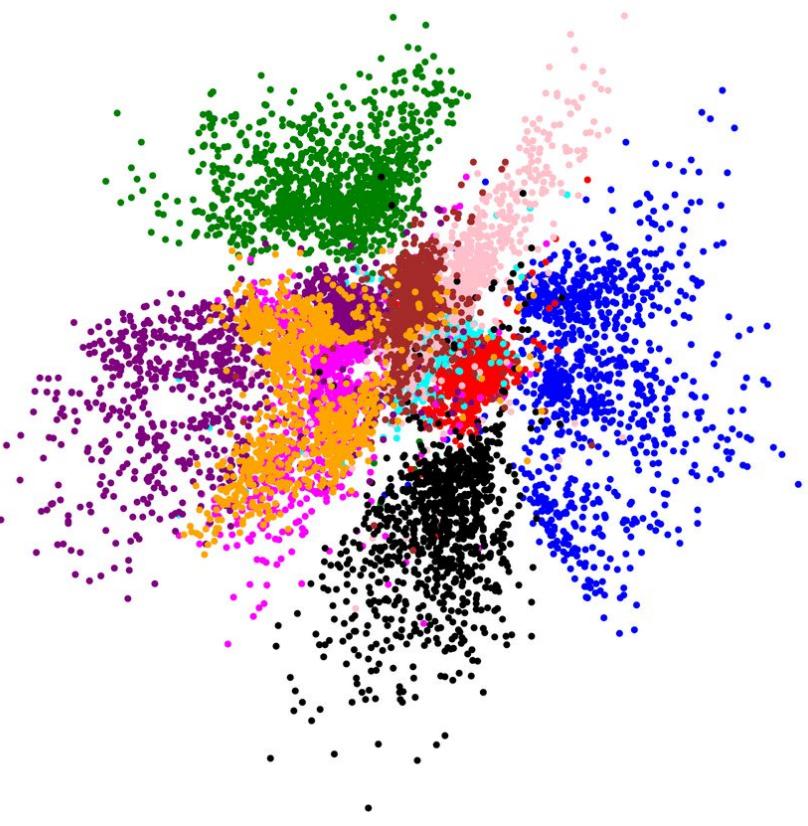
VAE: maximize variational lower bound



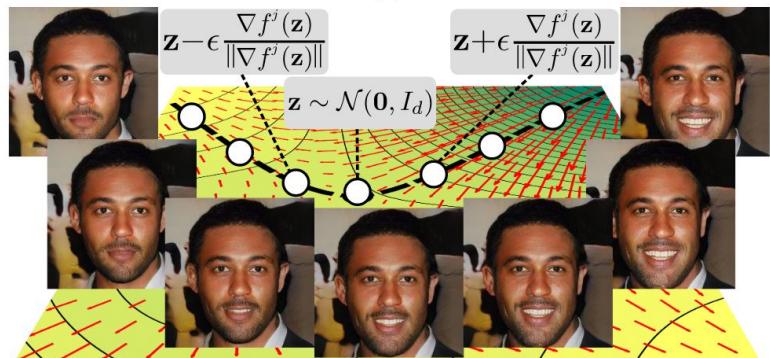
In short:

1. Easy to train and simple loss
2. VAE has limited latent space, which suffers low quality generation
3. Reversible
4. Diffusion model is updated version of VAE

Generative Model and Latent Space



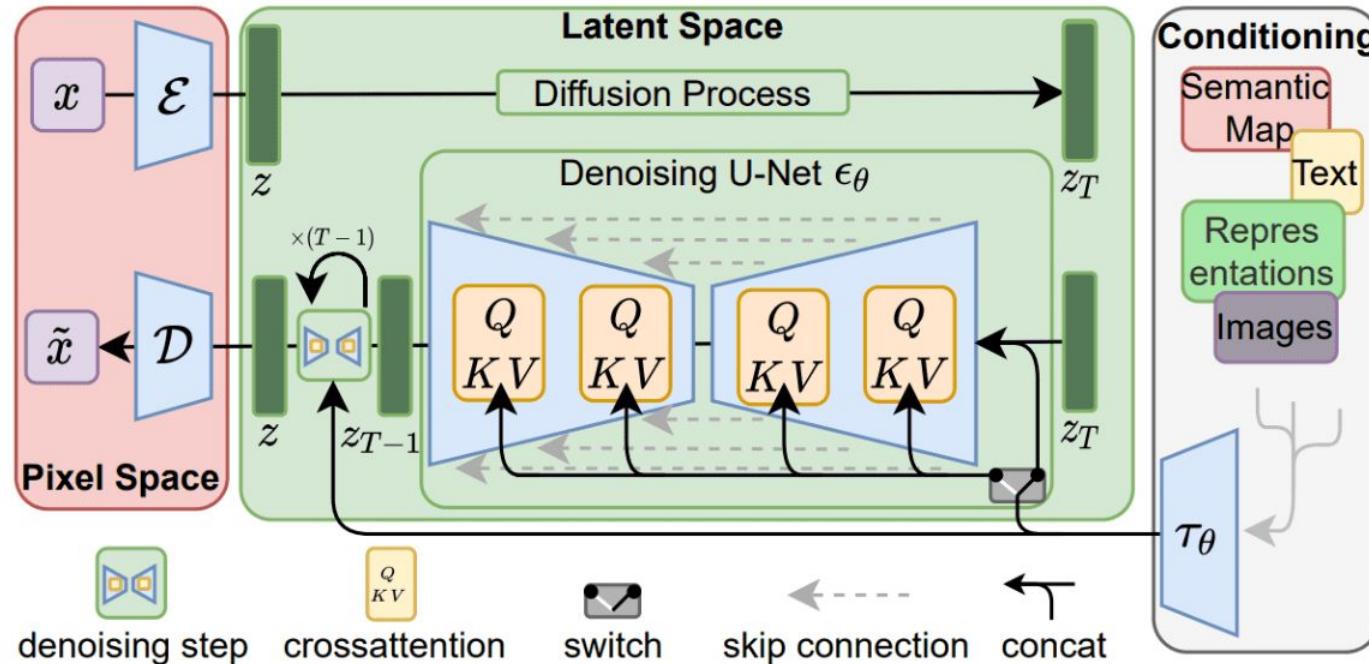
(a)



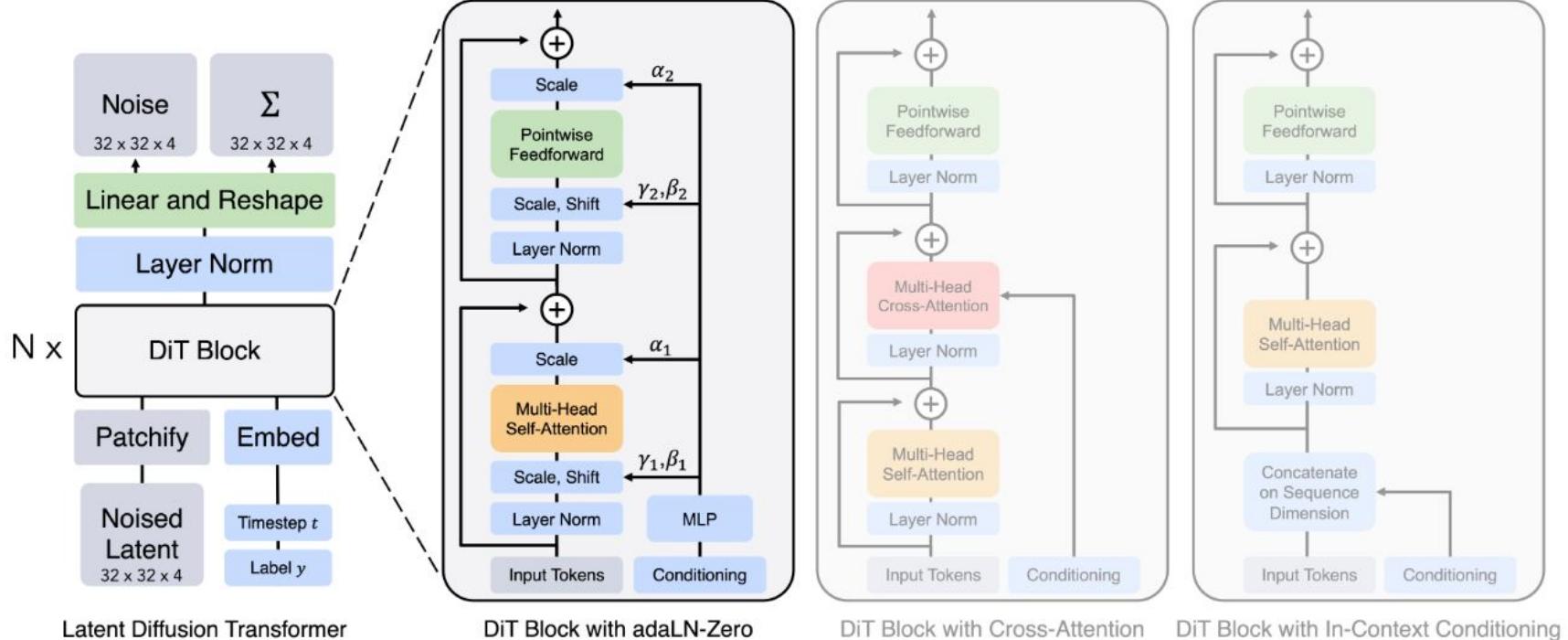
(b)

High-Resolution Image Synthesis with Latent Diffusion Models

- Stable Diffusion



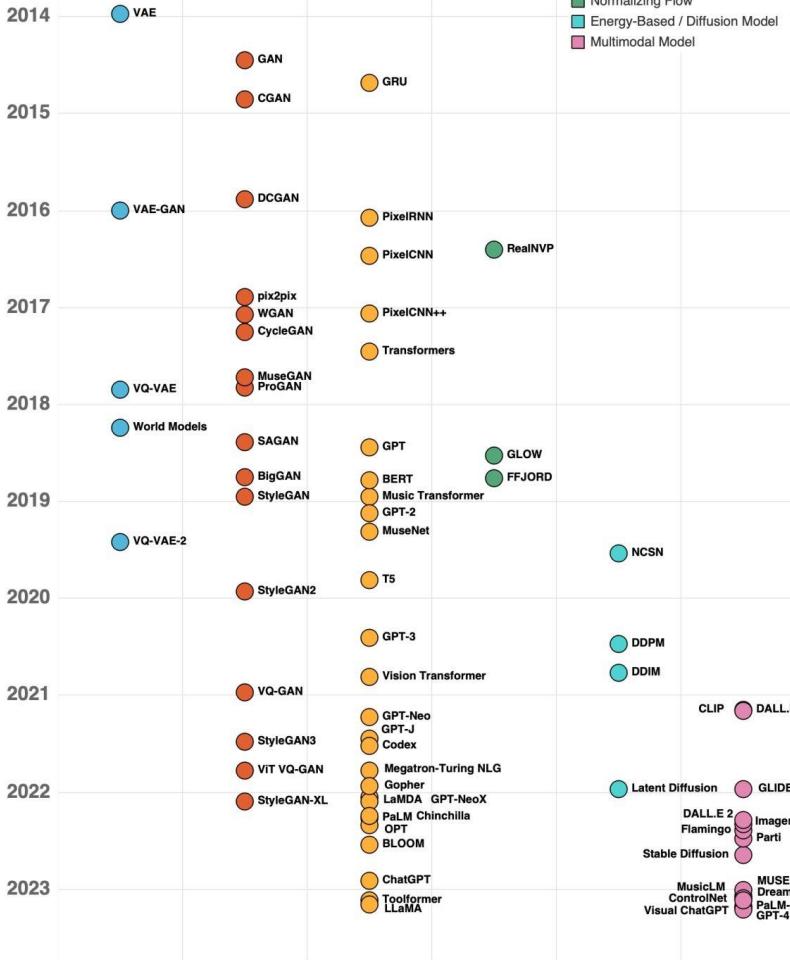
Scalable Diffusion Models with Transformers - DiT



Generative AI Timeline

Model Type

- Variational Autoencoder
- Generative Adversarial Network
- Autoregressive / Transformer
- Normalizing Flow
- Energy-Based / Diffusion Model
- Multimodal Model



Fast Sampling on Diffusion Models

1. Theory-based
2. Training-based
3. Optimization-based

METHOD	NFE (\downarrow)	FID (\downarrow)	IS (\uparrow)		METHOD	NFE (\downarrow)	FID (\downarrow)	Prec. (\uparrow)	Rec. (\uparrow)
Diffusion + Samplers									
DDIM (Song et al., 2020)	50	4.67			PD [†] (Salimans & Ho, 2022)	1	15.39	0.59	0.62
DDIM (Song et al., 2020)	20	6.84			DFNO* (Zheng et al., 2022)	1	8.35		
DDIM (Song et al., 2020)	10	8.23			CD[†]	1	6.20	0.68	0.63
DPM-solver-2 (Lu et al., 2022)	12	5.28			PD [†] (Salimans & Ho, 2022)	2	8.95	0.63	0.65
DPM-solver-3 (Lu et al., 2022)	12	6.03			CD[†]	2	4.70	0.69	0.64
3-DEIS (Zhang & Chen, 2022)	10	4.17			ADM (Dhariwal & Nichol, 2021)	250	2.07	0.74	0.63
Diffusion + Distillation									
Knowledge Distillation* (Luhman & Luhman, 2021)	1	9.36			EDM (Karras et al., 2022)	79	2.44	0.71	0.67
DFNO* (Zheng et al., 2022)	1	4.12			BigGAN-deep (Brock et al., 2019)	1	4.06	0.79	0.48
1-Rectified Flow (+distill)* (Liu et al., 2022)	1	6.18	9.08		CT	1	13.0	0.71	0.47
2-Rectified Flow (+distill)* (Liu et al., 2022)	1	4.85	9.01		CT	2	11.1	0.69	0.56
3-Rectified Flow (+distill)* (Liu et al., 2022)	1	5.21	8.79						
PD (Salimans & Ho, 2022)	1	8.34	8.69						
CD	1	3.55	9.48						
PD (Salimans & Ho, 2022)	2	5.58	9.05						
CD	2	2.93	9.75						
Direct Generation									
BigGAN (Brock et al., 2019)	1	14.7	9.22		PD [†] (Salimans & Ho, 2022)	1	16.92	0.47	0.27
CR-GAN (Zhang et al., 2019)	1	14.6	8.40		PD [†] (Salimans & Ho, 2022)	2	8.47	0.56	0.39
AutoGAN (Gong et al., 2019)	1	12.4	8.55		CD[†]	1	7.80	0.66	0.34
E2GAN (Tian et al., 2020)	1	11.3	8.51		CD[†]	2	5.22	0.68	0.39
ViTGAN (Lee et al., 2021)	1	6.66	9.30		DDPM (Ho et al., 2020)	1000	4.89	0.60	0.45
TransGAN (Jiang et al., 2021)	1	9.26	9.05		ADM (Dhariwal & Nichol, 2021)	1000	1.90	0.66	0.51
StyleGAN2-ADA (Karras et al., 2020)	1	2.92	9.83		EDM (Karras et al., 2022)	79	3.57	0.66	0.45
StyleGAN-XL (Sauer et al., 2022)	1	1.85			SS-GAN (Chen et al., 2019b)	1	13.3		
Score SDE (Song et al., 2021)	2000	2.20	9.89		PGGAN (Karras et al., 2018)	1	8.34		
DDPM (Ho et al., 2020)	1000	3.17	9.46		PG-SWGAN (Wu et al., 2019)	1	8.0		
LSGM (Vahdat et al., 2021)	147	2.10			StyleGAN2 (Karras et al., 2020)	1	2.35	0.59	0.48
PFGM (Xu et al., 2022)	110	2.35	9.68		CT	1	16.0	0.60	0.17
EDM (Karras et al., 2022)	36	2.04	9.84		CT	2	7.85	0.68	0.33
1-Rectified Flow (Liu et al., 2022)	1	378	1.13						
Glow (Kingma & Dhariwal, 2018)	1	48.9	3.92						
Residual Flow (Chen et al., 2019a)	1	46.4							
GLFlow (Xiao et al., 2019)	1	44.6							
DenseFlow (Grcić et al., 2021)	1	34.9							
DC-VAE (Parmar et al., 2021)	1	17.9	8.20						
CT	1	8.70	8.49						
CT	2	5.83	8.85						
LSUN Bedroom 256 × 256									
PD [†] (Salimans & Ho, 2022)	1	29.6	0.51	0.25					
PD [†] (Salimans & Ho, 2022)	2	15.5	0.59	0.36					
CD[†]	1	11.0	0.65	0.36					
CD[†]	2	8.84	0.66	0.40					
DDPM (Ho et al., 2020)	1000	17.1	0.53	0.48					
ADM (Dhariwal & Nichol, 2021)	1000	5.57	0.63	0.52					
EDM (Karras et al., 2022)	79	6.69	0.70	0.43					
PGGAN (Karras et al., 2018)	1	37.5							
StyleGAN2 (Karras et al., 2020)	1	7.25	0.58	0.43					
CT	1	20.7	0.56	0.23					
CT	2	11.7	0.63	0.36					
LSUN Cat 256 × 256									
PD [†] (Salimans & Ho, 2022)	1	29.6	0.51	0.25					
PD [†] (Salimans & Ho, 2022)	2	15.5	0.59	0.36					
CD[†]	1	11.0	0.65	0.36					
CD[†]	2	8.84	0.66	0.40					
DDPM (Ho et al., 2020)	1000	17.1	0.53	0.48					
ADM (Dhariwal & Nichol, 2021)	1000	5.57	0.63	0.52					
EDM (Karras et al., 2022)	79	6.69	0.70	0.43					
PGGAN (Karras et al., 2018)	1	37.5							
StyleGAN2 (Karras et al., 2020)	1	7.25	0.58	0.43					
CT	1	20.7	0.56	0.23					
CT	2	11.7	0.63	0.36					

Applications of Diffusion Models

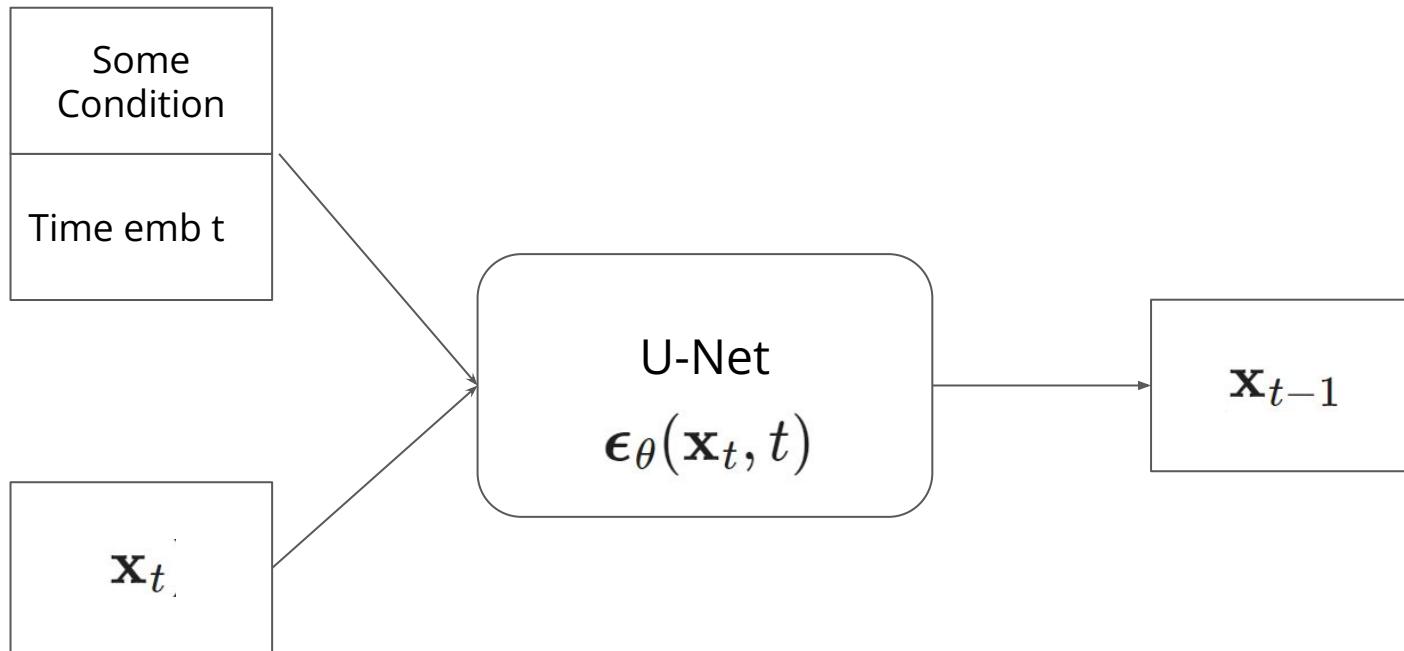
Summary

1. Unconditional Generation
2. Conditional Generation
3. Text-to-Image generation
4. Super-Resolution
5. Image Editing
6. Region Image Editing
7. Inpainting
8. Image-to-Image Translation
9. Image Segmentation
10. Multi-Task
11. Medical Image-to-Image Translation
12. Medical Image Generation
13. Medical Image Segmentation
14. Medical Image Anomaly Detection
15. Video Generation
16. Few-Shot Image Generation
17. Counterfactual Explanations and Estimations
18. Image Restoration
19. Image Registration
20. Adversarial Purification
21. Semantic Image Generation
22. Shape Generation and Completion
23. Classification
24. Point Cloud Generation
25. Theoretical
26. Graphs
27. Deblurring
28. Face Morphing Attack Detection

<https://github.com/CroitoruAlin/Diffusion-Models-in-Vision-A-Survey>

Conditional Diffusion Model

Network Architecture: Conditional U-Net



Guided Diffusion & Classifier-free Diffusion

Guided: <https://arxiv.org/abs/2105.05233> <https://arxiv.org/abs/2102.09672>

$$p_{\theta}(x_{t-1}|x_t) \longrightarrow p_{\theta,\phi}(x_{t-1}|x_t, y)$$

Train a classifier for noisy variable and apply on pre-trained ddpm

Classifier-free: <https://openreview.net/pdf?id=qw8AKxfYbl>

No extra classifier, update p process with

$$\hat{\epsilon}_{\theta}(x_t|y) = \epsilon_{\theta}(x_t) + s \cdot (\epsilon_{\theta}(x_t, y) - \epsilon_{\theta}(x_t))$$

Diffusion Models in Recent Conferences

Diffusion models got overwhelmed attention in the past year. At the time I worked on diffusion models during CVPR'22, only < 15 papers published in such topic. But now there are over 100 papers only on CVPR'23.

I selected some interesting work in CVPR'22/23 and ignored those on ICLR/NIPS/ICML since CVPR papers are more practical and less theoretical. I'm happy to elaborate if any interesting subject is favored in the discussion.

Some Surveys

A Comprehensive Survey of AI-Generated Content (AIGC): A History of Generative AI from GAN to ChatGPT

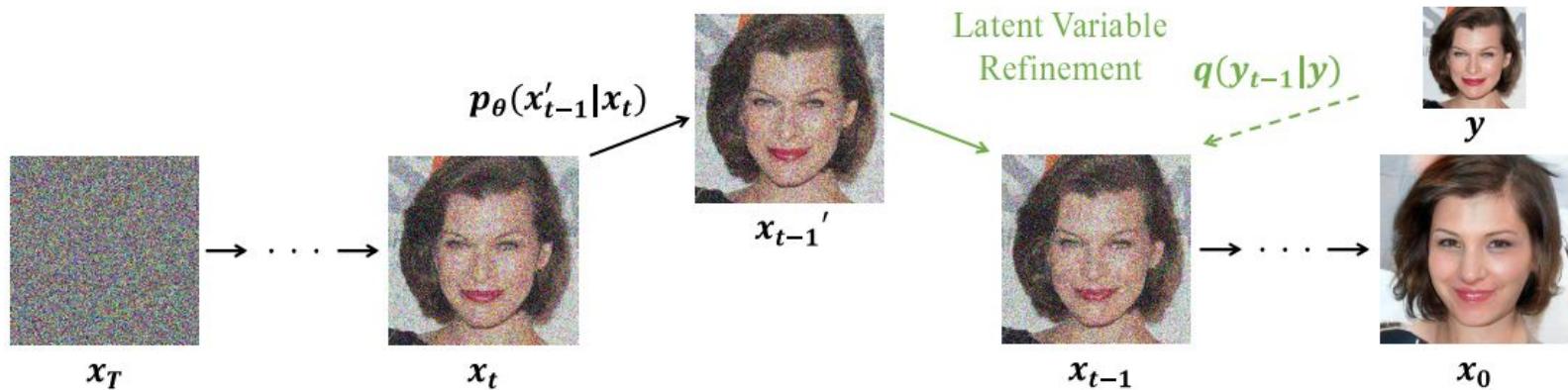
<https://arxiv.org/pdf/2303.04226.pdf>

<https://github.com/CroitoruAlin/Diffusion-Models-in-Vision-A-Survey>

ILVR (ICCV2021)

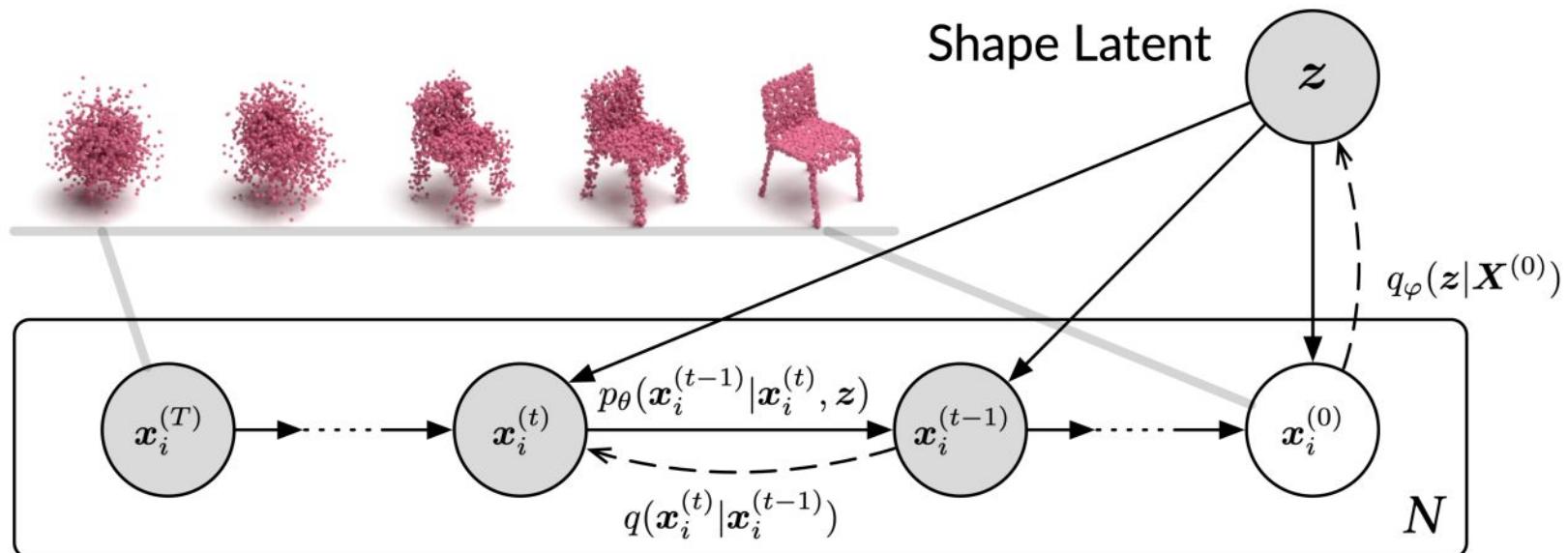
Very impressive work. Use filter to guide ddpm generate conditional images.

https://openaccess.thecvf.com/content/ICCV2021/papers/Choi_ILVR_Conditioning_Method_for_Denoising_Diffusion_Probabilistic_Models_ICCV_2021_paper.pdf



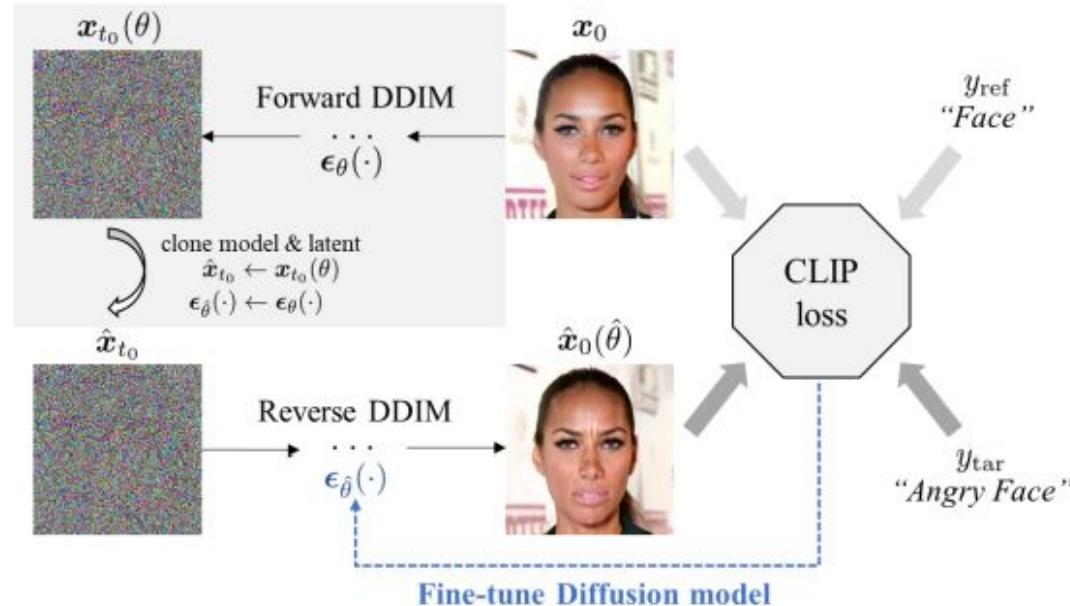
Diffusion Probabilistic Models for 3D Point Cloud Generation (ICCV2021)

Point Cloud Generation



DiffusionCLIP: Text-Guided Diffusion Models for Robust Image Manipulation (CVPR2022)

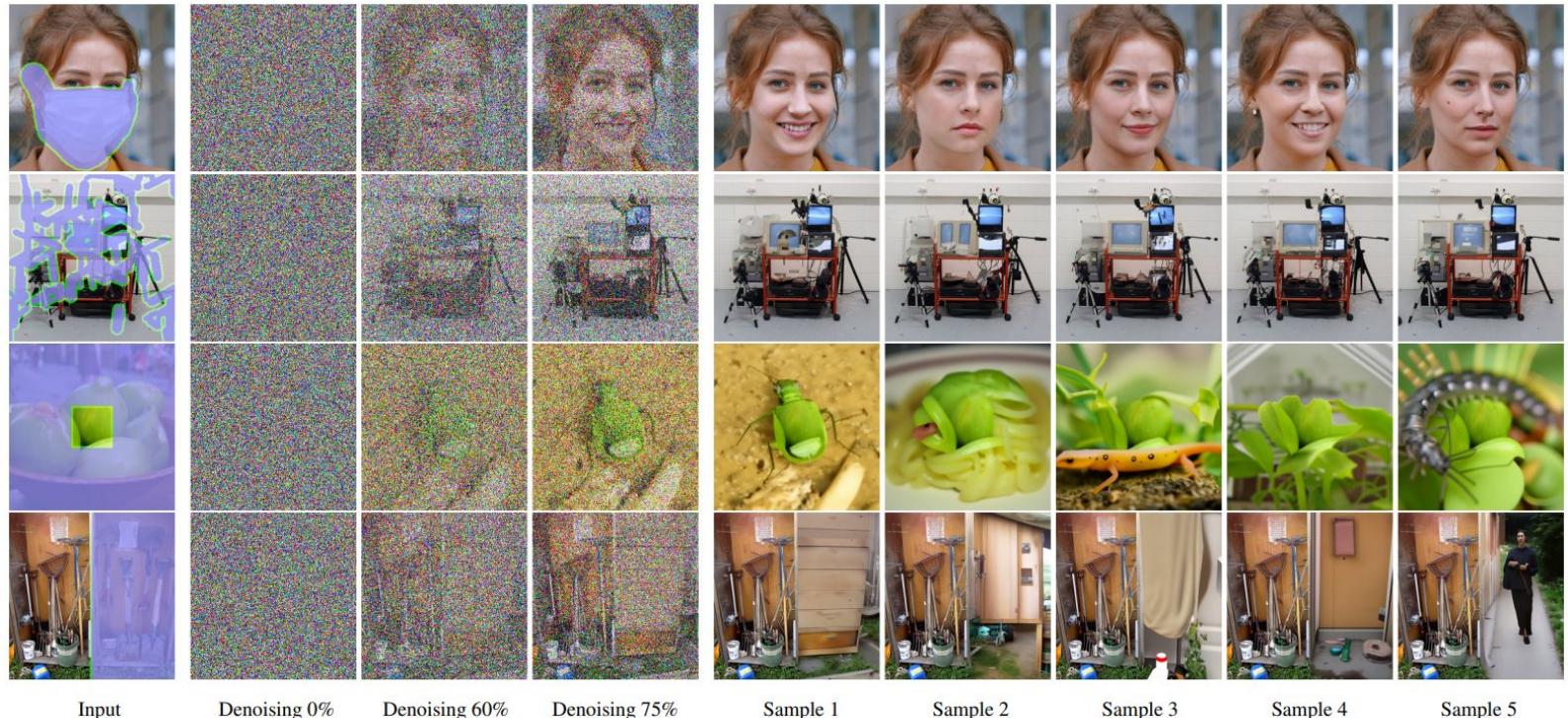
Image Editing



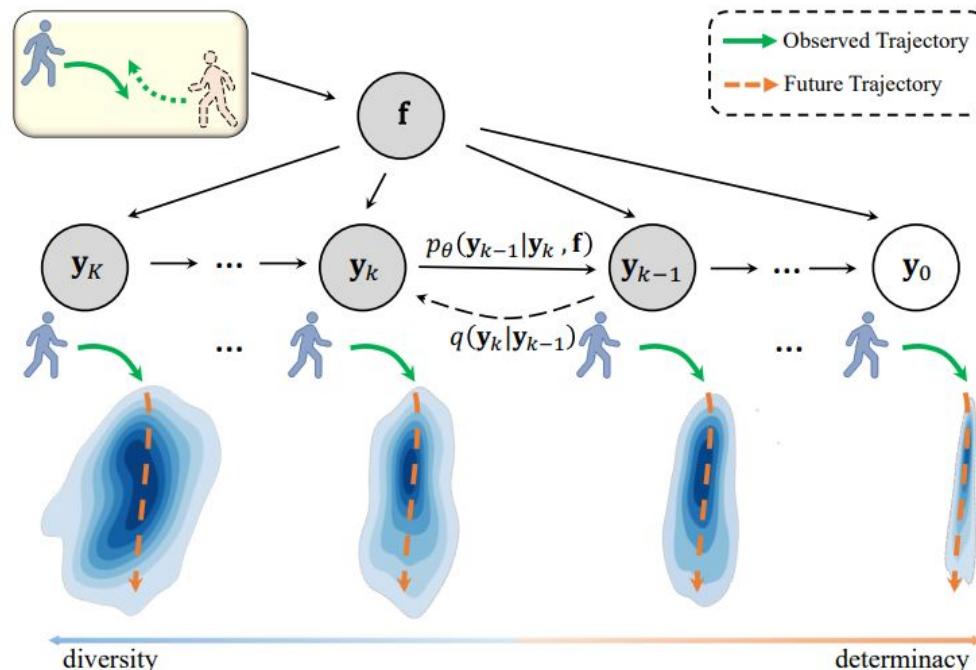
OpenAI

1. Glide
2. DALLE1
3. DALLE2

RePaint: Inpainting using Denoising Diffusion Probabilistic Models (CVPR2022)



Stochastic Trajectory Prediction via Motion Indeterminacy Diffusion (CVPR2022)



https://openaccess.thecvf.com/content/CVPR2022/papers/Gu_Stochastic_Trajectory_Prediction_via_Motion_Indeterminacy_Diffusion_CVPR_2022_paper.pdf

Blended Diffusion for Text-driven Editing of Natural Images (CVPR2022)



Input Image



Input Mask

+ “A man with a
yellow
sweater”



Result

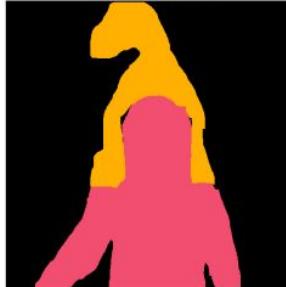
SpaText: Spatio-Textual Representation for Controllable Image Generation (CVPR2023)

“in the forest”



“a black cat with a red sweater and a blue jeans”

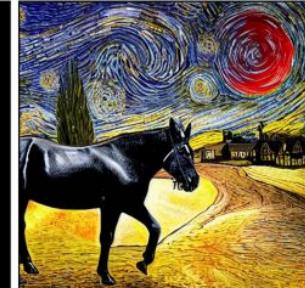
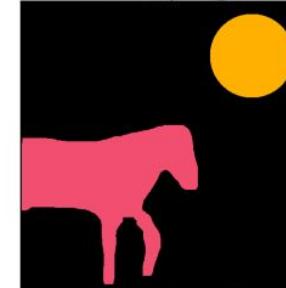
“on the moon”



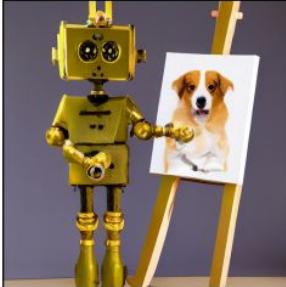
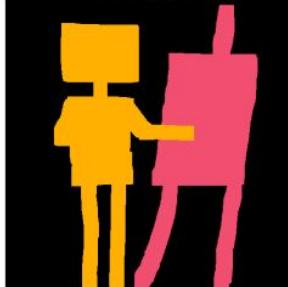
“an astronaut”
“a horse”



“in the style of
The Starry Night”

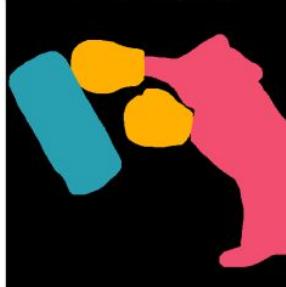


“in an empty room”



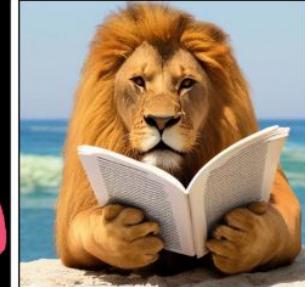
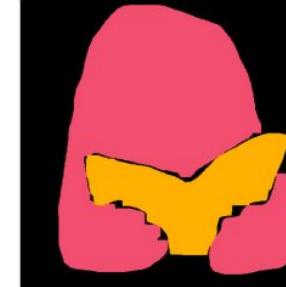
“a canvas with a painting
of a Corgi dog”
“a metallic yellow robot”

“on a snowy day”



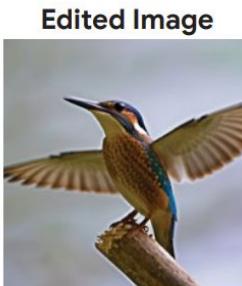
“a mouse”
“boxing gloves”
“a black punching bag”

“at the beach”



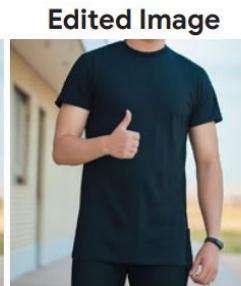
“a lion”
“a book”

Imagic: Text-Based Real Image Editing with Diffusion Models (CVPR2023)



Target Text:

“A bird spreading wings”



“A person giving the thumbs up”



“A goat jumping over a cat”

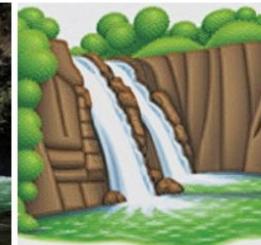


Target Text:

“A sitting dog”

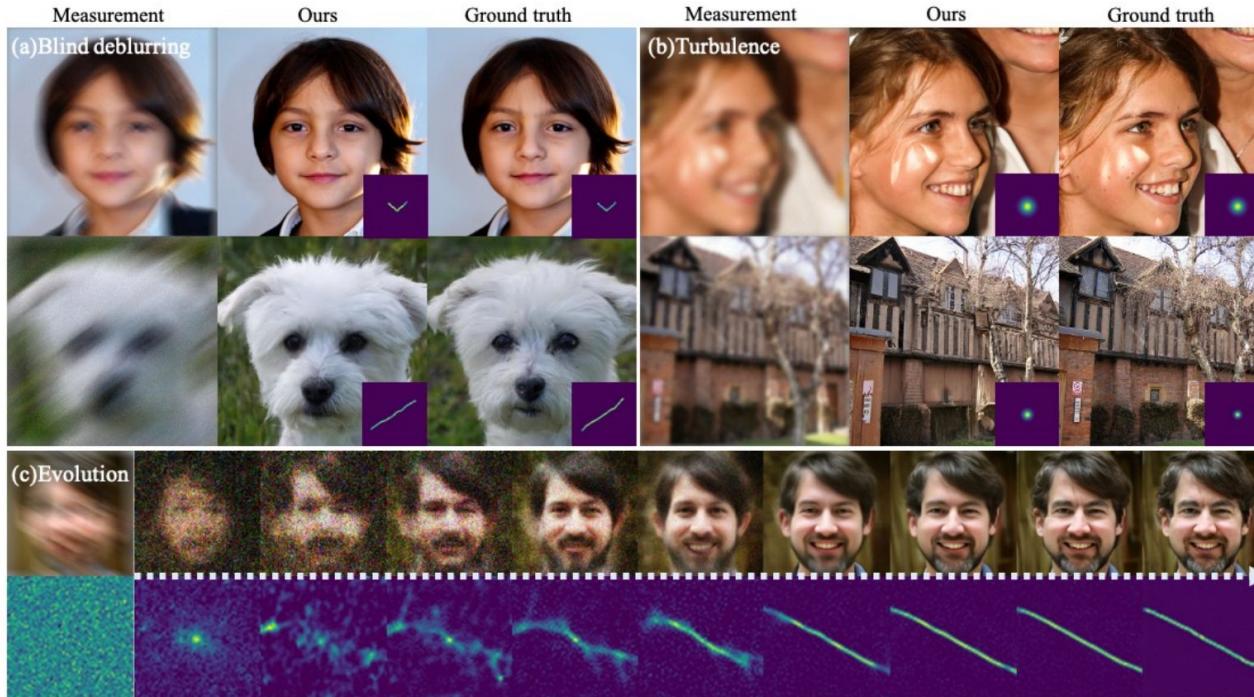


“Two kissing parrots”



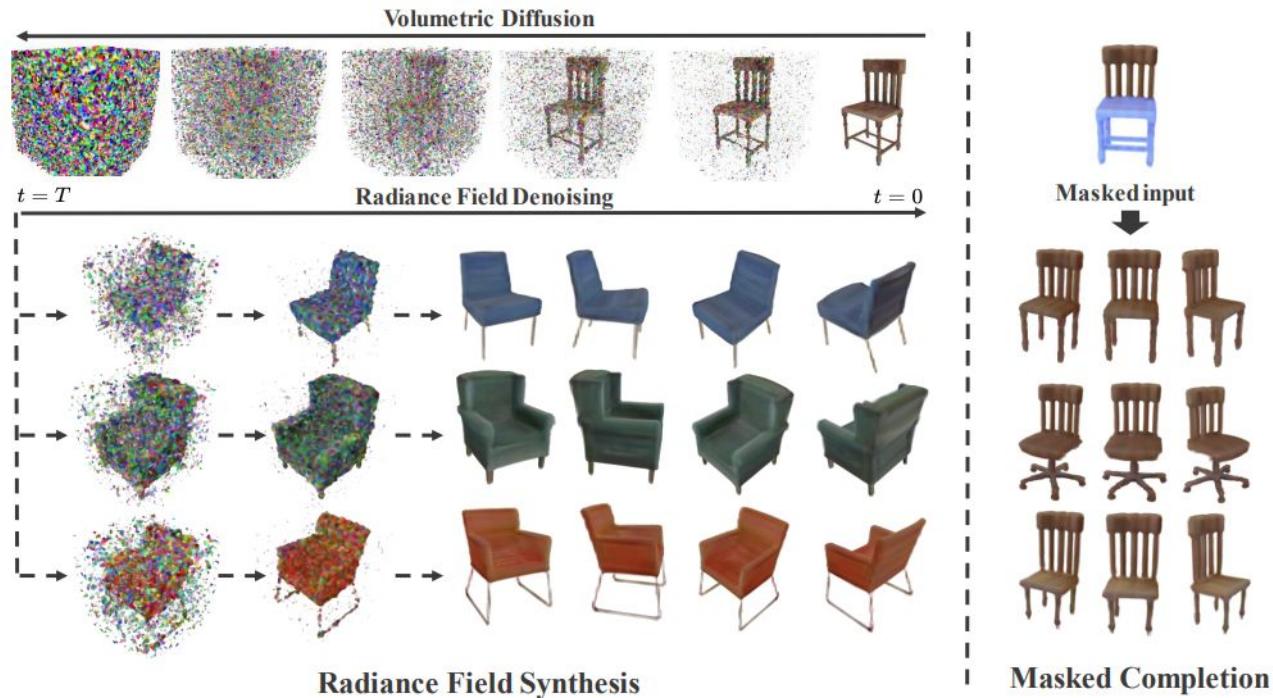
“A children’s drawing of a waterfall”

Parallel Diffusion Models of Operator and Image for Blind Inverse Problems (CVPR2023)

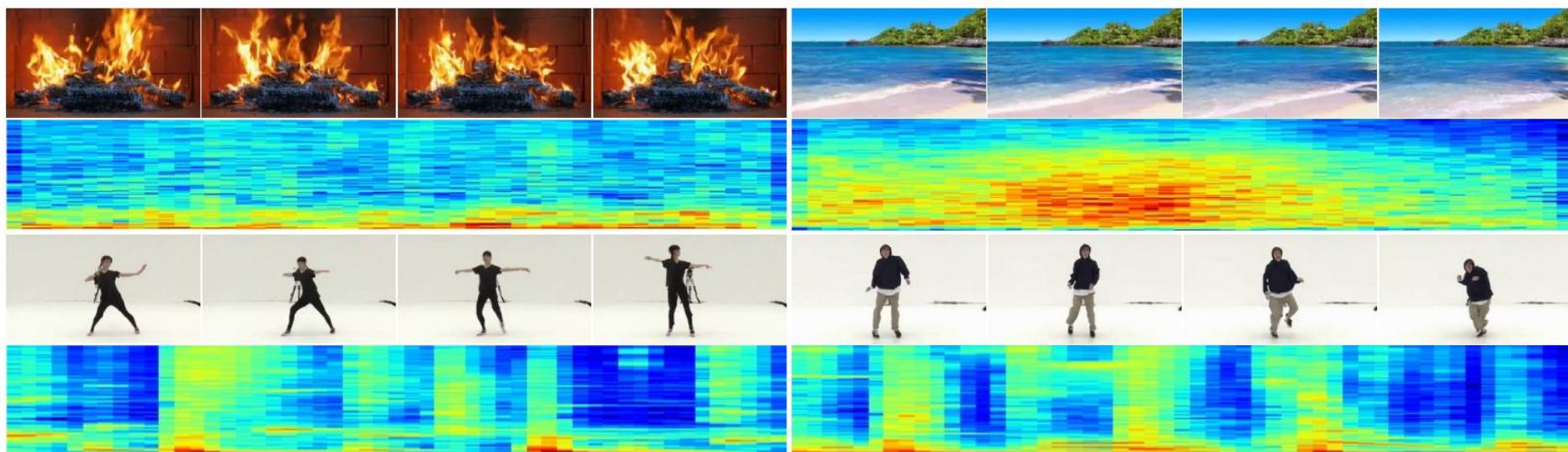


<https://arxiv.org/pdf/2211.10656.pdf>

DiffRF: Rendering-Guided 3D Radiance Field Diffusion (CVPR2023)



MM-Diffusion: Learning Multi-Modal Diffusion Models for Joint Audio and Video Generation (CVPR2023)



HouseDiffusion: Vector Floorplan Generation via a Diffusion Model with Discrete and Continuous Denoising (CVPR2023)

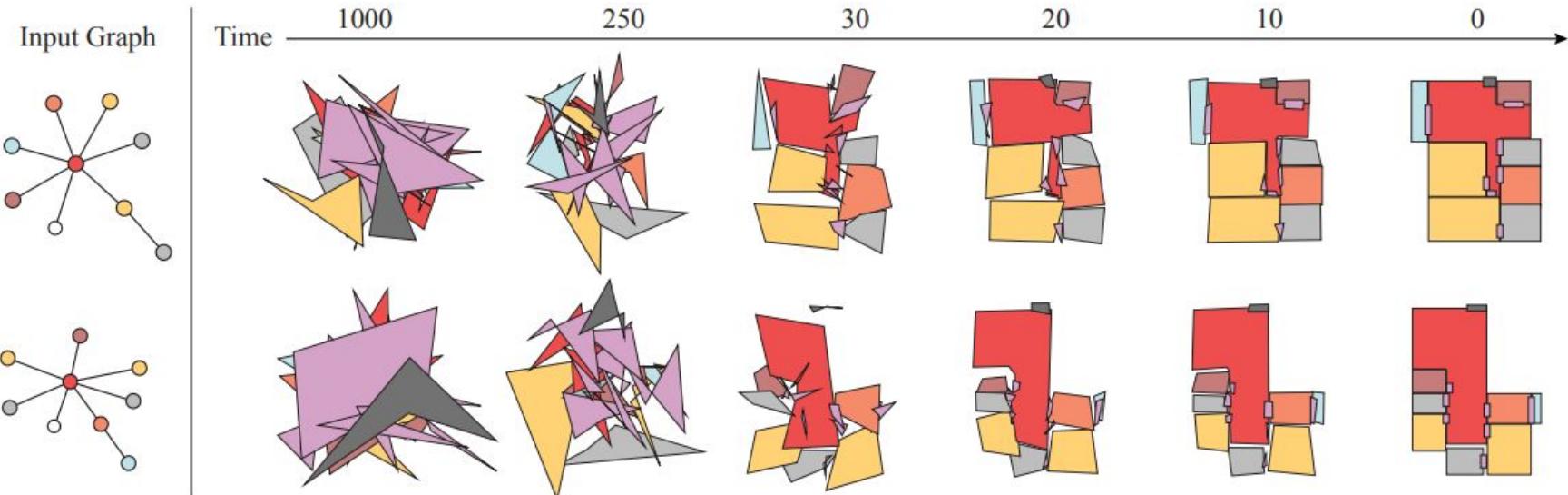
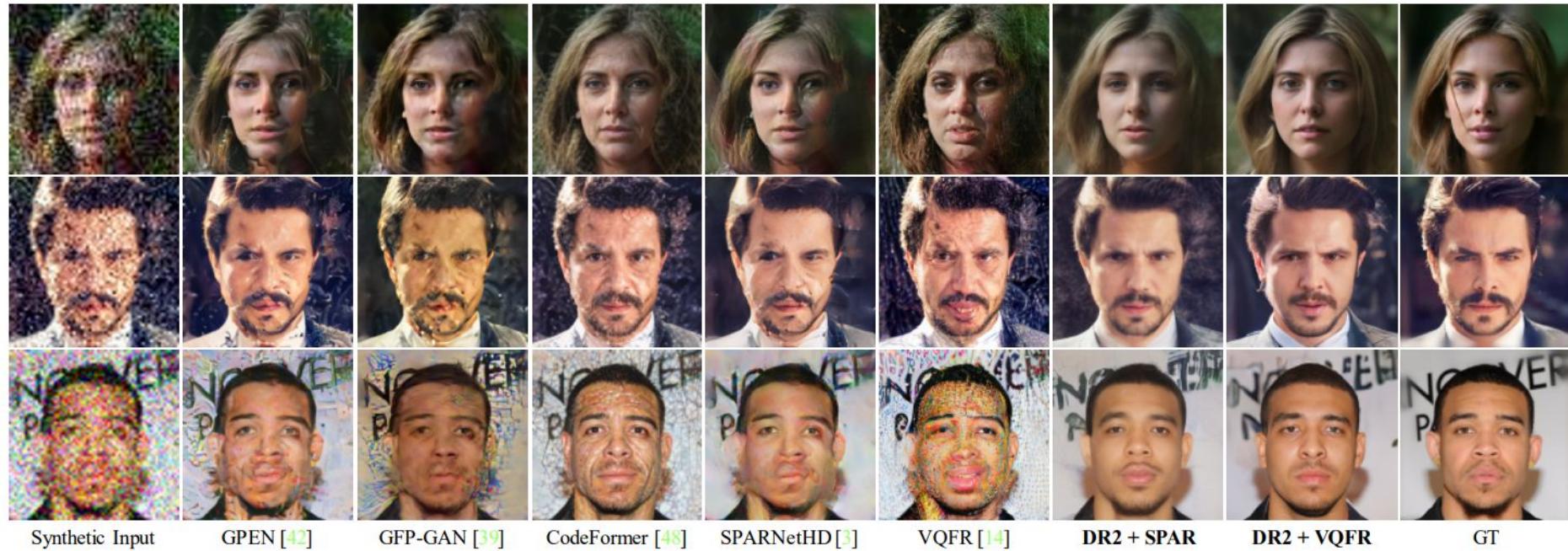


Figure 1. Given a bubble diagram as the input constraint, HouseDiffusion directly generates a vector floorplan by initializing the room/door coordinates with Gaussian noise and iteratively denoising them. Qualitative and quantitative evaluations demonstrate that HouseDiffusion significantly outperforms the current state-of-the-art with large margins.

<https://arxiv.org/pdf/2211.13287.pdf>

DR2: Diffusion-based Robust Degradation Remover for Blind Face Restoration (CVPR2023)



Video Probabilistic Diffusion Models in Projected Latent Space (CVPR2023)

<https://sihyun.me/PVDM/>

<https://arxiv.org/pdf/2302.07685.pdf>

Real-world Applications with Diffusion Models

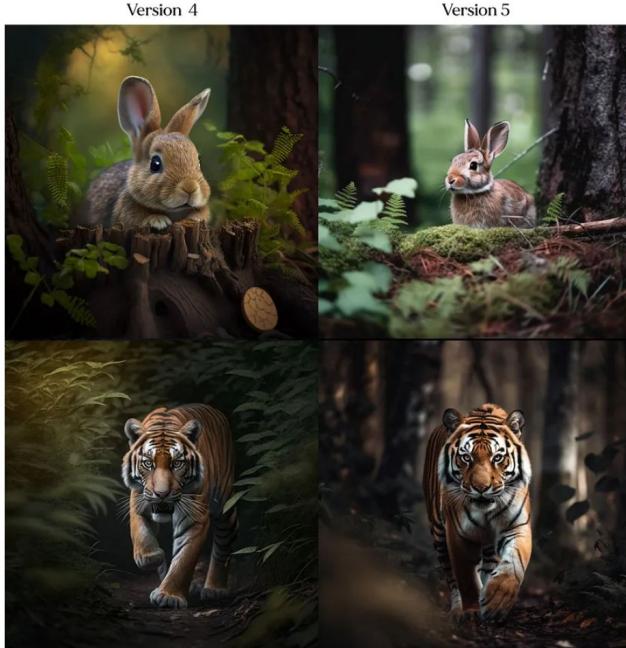
Most applications built upon Stable Diffusion (SD). Despite the most popular text-to-image generation, here are some interesting applications:

1. Image-to-image generation with text prompt (ControlNet)
2. Video generation
3. Image generation with consistency

In this section, I'll less focus on technical details but more on how the communities using diffusion models.

SD Family — Midjourney v5

Much more realistic generation



Wildlife photography created with Midjourney V4 and V5

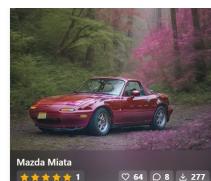
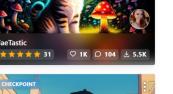
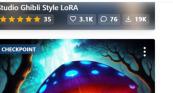
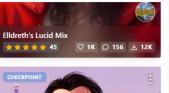
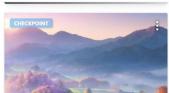
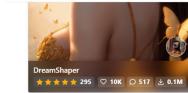
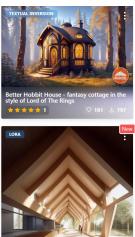
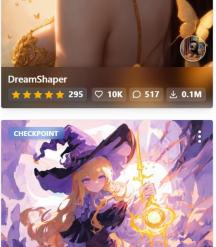
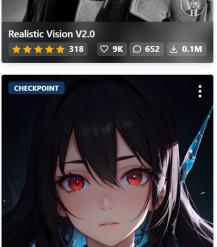
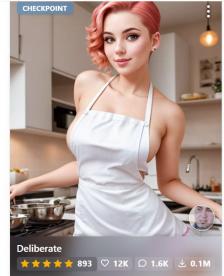


Photos of a dark purple Nissan taken at night and during golden hours, created with MidJourney



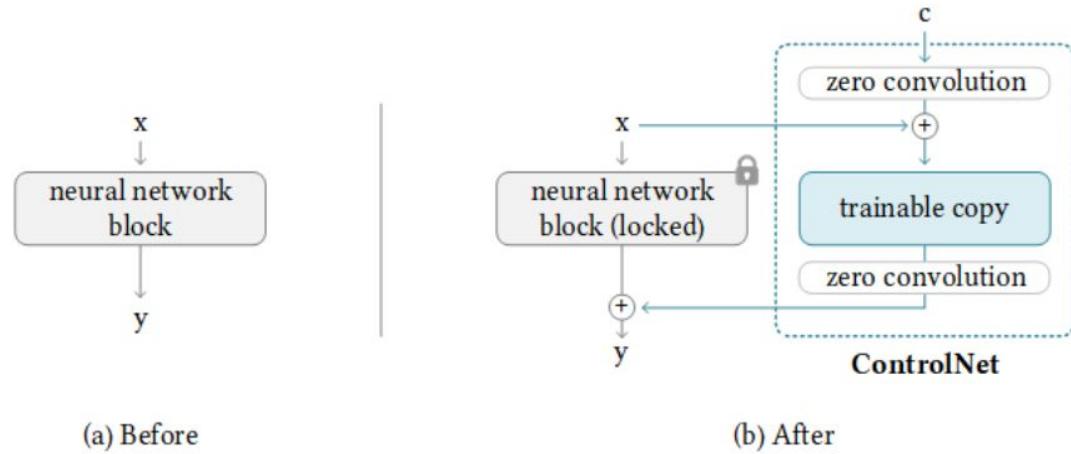
Underwater photography created with Midjourney V4 and V5

SD Family — Customer Tuned SD



SD Family — ControlNet

Github 15.5k starts



Very similar to GAN layer swap

The idea is to “lock” the pretrained knowledge and finetune a small scale of the network into user-defined tasks. The author claims the method can be run on personal devices.

<https://github.com/llyasviel/ControlNet>

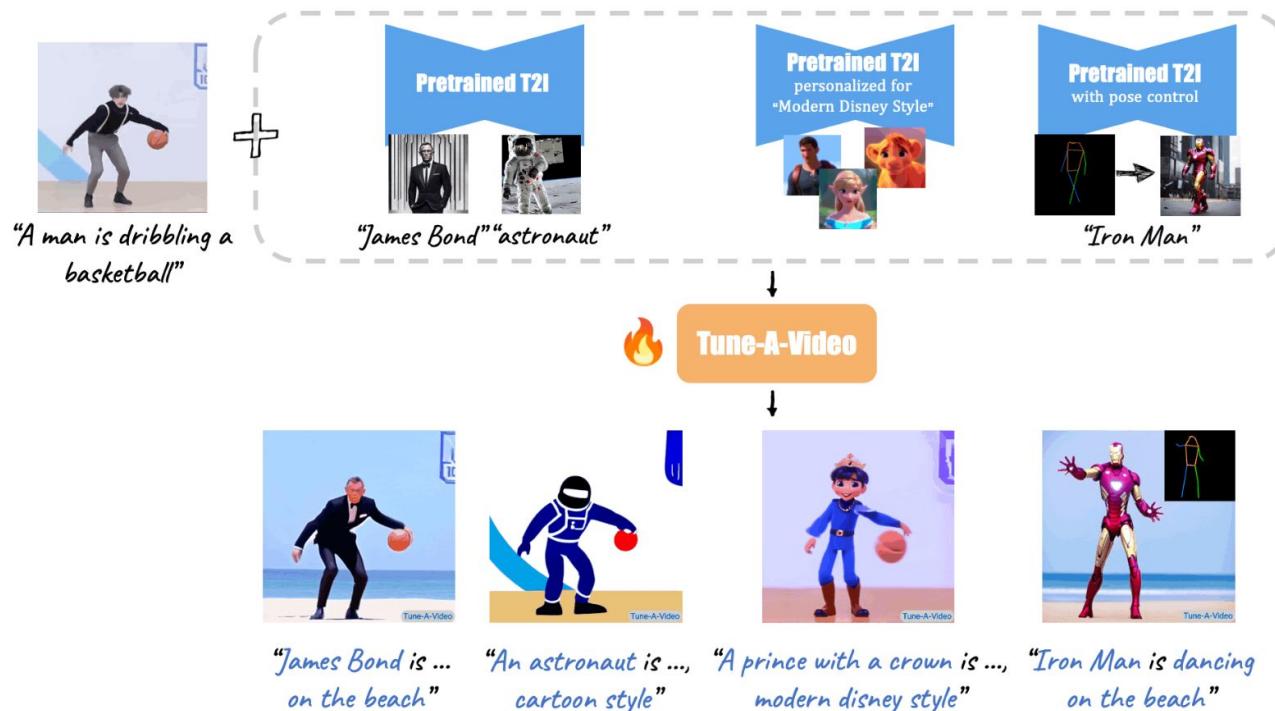
SD Family — Clipdrop

1. SR
2. Replace Background
3. Text Remover
4. Image-to-Image

The image displays a grid of nine AI-powered image editing tools from Clipdrop, each with a subgrid of three smaller images:

- Cleanup**: Remove objects, people, text and defects from your pictures automatically.
- Remove background**: Extract the main subject from a picture with incredible accuracy. It's like magic.
- Relight**: Relight your images with beautiful lights.
- Image upscaler**: Upscale your images by 2x or 4x in seconds. It can also remove noise and...
- Text to image**: Generate high-resolution realistic images with AI.
- Replace background**: Teleport anything, anywhere with AI.
- Text remover**: Remove text from any image.
- Stable diffusion...**: Create multiple variants of an image.

SD Family — Tune-A-Video



SD Family — D-ID

ChatGPT for text and Stable Diffusion for portrait.

I generated a video using this portrait with a description of Yongping Duan, the output is amazing.

SD as a small plugin.

<https://www.d-id.com/>



SD Family — Gen-2 (Video Generation)

<https://research.runwayml.com/gen2>



Input Image

A low angle shot of a man walking down a street, illuminated by the neon signs of the bars around him.

Driving Prompt



Output Video

SD Family — Image to Prompt

Featured Code Competition

Stable Diffusion - Image to Prompts

Deduce the prompts that generated our "highly detailed, sharp focus, illustration, 3d renders of majestic, epic" images

Kaggle · 785 teams · 2 months to go (a month to go until merger deadline)

\$50,000 Prize Money

Overview Data Code Discussion Leaderboard Rules Join Competition ...

Overview Data Code Discussion Leaderboard Rules Join Competition ...

Overview

Description	Goal of the Competition
Evaluation	The goal of this competition is to reverse the typical direction of a generative text-to-image model: instead of generating an image from a text prompt, can you create a model which can predict the text prompt given a generated image? You will make predictions on a dataset containing a wide variety of (prompt, image) pairs generated by Stable Diffusion 2.0, in order to understand how reversible the latent relationship is.
Timeline	
Prizes	
Code Requirements	<p>Context</p> <p>The popularity of text-to-image models has spurred an entire new field of prompt engineering. Part art and part unsettled science, ML practitioners and researchers are rapidly grappling with understanding the relationships between prompts and the images they generate. Is adding "4k" to a prompt the best way to make it more photographic? Do small perturbations in prompts lead to highly divergent images? How does the order of prompt keywords impact the resulting generated scene? This competition tasks you with creating a model that can reliably invert the diffusion process that generated to a given image.</p> <p>In order to calculate prompt similarity in a robust way—meaning that "epic cat" is scored as similar to "majestic kitten" in spite of character-level differences—you will submit embeddings of your predicted prompts. Whether you model the embeddings directly or first predict prompts and then convert to embeddings is up to you! Good luck, and may you create "highly quality, sharp focus, intricate, detailed, in the style of"</p>

SD Family — Image to Prompt

MJ released same feature



AI Generated Prompt



an aerial view of a city at night, a detailed matte painting by Feng Zhu, cgsociety, video art, playstation 5 screenshot, matte painting, reimagined by industrial light and magic

[Run Prompt on Stable Diffusion >](#)

[Try another image >](#)

<https://imagetoprompt.com/>

SD Family — Image to Prompt



AI Generated Prompt



a man wearing glasses and a red hoodie, a character portrait by Xia Shuwen, pexels contest winner, generative art, quantum wavetracing, 2d game art, physically based rendering

[Run Prompt on Stable Diffusion >](#)

[Try another image >](#)

SD Family — Multi Frame Rendering

Built only on SD. No finetune or extra model needed.



<https://xanthius.itch.io/multi-frame-rendering-for-stablediffusion>