Text to image generation with diffusion models

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panda mad scientist mixing sparkling chemicals, artstation

Outline

- Auto-regressive models
- CLIP
- Intro to DALL·E models
- Diffusion models (generative)
- Diffusion models (conditioned)
- Before DALL·E 2 there was GLIDE
- DALL·E 2
- Other diffusion-based models
- Other text to image models

vibrant portrait painting of Salvador Dalí with a robotic half face

Auto-regressive (or left-to-right) models

Auto-regressive DL models are trained to predict an element in a sequence based on its causal context:

$$
L_1(\mathcal{U}) = \sum_{i} \log P(u_i | u_{i-k}, \dots, u_{i-1}; \Theta)
$$

Given an unsupervised corpus of tokens $\mathcal{U} = \{u_1, \dots, u_n\}$

A token can be anything (word, image patches, speech, video frames, etc)

Modern versions of these models typically use a decoder-only transformer model with casual masking.

https://jalammar.github.io/illustrated-gpt2/

CLIP (Contrastive Language-Image Pre-Training)

- Pretraining maps images and text to the same embedding space.
- Trained of 400M crawled imagetext pairs.
- Text encoder is a regular decoder-only transformer.
- The image encoder has different variants: Resnet and Vision transformers (best).
- Zero-shot classification for "free"
- Powerful model for multimodal search, re-ranking, and more…
- Weights are open source!

Intro to DALL·E models

DALL·E "1" was introduced in 2021 by OpenAI, a transformer generates directly image tokens from both text and image tokens (more or less).

an armchair in the shape of an avocado....

DALL·E "2" was released in 2022, *it's more sophisticated*, and better at both quality and diversity.

"A teddybear on a skateboard in Times Square."

Intro to DALL·E models

DALL·E "1" was introduced in 2021 by OpenAI, it receives text and image tokens in a single sequence, and is trained to generate image tokens auto-regressively (12b transformer).

Denoised Diffusion Probabilistic Models

Given a data distribution $x_0 \sim q(x_0)$ we design a noising process which produces *latents* $x_1 ... x_T$

Few tricks:

- x_t can be sampled in closed-form using the reparameterization trick (conditioned on x_0).
- β_t defines a noise schedule, [1] uses a simple linear one.
- [2] found that learning $\Sigma_{\theta}(x_t, t)$ instead of fixing it to $\beta_t I$ results in a model that requires less step for sampling.

Forward: is defined such that x_T is a nearly isotropic Gaussian Distribution (cov matrix is $\sigma^2 I$): $q(x_t|x_{t-1}) \coloneqq \mathcal{N}(x_t; \sqrt{1-\beta_t}x_{t-1}, \beta_t \mathbf{I})$

Reverse: We can start from $x_T \sim \mathcal{N}(0, I)$, but $q(x_t|x_{t-1})$ depends on the entire distribution, so we approximate it with a model:

$$
p_{\theta}(x_{t-1}|x_t) \coloneqq \mathcal{N}(x_{t-1}; \mu_{\theta}(x_t, t), \Sigma_{\theta}(x_t, t))
$$

$$
\mu_{\theta}(x_t, t) = \frac{1}{\sqrt{\alpha_t}} \left(x_t - \frac{\beta_t}{\sqrt{1 - \bar{\alpha}_t}} \epsilon_{\theta}(x_t, t) \right)
$$
 Fixed to
\n
$$
\beta_t I \text{ in } [1]
$$

[1] trains the model θ to predict ε , $\alpha_t \coloneqq 1 - \beta_t$ instead of $\mu_{\theta}(x_t, t)$

 $[1]$

How do we learn this?

Denoised Diffusion Probabilistic Models

• **Problem**: how do we learn $p_{\theta}(x_{t-1} | x_t)$?

[\[2006.11239\] Denoising](https://arxiv.org/abs/2006.11239) Diffusion Probabilistic Models (arxiv.org)

[What are Diffusion Models?](https://www.youtube.com/watch?v=fbLgFrlTnGU&t=698s)

- Computing directly $p_{\theta}(x_0)$ requires to consider all possible forward-reversed trajectories - not feasible.
- Remember that $x_1 ... x_T$ are latent variables, similar to latent **Z** in VAE models
- What VAE optimizes: variational lower bound $\leq p_{\theta}(x_0)$

$$
\mathbb{E}\left[-\log p_{\theta}(\mathbf{x}_{0})\right] \leq \mathbb{E}_{q}\left[-\log p(\mathbf{x}_{T}) - \sum_{t\geq 1} \log \frac{p_{\theta}(\mathbf{x}_{t-1}|\mathbf{x}_{t})}{q(\mathbf{x}_{t}|\mathbf{x}_{t-1})}\right] =: L \text{ \quad \ \ \, \text{cones from KL} \atop \text{ \quad \ \ \, \text{cones from KL} \atop \text{ \quad \ \ \, \text{and } \, \, \text{p}}}
$$
\n
$$
\text{This can be further re-arranged to \atop \text{reduce variance during training} \atop \text{the variance of training training} \atop \text{the probability sample of pairs of} \atop \
$$

Comparison to other generative models

What [are Diffusion Models? | Lil'Log](https://lilianweng.github.io/posts/2021-07-11-diffusion-models/) (lilianweng.github.io)

Classifier Guidance – adding text conditioning

- [1] used an auxiliary image classifier p_{ϕ} (trained on noised ImageNet) to guide the generative process using its gradients.
- [1] beats BigGAN on FiD scores, with more 'diverse' samples. This is also done by tweaking the UNet architecture.

Algorithm 1 Classifier guided diffusion sampling, given a diffusion model $(\mu_{\theta}(x_t), \Sigma_{\theta}(x_t))$, classifier $p_{\phi}(y|x_t)$, and gradient scale s.

[\[1\] \[2105.05233\] Diffusion Models Beat GANs on Image Synthesis \(arxiv.org\)](https://arxiv.org/abs/2105.05233)

Classifier Guidance - adding text conditioning

- Increasing the classifier gradients hits a trade-off between fidelity and diversity.
- The 'optimal' value for FiD and sFID scores is in between.

A note in eval metrics:

- FID [3] compares the distribution of generated images (given by the Inception model) and the images in the training set.
- Precision and Recall here refer to [2]:

Figure 1: Definition of precision and recall for distributions [25]. (a) Denote the distribution of real images with P_r (blue) and the distribution of generated images with P_q (red). (b) Precision is the probability that a random image from P_q falls within the support of P_r . (c) Recall is the probability that a random image from P_r falls within the support of P_q .

[\[1\] \[2105.05233\] Diffusion Models Beat GANs on Image Synthesis \(arxiv.org\)](https://arxiv.org/abs/2105.05233) [2] [\[1904.06991\] Improved Precision and Recall Metric for Assessing Generative Models \(arxiv.org\)](https://arxiv.org/abs/1904.06991) [\[3\] \[1706.08500\] GANs Trained by a Two Time-Scale Update Rule Converge to a Local Nash Equilibrium \(arxiv.org\)](https://arxiv.org/abs/1706.08500)

Classifier Guidance – CLIP (image and text aligned embeddings)

- CLIP naturally resembles a 'zero-shot' classifier.
- Although in theory guidance requires a classifier trained in noised data, [2] used CLIP public models with some level of success.
- [2] and [3] are important since were projects developed "on the open", arguably influenced Stable Diffusion.

"A painting of an apple"

"a futuristic city in synthwave style" VQGAN-CLIP

We will talk about classifier-free guidance next

[2] CLIP Guided [Diffusion HQ 256x256.ipynb -](https://colab.research.google.com/drive/12a_Wrfi2_gwwAuN3VvMTwVMz9TfqctNj) Colaboratory (google.com)

[3] [\[2204.08583v1\] VQGAN-CLIP: Open Domain Image Generation and Editing with Natural Language Guidance \(arxiv.org\)](https://arxiv.org/abs/2204.08583v1)

[4] GitHub - [nerdyrodent/CLIP-Guided-Diffusion: Just playing with getting CLIP Guided Diffusion running locally, rather than having to use colab.](https://github.com/nerdyrodent/CLIP-Guided-Diffusion)

"Rice farming by Hokusai Gogh"

Classifier-free guidance (still text-conditioned)

- Classifier guidance was a great fix to be able to trade-off diversity by fidelity.
- Authors argue classifier guidance boosts FiD scores more-and-less artificially.
- [1] proposes a simple trick to still control diversity/fidelity without another model:
	- Used paired data (x, y) during training, typically image captions
	- Train an **unconditional** diffusion model $p_{\theta}(x)$, and a **conditional** diffusion model $p_{\theta}(x|y)$
	- Use a single network to represent $p_{\theta}(x)$ and $p_{\theta}(x|y)$. Note that p_{θ} is simply the same diffusion model.
	- Practically speaking, $p_{\theta}(x|y)$ is trained and periodically y is simply discarded (set to zeroes)
	- The parametrized outputs (ϵ_{θ}) of $p_{\theta}(x)$ and $p_{\theta}(x|y)$ are weighted:

$$
\tilde{\boldsymbol{\epsilon}}_{\theta}(\mathbf{z}_{\lambda}, \mathbf{c}) = (1+w)\boldsymbol{\epsilon}_{\theta}(\mathbf{z}_{\lambda}, \mathbf{c}) - w\boldsymbol{\epsilon}_{\theta}(\mathbf{z}_{\lambda})
$$
\nIncrease guidance, more fidelity

\nMore diversity

[\[1\] \[2207.12598\] Classifier-Free Diffusion Guidance](https://arxiv.org/abs/2207.12598) (arxiv.org)

GLIDE – a model before DALLE-2 *with diffusion*

- Rivals DALLE "1": 3b parameters vs DALLE's 12b.
- GLIDE reports that classifier-free is preferred over CLIP-guidance by *human evaluators*.
- The condition on text is done via attention and token embeddings from a text-transformer.
- GLIDE's samples are preferred over DALLE's by *human evaluators* (89% in photorealism, and 69% in caption similarity).
- Training details:
	- 3.5 billion parameter text-conditional diffusion model, at 64x64 resolution
		- 2.3b for the visual part
		- 1.5b for a transformer encoding the txt
	- 1.5 billion parameter up-sampling diffusion model, at 256x256 resolution
	- Same dataset as DALLE-1, and roughly the same compute
	- Extra fine-tuning for *unconditional* image generation, and for *inpainting (*random masks added in a 4th channel)

[\[1\] \[2112.10741\] GLIDE: Towards Photorealistic Image Generation and Editing with Text-Guided Diffusion Models \(arxiv.org\)](https://arxiv.org/abs/2112.10741)

GLIDE – a model before DALLE-2

Conditional generation:

"a small kitchen with a low ceiling"

Inpainting: iteratively add a mask to the model. The first image is generated from the prompt alone.

"a cozy living room"

"a painting of a corgi on the wall above

"a round coffee table in front of a couch"

"a vase of flowers on a coffee table"

"a couch in the corner of a room"

[\[1\] \[2112.10741\] GLIDE: Towards Photorealistic Image Generation and Editing with Text-Guided Diffusion Models \(arxiv.org\)](https://arxiv.org/abs/2112.10741)

Summary so far – before DALLE-2

- Diffusion models are generative models that generate noised latents, and then are denoised iteratively.
- Conditional generation can be enabled with classifiers (CLIP, etc)…
- ... However, classifier-free guidance is a established trick to do so without extra classifiers.
- DALLE-1 uses a discrete VAE and a transformer to generate image tokens.
- Diffusion-based conditional models beat GANs
- GLIDE, a diffusion-based model with the 'tricks' above, beats GANs and DALLE-1. However, it's resource-intensive at inference.

DALL·E 2 (finally) - Basics

- **CLIP**: a model that maps text and images to the same embedding space
- **Auto-regressive model (AR)**: A sequence model that generate tokens causally.
- **Diffusion model (DM)**: a generative model which uses iterative denoising to learn a data distribution (optionally conditioned)
- **Upsampling:** cascaded diffusion models can increase resolution, typically using U-net: 256×256

[\[2106.15282\] Cascaded Diffusion Models for High Fidelity Image Generation \(arxiv.org\)](https://arxiv.org/abs/2106.15282)

DALL·E 2 - unCLIP

embedding

unCLIP is a two stages model. The goal is to "invert" the CLIP text embeddings.

embedding.

CLIP's training objective only forces the embeddings to match an image to a caption, but it does not necessarily captures image features describable by text, relative positions, etc.

By training the prior now the produced embeddings should capture salient features of the image, rather than just its relationship to the caption.

The decoder has 2 upsamplers to go from 64x64->256x256->1024x1024

DALL·E 2 - unCLIP

Given image-caption pairs (x, y) :

- z_i = CLIP image embeddings
- Z_t = CLIP text embeddings
- **Prior**: $P(z_i, y)$ produces z_i , conditioned on captions *.* • z_i = CLIP image embeddings

• z_i = CLIP text embeddings

• **Prior**: $P(z_i, y)$ produces z_i , conditioned on

captions y.
 • Decoder: $P(x|z_i, y)$ produces an image x
 • Decoder: $P(x|z_i, y)$ produces an image x

the
	- **Decoder**: $P(x|z_i, y)$ produces an image x conditioned on z_i and (optionally) y .

Note that even though not explicitly mentioned here, the prior can be conditioned on z_t because z_t is a *deterministic function of .*

Model: $P(x|y) = P(x, z_i | y) = P(x|z_i, y)P(z_i, y)$

DALL \cdot E 2 – the prior

Two models were tried in the paper for the Prior (only one is used):

- **Auto-regressive (AR)**: The embeddings are discretized and generated left-ot-right conditioned on the captions.
- **Diffusion (DM)**: the embedding z_i is directly produced by a diffusion model conditioned on y .

The DM prior outperforms AR in human eval, FID (MS-COCO), and "aesthetic".

AR: dot product is added during training and in inferencing a fixed value is hardcoded in the input.

DM: instead of adding dot product, two z_i are sampled and the best is chosen.

Why do we need a prior? CLIP's objective is not enough to align the embeddings for image generation, the prior aligns them correctly for this.

 $decoder(y)$ no prior

 $decoder(y, z_t)$ no prior

 $decoder(y, \hat{z_i})$ with prior

[&]quot;an oil painting of a corgi wearing a party hat"

DALL·E 2 – the decoder

The decoder is very similar to GLIDE, except that it includes the CLIP embeddings \widehat{Z}_i generated by the prior.

Classifier-free guidance:

- The CLIP embeddings are dropped 10% of the time.
- The text caption is dropped 50% of the time.

The projected image embedding is added to the diffusion step embedding too.

decoder

 $\widehat{z_i}W_2$

[\[2204.06125\] Hierarchical Text-Conditional Image Generation with CLIP Latents \(arxiv.org\)](https://arxiv.org/abs/2204.06125)

Image manipulations – Inverse DDIM

- $x =$ any image
- z_i = CLIP_image_encoder(x)
- X_T : latent given X, computed as:

We can now generate images conditioned on z_i and also manipulate the latent x_T (next).

The unclip decoder inverse DDIM

[\[1\] \[2204.06125\] Hierarchical Text-Conditional Image Generation with CLIP Latents \(arxiv.org\)](https://arxiv.org/abs/2204.06125)

[\[2\] \[2105.05233\] Diffusion Models Beat GANs on Image Synthesis \(arxiv.org\)](https://arxiv.org/abs/2105.05233)

Image manipulations – Inverse DDIM

Variation: fix z_i , change noise in x_T :

Interpolation: Interpolate between $z_i \rightarrow z'_i$ and $x_T \rightarrow x'_T$:

a photo of a landscape in winter \rightarrow a photo of a landscape in fall

How good is DALLE-2? - Metrics

% of human evaluators that prefer unCLIP over GLIDE

Table 2: Comparison of FID on MS-COCO 256×256 . We use guidance scale 1.25 for the decoder for both the AR and diffusion prior, and achieve the best results using the diffusion prior.

Limitations

"A corgi with a green bow tie and a red party hat"

- DALLE-2 has a hard time with variable binding.
- CLIP's inductive bias doesn't bind linguistic properties from image to text, and neither does unCLIP.
- The authors hoped that conditioning on encoded text in diffusion would help, but it didn't

"a sign that says deep learning" . It's possible the text tokenization procedure makes it very difficult for clip.

High-level details are hard, likely due to the resolution up-sampling.

(b) A high quality photo of Times Square.

Risks, datasets, and copyright

- As usual with large-pretrained models, the datasets used are very large and with a lot of problematic content (hate-speech, stereotypes, etc), usually with poor/none filtering. These models can be easily/cheaply weaponized. Images have higher impact than text?
- Available models (DALLE-2, SD) "patch" this with a binary classifier to prevent misuse, it's effectiveness is debatable.
- There's a growing concern on artists about having their work used on these massive models without consent -> legal loophole. Some US rulings argue scrapped datasets are fair-use [2].

Very important problems, **no effective solutions thus far**, *models are in the open now*… [1] [Multimodal datasets: misogyny, pornography, and malignant](https://arxiv.org/abs/2110.01963) stereotypes [2] https://en.wikipedia.org/wiki/Authors Guild, Inc. v. Google, Inc.

Stable Latent Diffusion (SD)

- Trained on "LAION Aesthetics"
- Re-introduces latent spaces, similar to the codebook of DALLE-1, which used a dVAE.
- The encoder ϵ has a strong image inductive bias (~VQGAN).
- The latent now can exploit these inductive biases to do diffusion more effectively, instead of "naïve" AR (DALLE-1).
- Models are $0.8b 1.45B$ parameters, noticeably smaller than DALLE (3.5b and 12b).
- **Model weights and code are opensource.**

1) Diffusion is done **in a compressed** space (down sampling factor of 4), not pixel space. This is why SD is noticeably faster.

4 in the paper, 8 in the widely

released "stable diffusion"

2) Cross-attention coming from the output of a domain-specific encoder τ_A and the UNet backbone.

Stable Latent Diffusion (SD)

Conditioning in Latent spaces:

The Unet architecture. Cross-attention is applied on intermediate flattened representations.

SOTA on Diffusion Models – Google's Imagen

Dramatically simpler architecture, it simply uses a very big text encoder followed by diffusion for superresolution, for a total of 14b parameters.

> Beats GLIDE and DALL-E 2 in human evaluation benchmarks and FiD.

SOTA on Diffusion Models – Google's Imagen

A panda making latte art.

Parti – scratch Diffusion Models

The bitter lesson strikes again: scratch diffusion models and use a simpler, and scaled-up architecture (20B model): . Stage 1: tokenize images using ViT-

- VQGAN (30M parameters).
- Stage 2: Train an encoder-decoder model to generate image tokens given text tokens.
- Stage 3: Scale up the image detokenizer to 600M parameters and higher resolution.
- Add a super-resolution model on top to reach 1024x1024.
- Also uses classifier-free guidance.

Pre-Trained with BERT's loss and contrastive image-text data

Parti – scratch diffusion models

Takeaway: scale matters! But diffusion is *probably* a good generic compression technique, time will tell (as with the transformer… or with GANs)

"a propaganda poster depicting a cat dressed as french emperor napoleon holding a piece of cheese"

Can you tell which one is Parti and which one is DALLE-2? ☺

Summary

- Diffusion models are a powerful compression/denoising technique
- Small diffusion models seem to be very good, but scaling up still pays off.
- Classifier-free guidance > CLIP-guidance
- Aligning text and image embedding spaces is probably easier than what we thought before.
- No one has solved variable binding, counting, negation, spatial relations, grounding, etc… Language is still unsolved \odot

A red cube **on top** of a blue cube

a plate that has **no bananas** on it. there is a glass **without** orange juice next to it

two baseballs to the left of **three** tennis balls

rhino beetle this size of a tank grapples a real life passenger airplane on the tarmac

