

# Diffusion Autoencoders: Toward a Meaningful and Decodable Representation

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Figure 1. Attribute manipulation and interpolation on real images. Diffusion autoencoders can encode any image into a two-part latent code that captures both semantics and stochastic variations and allows near-exact reconstruction. This latent code can be interpolated or modified by a simple linear operation and decoded back to a highly realistic output for various downstream tasks.

## Abstract

Diffusion probabilistic models (DPMs) have achieved remarkable quality in image generation that rivals GANs'. But unlike GANs, DPMs use a set of latent variables that lack semantic meaning and cannot serve as a useful representation for other tasks. This paper explores the possibility of using DPMs for representation learning and seeks to extract a meaningful and decodable representation of an input image via autoencoding. Our key idea is to use a learnable encoder for discovering the high-level semantics, and a DPM as the decoder for modeling the remaining stochastic variations. Our method can encode any image into a two-part latent code where the first part is semantically meaningful and linear, and the second part captures stochastic details, allowing near-exact reconstruction. This capability enables challenging applications that currently foil GAN-based methods, such as attribute manipulation on real images. We also show that this two-level encoding improves denoising efficiency and naturally facilitates various downstream tasks including few-shot conditional sampling. Please visit our page <https://Diff-AE.github.io/>

## 1. Introduction

Diffusion-based (DPMs) [22, 46] and score-based [49] generative models have recently succeeded in synthesizing realistic and high-resolution images, rivaling those from GANs [11, 15, 23]. These two models are closely related

and, in practice, optimize similar objectives. Numerous applications have emerged notably in the image domain, such as image manipulation, translation, super-resolution [8, 32, 35, 43], in speech and text domains [5, 6], or 3D point cloud [34]. Recent studies have improved DPMs further in both theory and practice [25, 29, 31]. In this paper, however, we question whether DPMs can serve as a good representation learner. Specifically, we seek to extract a meaningful and decodable representation of an image that contains an input image via autoencoding. Our exploration focuses on diffusion models, but the contributions are applicable also to score-based models.

One way to learn a representation is through an autoencoder. There exists a certain kind of DPM [47] that can act as an encoder-decoder that converts any input image into a spatial latent variable  $z_T$  by running the generative process backward. However, the resulting latent variable lacks high-level semantics and other desirable properties, such as disentanglement, compactness, or the ability to perform meaningful linear interpolation in the latent space. Alternatively, one can use a trained GAN for extracting a representation using the so-called GAN inversion [28, 58], which optimizes for a latent code that reproduces the given input. Even though the resulting code carries rich semantics, this technique struggles to faithfully reconstruct the input image. To overcome these challenges, we propose a diffusion-based autoencoder that leverages the powerful DPMs for decodable representation learning.

Finding a meaningful representation that is decodable re-using a UNet [40]. The model is trained with a loss function that captures both the high-level semantics and low-level stochastic variations. Our key idea is to learn both levels of representation by utilizing a learnable encoder for discovering high-level semantics and utilizing a DPM for decoding and modeling stochastic variations. In particular, we use our conditional variant of the Denoising Diffusion Implicit Model (DDIM) [47] as the decoder and separate the latent code into two subcodes. The first “semantic” subcode is compact and inferred with a CNN encoder, whereas the second “stochastic” subcode is inferred by reversing the generative process of our DDIM variant conditioned on the semantic subcode. In contrast to other DPMs, DDIM modifies the forward process to be non-Markovian while preserving the training objectives of DPMs. This modification allows deterministically encoding an image to its corresponding initial noise, which represents our stochastic subcode.

The implication of this framework is two-fold. First, by conditioning DDIM on the semantic information of the target output, denoising becomes easier and faster. Second, this design produces a representation that is linearly semantically meaningful, and decodable—a novel property for DPMs’ latent variables. This crucial property allows harnessing DPMs for many tasks including those that are highly challenging for any GAN-based methods, such as interpolation and attribute manipulation on real images. Unlike GANs, which rely on error-prone inversion before operating on real images, our method requires no optimization to encode the input and produces high-quality output with original details preserved.

Despite being an autoencoder, which is generally not designed for unconditional generation, our framework can be used to generate image samples by conditioning another DPM to the semantic subcode distribution. This combination achieves competitive FID scores on unconditional generation compared to a vanilla DPM. Moreover, the ability to sample from our compact and meaningful latent space also enables few-shot conditional generation (i.e., generate images with similar semantics to those of a few examples). Compared to other DPM-based techniques for the few-shot setup, our method produces convincing results with only a handful labeled examples without additional contrastive learning used in prior work [45].

## 2. Background

Diffusion-based (DPMs) and score-based generative models belong to a family of generative models that model the target distribution by learning a denoising process of varying noise levels. A successful process can denoise an arbitrary Gaussian noise map from the prior  $N(0; I)$  to a clean image sample after successive denoising passes. Ho et al. [22] proposed to learn a function  $q(x_t; t)$  that takes a noisy image  $x_t$  and predicts its noise

More formally, we define a Gaussian diffusion process at time  $t$  (out of  $T$ ) that increasingly adds noise to an input image  $x_0$  as  $q(x_t|x_{t-1}) = N(\sqrt{1-\beta_t}x_{t-1}; \beta_t I)$ , where  $\beta_t$  are hyperparameters representing the noise levels. With the Gaussian diffusion, the noisy version of an image at time  $t$  is another Gaussian  $q(x_t|x_0) = N(\sqrt{\alpha_t}x_0; (1-\alpha_t)I)$  where  $\alpha_t = \prod_{s=1}^t (1-\beta_s)$ . We are interested in learning the reverse process of this, i.e., the distribution  $p(x_{t-1}|x_t)$ . This probability function is likely a complex one unless the gap between  $t-1$  and  $t$  is infinitesimally small ( $T-t \rightarrow 1$ ) [46]. In such a case,  $p(x_{t-1}|x_t)$  can be modeled as  $N(x_t; t; \beta_t)$  [22]. There are many ways to model this distribution, one of which is via  $(x_t; t)$  mentioned earlier. In practice, the assumption  $\beta_t = 1$  is never satisfied; hence, DPMs are only approximations.

As latent-variable models, DPMs can naturally yield the latent variables  $x_{1:T}$  through its forward process; however, these variables are stochastic and only representing a sequence of image degradation by Gaussian noise, which does not contain much semantics. Song et al. [47] proposed another kind of DPM called Denoising Diffusion Implicit Model (DDIM) that enjoys the following deterministic generative process:

$$x_{t-1} = \sqrt{\frac{\alpha_{t-1}}{\alpha_t}} x_t + \sqrt{1 - \frac{\alpha_{t-1}}{\alpha_t}} \epsilon_t \quad (1)$$

and the following novel inference distribution:

$$q(x_{t-1}|x_t; x_0) = N\left(\sqrt{\frac{\alpha_{t-1}}{\alpha_t}} x_t + \sqrt{1 - \frac{\alpha_{t-1}}{\alpha_t}} \sqrt{\frac{\alpha_t}{1-\alpha_t}} \epsilon_0; 0\right) \quad (2)$$

while maintaining the original DDPM marginal distribution  $q(x_t|x_0) = N(\sqrt{\alpha_t}x_0; (1-\alpha_t)I)$ . By doing so, DDIM shares both the objective and solution with DDPM and only differs in how samples are generated.

With DDIM, it is possible to run the generative process backward deterministically to obtain the noise map which represents the latent variable or encoding of a given image  $x_0$ . In this context, DDIM can be thought of as an image decoder that decodes the latent code back to the

input image. This process can yield a very accurate reconstruction; however,  $x_T$  still does not contain high-level semantics as would be expected from a meaningful representation. We show in Figure 4c that the interpolation between two latent variables  $x_T$ ’s does not correspond to a semantically-smooth change in the resulting images. The images only share the overall composition and background colors but do not resemble the identity of either person. This

### 3.1. Diffusion-based Decoder

Our conditional DDIM decoder receives as input  $(z_{\text{sem}}; x_T)$  to produce the output image. This decoder is a conditional DDIM that models  $p(x_{t-1}|x_t; z_{\text{sem}})$  to match the inference distribution  $q(x_{t-1}|x_t; x_0)$  defined in Equation 2, with the following reverse (generative) process:

$$p(x_{0:T} | z_{\text{sem}}) = p(x_T) \prod_{t=1}^T p(x_{t-1} | x_t; z_{\text{sem}}) \quad (3)$$

$$p(x_{t-1} | x_t; z_{\text{sem}}) = \begin{cases} N(f(x_1; 1; z_{\text{sem}}); 0) & \text{if } t = 1 \\ q(x_{t-1}|x_t; f(x_t; t; z_{\text{sem}})) & \text{otherwise} \end{cases} \quad (4)$$

Figure 2. Overview of our diffusion autoencoder. The autoencoder consists of a “semantic” encoder that maps the input image  $x_0$  to the semantic subcode  $z_{\text{sem}}$ , and a conditional DDIM that acts both as a “stochastic” encoder  $((z_{\text{sem}}; x_T) \rightarrow x_0)$  and a decoder  $((z_{\text{sem}}; x_T) \rightarrow x_0)$ . Here,  $z_{\text{sem}}$  captures the high-level semantics while  $x_T$  captures low-level stochastic variations, and together they can be decoded back to the original image with high fidelity. To sample from this autoencoder, we first sample a latent DDIM to the distribution of  $z_{\text{sem}}$  and sample  $x_T \sim N(0; I)$  for decoding.

Following Song et al. [47], we parameterize in Equation 4 as a noise prediction network  $f(x_t; t; z_{\text{sem}})$ :

$$f(x_t; t; z_{\text{sem}}) = \frac{1}{\alpha_t} x_t - \frac{\beta_t}{1 - \alpha_t} (x_t; t; z_{\text{sem}}) \quad (5)$$

This network is a modified version of the UNet of a recent DPM from Dhariwal et al. [11]. Training is done by optimizing  $L_{\text{simple}}$  [22] loss function with respect to  $\theta$ .

$$L_{\text{simple}} = \sum_{t=1}^T \mathbb{E}_{x_0; t} \left[ \frac{1}{2} \| \hat{x}_t - x_t \|^2 \right] \quad (6)$$

where  $t \in \mathbb{R}^3 \times \mathbb{R}^h \times \mathbb{R}^w \sim N(0; I)$ ,  $x_t = \frac{\beta_t}{1 - \alpha_t} x_0 + \frac{1 - \beta_t}{1 - \alpha_t} \epsilon_t$ , and  $T$  is set to some large number, e.g., 1,000. Note that this simplified loss function has been shown to optimize both DDPM [22] and DDIM [47], though not the actual variational lower bound. For training, the stochastic subcode  $x_T$  is not needed. We condition the UNet using adaptive group normalization layers (AdaGN), following Dhariwal et al. [11], which extend group normalization [56] by applying channel-wise scaling and shifting on the normalized feature maps  $\mathbf{h} \in \mathbb{R}^{c \times h \times w}$ . Our AdaGN is conditioned on  $t$  and  $z_{\text{sem}}$ :

$$\text{AdaGN}(\mathbf{h}; t; z_{\text{sem}}) = z_s(t_s \text{GroupNorm}(\mathbf{h}) + t_b) \quad (7)$$

where  $z_s \in \mathbb{R}^c = \text{Affine}(z_{\text{sem}})$  and  $(t_s; t_b) \in \mathbb{R}^{2 \times c} = \text{MLP}(\epsilon_t)$  is the output of a multilayer perceptron with a sinusoidal encoding function. These layers are used throughout the UNet. Please see details in Appendix A.

### 3.2. Semantic encoder

The goal of the semantic encoder  $\text{Enc}$  is to summarize an input image into a descriptive vector  $z_{\text{sem}} = \text{Enc}(x_0)$  with necessary information to help the decoder  $p(x_{t-1}|x_t; z_{\text{sem}})$  denoise and predict the output image. We do not assume any particular architecture for this encoder; however, in our experiments, this encoder shares the same architecture as the first half of our UNet decoder. One benefit that allows smooth interpolation, similar to those learned by GANs, while keeping the reconstruction capability that diffusion models excel.

## 3. Diffusion autoencoders

In the pursuit of a meaningful latent code, we design a conditional DDIM image decoder  $p(x_{t-1}|x_t; z_{\text{sem}})$  that is conditioned on an additional latent variable  $z_{\text{sem}}$  and a semantic encoder  $z_{\text{sem}} = \text{Enc}(x_0)$  that learns to map an input image  $x_0$  to a semantically meaningful  $z_{\text{sem}}$ . Here, the conditional DDIM decoder takes as input a latent variable  $z = (z_{\text{sem}}; x_T)$ , which consists of the high-level “semantic” subcode  $z_{\text{sem}}$  and a low-level “stochastic” subcode  $x_T$ , inferred by reversing the generative process of DDIM. In this framework, DDIM acts as both the decoder and the stochastic encoder. The overview is shown in Figure 2.

Unlike in other conditional DPMs [23, 32, 45] that use spatial conditional variables (e.g., 2D latent maps),  $z_{\text{sem}}$  is a non-spatial vector of dimension  $d = 512$ , which resembles the style vector in StyleGAN [27, 28] and allows us to encode global semantics not specific to any spatial regions. One of our goals is to learn a semantically rich latent space architecture as the first half of our UNet decoder. One benefit that allows smooth interpolation, similar to those learned by GANs, while keeping the reconstruction capability that diffusion models excel.

### 3.3. Stochastic encoder

Besides decoding, our conditional DDIM can also be used to encode an input image to the stochastic subcode  $x_T$  by running its deterministic generative process backward (the reverse of Equation 1):

$$x_{t+1} = p_{t+1} f(x_t; t; z_{\text{sem}}) + p_{1-t+1} (x_t; t; z_{\text{sem}}) \quad (8)$$

We can think of this process as a stochastic encoder because  $x_T$  is encouraged to encode only the information left out by  $z_{\text{sem}}$ , which has a limited capacity for compressing stochastic details. By utilizing both semantic and stochastic encoders, our autoencoder can capture an input image to the very last detail while also providing a high-level representation  $z_{\text{sem}}$  for downstream tasks. Note that the stochastic encoder is not used during training (Equation 6) and is used to compute  $x_T$  for tasks that require exact reconstruction or inversion, such as real-image manipulation.

### 4. Sampling with diffusion autoencoders

By conditioning the decoder on  $z_{\text{sem}}$ , diffusion autoencoders are no longer generative models. So, to sample from our autoencoder, we need an additional mechanism to sample  $z_{\text{sem}} \in \mathbb{R}^d$  from the latent distribution. While VAE is an appealing choice for this task, balancing between retaining rich information in the latent code and maintaining the sampling quality in VAE is hard [41, 42, 45, 52]. GAN is another choice, though it complicates training stability, which is one main strength of DPMs. Here, we choose to train another DDIM, called latent DDIM  $p_t(z_{\text{sem}}; t)$ , to the latent distribution of  $z_{\text{sem}} = \text{Enc}(x_0); x_0 \sim p(x_0)$ . Analogous to Equation 5 and 6, training the latent DDIM is done by optimizing  $\mathcal{L}_{\text{latent}}$  with respect to :

$$\mathcal{L}_{\text{latent}} = \sum_{t=1}^T \mathbb{E}_{z_{\text{sem}} \sim p_t(z_{\text{sem}}; t)} \left[ \frac{1}{2} \|x_t - \text{Dec}(z_{\text{sem}}; x_T)\|^2 \right] \quad (9)$$

where  $t \in [1, T]$ ,  $z_{\text{sem}} \sim p(z_{\text{sem}})$ , and  $T$  is the same as in the DDIM image decoder. For  $\mathcal{L}_{\text{latent}}$ , we empirically found that  $\ell_1$  works better than  $\ell_2$  loss. Unlike for 1D/2D images, there is no well-established DPM architecture for non-spatial data, but we have found that deep MLPs (10-20 layers) with skip connections perform reasonably well. The details are provided in Appendix A.1.

We first train the semantic encoder  $\chi$  and the image decoder  $\phi$  via Equation 6 until convergence. Then, we train the latent DDIM  $\psi$  via Equation 9 with the semantic encoder fixed. In practice, the latent distribution modeled by the latent DDIM is first normalized to have zero mean and unit variance. Unconditional sampling from a diffusion autoencoder is thus done by sampling  $z_{\text{sem}}$  from the latent DDIM and unnormalizing it, then sampling  $x_T \sim N(0; I)$ , and finally decoding  $z = (z_{\text{sem}}; x_T)$  using the decoder.

Our choice of training the latent DDIM post-hoc has a few practical reasons. First, since training the latent DDIM takes only a fraction of the full training time, post-hoc training enables quick experiments on different latent DDIMs with the same diffusion autoencoder. Another reason is to keep  $z_{\text{sem}}$  as expressive as possible by not imposing any constraints, such as the prior loss in VAE [30], that can compromise the quality of the latent variables.

## 5. Experiments

We now turn to assessing the properties of our learned latent space and demonstrating new capabilities, such as attribute manipulation and conditional generation. For fair comparison, the DDIM baseline in our experiments refers to our reimplementation of DDIM [47] based on an improved architecture of Dhariwal et al. [11] with the same UNet hyperparameters as our decoder. In short, the DDIM baseline is similar to our decoder except that it does not take

### 5.1. Latent code captures both high-level semantics and low-level stochastic variations

To demonstrate that high-level semantics are mostly captured in  $z_{\text{sem}}$  and very little in  $x_T$ , we first compute the semantic subcode  $z_{\text{sem}} = \text{Enc}(x_0)$  from an input image  $x_0$ . For the stochastic subcode, instead of inferring it from the input, we will sample this subcode multiple times  $x_T^i \sim N(0; I)$  and decode multiple  $z^i = (z_{\text{sem}}; x_T^i)$ . Figure 3 shows the variations induced by varying given the same  $z_{\text{sem}}$ , as well as the variations from different  $z_{\text{sem}}$ .

The result shows that with a fixed  $z_{\text{sem}}$ , the stochastic subcode  $x_T$  only affects minor details, such as the hair and skin details, the eyes, or the mouth, but does not change the overall global appearance. And by varying  $z_{\text{sem}}$ , we obtain completely different people with different facial shapes, illuminations, and overall structures. Quantitative results are discussed in Section 5.4 and Table 2.

### 5.2. Semantically meaningful latent interpolation

One desirable property of a useful latent space is the ability to represent semantic changes in the image by a simple linear change in the latent space. For example, by moving along a straight line connecting any two latent codes, we expect a smooth morphing between the corresponding two images. In Figure 4d and Figure 1, we show our interpolation results by encoding two input images into  $(z_{\text{sem}}^1; x_T^1)$  and  $(z_{\text{sem}}^2; x_T^2)$ , then decoding  $z(t) = (\text{Lerp}(z_{\text{sem}}^1; z_{\text{sem}}^2; t); \text{Slerp}(x_T^1; x_T^2; t))$  for various values of  $t \in [0, 1]$ , where linear interpolation is used for  $z_{\text{sem}}$  and spherical linear interpolation is used for  $x_T$ , following [47].

Compared to DDIM, which produces non-smooth transitions, our method gradually changes the head pose, background, and facial attributes between the two endpoints. The interpolation results from StyleGAN in both  $W$  and

Figure 3. Reconstruction results and the variations induced by changing the stochastic subcode. Each row corresponds to a different  $z_{\text{sem}}$  which completely changes the person, whereas changing the stochastic subcode only affects minor details.

(a) StyleGAN2 interpolation after  $W$  space inversion.

(b) StyleGAN2 interpolation after  $W +$  space inversion.

(c) DDIM interpolation.

(d) Our diffusion autoencoder interpolation.

Figure 4. Interpolation between two real images. In contrast to StyleGAN2 and DDIM, our method produces smooth and consistent results with well-preserved original details from both images.

$W +$  spaces are smooth, but the two endpoints do not resemble the input images, whereas ours and DDIM's match the real input images almost exactly. We quantitatively evaluate how smooth our interpolation is in Appendix F.

### 5.3. Attribute manipulation on real images

Another way to assess the relationship between image semantics and linear motion or separability in the latent space is by moving the latent  $z_{\text{sem}}$  of an image in a particular direction and observing changes in the image [44]. By finding such a direction from the weight vector of a linear classifier trained on latent codes of negative and positive im-

ages of a target attribute, e.g., smiling, this operation consequently changes the semantic attribute in the image. There exists specialized techniques for this task [3, 37, 44, 57], but here we aim to showcase the quality and applicability of our latent space by using the simplest linear operation.

We trained linear classifiers using images and attribute labels from CelebA-HQ [26] and tested on CelebA-HQ and FFHQ [27] in Figure 5. Implementation details and more results can be found in Appendix G. Note that our autoencoder was trained on FFHQ but can generalize to CelebA-HQ without re-tuning the autoencoder. Our method is able to change local features, such as the mouth for smiling, while keeping the rest of the image and details mostly stationary. For global attributes that involve changing multiple features at the same time, such as aging, our results look highly plausible and realistic. Additionally, we compare the accuracy of these linear classifiers (Appendix E) that take  $z_{\text{sem}}$  versus those taking StyleGAN's inverted  $W$  as input. The AUROC over 40 attributes of our method is 0.925 and of StyleGAN's is 0.891. And we test how much the input's identity is preserved via ArcFace [10] and quantify the manipulation quality in Appendix G.

One notable advantage of diffusion autoencoders over GAN-based manipulation techniques is the ability to manipulate real images while preserving details irrelevant to the manipulation (e.g., keeping the original hair and background when manipulating facial expression). When GANs are used for such tasks, the details are often altered because real images cannot be faithfully inverted back to the GAN's latent space. Compared to a recent score-based manipulation technique SDEdit [35], which focuses on local edits or translating images from another domain using a forward-backward sampling trick, our method solves changing semantic attributes by simply modifying the latent code. We also compare qualitatively to D2C [45], which uses NVAE [50] decoder to perform a similar task in our Appendix H.

### 5.4. Autoencoding reconstruction quality

Although good reconstruction quality of an autoencoder may not necessarily be an indicator of good representation learning, this property plays an important role in many applications, such as compression or image manipulation that requires accurate encoding-decoding abilities. For these tasks, traditional autoencoders that rely on MSE or  $\ell_1$  loss functions perform poorly and produce blurry results. More advanced autoencoders combine perceptual loss and adversarial loss, e.g., VQGAN [13], or rely on a hierarchy of latent variables, e.g., NVAE [50], VQ-VAE2 [38]. Our design is an alternative that produces a latent code with compact semantics and performs competitively with state-of-the-art autoencoders. The key is our two-level encoding that delegates the reconstruction of less compressible stochastic details to our conditional DDIM.

Figure 5. Real-image attribute manipulation results on two global attributes (gender, age) and two local attributes (smile, wavy hair) by moving  $z_{\text{sem}}$  along the positive or negative direction found by linear classifiers. The top two are from FFHQ [27] and the bottom two are from CelebA-HQ [26]. Our method synthesizes highly-plausible and realistic results that preserve an unprecedented level of detail.

In Table 1, we evaluate the reconstruction quality of 1) our diffusion autoencoder, 2) DDIM [47], 3) a pretrained StyleGAN2 [28] (via two types of inversion), 4) VQ-GAN [13], 5) VQ-VAE2 [38], 6) NVAE [50]. Both DDIM and ours were trained on 130M images and used  $T=100$  for decoding. All these models were trained on FFHQ [27] and tested on 30k images from CelebA-HQ [26]. For our method and DDIM, we encoded downsampled test images of size  $128 \times 128$  and decoded them back. For the others, we used publicly available pretrained networks for  $256 \times 256$  and downsampled the results to the same  $128 \times 128$  before comparison. For StyleGAN2, we performed inversion in  $W$  [28] and  $W+$  [1, 2] spaces on the test images and used the optimized codes for reconstruction. The evaluation metrics are SSIM [53] (%), LPIPS [61] (#), and MSE. NVAE [50] achieves the lowest LPIPS and MSE scores, though it requires orders of magnitude larger latent dimension compared to others. Besides NVAE, our diffusion autoencoders outperform other models on all metrics, and only require  $T=20$  steps to surpass DDIM with  $T=100$  steps (Table 2).

Furthermore, we performed ablation studies to investigate 1) the reconstruction quality when only  $z_{\text{sem}}$  is encoded from the input but  $x_T$  is sampled from  $N(0, I)$  for decoding (Table 2.a), and 2) the effects of varying the dimension of  $z_{\text{sem}}$  from 64 to 512 (Table 2.b-e) on our autoencoder trained with 48M images for expedience. All configurations a-e) produce realistic results but differ in the degree of detail, where higher latent dimensions are better. For configuration a) with 512D  $z_{\text{sem}}$ , even though  $x_T$  is random, the reconstructions still look perceptually close to the input images as measured by LPIPS (also Figure 3). Our reconstruction with a small

Table 1. Autoencoding reconstruction quality of models trained on FFHQ [27] and tested on unseen CelebA-HQ [26]. Our model is competitive with state-of-the-art NVAE while producing readily useful high-level semantics in a compact 512D.

Model	Latent dim	SSIM %	LPIPS #	MSE #
StyleGAN2 ( $W$ ) [28]	512	0.677	0.168	0.016
StyleGAN2 ( $W+$ ) [28]	7,168	0.827	0.114	0.006
VQ-GAN [13]	65,536	0.782	0.109	3.61e-3
VQ-VAE2 [38]	327,680	0.947	0.012	4.87e-4
NVAE [50]	6,005,760	0.984	0.001	4.85e-5
DDIM ( $T=100, 128^2$ ) [47]	49,152	0.917	0.063	0.002
Ours ( $T=100, 128^2, \text{no } x_T$ )	512	0.677	0.073	0.007
Ours ( $T=100, 128^2$ )	49,664	0.991	0.011	6.07e-5

64D  $z_{\text{sem}}$  is already on par with StyleGAN2 inversion in 512D  $W$  latent space, suggesting that our diffusion autoencoders are proficient in compression.

### 5.5. Faster denoising process

One useful benefit of conditioning the denoising process with semantic information from  $z_{\text{sem}}$  is faster generation. One main reason DPMs require many generative steps is because DPMs can only use a Gaussian distribution to approximate  $p(x_{t-1}|x_t)$  when  $T$  is sufficiently large ( $> 1000$ ). Recent attempts to improve sampling speed focus on finding a better sampling interval or noise schedule [25, 29, 31, 36], or using more efficient solvers to solve the score-based ODE counterpart [25]. Our diffusion autoencoders do not aim to solve this problem directly, nor can they be compared in the same context as generative models that lack access to the target samples. It is, however, worth mentioning the effects they have within the DPM framework.

Table 2. Ablation study results for a) autoencoding reconstruction quality where  $z_{sem}$  is not encoded from the input but sampled from  $N(0; I)$ , and b-e) the effects of varying the dimension of  $z_{sem}$  from 64 to 512 on our autoencoder trained with 48M images for expedience. In a), our reconstruction is perceptually close to the input images (LPIPS=0.073) even when random. b-e) suggest that higher  $z_{sem}$  dimensions lead to higher fidelity reconstruction. Our diffusion autoencoders with  $T=20$  steps also surpass DDIM with  $T=100$  steps.

Model	SSIM %				LPIPS #				MSE #			
	T=10	T=20	T=50	T=100	T=10	T=20	T=50	T=100	T=10	T=20	T=50	T=100
DDIM (@130M) [47]	0.600	0.760	0.878	0.917	0.227	0.148	0.087	0.063	0.019	0.008	0.003	0.002
Ours (@130M, 512D $z_{sem}$ )	0.827	0.927	0.978	0.991	0.078	0.050	0.023	0.011	0.001	0.001	0.000	0.000
a) No encoded $z_T$	0.707	0.695	0.683	0.677	0.085	0.078	0.074	0.073	0.006	0.007	0.007	0.007
b) No encoded $z_T$ , @48M, 512D $z_{sem}$	0.662	0.650	0.637	0.631	0.102	0.096	0.093	0.092	0.009	0.009	0.009	0.010
c) No encoded $z_T$ , @48M, 256D $z_{sem}$	0.637	0.624	0.612	0.606	0.116	0.109	0.106	0.105	0.010	0.011	0.011	0.011
d) No encoded $z_T$ , @48M, 128D $z_{sem}$	0.613	0.600	0.588	0.582	0.133	0.127	0.125	0.124	0.012	0.012	0.013	0.013
e) No encoded $z_T$ , @48M, 64D $z_{sem}$	0.551	0.538	0.527	0.521	0.168	0.165	0.163	0.162	0.018	0.019	0.020	0.020

(a) DDIM predicting  $x_0$ .

(b) Our diffusion autoencoder predicting  $x_0$ .

Figure 6. Predicted  $x_0$  at  $t_{9;8;7;5;2;0}$  ( $T=10$ ). By conditioning on  $z_{sem}$ , our method predicts images that resemble much faster.

Consider a scenario where  $x_0$  is known to the denoising network. The noise prediction task will become trivial, and  $q(x_{t-1}|x_t; x_0)$  is a Gaussian distribution regardless of the number of timesteps [22]. Since our diffusion autoencoders model the distribution  $q(x_{t-1}|x_t; z_{sem})$ , it follows that  $p(x_{t-1}|x_t; z_{sem})$  is a better approximation to  $q(x_{t-1}|x_t; x_0)$  than  $p(x_{t-1}|x_t)$  when  $z_{sem}$  has captured much information about  $x_0$ . Figure 6 shows that a diffusion autoencoder is able to predict  $x_0$  more accurately in fewer steps than DDIM and yield better image quality on four different datasets with the same timesteps in Table 4.

## 5.6. Class-conditional sampling

This experiment demonstrates how our framework can be used for few-shot conditional generation and compares to D2C [45], a state-of-the-art DPM-based method for this setup. We follow the problem setup in D2C where the goal is to generate a diverse set of images of a target class, such as female, by utilizing a small number of labeled examples (e.g., 100). The labels can specify both the positives and negatives with respect to the target class (binary scenario) or only the positives (positive and unlabeled, or PU scenario). Given a latent class  $c_j$  (e.g., female), one simple way to do class-conditional sampling is with rejection sampling, as used by D2C. That is, we sample from our latent DDIM and accept this sample with probability  $p(c_j|z_{sem})$ . We followed D2C’s methodology and conditionally sampled 5k images, then computed FID scores

Table 3. FID scores for class-conditional generation on CelebA 64 dataset computed between 5k sampled images and the target subset.  $\sigma$  represents one standard deviation. D2C [45] results come from their paper (one run of FID computation on 5k samples). Binary classifier was trained with 50 positives and 50 negatives. Positive-unlabeled (PU) classifier was trained with 100 positives and 10,000 unlabeled examples (as negatives). Naive FIDs were computed between all images and the target subset.

Scenario	Classes	Ours		D2C [45]		Naive
Binary	Male	11.52	1.19	13.44		23.83
	Female	7.29	0.44	9.51		13.64
	Blond	16.10	2.00	17.61		25.62
	Non-Blond	8.48	0.52	8.94		0.96
PU	Male	9.54	0.54	16.39		23.83
	Female	9.21	0.19	12.21		13.64
	Blond	7.01	0.25	10.09		25.62
	Non-Blond	7.91	0.15	9.09		0.96

between these images and all the images of the same target class in CelebA dataset (with the same crop used by D2C). We used  $T = 100$  for both latent and image generations. Table 3 shows that our method achieves comparable FID scores to D2C, despite not using any self-supervised contrastive learning used in D2C.

## 5.7. Unconditional sampling

To evaluate the quality of our unconditional samples from diffusion autoencoders, we first sample  $z_{sem}$  from the latent DDIM, then decode  $z = (z_{sem}; x_T \sim N(0; I))$  using our decoder. We trained our autoencoders on FFHQ [27], LSUN Horse & Bedroom [60], and CelebA [33]. For each dataset, we computed FID scores between 50k randomly sampled images from the dataset and our 50k generated images. We also varied the timesteps (10, 20, 50, 100) used in both latent DDIM and our main decoder.

As shown in Table 4, our diffusion autoencoders are competitive with DDIM baselines and produce higher FID scores in most cases across numbers of timesteps. We also provide as reference our diffusion autoencoders trained with ground-truth latent variables encoded from the test images,

Table 4. FID scores<sup>#</sup> for unconditional generation. Our method is competitive with DDIM baselines. “+ autoencoding” refers to diffusion autoencoders that infer ground-truth semantic subcode from the test set and do not sample from the latent DDIM.

Dataset	Model	FID #			
		T=10	T=20	T=50	T=100
FFHQ 128	DDIM	29.56	21.45	15.08	12.03
	Ours	20.80	16.70	12.57	10.59
	+ autoencoding	14.43	10.70	6.69	4.56
Horse 128	DDIM	22.17	12.92	7.92	5.97
	Ours	11.97	9.37	7.44	6.71
	+ autoencoding	9.27	6.23	3.87	2.92
Bedroom 128	DDIM	13.70	9.23	7.14	5.94
	Ours	10.69	8.19	6.50	5.70
	+ autoencoding	6.36	4.88	3.61	2.88
CelebA 64	DDIM	16.38	12.70	8.52	5.83
	Ours	12.92	10.18	7.05	5.30
	+ autoencoding	12.78	9.06	5.15	3.11

labeled “+autoencoding.” In every dataset, perhaps unsurprisingly, conditioning the DDIM decoder with  $z_{sem}$  significantly improves the quality with smalls. In Appendix C, we show qualitative results and an additional experiment to verify that the latent DDIM does not memorize its input.

## 6. Related work

Denoising diffusion-based generative models [22,46] are closely related to denoising score-based generative models [48]. Models under this family have been shown to produce images with high quality rivaling those of GANs [11] without using the less stable adversarial training. They are also used widely for multiple conditional generation tasks, such as image super-resolution [32,43], image conditional generation [8,35], class-conditional generation in ImageNet dataset [11], and mel-spectrogram conditional speech synthesis [6]. Similar to our work, these methods rely on conditional DPMs; however, most conditioning signals in prior work are known a priori and fixed, while our diffusion autoencoder augments the latent variable with an end-to-end learnable signal that the CNN encoder discovers. This puts our work closer to VAE [30], particularly Wehenkel et al. [54] and D2C [45]. While these only utilize DPMs to model the prior distribution or latent representation for another generative model [12], our focus is on how DPMs can be augmented with meaningful latent codes.

Our diffusion autoencoders share common goals with other kinds of autoencoders such as VAE [30], NVAE [50], and VQ-VAE [52] and VQ-VAE2 [38]. While VAEs provide reasonable latent quality and sample quality, they are subject to posterior collapse [52] and prior holes problems [45], whereas DPMs are not. VQ-VAE with discrete latent variables was proposed to deal with these problems by fitting an autoregressive Pixel-CNN model to the latent variable post-hoc [51]. Fitting the latent variable post-hoc is also used in our work, but we utilize another DPM in-

stead of an autoregressive model. Rich image representations are useful for many downstream tasks; for example, VAE are often used in model-based reinforcement learning [14,17,18] for predicting future outcomes of the environment. VQ-VAE’s latent variables are used as a means for video generation tasks [59]. Our diffusion autoencoders also provide useful representations with an added ability to decode the representations back near perfectly.

Besides producing impressive image samples, GANs [15] have been shown to learn meaningful latent spaces [27] with extensive studies on multiple derived spaces [24,57] and various knobs and controls for conditional human face generation [21,37,55]. Encoding an image to the GAN’s latent space requires an optimization-based inversion process [28,58] or an external image encoder [39], which has limited reconstruction fidelity (or yields high-dimensional codes outside the learned manifold). This problem may be related to the GAN’s limited latent size and mode-collapse problem, where the latent space only partially covers the support of training samples. Diffusion autoencoders do not have this problem and can readily encode any image without any additional error-prone optimization.

## 7. Limitations & Discussion

When encoding images that are out of the training distribution, our diffusion autoencoders can still reconstruct the images well, owing to the high-dimensional stochastic subcode from DDIM. However, both the inferred semantic and stochastic subcodes may fall outside the learned distributions, resulting in a poor representation that can no longer be interpreted or interpolated. While our choice of using non-spatial latent code is suitable for learning global semantics, certain image and spatial reasoning tasks may require more precise local latent variables. For these tasks, incorporating 2D latent maps can be beneficial. For image generation, one unique feature of StyleGAN that is lacking from our diffusion autoencoders is the ability to control scale-specific generation. In terms of generation speed, our framework has significantly reduced the timesteps needed to achieve high-quality samples from our DDIM but still lacks behind GANs, which only require a single generator’s pass. Furthermore, we discuss the potential for generating deepfakes and possible technical safeguards in Appendix L.

In conclusion, we have presented diffusion autoencoders that can separately infer both semantics and stochastic information from an input image. In contrast to DPMs and high-fidelity autoencoders like NVAE, our latent representation allows near-exact decoding while containing compact semantics readily useful for downstream tasks. These properties enable simple solutions to various real-image editing tasks without requiring GANs and their error-prone inversion. Our framework also improves denoising efficiency and retains competitive unconditional sampling of DPMs.

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